# Well-posedness of a Newtonian two-phase flow with Boussinesq-Scriven surface fluid 

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Dedicated to Katarina<br>and my parents<br>Simone $\mathcal{E}$ Michael

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## Introduction

The motion of two immiscible fluids like oil and water can be modeled as a moving boundary problem for the two-phase Navier-Stokes equations. In a sharp-interface model, the interface between both fluid phases is considered as a geometric hypersurface. Physicists expect that interfacial properties such as surface tension play a prominent role when the interfacial area is large compared to the fluid volume. In this regard, Boussinesq [Bou13] proposed to consider certain surface viscosities that are related to intrinsic frictional forces within the interface. Several decades later, Scriven [Scr60] generalized Boussinesq's approach and obtained a model for arbitrary coordinate systems. This model is nowadays called the two-phase Navier-Stokes equations with Boussinesq-Scriven surface fluid and is denoted by ( N ) in this thesis. From a mathematician's point of view, it is fundamental to investigate whether this problem is wellposed; that is, whether it admits a uniquely determined solution that depends continuously on the initial state. Such a theory also has practical advantages. In particular, it can clarify admissible ranges of relevant parameters and indicate general limitations of the model that might be difficult to explore with experiments or numerical simulations alone. In this spirit, Bothe and Prüss [BP10] formally analyzed a related linear model problem and proved that its wellposedness depends on a condition for the interfacial velocity. The purpose of the present thesis is to extend their work and to investigate whether the original nonlinear problem is well-posed.

Let us formulate the model ( N ). We assume that the adjacent fluid phases occupy timedependent disjoint open subsets $\Omega_{+}(t)$ and $\Omega_{-}(t)$ in $\mathbb{R}^{n}(n \geq 2)$, which are separated by the sharp interface $\Gamma(t)=\partial \Omega_{+}(t) \cap \partial \Omega_{-}(t)$. Both bulk phases $\Omega_{ \pm}(t)$ and the interface $\Gamma(t)$ fill a rigid container $\Omega=\Omega_{+}(t) \cup \Gamma(t) \cup \Omega_{-}(t)$, which is a stationary domain. We employ the mass densities $\rho_{ \pm}$, the velocity fields $u_{ \pm}$, and the stress tensor $T_{ \pm}=S_{ \pm}-\pi_{ \pm} I$ with viscous stress tensor $S_{ \pm}$ and pressure $\pi_{ \pm}$. With the characteristic function $\chi_{ \pm}$of $\Omega_{ \pm}$, we put $\rho=\rho_{+} \chi_{+}+\rho_{-} \chi_{-}$and analogously for the other quantities. The principles of conservation of mass and momentum in $\Omega_{ \pm}$yield the continuity equation and the Navier-Stokes equation

$$
\partial_{t} \rho+\operatorname{div}(\rho u)=0, \quad \partial_{t}(\rho u)+\operatorname{div}(\rho u \otimes u-T)=\rho f .
$$

We restrict our considerations to incompressible Newtonian flows for which $\rho_{ \pm}$are positive constants and the viscous stress tensor $S_{ \pm}=2 \mu_{ \pm} D_{ \pm}$depends linearly on the rate-of-strain tensor $D_{ \pm}=\left(\nabla u_{ \pm}+\left[\nabla u_{ \pm}\right]^{\top}\right) / 2$ with constant positive shear viscosities $\mu_{ \pm}$. By putting also $f_{ \pm}=0$, we neglect external forces like gravity.

Additional conditions must be imposed on the fluid-solid boundary $\partial \Omega$ and the interface $\Gamma$. For simplicity, the latter should not touch the boundary $\partial \Omega$ and hence one of the bulk phases, say $\Omega_{-}$, should have its boundary $\partial \Omega_{-}=\Gamma$ in $\Omega$. Furthermore, we let the flow satisfy the noslip conditions $u_{+}=0$ on $\partial \Omega$ and $\llbracket u \rrbracket=0$ on $\Gamma$, where $\llbracket u \rrbracket:=\left.u_{+}\right|_{\Gamma}-\left.u_{-}\right|_{\Gamma}$ denotes the jump of $u$ across $\Gamma$. We exclude phase transitions and assume that the interface is material in the sense that the normal velocity $V_{\Gamma}$ of $\Gamma$ is given by $V_{\Gamma}=\left.u_{ \pm}\right|_{\Gamma} \cdot \nu_{\Gamma}$, where $\nu_{\Gamma}$ denotes the unit normal directed into $\Omega_{+}$. Thus, $\Gamma$ is advected with the flow of the bulk phases. Conservation of momentum also yields the interfacial momentum balance

$$
-\llbracket T \rrbracket \nu_{\Gamma}=\operatorname{div}_{\Gamma} T_{\Gamma},
$$

where $T_{\Gamma}$ is the surface stress tensor and $\operatorname{div}_{\Gamma} T_{\Gamma}$ denotes its surface divergence. When surface viscosities are negligible, we can put $T_{\Gamma}=\sigma P_{\Gamma}$ with the surface tension coefficient $\sigma$ and the tangential projection $P_{\Gamma}=I-\nu_{\Gamma} \otimes \nu_{\Gamma}$. With the $(n-1)$-fold mean curvature $H_{\Gamma}=-\operatorname{div}_{\Gamma} \nu_{\Gamma}$, this yields the well-known Laplace-Young law $-\llbracket T \rrbracket \nu_{\Gamma}=\sigma H_{\Gamma} \nu_{\Gamma}$ if $\sigma$ is constant, and otherwise $-\llbracket T \rrbracket \nu_{\Gamma}=\sigma H_{\Gamma} \nu_{\Gamma}+\nabla_{\Gamma} \sigma$ with Marangoni force $\nabla_{\Gamma} \sigma$, where $\nabla_{\Gamma}$ denotes the surface gradient. In order to incorporate surface viscosities, we assume that $T_{\Gamma}$ is given by the Boussinesq-Scriven constitutive law [cf. SSO07]

$$
T_{\Gamma}=\sigma P_{\Gamma}+\left(\lambda_{s}-\mu_{s}\right) \operatorname{div}_{\Gamma} u P_{\Gamma}+2 \mu_{s} D_{\Gamma},
$$

where $\lambda_{s}$ and $\mu_{s}$ are the surface viscosities and $D_{\Gamma}$ is the interfacial rate-of-strain tensor

$$
D_{\Gamma}=2^{-1} P_{\Gamma}\left(\nabla_{\Gamma} u+\left[\nabla_{\Gamma} u\right]^{\top}\right) P_{\Gamma} .
$$

We can decompose $T_{\Gamma}$ into

$$
T_{\Gamma}=\left\{\sigma+\left(\lambda_{s}+(3-n) \mu_{s} /(n-1)\right) \operatorname{div}_{\Gamma} u\right\} P_{\Gamma}+2 \mu_{s}\left[D_{\Gamma}-\left(\operatorname{tr} D_{\Gamma} /(n-1)\right) P_{\Gamma}\right],
$$

where the first summand is an isotropic tensor field and the second one has vanishing trace. Thus, we call $\mu_{s}$ the surface shear viscosity and

$$
\kappa_{s}=\lambda_{s}+(3-n) \mu_{s} /(n-1)
$$

the surface dilational viscosity. The latter equals $\lambda_{s}$ in the case $n=3$.
Bothe and Prüss [BP10] already noticed that the tangential part of the interfacial force $\operatorname{div}_{\Gamma} T_{\Gamma}$ is of second order in $v$ but only of first order in $w$. Accordingly, when we reformulate problem ( N ), we should handle the tangential and normal components separately; a complication that is not present in the situation without surface viscosities that was investigated by Köhne, Prüss, and Wilke [KPW13], where simply $\operatorname{div}_{\Gamma} T_{\Gamma}=\sigma H_{\Gamma} \nu_{\Gamma}$ with ( $n-1$ )-fold mean curvature $H_{\Gamma}$. In our situation, we decompose the velocity field $u$ near $\Gamma$ into $u=v+w \nu_{\Gamma}$ with tangential component $v:=P_{\Gamma} u$ and normal component $w:=\nu_{\Gamma} \cdot u$ and decompose the vector field $\operatorname{div}_{\Gamma} T_{\Gamma}$ accordingly. Then it can be shown that $\operatorname{div}_{\Gamma} T_{\Gamma}$ has the following structure.

$$
\begin{aligned}
\operatorname{div}_{\Gamma} T_{\Gamma}= & \mu_{s} \widetilde{\Delta}_{\Gamma} v+\lambda_{s} \nabla_{\Gamma} \operatorname{div}_{\Gamma} v+\mu_{s} H_{\Gamma}\left[\nabla_{\Gamma} v\right] \nu_{\Gamma}-\mu_{s} L_{\Gamma}^{2} v \\
& -2 \mu_{s} L_{\Gamma} \nabla_{\Gamma} w+\left(\mu_{s}-\lambda_{s}\right) \nabla_{\Gamma} w H_{\Gamma}-\left(\mu_{s}+\lambda_{s}\right) w \nabla_{\Gamma} H_{\Gamma} \\
& +\left[\left(\lambda_{s}-\mu_{s}\right) H_{\Gamma} \operatorname{div}_{\Gamma} v+2 \mu_{s} \operatorname{tr}\left(L_{\Gamma} D_{\Gamma}(v)\right)\right] \nu_{\Gamma} \\
& +\left[\sigma H_{\Gamma}-\left(\lambda_{s}-\mu_{s}\right) H_{\Gamma}^{2} w-2 \mu_{s} \operatorname{tr}\left(L_{\Gamma}^{2}\right) w\right] \nu_{\Gamma} .
\end{aligned}
$$

Here we employ a Laplace-Beltrami operator $\widetilde{\Delta}_{\Gamma}$ that acts on tangential vector fields, the scalar Laplace-Beltrami operator $\Delta_{\Gamma}=\operatorname{div}_{\Gamma} \nabla_{\Gamma}$, and the Weingarten tensor $L_{\Gamma}$.

We summarize these considerations in the aforementioned free boundary problem

$$
\left\{\begin{align*}
\partial_{t}(\rho u)+\operatorname{div}(\rho u \otimes u-T) & =0 & & \text { in } \Omega \backslash \Gamma(t),  \tag{N}\\
\operatorname{div} u & =0 & & \text { in } \Omega \backslash \Gamma(t), \\
\llbracket u \rrbracket & =0 & & \text { on } \Gamma(t), \\
-\llbracket T \rrbracket \nu_{\Gamma}-\operatorname{div}_{\Gamma} T_{\Gamma} & =0 & & \text { on } \Gamma(t), \\
V_{\Gamma}-u \cdot \nu_{\Gamma} & =0 & & \text { on } \Gamma(t), \\
\left.u\right|_{\partial \Omega} & =0 & & \text { on } \partial \Omega, \\
\Gamma(0) & =\Gamma_{0}, & & \\
\left.u\right|_{t=0} & =u_{0} & & \text { in } \Omega \backslash \Gamma_{0} .
\end{align*}\right.
$$

This model is considered as an initial value problem for a given initial velocity $u_{0}: \Omega \rightarrow \mathbb{R}^{n}$ and a given initial interface $\Gamma_{0} \subset \Omega$ and we ask for short-time existence and uniqueness of the unknown solution ( $u, \pi, \Gamma$ ) and its continuous dependence with respect to $\left(u_{0}, \Gamma_{0}\right)$. More information related to this model is given in the monographs of Aris [Ari89]; Edwards, Brenner, and

Wasan [EBW91]; and Slattery, Sagis, and Oh [SSO07]. A more recent survey on related models is given by Sagis [Sag11]. These authors mainly deal with theoretical properties in special situations and with experimental results. Furthermore, Barrett, Garcke, and Nürnberg [BGN14] analyzed a semi-discretized version of ( N ), where surface tension and surface viscosities depend on the concentration of a surfactant, whose distribution is governed by a convection-diffusion equation on the interface. For the simplified situation of a spherical droplet $\Omega_{-}$in a StokesPoiseuille flow, Reusken and Zhang [RZ13] carried out numerical experiments and studied the migration velocity of that droplet.

On the other hand, the theoretical understanding of problem ( N ) in general bounded configurations is still limited. Bothe and Prüss [BP10] have shown that the energy functional

$$
\frac{1}{2} \int_{\Omega} \rho|u(t, x)|^{2} d x+\sigma|\Gamma(t)|
$$

is always a strict Ljapunov functional for sufficiently smooth solutions and that its critical points for constant phase volumes $\left|\Omega_{ \pm}\right|$are precisely the stationary states of (N). However, the well-posedness of problem ( N ) has not been proved by rigorous mathematical analysis. Even worse, they found an additional condition that determines the well-posedness of a linear model problem in the whole space $\Omega=\mathbb{R}^{n}$ with flat reference interface $\Sigma=\mathbb{R}^{n-1} \times\{0\}$. In terms of some reference velocity $u_{*}$ related to $u_{0}$, this condition is given by

$$
d_{0}^{\mathrm{BP}}:=\sigma+\left(\lambda_{s}-\mu_{s}\right) \operatorname{div}_{\Sigma}\left(P_{\Sigma} u_{*}\right)+2 \mu_{s} \min _{\zeta \in \mathbb{R}^{n},|\zeta|=1} \zeta \cdot\left[\nabla_{\Sigma}\left(P_{\Sigma} u_{*}\right)\right] \zeta>0 .
$$

In case $d_{0}^{\mathrm{BP}}<0$, the interface symbol is not invertible. Hence it is not clear whether problem $(\mathrm{N})$ is well-posed for arbitrary velocities $u_{0}$, not even for short times.

This thesis attempts to fill this gap. We will reformulate problem (N) as an equivalent transformed problem (T) where the unkown interface $\Gamma(t)$ is replaced by a stationary interface $\Sigma$ and a height fucntion $h(t, \cdot): \Sigma \rightarrow \mathbb{R}$. As our main result, we prove that problem (T) is locally well-posed for initial velocities subject to the following well-posedness condition:

$$
\begin{equation*}
\inf _{\Sigma}\left(\sigma+\left(\lambda_{s}-\mu_{s}\right) \operatorname{div}_{\Sigma} u_{0}+2 \mu_{s} \min _{\zeta \in \mathbb{R}^{n},|\zeta|=1} \zeta \cdot\left[\nabla_{\Sigma} u_{0}\right] \zeta\right)>0 . \tag{WPC}
\end{equation*}
$$

Thus, compared to the linear model problem of Bothe and Prüss, not only the tangential velocity $\left.P_{\Sigma} u_{0}\right|_{\Sigma}$, but the full velocity is important for the well-posedness of the nonlinear problem. We further show that the corresponding condition is not only sufficient, but also necessary for the invertibility of the interface symbol of a corresponding linear model problem.

We mainly follow the strategy of Köhne, Prüss, and Wilke [KPW13] and employ a timedependent diffeomorphism $\Theta(t, \cdot)$ of the underlying domain $\Omega$, which maps a fixed hypersurface $\Sigma \subset \Omega$ onto $\Gamma(t)=\Theta(\{t\} \times \Sigma)$. One such diffeomorphism is the well-known Hanzawa map $\Theta_{h}^{\text {Han }}$ [Han81, (2.1)], which was first used by Hanzawa for transforming the one-phase Stefan problem. It is an extension to $\Omega$ of the parametrization

$$
\theta_{h}(t, x)=x+h(t, x) \nu_{\Sigma}(x) \in \Gamma(t) \quad \text { for } t \in J, x \in \Sigma
$$

The Hanzawa map was also applied by Escher, Prüss, and Simonett [EPS02] for transforming a two-phase Stefan problem and by Köhne, Prüss, and Wilke [KPW13] for transforming the two-phase Navier-Stokes equations with surface tension. For the latter, the authors considered the transformed functions

$$
\bar{u}(t, x)=u\left(t, \Theta_{h}(t, x)\right), \quad \bar{\pi}(t, x)=\pi\left(t, \Theta_{h}(t, x)\right),
$$

and reformulated their original problem for $(u, \pi, \Gamma)$ as a transformed problem for $(\bar{u}, \bar{\pi}, h)$.
However, this velocity transformation does not seem appropriate for transforming our problem (N) with additional surface viscosities, since, on the one hand, both $v=P_{\Gamma} u$ and
$w=\nu_{\Gamma} \cdot u$ would depend on both $\bar{v}=P_{\Sigma} \bar{u}$ and $\bar{w}=\nu_{\Sigma} \cdot \bar{u}$, but on the other hand, the interface momentum balance requires different orders of differentiability for $v$ and $w$. We therefore employ both a different diffeomorphism and a different velocity transformation for ensuring that these velocity components are transformed separately. We consider a class of maps $\Theta: J \times \bar{\Omega} \rightarrow \bar{\Omega}$ that we call the normal-preserving admissible maps. For such a map, the normal derivative $\partial_{\nu_{\Sigma}} \Theta(t, x)$ is a multiple of the original normal vector field $\nu_{\Gamma}(t, \Theta(t, x))$, whereas the Hanzawa map satisfies $\partial_{\nu_{\Sigma}} \Theta_{h}^{\mathrm{Han}}=\nu_{\Sigma}$. Moreover, the Jacobian $\partial_{x} \Theta(t, x)$ maps the normal space $\mathbb{R} \nu_{\Sigma}(x)$ onto the normal space $\mathbb{R} \nu_{\Gamma}(t, \Theta(t, x))$ and the tangent space $T_{x} \Sigma$ onto $T_{\Theta(t, x)} \Gamma(t)$. We will construct such a map $\Theta_{h}$ in terms of a height function $h$ by using a similar method as Abels and Terasawa [AT09], who transformed a Stokes problem with variable viscosity in a bent halfspace. Our map $\Theta_{h}$ has several advantages when we consider the velocity transformation

$$
u\left(t, \Theta_{h}(t, x)\right)=\left[\partial_{x} \Theta_{h}(t, x)\right] \bar{u}(t, x)
$$

First, we have

$$
v\left(t, \Theta_{h}(t, x)\right)=\left[\partial_{x} \Theta_{h}(t, x)\right] \bar{v}(t, x), \quad w\left(t, \Theta_{h}(t, x)\right)=\nu_{\Gamma}\left(t, \Theta_{h}(t, x)\right) \cdot \nu_{\Sigma}(x) \bar{w}(t, x)
$$

and thus the velocity components are transformed separately. Second, the advected moving interface condition $V_{\Gamma}=\nu_{\Gamma} \cdot u$ is transformed to the simple identity

$$
\partial_{t} h=\bar{w}
$$

Thus, compared to Prüss and Simonett [PS11], we can avoid perturbations in this equation.
In this way, problem (N) can be reformulated as a transformed problem

$$
\left\{\begin{align*}
\rho \partial_{t} \bar{u}-\mu \Delta \bar{u}+\nabla \bar{\pi} & =F_{u}(\bar{u}, \bar{\pi}, h) & & \text { in } J \times \Omega \backslash \Sigma,  \tag{T}\\
\operatorname{div} \bar{u} & =F_{d}(\bar{u}, h) & & \text { in } J \times \Omega \backslash \Sigma, \\
\llbracket \bar{u} \rrbracket & =0 & & \text { on } J \times \Sigma \\
L_{u}\left(\bar{u}, \bar{\pi}, h ; \bar{u}_{*}\right) & =G_{u}\left(\bar{u}, \bar{\pi}, h ; \bar{u}_{*}, \bar{\pi}_{*}, h_{*}\right) & & \text { on } J \times \Sigma \\
\partial_{t} h-\bar{u} \cdot \nu_{\Sigma} & =0 & & \text { on } J \times \Sigma \\
\left.\bar{u}\right|_{\partial \Omega} & =0 & & \text { on } J \times \partial \Omega \\
\left.h\right|_{t=0} & =h_{0} & & \text { on } J \times \Sigma \\
\left.\bar{u}\right|_{t=0} & =\bar{u}_{0} & & \text { in } \Omega \backslash \Sigma
\end{align*}\right.
$$

Here the left-hand sides are linear with respect to $(\bar{u}, \bar{\pi}, h)$, the functions $\bar{u}_{*}$, $\bar{\pi}_{*}$, and $h_{*}$ are chosen according to the initial data, and $F_{u}, F_{d}$, and $G_{u}$ are nonlinear perturbations that have to be controlled in a suitable way. In the following, we omit the bars over $\bar{u}, \bar{u}_{*}, \bar{\pi}$, and $\bar{\pi}_{*}$. For solving problem (T), we also employ its principal linearization

$$
\left\{\begin{array}{rlrl}
\rho \partial_{t} u-\mu \Delta u+\nabla \pi & =f_{u} & & \text { in } J \times \Omega \backslash \Sigma,  \tag{PL}\\
\operatorname{div} u=f_{d} & & \text { in } J \times \Omega \backslash \Sigma, \\
\llbracket u \rrbracket & =0 & & \text { on } J \times \Sigma, \\
L_{u}\left(u, \pi, h ; u_{*}\right)=g_{u} & & \text { on } J \times \Sigma, \\
\partial_{t} h-u \cdot \nu_{\Sigma}=0 & & \text { on } J \times \Sigma, \\
\left.u\right|_{\partial \Omega}=0 & & \text { on } J \times \partial \Omega, \\
\left.h\right|_{t=0}=0 & & \text { on } \Sigma, \\
\left.u\right|_{t=0}=0 & & \text { in } \Omega \backslash \Sigma .
\end{array}\right.
$$

In order to define the operator $L_{u}$, we decompose $u_{*}=v_{*}+w_{*} \nu_{\Sigma}$ and $g_{u}=g_{v}+g_{w} \nu_{\Sigma}$ as well as $L_{u}\left(u, \pi, h ; u_{*}\right)=L_{v}\left(u, h ; u_{*}\right)+L_{w}\left(u, \pi, h ; u_{*}\right) \nu_{\Sigma}$. Then we have

$$
\begin{aligned}
L_{v}\left(u, h ; u_{*}\right)= & -\mu_{s} \widetilde{\Delta}_{\Sigma} v-\lambda_{s} \nabla_{\Sigma} \operatorname{div}_{\Sigma} v-\llbracket \mu \nabla_{\Sigma} w \rrbracket-\llbracket \mu \partial_{\nu} v \rrbracket+\left(\lambda_{s}+\mu_{s}\right) w_{*} \nabla_{\Sigma} \Delta_{\Sigma} h, \\
L_{w}\left(u, \pi, h ; u_{*}\right)= & -\operatorname{tr}\left(\left[\left(\lambda_{s}-\mu_{s}\right) H_{\Sigma}+2 \mu_{s} L_{\Sigma}\right] \nabla_{\Sigma} v\right)-2 \llbracket \mu \partial_{\nu} w \rrbracket+\llbracket \pi \rrbracket \\
& -\operatorname{tr}\left(\left[\sigma+\left(\lambda_{s}-\mu_{s}\right)\left(\operatorname{div}_{\Sigma} v_{*}-2 H_{\Sigma} w_{*}\right)+2 \mu_{s} D_{\Sigma}\left(v_{*}\right)-4 \mu_{s} w_{*} L_{\Sigma}\right] \nabla_{\Sigma}^{2} h\right) .
\end{aligned}
$$

A crucial task is to verify that problem (PL) has optimal regularity, which means that the solution-to-data map $(u, \pi, h) \mapsto\left(f_{u}, f_{d}, g_{u}\right)$ is a topological linear isomorphism between suitable function spaces. Hence the regularity conditions on the data must be both necessary and sufficient for the existence and regularity of the solution. In this case, the well-posedness of the nonlinear problem ( T ) can be proved simply by Banach's fixed-point theorem. We are interested in spaces for which the velocity $u(t, x)$ and pressure $\pi(t, x)$ satisfy the regularity conditions

$$
u \in H_{p}^{1}\left(J ; L_{p}(\Omega)^{n}\right) \cap L_{p}\left(J ; H_{p}^{2}(\Omega \backslash \Sigma)^{n}\right), \quad \pi \in L_{p}\left(J ; \dot{H}_{p}^{1}(\Omega \backslash \Sigma)\right),
$$

where the Lebesgue exponent $p \in(1, \infty)$ will be chosen sufficiently large for controlling the nonlinear perturbations in problem (T). In order to construct such spaces, we solve a linear model problem for (PL) in the whole space $\Omega=\mathbb{R}^{n+1}$ with a flat interface $\Sigma=\mathbb{R}^{n} \times\{0\}$ under the restriction $\left(f_{u}, f_{d}, u_{0}, h_{0}\right)=0$. The generic element of $\mathbb{R}^{n+1}$ is denoted by $(x, y)$ with $x \in \mathbb{R}^{n}$ and $y \in \mathbb{R}$. Let $\vartheta_{w} \in \mathbb{R}, \vartheta_{L} \in \mathbb{R}^{n \times n}$, and $\vartheta_{D v} \in \mathbb{R}^{n \times n}$ denote the values of $w_{*}, L_{\Sigma}$, and $D_{\Sigma}\left(v_{*}\right)$ at some fixed position and define the parameters

$$
\begin{aligned}
c_{1} & :=\left(\lambda_{s}+\mu_{s}\right) \vartheta_{w}, & & c_{2}:=\left(\lambda_{s}-\mu_{s}\right) \operatorname{tr} \vartheta_{L}, \\
C_{3} & :=\mu_{s} \vartheta_{L}, & & C_{4}:=2 \mu_{s}\left(\vartheta_{D v}-2 \vartheta_{w} \vartheta_{L}\right), \\
c_{5,6} & \in\{0,1\}, & & c_{\sigma}:=\sigma+\left(\lambda_{s}-\mu_{s}\right) \operatorname{tr}\left(\vartheta_{D v}-2 \vartheta_{w} \vartheta_{L}\right) .
\end{aligned}
$$

Then the aforementioned model problem is given by
(MP)

$$
\begin{array}{rlrl}
\rho\left(\tau+\partial_{t}\right) u-\mu \Delta u+\nabla \pi & =0 & & \text { in } \mathbb{R}_{+} \times \dot{R}^{n+1}, \\
\operatorname{div} u & =0 & & \text { in } \mathbb{R}_{+} \times \dot{\mathbb{R}}^{n+1}, \\
\llbracket u \rrbracket & =0 & & \text { on } \mathbb{R}_{+} \times \mathbb{R}^{n}, \\
-\mu_{s} \Delta_{x} v-\lambda_{s} \nabla_{x} \operatorname{div}_{x} v-c_{5} \llbracket \mu \nabla_{x} w \rrbracket-c_{6} \llbracket \mu \partial_{y} v \rrbracket+c_{1} \nabla_{x} \Delta_{x} h & =g_{v} & & \text { on } \mathbb{R}_{+} \times \mathbb{R}^{n}, \\
-\operatorname{tr}\left(\left(c_{2}+2 C_{3}\right) \nabla_{x} v\right)-2 \llbracket \mu \partial_{y} w \rrbracket+\llbracket \pi \rrbracket-\operatorname{tr}\left(\left(c_{\sigma}+C_{4}\right) \nabla_{x}^{2} h\right) & =g_{w} & & \text { on } \mathbb{R}_{+} \times \mathbb{R}^{n}, \\
\left(\tau+\partial_{t}\right) h-w & =g_{h} & & \text { on } \mathbb{R}_{+} \times \mathbb{R}^{n}, \\
\left.h\right|_{t=0} & =0 & & \text { on } \mathbb{R}^{n}, \\
\left.u\right|_{t=0}=0 & & \text { in } \mathbb{R}^{n+1} .
\end{array}
$$

Here, $\tau>0$ will be a sufficiently large number and we allow for $g_{h} \neq 0$. The term $c_{5} \llbracket \mu \nabla_{x} w \rrbracket$ is of lower order in our functional analytic setting and therefore negligible, in contrast to the situation without surface viscosities. Moreover, we will choose $c_{6}=1$ for proving well-posedness, but also allow for $c_{6} \in\{0,1\}$ in the symbolic calculations.

A basic version of problem (MP) without surface viscosities was solved in an $L_{p}$-setting by Prüss and Simonett [PS10] for the parameters $\left(c_{1}, c_{2}, C_{3}, C_{4}\right)=0, c_{5,6}=1$, and $c_{\sigma}=\sigma$. They also included gravity acting in the negative $x_{n+1}$-direction and studied the modified equation $\partial_{t} h-w+b \cdot \nabla h=g_{h}$ in [PS11]. Here the additional term $b \cdot \nabla h$ with $b \in \mathbb{R}^{n}$ arises when the free-interface problem is transformed by means of the Hanzawa diffeomorphism $\Theta_{h}$ and the velocity transformation $u\left(t, \Theta_{h}(t, x)\right)=\bar{u}(t, x)$ and when the transformed problem is linearized at a non-trivial reference velocity. In this thesis we can neglect the term $b \cdot \nabla h$.

A linear problem including surface viscosities $\lambda_{s}, \mu_{s}>0$ was derived and analyzed by Bothe and Prüss [BP10]. Roughly speaking, their model corresponds to (MP) with $c_{1}=0$, $c_{2}=0, C_{3}=0, c_{\sigma}=\sigma+\left(\lambda_{s}-\mu_{s}\right) \vartheta_{d v}, C_{4}=2 \mu_{s} \vartheta_{D v}, c_{5,6}=1$. In particular, all terms arising for
non-trivial $\vartheta_{w}$ are not present and their $g_{v}$-equation is only of second order in $h$. As mentioned above, we require the well-posedness condition

$$
d_{0}\left(\vartheta_{D u}\right):=\sigma+\left(\lambda_{s}-\mu_{s}\right) \operatorname{tr} \vartheta_{D u}+2 \mu_{s} \min _{\xi \in \mathbb{R}^{n} \backslash\{0\}} \xi \cdot\left[\vartheta_{D u}\right] \xi|\xi|^{-2}>0 .
$$

This condition is also necessary for the invertibility of the associated interface symbol.
Denk and Kaip [DK13, Section 4.7] solved a variant of (MP) for vanishing $c_{1}, c_{2}, C_{3}, C_{4}$ which combines the models of [BP10] and [PS11] for surface viscosities and gravity. They derived function spaces for the interface quantities for which the corresponding interface operator is an isomorphism and their results cover both cases $\lambda_{s}, \mu_{s}=0$ and $\lambda_{s}, \mu_{s}>0$. Fortunately, we can adapt their method to the present situation, but we shall employ somewhat different function spaces, due to the additional leading order term $c_{1} \nabla_{x} \Delta_{x} h$. We will compute the interface symbol and prove that it is invertible for all cases $c_{5,6} \in\{0,1\}$. Moreover, we will see that the order structure of the system and hence also the spaces for optimal regularity depend on $c_{6}$ but not on $c_{5}$. Suitable function spaces for solving problem (MP) are only constructed in the case $c_{6}=1$ since these spaces allow for better time regularity than the case $c_{6}=0$. Unfortunately, the $g_{v}$-equation is not invariant under the parabolic scaling $v(t, x)=v_{\zeta}(\zeta t, \sqrt{\zeta} x)$ and in this situation the author does not know how to perform perturbation theory on $J=(0, \infty)$ for arbitrary initial states. Therefore we deal with short time intervals $J=(0, T)$ and use a small end time $T$ instead of a large number $\tau$ as a perturbation parameter.

To transfer optimal regularity of the model problem to the principal linearization (PL), we adapt the localization procedures of Köhne, Prüss, and Wilke [KPW13]; Abels and Terasawa [AT09]; Denk, Hieber, and Prüss [DHP03]; Amann, Hieber, and Simonett [AHS94]; and Ladyzhenskaya, Solonnikov, and Ural'tseva [LSU68]. With also provide a theory on an elliptic transmission problem

$$
\left\{\begin{align*}
\operatorname{div}(\mu \nabla \psi) & =\operatorname{div} u & & \text { in } \Omega \backslash \Sigma,  \tag{TP}\\
\partial_{\nu} \psi & =\nu \cdot u & & \text { on } \partial \Omega, \\
\llbracket \mu \partial_{\nu} \psi \rrbracket & =\llbracket \mu \nu \cdot u \rrbracket & & \text { on } \Sigma, \\
\llbracket \psi \rrbracket & =0 & & \text { on } \Sigma,
\end{align*}\right.
$$

and its weak version

$$
\int_{\Omega} \mu \nabla \psi \cdot \nabla \varphi d x=\int_{\Omega} u \cdot \nabla \varphi d x \quad \text { for } \varphi \in C_{c}^{\infty}\left(\mathbb{R}^{n}\right), \quad \llbracket \psi \rrbracket=0 \text { on } \Sigma .
$$

This theory suffices to determine the bulk pressure $\pi$ and to handle the inhomogeneity $f_{d}$ in the divergence equation. For these problems we prove the optimal a priori estimates

$$
\|\nabla \psi\|_{H_{p}^{k}(\Omega \backslash \Sigma)} \leq C\|u\|_{H_{p}^{k}(\Omega \backslash \Sigma)} \quad \text { for } k \in\{0,1,2\}
$$

by means of a localization procedure based on the methods of Simader and Sohr [SS92] and Köhne, Prüss, and Wilke [KPW13].

In this way we can conclude that (PL) induces a topological isomorphism and that the linear solution operator corresponding to (PL) is uniformly bounded with respect to the length of the time interval $T \rightarrow 0+$ and certain reference velocities $u_{*}$ which satisfy (WPC). By means of Banach's fixed point theorem we show that problem (T) is well-posed for small $T$, for small $h_{0}$ and for possibly large $u_{0}$ that satisfies (WPC).

This thesis is organized as follows. In Chapter 1, we derive problem (N) in a mathematically rigorous way and study some properties of this model. Chapter 2 provides an optimal regularity theory for the transmission problem (TP), which is employed later on. Optimal regularity for the principal linearization (PL) is proved in Chapter 3. Finally, we establish well-posedness for the transformed problem (T) in Chapter 4. For keeping this thesis self-contained, we provide relevant results on differential geometry and functional analysis in Appendices A and B.

## CHAPTER 1

## Modeling of moving interface flows

In this chapter we derive the model $(\mathrm{N})$ in a rigorous way from basic principles and constitutive assumptions. To this end, we also study the concepts of moving domains and moving hypersurfaces and recall important divergence theorems and transport theorems.

Basic notation. Throughout this thesis, the symbols $\mathbb{N}=\{1,2,3, \ldots\}, \mathbb{Z}, \mathbb{R}$, and $\mathbb{C}$ denote the sets of the positive integers, the integers, the real numbers, and the complex numbers, and we let $\mathbb{K}$ denote either $\mathbb{R}$ or $\mathbb{C}$. We also put $\mathbb{N}_{0}:=\mathbb{N} \cup\{0\}, \mathbb{R}_{+}:=[0, \infty), \mathbb{R}_{-}:=(-\infty, 0]$, and $\mathbb{C}_{+}:=\{z \in \mathbb{C}: \operatorname{Re} z \geq 0\}$. The imaginary unit is denoted by $i$. For a real number $x$ we let $\lfloor x\rfloor:=[x]:=\max \{k \in \mathbb{Z}: k \leq x\},\lceil x\rceil:=\min \{k \in \mathbb{Z}: k \geq x\}$, and $\{x\}:=x-\lfloor x\rfloor$.

The $n$-dimensional Euclidean space $\mathbb{R}^{n}(n \in \mathbb{N})$ is equipped with the scalar product $v$. $w=(v \mid w)=v_{1} w_{1}+v_{2} w_{2}+\cdots+v_{n} w_{n}$ and the norm $|v|=\sqrt{v \cdot v}$ for $v=\left(v_{1}, v_{2}, \ldots, v_{n}\right)$ and $w=\left(w_{1}, w_{2}, \ldots, w_{n}\right)$. The vector space $\mathbb{C}^{n}$ is equipped with the scalar product $(v \mid w)=$ $v_{1} \bar{w}_{1}+v_{2} \bar{w}_{2}+\cdots+v_{n} \bar{w}_{n}$, where the bar denotes complex conjugation. We let $\langle v, w\rangle=v \cdot w=$ $v_{1} w_{1}+v_{2} w_{2}+\cdots+v_{n} w_{n}$ denote the bilinear product of two vectors $v, w \in \mathbb{C}^{n}$. The canonical basis of $\mathbb{K}^{n}$ as a $\mathbb{K}$-vector space consists of the unit vectors $e^{j}=e_{j}=\left(\delta_{i j}\right)_{i}$, where $\delta_{i j}, \delta_{i}^{j}$, and $\delta^{i j}$ denote the Kronecker delta.

Matrices are denoted by $A=\left[a_{i j}\right]_{i j} \in \mathbb{K}^{n \times m}$ for $i \in\{1, \ldots, n\}$ and $j \in\{1, \ldots, m\}$. The transposed matrix of $A$ is given by $A^{\top}=\left[a_{j i}\right]_{i j}$. The symmetric part of a quadratic matrix $A \in \mathbb{K}^{n \times n}$ is $\operatorname{sym} A:=2^{-1}\left(A+A^{\top}\right)$ and the Kronecker product of two vectors $v \in \mathbb{K}^{n}$ and $w \in \mathbb{K}^{m}$ is defined by $v \otimes w:=\left[v_{i} w_{j}\right]_{i j}$. The symbol $|A|$ denotes the induced matrix norm of the Euclidean vector norm; that is, $|A|=\max \left\{|A v|: v \in \mathbb{K}^{n}\right.$ with $\left.|v|=1\right\}$. The trace of $A \in \mathbb{K}^{n \times n}$ is $\operatorname{tr} A=a_{11}+\cdots+a_{n n}$. Using Einstein's summation convention, we write $\operatorname{tr} A=e_{j} \cdot A e^{j}$. For two matrices $A, B \in \mathbb{R}^{n \times n}$ we put $A: B:=\operatorname{tr}\left(A^{\top} B\right)=a_{i j} b_{i j}$.

For two sets $U$ and $V$, we write $U \subset V$, if $U$ is a subset of $V$. We also write $U \dot{U} V$ for the union $U \cup V$ of disjoint sets $U$ and $V$. The power set of $U$, which consists of all subsets of $U$, is denoted by $2^{U}$. We write $U \subset \subset V$ if $U$ and $V$ are subsets of some metric space such that $U$ is bounded and its closure $\bar{U}$ is contained in $V$.

The notation $f: X \supset U \rightarrow V \subset Y$ or $f: U \subset X \rightarrow V \subset Y$ indicates that $f$ is a mapping from the subset $U$ of the set $X$ into the subset $V$ of the set $Y$. The set gr $f=\{(x, f(x)): x \in U\}$ is the graph of $f$. For a set-valued map $F: U \rightarrow 2^{Y}$ we put gr $F:=\cup_{x \in U}(\{x\} \times F(x)) \subset U \times Y$. If $U$ and $V$ are subspaces of topological spaces $X$ and $Y$, then the vector space $C(U ; V)$ contains all maps $f: U \rightarrow V$ that are continuous with respect to the topologies induced by $X$ and $Y$. We will abbreviate $C(U ; \mathbb{K})=: C(U)$.

The partial derivatives of a $C^{1}$-map $f$ defined in $U \subset \mathbb{R}^{n}$ are denoted by $\partial_{i} f=\partial f / \partial x_{i}$ and the (Fréchet) derivative $\partial f$ of $f$ at $x_{*} \in U$ is the linear map $v \mapsto\left[\partial f\left(x_{*}\right)\right] v=(d / d h) f\left(x_{*}+\right.$ $h v)\left.\right|_{h=0}$. The nabla operator $\nabla=\left(\partial_{1}, \partial_{2}, \ldots, \partial_{n}\right)^{\top}$ is defined by $\nabla f=\left(\partial_{1} f, \ldots, \partial_{n} f\right)^{\top}$ for a scalar field $f$ and $\nabla v=e^{j} \otimes \partial_{j} v=\left[\partial_{i} v_{j}\right]_{i j}$ for a vector field $v$. The divergence is defined by $\operatorname{div} v=\partial_{i} v_{i}$ for a vector field $v$ and $\operatorname{div} S=\left(\partial_{j} S_{i j}\right)_{i}$ for a symmetric matrix field $S$. Thus, $\operatorname{div}(S v)=\operatorname{div} S \cdot v+S: \nabla v$.

Let $\Sigma$ be a $C^{1}$-hypersurface in $\mathbb{R}^{n}$ with local parametrization $U \subset \mathbb{R}^{n-1} \rightarrow \Sigma, u \mapsto x=\phi(u)$. We employ the tangent vectors $\tau_{j}(x)=\partial_{j} \phi(u)$, which span the tangent space $T_{x} \Sigma$, and the cotangent vectors $\tau^{k}(x)$, which are uniquely determined by the relations $\tau_{j}(x) \cdot \tau^{k}(x)=\delta_{j}^{k}$. The
partial derivatives of a $C^{1}$-map $f: \Sigma \rightarrow X$ are denoted by $\partial_{i} f(x):=\partial_{i}(f \circ \phi)(u)$. We define the surface gradient $\nabla_{\Sigma} f=\tau^{j} \partial_{j} f$ for a scalar field $f$ and $\nabla_{\Sigma} u=\tau^{j} \otimes \partial_{j} u$ for a (not necessarily tangential) vector field $u$. Moreover, we define the surface divergence $\operatorname{div}_{\Sigma} u=\partial_{j} u \cdot \tau^{j}$ and $\operatorname{div}_{\Sigma} S=\left(\partial_{j} S\right) \tau^{j}$ for a symmetric matrix field $S$. Thus, $\operatorname{div}_{\Sigma}(S u)=\operatorname{div}_{\Sigma} S \cdot u+S: \nabla_{\Sigma} u$. Moreover, $\nu_{\Sigma}: \Sigma \rightarrow \mathbb{R}^{n}$ is a unit normal field of $\Sigma$ and, if $\Sigma$ is of class $C^{2}, L_{\Sigma}=-\nabla_{\Sigma} \nu_{\Sigma}$ denotes the Weingarten tensor and $H_{\Sigma}=-\operatorname{div}_{\Sigma} \nu_{\Sigma}=\operatorname{tr} L_{\Sigma}$ denotes the ( $(n-1)$-fold) mean curvature.

For a metric space $\left(X, d_{X}\right)$, the symbols $B_{R}^{X}(x)$ or $\mathbb{B}_{R}^{X}(x)$ denote the open ball $\{y \in X$ : $\left.d_{X}(x, y)<R\right\}$ of radius $R$ and center $x \in X$. If $\left(X,\|\cdot\|_{X}\right)$ is a normed vector space, we abbreviate $B_{R}^{X}:=B_{R}^{X}(0):=\left\{x \in X:\|x\|_{X}<R\right\}$. We will write $B_{R}(x)$ instead of $B_{R}^{X}(x)$ if $X$ is known from the context. For a subset $M$ of $X$ we define $B_{R}(M)=\{y \in X: \operatorname{dist}(x, M)<R\}$, where $\operatorname{dist}(x, M):=\inf \{\operatorname{dist}(x, y): y \in M\}$. Two normed vector spaces $X$ and $Y$ are equal if they coincide as sets and have equivalent norms. We write $Y \hookrightarrow X$, if $Y$ is continuously embedded into $X$ and we write $Y \hookrightarrow^{d} X$, if the embedding is also dense. The complexification of a real vector space $X=X_{\mathbb{R}}$ is denoted by $X_{\mathbb{C}}=\left\{x_{1}+i x_{2}: x_{1}, x_{2} \in X\right\}$.

The vector space of all linear operators $A: X \rightarrow Y$ between vector spaces $X$ and $Y$ is denoted by $\mathcal{L}(X ; Y)$ and we abbreviate $\mathcal{L}(X):=\mathcal{L}(X ; X)$. We let $N(A)=\{x \in X: A x=0\}$ and $R(A)=\{A x: x \in X\}$ denote the null space an range of a linear operator $A: X \rightarrow Y$. The complexification of an $\mathbb{R}$-linear operator $A: X \rightarrow Y$ is given by $A_{\mathbb{C}}: X_{\mathbb{C}} \rightarrow Y_{\mathbb{C}},\left(x_{1}+i x_{2}\right) \mapsto$ $A x_{1}+i A x_{2}$. The space of bounded linear operators $A: X \rightarrow Y$ between normed vector spaces $X$ and $Y$ is denoted by $\mathcal{B}(X ; Y)$, and it is equipped with the operator norm $\|A\|_{\mathcal{B}(X ; Y)}=\|A\|_{X \rightarrow Y}$. The space of bounded $k$-linear maps $A: X^{k} \rightarrow Y$ for $k \in \mathbb{N}$ is denoted by $\mathcal{B}^{k}\left(X^{k} ; Y\right)$, and its norm is denoted by

$$
\|A\|_{\mathcal{B}^{k}\left(X^{k} ; Y\right)}=\sup \left\{\left\|A\left(x_{1}, \ldots, x_{k}\right)\right\|_{Y}: x_{1}, \ldots, x_{k} \in X \text { with }\left\|x_{1}\right\|=\cdots=\left\|x_{k}\right\|=1\right\} .
$$

The space of bounded linear isomorphisms from $X$ to $Y$ is denoted by $\mathcal{B}_{\text {isom }}(X ; Y)$ and that of linear isomorphisms by $\mathcal{L}_{\text {isom }}(X ; Y)$. We let $I_{X}: x \mapsto x$ denote the identity on $X$ and $\left\langle x^{*}, x\right\rangle_{X^{*} \times X}=\left\langle x^{*}, x\right\rangle=x^{*}(x)$ denote the duality pairing for $x^{*} \in X^{*}$ and $x \in X$.

We employ the theory of moving hypersurfaces and Riemannian manifolds as given by do Carmo [Car92], Kimura [Kim08], and Prüss and Simonett [PS13]. More background information on differential geometry and the theory of function spaces is given in Appendices A and B.1.

### 1.1. Moving hypersurfaces and integral theorems

In order to define moving domains and hypersurfaces, we consider the initial-value problem

$$
\begin{equation*}
\dot{x}(t)=u(t, x(t)) \text { for } t \in J, \quad x\left(t_{0}\right)=x_{0}, \tag{1.1}
\end{equation*}
$$

where $J$ is an open interval and $u: G \rightarrow \mathbb{R}^{n}(n \in \mathbb{N})$ is a given vector field on an open subset $G$ of $J \times \mathbb{R}^{n}$. It is custom to understand the map $t \mapsto x(t)$ as the trajectory of a moving particle that starts at position $x_{0}$ at time $t_{0}$ and moves with velocity $u(t, x(t))$. We say that $x_{0}$ is the convected coordinate of the moving particle [cf. Old50; Scr60].

A local solution of (1.1) is a $C^{1}$-map $x: J\left(t_{0}, x_{0}\right) \rightarrow \mathbb{R}^{n}$ on some interval $J\left(t_{0}, x_{0}\right) \subset \mathbb{R}$ that contains $t_{0}$, such that $(t, x(t))$ belongs to $G$ for all $t \in J\left(t_{0}, x_{0}\right)$ and such that (1.1) is satisfied on $J\left(t_{0}, x_{0}\right)$. If $(t, x) \mapsto u(t, x)$ is continuous on $G$ and locally Lipschitz with respect to $x$, then the Picard-Lindelöf theorem implies that for every $\left(t_{0}, x_{0}\right) \in G$, there exists a unique local solution on some interval $\left(t_{0}-\delta, t_{0}+\delta\right)$. Moreover, the solution has a unique extension to a maximal interval of existence, which is again denoted by $J\left(t_{0}, x_{0}\right)$. This interval is open and for any finite $t_{*} \in \partial J\left(t_{0}, x_{0}\right)$, the function $(t, x(t))$ tends to $\partial G$ or it blows up as $t \rightarrow t_{*}$, in the sense that $\operatorname{dist}((t, x(t)), \partial G) \rightarrow 0$ or $|x(t)| \rightarrow \infty$.
1.1. Proposition. Let $J \subset \mathbb{R}$ be an open interval, let $G \subset J \times \mathbb{R}^{n}$ be open, let $u \in C(G)^{n}$ be locally Lipschitz with respect to $x$, and let $t \mapsto x_{\left(t_{0}, x_{0}\right)}(t)$ denote the unique solution to (1.1) for $\left(t_{0}, x_{0}\right) \in G$.

Then the map

$$
\Phi:\left(t, t_{0}, x_{0}\right) \mapsto x_{\left(t_{0}, x_{0}\right)}(t), \quad\left\{\left(t, t_{0}, x_{0}\right) \in J \times G: t \in J\left(t_{0}, x_{0}\right)\right\} \rightarrow \mathbb{R}^{n}
$$

has the following properties.
(i) $(t, \Phi(t, s, x))$ belongs to $G$ for all $t \in J(s, x)$ and $(s, x) \in G$.
(ii) $\Phi(t, t, x)=x$ for all $(t, x) \in G$.
(iii) $\Phi(t, s, \Phi(s, r, x))=\Phi(t, r, x)$ for all $(r, x) \in G$ and $t, s \in J(r, x)$.
(iv) $\Phi(\cdot, s, x)$ is continuously differentiable in $J(s, x)$ for all $(s, x) \in G$.
(v) $\Phi(t, \cdot, \cdot)$ is locally Lipschitz in $\{(s, x) \in G: t \in J(s, x)\}$ for all $t \in J$.

Proof. Since $\Phi(\cdot, t, x)$ is a solution, it satisfies (i) and (ii). Next, the functions $\Phi(\cdot, s, \Phi(s, r, x))$ and $\Phi(\cdot, r, x)$ are solutions to (1.1) and coincide at $t=s$, since $\Phi(s, r, x)=\Phi(s, s, \Phi(s, r, x))$. By uniqueness, we have $\Phi(\cdot, s, \Phi(s, r, x))=\Phi(\cdot, r, x)$ on $J(r, x)$. The $C^{1}$-regularity of $\Phi(\cdot, s, x)$ follows from $\partial_{t} \Phi(t, s, x)=u(t, \Phi(t, s, x))$, and the local Lipschitz condition is a consequence of Gronwall's Lemma [see PW10, Satz 4.1.2].
1.2. Remark. We can guarantee that every solution exists for all $t \in J$, when we also assume that $G=J \times \mathbb{R}^{n}$ and that $u$ is linearly bounded with respect to $x$; that is, there are $a, b \in C\left(J ; \mathbb{R}_{+}\right)$ such that $|u(t, x)| \leq a(t)+b(t)|x|$ for all $t \in J, x \in \mathbb{R}^{n}$ [see PW10, Korollar 2.5.1].

The map $\Phi$ induces a local flow in $G$ in the following sense.
1.3. Definition. Let $G$ be a topological space and $U$ be an open subset of $\mathbb{R} \times G$ that contains $\{0\} \times G$. A continuous map $\tilde{\Phi}: U \subset \mathbb{R} \times G \rightarrow G$ is called a local flow in $G$, if
(i) $\tilde{\Phi}(0, z)=z$ for all $z \in G$.
(ii) $\tilde{\Phi}(t+s, z)=\tilde{\Phi}(t, \tilde{\Phi}(s, z))$ for all $(s, z) \in U$ and $t \in \mathbb{R}$ with $(t, \tilde{\Phi}(s, z)) \in U$.

If $\mathcal{U}=\mathbb{R} \times G$, in addition, then we call $\tilde{\Phi}$ a (global) flow in $G$.
1.4. Corollary. In the situation of Proposition 1.1, the mapping

$$
\begin{align*}
\tilde{\Phi}: & \left(s,\left(t_{0}, x_{0}\right)\right) \mapsto\left(t_{0}+s, \Phi\left(t_{0}+s, t_{0}, x_{0}\right)\right), \\
& \left\{\left(s, t_{0}, x_{0}\right) \in \mathbb{R} \times \operatorname{gr} \Omega: t_{0}+s \in J\left(t_{0}, x_{0}\right)\right\} \rightarrow \operatorname{gr} \Omega \tag{1.2}
\end{align*}
$$

is a local flow in $\mathrm{gr} \Omega$. We also call $\Phi$ the flow in $\mathrm{gr} \Omega$ induced by the velocity field $u$.
1.5. Definition. Let $J \subset \mathbb{R}$ be an open interval, $n \in \mathbb{N}$, and $\Omega: J \ni t \mapsto \Omega(t)$ be a set-valued map such that each $\Omega(t)$ is a domain in $\mathbb{R}^{n}$. We call $\Omega$ a moving domain, if there is a flow $\Phi: J \times \operatorname{gr} \Omega \rightarrow$ $\operatorname{gr} \Omega$ induced by some velocity field $u: \operatorname{gr} \Omega \rightarrow \mathbb{R}^{n}$ such that

$$
\Omega(t)=\Phi\left(t, t_{0}, \Omega\left(t_{0}\right)\right):=\left\{\Phi\left(t, t_{0}, x\right): x \in \Omega\left(t_{0}\right)\right\} \quad \text { for all } t, t_{0} \in J .
$$

This definition allows to describe fluid volumes, since $\Omega(t)$ can be obtained by following the trajectories $\Phi\left(\cdot, t_{0}, x_{0}\right)$ of the particles with initial position $x_{0} \in \Omega\left(t_{0}\right)$. We note that there may be different velocity fields that describe the same moving domain; for instance $\Omega(t):=(-t, t)$ moves according to the velocity field $u(t, x)=x / t$, but also according to $u(t, x)=x^{3} / t^{3}$ for $|x| \leq t$. However, the normal component $u(t, x) \cdot \nu_{\partial \Omega(t)}(x)$ at $\partial \Omega(t)$ does not depend on the choice of such a velocity field; a property that holds true for general moving hypersurfaces, which are defined as follows [cf. Kim08, Definition 5.1].
1.6. Definition. Let $J \subset \mathbb{R}$ be an open interval, $n \geq 2$, and $k, l \in \mathbb{N}_{0}$.
(i) A set-valued map $\Gamma$ : $J \ni t \mapsto \Gamma(t)$ is called a moving hypersurface (of class $C^{1}$ ), if each $\Gamma(t)$ is an oriented $C^{1}$-hypersurface in $\mathbb{R}^{n}$ and its graph gr $\Gamma$ is a $C^{1}$-hypersurface in $\mathbb{R}^{1+n}$.
(ii) A moving hypersurface $\Gamma$ is of class $C^{k}\left[C^{(k, l)}\right]$, if all its local height functions $h: J^{\prime} \times$ $U \subset J \times \nu_{0}^{\perp} \rightarrow \operatorname{gr} \Gamma$ in the sense of Definition A. 1 on page 129 are of class $C^{k}\left(J^{\prime} \times U\right)$ $\left[C^{(k, l)}\left(J^{\prime} \times U\right)\right]$.
(iii) A moving hypersurface $\Gamma$ is compact, if each $\Gamma(t)$ is compact.
(iv) A moving hypersurface $\Gamma$ is induced by a moving domain $\Omega: J \in t \mapsto \Omega(t)$ with flow $\Phi: J \times \operatorname{gr} \Omega \rightarrow \operatorname{gr} \Omega$, if $\operatorname{gr} \Gamma \subset \operatorname{gr} \Omega$ and $\Gamma(t)=\Phi\left(t, t_{0}, \Gamma\left(t_{0}\right)\right)$ for all $t, t_{0} \in J$.
Proposition A. 5 implies that for every moving $C^{k+1}$-hypersurface $\Gamma$, the $(n+1)$-dimensional normal $\nu_{\mathrm{gr} \Gamma}$ of $\operatorname{gr} \Gamma$ belongs to the class $C^{k}(\mathrm{gr} \Gamma)^{1+n}$ and the $n$-dimensional normal $\nu_{\Gamma(t)}$ of $\Gamma(t)$ belongs to $C^{k}(\Gamma(t))^{n}$. Later on, we will show that every compact moving $C^{2}$-hypersurface is induced by some flow.
1.7. Proposition ([cf. Kim08, Definition 5.4]). Suppose that $\Gamma$ is a moving $C^{1}$-hypersurface in $\mathbb{R}^{n}$. Then there exists a unique function $V_{\Gamma}: \operatorname{gr} \Gamma \rightarrow \mathbb{R}$, called the normal velocity of $\Gamma$, which satisfies the identity

$$
\begin{equation*}
V_{\Gamma}(t, x)=\gamma^{\prime}(t) \cdot \nu_{\Gamma(t)}(x) \quad \text { for all } t \in J, x \in \Gamma(t), \tag{1.3}
\end{equation*}
$$

for every $C^{1}$-path $(t-\delta, t+\delta) \ni s \mapsto \gamma(s) \in \Gamma(s)$ with $\gamma(t)=x$.
Moreover, the unit normal $\nu_{\operatorname{gr} \Gamma}=\left(\nu_{\mathrm{gr} \Gamma, t}, \nu_{\mathrm{gr} \Gamma, x}\right) \in \mathbb{R}^{1+n}$ of $\mathrm{gr} \Gamma$ is given by

$$
\begin{equation*}
\nu_{\mathrm{gr} \Gamma}=\left(1+V_{\Gamma}^{2}\right)^{-1 / 2}\left(-V_{\Gamma}, \nu_{\Gamma}\right) \quad \text { with } V_{\Gamma}=-\nu_{\mathrm{gr} \Gamma, t}\left(1-\nu_{\mathrm{gr} \Gamma, t}^{2}\right)^{-1 / 2} . \tag{1.4}
\end{equation*}
$$

In particular, if $\Gamma$ is induced by a flow with velocity $u$, then

$$
V_{\Gamma}(t, x)=u(t, x) \cdot \nu_{\Gamma(t)}(x) \quad \text { for all } t \in J, x \in \Gamma(t) .
$$

Proof. Since $\operatorname{gr} \Gamma$ has dimension $n$ and each $\Gamma(t)$ has dimension $n-1$, we conclude that every tangent space $T_{(t, x)} \operatorname{gr} \Gamma$ must have the form $T_{(t, x)} \operatorname{gr} \Gamma=\left\{\nu_{\mathrm{gr} \Gamma}(t, x)\right\}^{\perp}=\mathbb{R} \times \nu_{\mathrm{gr} \Gamma, x}(t, x)^{\perp}$ with $\nu_{\mathrm{gr} \Gamma, x}(t, x) \neq 0$, and hence $\left|\nu_{\operatorname{gr} \Gamma, t}(t, x)\right|<1$. Moreover, for every $(t, x) \in \operatorname{gr} \Gamma$ and $\tau \in T_{x} \Gamma(t)$, the vector $(0, \tau)$ belongs to $T_{(t, x)} \operatorname{gr} \Gamma$, and hence $\nu_{\mathrm{gr} \Gamma, x}(t, x)$ must be parallel to $\nu_{\Gamma(t)}(x)$. Therefore the identity $\left|\nu_{\mathrm{gr} \Gamma}\right|^{2}=\nu_{\mathrm{gr} \Gamma, t}^{2}+\left|\nu_{\mathrm{gr} \Gamma, x}\right|^{2}=1$ yields $\nu_{\mathrm{gr} \Gamma, x}=\left|\nu_{\mathrm{gr} \Gamma, x}\right| \nu_{\Gamma(t)}=\left(1-\nu_{\mathrm{gr} \Gamma, t}^{2}\right)^{1 / 2} \nu_{\Gamma(t)}$.

Uniqueness of $V_{\Gamma}$. Let $V_{\Gamma}$ satisfy (1.3) and consider a $C^{1}$-path $s \mapsto(s, \gamma(s))$ in $\mathrm{gr} \Gamma$ with $\gamma(t)=x$. Then its derivative $\left(1, \gamma^{\prime}(t)\right)$ belongs to $T_{(t, x)} g r \Gamma$ and we have

$$
\begin{equation*}
0=\left(1, \gamma^{\prime}(t)\right) \cdot \nu_{\mathrm{gr} \Gamma}(t, x)=\nu_{\mathrm{gr} \Gamma, t}(t, x)+\left(1-\nu_{\mathrm{gr} \Gamma, t}(t, x)^{2}\right)^{1 / 2} \gamma^{\prime}(t) \cdot \nu_{\Gamma(t)}(x) . \tag{1.5}
\end{equation*}
$$

Since $\left|\nu_{\mathrm{gr} \Gamma, t}\right|$ is smaller than 1 , we obtain $V_{\Gamma}=\gamma^{\prime}(t) \cdot \nu_{\Gamma(t)}(x)=-\nu_{\mathrm{gr} \Gamma, t}\left(1-\nu_{\mathrm{gr} \Gamma, t}^{2}\right)^{-1 / 2}$. Therefore $V_{\Gamma}(t, x)$ is uniquely determined.

Existence of $V_{\Gamma}$. The function $V_{\Gamma}=-\nu_{\mathrm{gr} \Gamma, t}\left(1-\nu_{\mathrm{gr} \Gamma, t}^{2}\right)^{-1 / 2}$ is well-defined and thus (1.5) implies (1.3). Finally, the identity (1.4) follows from those of $\nu_{\mathrm{gr} \Gamma, x}$ and $V_{\Gamma}$ in terms of $\nu_{\mathrm{gr} \Gamma, t}$.
1.8. Proposition. Every compact moving $C^{2}$-hypersurface is induced by some flow.

Proof. Let $\Gamma$ : $J \ni t \mapsto \Gamma(t)$ be a moving $C^{2}$-hypersurface in $\mathbb{R}^{n}$ with normal velocity $V_{\Gamma}$. From $\nu_{\mathrm{gr} \Gamma} \in C^{1}(\mathrm{gr} \Gamma)$ and (1.4) we infer that $u:=V_{\Gamma} \nu_{\Gamma}$ is of class $C^{1}(\mathrm{gr} \Gamma)^{n}$. By compactness of $\Gamma(t)$ and Proposition A.12, the vector field $u(t, \cdot)$ is $L(t)$-Lipschitz on $\Gamma(t)$; that is, we have $|u(t, x)-u(t, y)| \leq L(t)|x-y|$ for all $x, y \in \Gamma(t), t \in J$. Then the McShane-Whitney extension [cf. Hei05, p. 5]

$$
\tilde{u}^{j}(t, x):=\inf _{y \in \Gamma(t)}\left(u^{j}(t, y)+L(t)|x-y|\right) \quad \text { for } j \in\{1, \ldots, n\}, x \in \mathbb{R}^{n}, t \in J
$$

of $u=\left(u^{j}\right)_{j}$ is $\sqrt{n} L(t)$-Lipschitz on $\mathbb{R}^{n}$ and linearly bounded. According to Proposition 1.1 and Remark 1.2, there is a flow $\Phi: J \times J \times \mathbb{R}^{n} \rightarrow J \times \mathbb{R}^{n}$ with velocity $\tilde{u}$, which induces $\Gamma$.

We will frequently employ the following version of the divergence theorem.
1.9. Theorem (Divergence theorem). Let $\Omega \subset \mathbb{R}^{n}(n \in \mathbb{N})$ be a bounded open set with $C^{1}$-boundary $\partial \Omega$ or a bent half-space $\left\{\left(x^{\prime}, x_{n}\right) \in \mathbb{R}^{n}: x_{n}>\omega\left(x^{\prime}\right)\right\}$ with $\omega \in C_{c}^{1}\left(\mathbb{R}^{n-1}\right)$. Then

$$
\begin{equation*}
\int_{\Omega} \operatorname{div} u d x=\int_{\partial \Omega} u \cdot \nu_{\partial \Omega} d(\partial \Omega) \quad \text { for all } u \in H_{1}^{1}\left(\Omega ; \mathbb{R}^{n}\right) \tag{1.6}
\end{equation*}
$$

Here $\int_{\partial \Omega} \ldots d(\partial \Omega)$ denotes integration with respect to the $(n-1)$-dimensional Hausdorff measure on $\partial \Omega$, and $H_{p}^{k}=W_{p}^{k}$ denotes the Sobolev space of order $k \in \mathbb{N}_{0}$ and power $p \in[1, \infty)$.
Proof of Theorem 1.9. The assertion for the case $\partial \Omega \in C^{2}$ and $u \in C_{c}^{1}\left(\mathbb{R}^{n} ; \mathbb{R}^{n}\right)$ is well-known. In the general case, it follows from an approximation argument.

Next, we state the surface divergence theorem for tangential vector fields of class. The Sobolev space $H_{p}^{k}(\Gamma ; T \Gamma)$ of tangential vector fields is defined by means of trivializing coordinate systems for the tangent bundle $T \Gamma$ (see page 163).
1.10. Theorem (Surface divergence theorem [cf. BPS05, Theorem A]). Let $\Gamma \subset \mathbb{R}^{n}$ be a compact $C^{2}$-hypersurface with boundary $\partial \Gamma$ of class $C^{2}$, whose normal within $T \Gamma$ is denoted by $n_{\partial \Gamma}$. Then

$$
\int_{\Gamma} \operatorname{div}_{\Gamma} v d \Gamma=\int_{\partial \Gamma} v \cdot n_{\partial \Gamma} d(\partial \Gamma) \quad \text { for all } v \in H_{1}^{1}(\Gamma ; T \Gamma) .
$$

Proof. By [BPS05, Theorem A], the surface divergence theorem applies to tangential vector fields $v$ of class $C^{1}$, and hence also to $v \in H_{1}^{1}(\Gamma ; T \Gamma)$ by approximation.

The next theorem allows to differentiate integrals $\int_{\Omega(t)} \psi(t, x) d x$ with respect to time. Assume that the velocity of a moving domain $\Omega$ belongs to the Banach space $B U C^{(0,1)}(\operatorname{gr} \Omega)^{n}$ of all bounded, uniformly continuous vector fields $u: \operatorname{gr} \Omega \rightarrow \mathbb{R}^{n}$ whose first-order spatial derivatives are bounded and uniformly continuous on $\operatorname{gr} \Omega$. Then the induced flow $\Phi: J \times \operatorname{gr} \Omega \rightarrow \operatorname{gr} \Omega$ is continuously differentiable [cf. PW10, Satz 4.3.1], its Jacobian $\partial_{x} \Phi$ with respect to the spatial variables is invertible, and we have $\operatorname{det} \partial_{x} \Phi>0$ on $J \times \operatorname{gr} \Omega$. Hence $\Phi\left(t, t_{0}, \cdot\right): \Omega\left(t_{0}\right) \rightarrow \Omega(t)$ is a $C^{1}$-diffeomorphism; that is, a bijective $C^{1}$-map whose Jacobian is invertible everywhere in $\Omega\left(t_{0}\right)$. For a $C^{1}$-function $\psi$ on $\operatorname{gr} \Omega$, the material derivative $D \psi / D t$ with respect to the flow $\Phi$ with velocity $u$ is defined by

$$
\begin{equation*}
\frac{D \psi(t, x)}{D t}:=\left.\frac{d}{d s} \psi(t+s, \Phi(t+s, t, x))\right|_{s=0}=\partial_{t} \psi(t, x)+\left[\partial_{x} \psi(t, x)\right] u(t, x) . \tag{1.7}
\end{equation*}
$$

1.11. Theorem (Reynolds transport theorem). Let $\Omega: J \ni t \mapsto \Omega(t)$ be a moving domain in $\mathbb{R}^{n}$ with velocity $u \in B U C^{(0,1)}(\mathrm{gr} \Omega)^{n}$. Then, given a function $\psi \in H_{1}^{1}(\mathrm{gr} \Omega)$, we have

$$
\begin{equation*}
\frac{d}{d t} \int_{\Omega} \psi d x=\int_{\Omega}\left(\frac{D \psi}{D t}+\psi \operatorname{div} u\right) d x \quad \text { a.e. in } J . \tag{1.8}
\end{equation*}
$$

Proof. Let $\Phi$ denote the flow induced by $u$. For fixed $t$ and $x$, the matrices $Y(s):=\partial_{x} \Phi(t+s, t, x)$ and $A(s):=\partial_{x} u(t+s, \Phi(t+s, t, x))$ satisfy $Y^{\prime}(s)=A(s) Y(s)$ by the chain rule. A well-known identity [see e. g. PW10, Lemma 3.1.2] yields $\operatorname{det} Y^{\prime}(s)=\operatorname{tr} A(s) \operatorname{det} Y(s)$. Thus,

$$
(d / d s) \operatorname{det} \partial_{x} \Phi(t+s, t, x)=\operatorname{div} u(t+s, \Phi(t+s, t, x)) \operatorname{det} \partial_{x} \Phi(t+s, t, x)
$$

Having in mind that $\partial_{x} \Phi(t, t, x)=1$ and that $\Phi(t+s, t, \cdot): \Omega(t) \rightarrow \Omega(t+s)$ is bijective, we conclude that $\Phi(t+s, t, \cdot)$ is a diffeomorphism. Therefore the change of variables formula gives

$$
\begin{equation*}
\int_{\Omega(t+s)} \psi(t+s, y) d y=\int_{\Omega(t)} \psi(t+s, \Phi(t+s, t, x)) \operatorname{det} \partial_{x} \Phi(t+s, t, x) d x . \tag{1.9}
\end{equation*}
$$

By differentiating (1.9) with respect to $s$ at $s=0$, we obtain (1.8).
In Theorem 1.11, the Sobolev space $H_{1}^{1}(\mathrm{gr} \Omega)$ has the usual meaning, since $\mathrm{gr} \Omega$ is an open subset of $\mathbb{R}^{1+n}$; a fact that does not hold true for $\mathrm{gr} \Gamma$ and can not be used for defining anisotropic spaces. Therefore, we employ the diffeomorphism

$$
\tilde{\Phi}_{t_{0}}: J \times \Omega\left(t_{0}\right) \rightarrow \operatorname{gr} \Omega, \quad \tilde{\Phi}_{t_{0}}(t, x)=\left(t, \Phi\left(t, t_{0}, x\right)\right)
$$

and the pull-back $\left(\tilde{\Phi}_{t_{0}}^{*} \psi\right)(t, x):=\left(\psi \circ \tilde{\Phi}_{t_{0}}\right)(t, x)=\psi\left(t, \Phi\left(t, t_{0}, x\right)\right)$, and we assume that $J$ is bounded. Having in mind that $\partial_{x} \Phi$ is bounded and $\operatorname{det} \partial_{x} \Phi$ is strictly positive on gr $\Omega$, we conclude that $\tilde{\Phi}_{t_{0}}^{*}: H_{1}^{1}(\operatorname{gr} \Omega) \rightarrow H_{1}^{1}\left(J \times \Omega\left(t_{0}\right)\right)$ is a topological linear isomorphism. This motivates the definitions

$$
\begin{aligned}
H_{p}^{k}(\operatorname{gr} \Omega) & :=\tilde{\Phi}_{t_{0}}^{*} H_{p}^{k}\left(J \times \Omega\left(t_{0}\right)\right), & H_{p}^{(k, l)}(\operatorname{gr} \Omega):=\tilde{\Phi}_{t_{0}}^{*} H_{p}^{(k, l)}\left(J \times \Omega\left(t_{0}\right)\right), \\
H_{p}^{k}(\operatorname{gr} \Gamma) & :=\tilde{\Phi}_{t_{0}}^{*} H_{p}^{k}\left(J \times \Gamma\left(t_{0}\right)\right), & H_{p}^{(k, l)}(\operatorname{gr} \Gamma):=\tilde{\Phi}_{t_{0}}^{*} H_{p}^{(k, l)}\left(J \times \Gamma\left(t_{0}\right)\right),
\end{aligned}
$$

where $H_{p}^{(k, l)}(J \times X)=H_{p}^{k}\left(J ; L_{p}(X)\right) \cap L_{p}\left(J ; H_{p}^{l}(X)\right)$ for $k, l \in \mathbb{N}_{0}, p \in[1, \infty)$. Their vectorvalued versions are defined as on page 163.

The following theorem allows to differentiate integrals over moving hypersurfaces.
1.12. Theorem (Surface transport theorem [cf. BPS05, Theorem B]). Let $\Gamma: J \ni t \mapsto \Gamma(t)$ be a compact moving $C^{2}$-hypersurface with velocity $u \in B U C^{(0,1)}(\mathrm{gr} \Gamma)^{n}$ and let $\psi \in H_{1}^{1}(\mathrm{gr} \Gamma)$. Then

$$
\frac{d}{d t} \int_{\Gamma} \psi d \Gamma=\int_{\Gamma}\left(\frac{D \psi}{D t}+\psi \operatorname{div}_{\Gamma} u\right) d \Gamma \quad \text { a.e. in } J .
$$

Proof. The special case $u \in B U C^{1}(\operatorname{gr} \Gamma)^{n}$ and $\psi \in C^{1}(\operatorname{gr} \Gamma)$ is treated in [BPS05, Theorem B] and therefore our assertion follows from a straightforward approximation argument.

### 1.2. Derivation of the model

In this section we derive problem ( N ) from integral balance equations and constitutive assumptions. More information on the mathematical modeling of fluid dynamics can be found for instance in [Ari89; And+07; BP10; BPS05; Den94; DS95; Old50; Scr60; SS82; SSO07; Tan93; Tan95].

We consider a bounded domain $\Omega \subset \mathbb{R}^{n}$ that contains a compact moving $C^{2}$-hypersurface $\Gamma(t)$ on a bounded open interval $J \subset \mathbb{R}$. Then we can decompose $\Omega=\Omega_{+}(t) \dot{\cup} \Gamma(t) \dot{\cup} \Omega_{-}(t)$ with moving domains $\Omega_{ \pm}(t)$ (see Corollary A. 19 on page 138). In particular, each $\Gamma(t)$ is a compact subset of $\Omega$ and therefore the interface does not touch the boundary. We may assume that $\partial \Omega_{-}(t)=\Gamma(t)$, and hence $\partial \Omega \subset \partial \Omega_{+}$. Let $u_{ \pm} \in B U C\left(\operatorname{gr} \Omega_{ \pm}\right)^{n}$ be corresponding velocity fields and define

$$
u(t, \cdot): \Omega \backslash \Gamma(t) \rightarrow \mathbb{R}^{n}, \quad u(t, x):=u_{ \pm}(t, x) \quad \text { for } x \in \Omega_{ \pm}(t), t \in J
$$

For the sake of brevity, we omit the argument $t$ if no confusion seems likely; that is, we write $\Omega \backslash \Gamma$ and $\Gamma$ instead of $\Omega \backslash \Gamma(t)$ and $\Gamma(t)$ when we consider some fixed $t$, and we understand that

$$
\left.u\right|_{\Gamma}(t, x)=\left.u(t, \cdot)\right|_{\Gamma(t)}(x)=\left.u\right|_{\operatorname{gr~} \Gamma}(t, x) \quad \text { for } x \in \Gamma(t), t \in J .
$$

Let $\nu_{ \pm}$denote the outward normal on $\partial \Omega_{ \pm}$and let $\nu_{\Gamma}=\nu_{-}=-\nu_{+}$denote the normal at $\Gamma$. With the Sobolev space $H_{p}^{k}=W_{p}^{k}$ of order $k \in \mathbb{N}_{0}$ and exponent $p \in[1, \infty)$, we write

$$
u \in H_{p}^{k}\left(\Omega \backslash \Gamma ; \mathbb{R}^{n}\right) \quad \text { if and only if } u_{+} \in H_{p}^{k}\left(\Omega_{+} ; \mathbb{R}^{n}\right) \text { and } u_{-} \in H_{p}^{k}\left(\Omega_{-} ; \mathbb{R}^{n}\right)
$$

Other function spaces on $\Omega \backslash \Gamma$ are defined analogously. The jump of $u \in H_{p}^{1}\left(\Omega \backslash \Gamma ; \mathbb{R}^{n}\right)$ on $\Gamma$,

$$
\llbracket u \rrbracket:=\left.u_{+}\right|_{\Gamma}-\left.u_{-}\right|_{\Gamma},
$$

is well-defined in the sense of traces. Then the following divergence theorem applies.
1.13. Theorem (Divergence theorem with interface). Let $\Omega \subset \mathbb{R}^{n}$ be an open set with $C^{1}$-boundary such that the divergence theorem (1.6) is valid and let $\Gamma \subset \Omega$ be a $C^{1}$-hypersurface. Then

$$
\int_{\Omega \backslash \Gamma} \operatorname{div} u d x=\int_{\partial \Omega} \nu_{\partial \Omega} \cdot u d(\partial \Omega)-\int_{\Omega \cap \Gamma} \nu_{\Gamma} \cdot \llbracket u \rrbracket d \Gamma \quad \text { for all } u \in H_{1}^{1}\left(\Omega \backslash \Gamma ; \mathbb{R}^{n}\right) .
$$

Proof. This follows by separating the integral over $\Omega \backslash \Gamma$ into integrals over $\Omega_{+}$and $\Omega_{-}$, and by applying the divergence theorem to the separate integrals.
1.2.1. Balance equations. Our next goal is to derive differential balance equations for a scalar quantity $\psi: J \times \Omega \rightarrow \mathbb{R}$ that satisfies certain integral balance equations. In order to apply the previous integral theorems, we assume that $\psi$ is of class $H_{1}^{1}(J \times \Omega)$ and that

$$
u_{ \pm} \in B U C^{(0,1)}\left(\operatorname{gr} \Omega_{ \pm}\right)^{n} ;
$$

that is, the vector fields $u_{+}$and $u_{-}$and their first-order spatial derivatives are bounded and uniformly continuous. We further assume that $u$ is continuous across $\Gamma$ and that $\Gamma$ is advected with the flow induced by $u$; that is,

$$
\llbracket u \rrbracket=0 \text { on } \Gamma, \quad V_{\Gamma}=\left.u\right|_{\Gamma} \cdot \nu_{\Gamma} .
$$

We also assume that $\left.\nu_{\partial \Omega} \cdot u\right|_{\partial \Omega}=0$, so that $\Omega$ is a trivial moving domain with velocity $u$.
We consider the density $\psi(t, x)$ of an extensive scalar quantity like the mass density $\rho$ or the kinetic energy density $\rho|u|^{2}$. Let $V$ be a control volume in $\Omega$; that is, a moving domain $V: J \ni t \mapsto V(t) \subset \Omega$ with the same velocity $u$. Suppose that $\psi \in H_{1}^{1}(J \times \Omega)$ satisfies an integral balance equation

$$
\begin{equation*}
\frac{d}{d t} \int_{V} \psi d x=\int_{V} g d x+\int_{V \cap \Gamma} g_{\Gamma \rightarrow \Omega} d \Gamma-\int_{\partial V} j \cdot \nu_{\partial V} d(\partial V) \quad \text { a. e. in } J \tag{1.10}
\end{equation*}
$$

for every control volume $V$ with appropriate quantities $g, g_{\Gamma \rightarrow \Omega}$, and $j$. Here
(i) $\int_{V} g d x$ are the sources of $\psi$ in $V$ with volume density $g$,
(ii) $\int_{\Gamma \cap V} g_{\Gamma \rightarrow \Omega} d \sigma_{\Gamma}$ are the sources of $\psi$ on $\Gamma \cap V$ with surface density $g_{\Gamma \rightarrow \Omega}$, and
(iii) $\int_{\partial V} j \cdot \nu_{\partial V} d(\partial V)$ is the molecular flow of $\psi$ through $\partial V$ with flux $j$.

It is sufficient to impose the regularity assumptions

$$
\begin{equation*}
\psi \in H_{1}^{1}(J \times \Omega), j \in H_{1}^{(0,1)}\left(J \times \Omega ; \mathbb{R}^{n}\right), g \in L_{1}(J \times \Omega), g_{\Gamma \rightarrow \Omega} \in L_{1}(\operatorname{gr} \Gamma) . \tag{1.11}
\end{equation*}
$$

We wish to derive a differential balance from (1.10). First, Theorem 1.13 yields

$$
\frac{d}{d t} \int_{V} \psi d x=\int_{V}(g-\operatorname{div} j) d x+\int_{V \cap \Gamma}\left(g_{\Gamma \rightarrow \Omega}-\llbracket j \rrbracket \cdot \nu_{\Gamma}\right) d \Gamma .
$$

With the transport theorem (1.8) we obtain the identity

$$
\begin{equation*}
\int_{V}\left(\frac{\partial \psi}{\partial t}+\operatorname{div}(\psi u+j)-g\right) d x+\int_{V \cap \Gamma}\left(\llbracket j \rrbracket \cdot \nu_{\Gamma}-g_{\Gamma \rightarrow \Omega}\right) d \Gamma=0 . \tag{1.12}
\end{equation*}
$$

For fixed $t$, equation (1.12) is valid for every bounded smooth subset $V(t)$ of $\Omega \backslash \Gamma(t)$. From the Lebesgue's integration theory we infer that the first integrand must vanish almost everywhere in $\Omega \backslash \Gamma(t)$. Therefore the following differential balance equation is valid a. e. in $J \times \Omega$.

$$
\partial_{t} \psi+\operatorname{div}(\psi u+j)=g \quad \text { in } \Omega \backslash \Gamma .
$$

Hence the surface integral in (1.12) vanishes for every time $t$ and every control volume $V$ in $\Omega$. It is not difficult to show that every domain in $\Gamma(t)$ with $C^{2}$-boundary can be represented as $V(t) \cap \Gamma$ with some control volume $V$. Therefore the following jump condition is satisfied.

$$
\llbracket j \rrbracket \cdot \nu_{\Gamma}=g_{\Gamma \rightarrow \Omega} \quad \text { on } \Gamma .
$$

Next, assume that there are a scalar surface density $\psi_{\Gamma}(t, x)$ for $x \in \Gamma(t)$ and quantities $g_{\Gamma}$ and $j_{\Gamma}$ such that the following surface integral balance equation is valid for every control volume.

$$
\begin{equation*}
\frac{d}{d t} \int_{V \cap \Gamma} \psi_{\Gamma} d \Gamma=\int_{V \cap \Gamma}\left(g_{\Gamma}-g_{\Gamma \rightarrow \Omega}\right) d \Gamma-\int_{C=\partial V \cap \Gamma} j_{\Gamma} \cdot n_{C} d C . \tag{1.13}
\end{equation*}
$$

Here $g_{\Gamma}$ is the interface source density, the integral $\int_{C=\partial V \cap \Gamma} j_{\Gamma} \cdot n_{C} d C$ is the molecular flow through the ( $n-2$ )-dimensional surface $C=\Gamma \cap \partial V$ with outward normal $n_{C}(t, x) \in T_{x} \Gamma(t)$ and the interface flux $j_{\Gamma}$ is tangential vector field on $\Gamma$. Sufficient regularity conditions are

$$
\begin{equation*}
\psi_{\Gamma} \in H_{1}^{1}(\mathrm{gr} \Gamma), j_{\Gamma} \in H_{1}^{(0,1)}(\operatorname{gr} \Gamma ; T \Gamma), g_{\Gamma} \in L_{1}(\mathrm{gr} \Gamma), \tag{1.14}
\end{equation*}
$$

We combine (1.10) and (1.13) to

$$
\begin{align*}
\frac{d}{d t}\left(\int_{V} \psi d x+\int_{V \cap \Gamma} \psi_{\Gamma} d \Gamma\right)= & \int_{V} g d x+\int_{V \cap \Gamma} g_{\Gamma} d \Gamma  \tag{1.15}\\
& -\int_{\partial V} j \cdot \nu_{\partial V} d(\partial V)-\int_{\partial V \cap \Gamma} j_{\Gamma} \cdot n_{C} d C .
\end{align*}
$$

Again, we wish to derive the differential balance equation that corresponds to (1.13). First, the surface divergence theorem yields

$$
\frac{d}{d t} \int_{V \cap \Gamma} \psi_{\Gamma} d \Gamma=\int_{V \cap \Gamma}\left(-\operatorname{div}_{\Gamma} j_{\Gamma}+g_{\Gamma}-g_{\Gamma \rightarrow \Omega}\right) d \Gamma,
$$

and the surface transport theorem (Theorem 1.12) implies

$$
\int_{V \cap \Gamma}\left(\frac{D \psi_{\Gamma}}{D t}+\psi_{\Gamma} \operatorname{div}_{\Gamma} u+\operatorname{div}_{\Gamma} j_{\Gamma}-g_{\Gamma}+g_{\Gamma \rightarrow \Omega}\right) d \Gamma=0 .
$$

Since $V$ is arbitary, we obtain the surface differential balance equation

$$
D \psi_{\Gamma} / D t+\psi_{\Gamma} \operatorname{div}_{\Gamma} u+\operatorname{div}_{\Gamma} j_{\Gamma}=g_{\Gamma}-g_{\Gamma \rightarrow \Omega} \quad \text { on } \Gamma .
$$

Consequently, we have shown that if the quantities $\psi, \psi_{\Gamma}, j, j_{\Gamma}, g$, and $g_{\Gamma}$ satisfy the integral balance equations (1.10) and (1.13) and the regularity conditions (1.11) and (1.14), then these quantities also satisfy the differential balance equations

$$
\begin{align*}
\partial_{t} \psi+\operatorname{div}(\psi u+j) & =g & & \text { in } \Omega \backslash \Gamma,  \tag{1.16a}\\
\llbracket j \rrbracket \cdot \nu_{\Gamma} & =g_{\Gamma \rightarrow \Omega} & & \text { on } \Gamma,  \tag{1.16b}\\
D \psi_{\Gamma} / D t+\psi_{\Gamma} \operatorname{div}_{\Gamma} u+\operatorname{div}_{\Gamma} j_{\Gamma} & =g_{\Gamma}-g_{\Gamma \rightarrow \Omega} & & \text { on } \Gamma . \tag{1.16c}
\end{align*}
$$

1.2.2. Balance of mass. In order to derive the balance equations for the mass from the differential balances (1.16), we let $\psi=\rho$ and $\psi_{\Gamma}=\rho_{\Gamma}$ and obtain the continuity equation

$$
\partial_{t} \rho+\operatorname{div}(\rho u+j)=g \quad \text { in } \Omega \backslash \Gamma .
$$

In this thesis we study the incompressible case $\rho=$ constant and $j(\rho)=0$. We also neglect interface mass and therefore let $\rho_{\Gamma}=0$ and $j_{\Gamma}(\rho)=0$. Assuming that $\Omega$ represents a closed system, we further neglect sources of mass; that is, $g(\rho)=0$ and $g_{\Gamma}(\rho)=0$. Hence

$$
\operatorname{div} u=0 \quad \text { in } \Omega \backslash \Gamma .
$$

1.2.3. Balance of momentum. The momentum density $\psi=\rho u$ is not scalar and thus we can not apply (1.16) directly. Instead, we consider the scalar densities $\psi(e):=\psi \cdot e=\rho e \cdot u$ for suitable vector fields $e$. This well-known approach was modified by Scriven [Scr60] for deriving the Boussinesq-Scriven law. For every constant vector $e$ we have $\operatorname{div}(\psi(e) u)=\partial_{i}\left(\left(\rho e_{j} u_{j}\right) u_{i}\right)=$ $e_{j} \partial_{i}\left(\rho u_{i} u_{j}\right)=e \cdot \operatorname{div}(\rho u \otimes u)$. We neglect external forces such as gravity and therefore let $g(e)=0$. It will suffice to assume that

$$
\begin{equation*}
u_{ \pm} \in B U C^{(0,1)}\left(\operatorname{gr} \Omega_{ \pm}\right)^{n} \cap H_{1}^{(1,0)}\left(J \times \Omega_{ \pm}\right)^{n} \cap H_{1}^{(0,2)}\left(\operatorname{gr} \Omega_{ \pm}\right)^{n} \tag{1.17}
\end{equation*}
$$

Then $\psi(e)$ belongs to $H_{1}^{1}(J \times \Omega)$ and (1.16a) implies

$$
e \cdot \partial_{t}(\rho u)+e \cdot \operatorname{div}(\rho u \otimes u)=\operatorname{div} j(e) \quad \text { in } \Omega \backslash \Gamma,
$$

for every constant vector $e$. We shall prescribe a flux of the form $j(e)=j_{\pi}(e)+j_{S}(e)=e \cdot T$ that consists of a pressure part $j_{\pi}(e)$ and a viscous part $j_{S}(e)$. The quantity $T$ is the stress tensor.

Consider a control volume $V$ in $\Omega$ with $V(t) \subset \Omega \backslash \Gamma(t)$. Then either $\bar{V}(t) \subset \Omega_{+}(t)$ or $\bar{V}(t) \subset \Omega_{-}(t)$. One force acting on $V$ is the pressure force $f_{\pi}=-\int_{\partial V} \pi \nu_{\partial V} d(\partial V)$ with pressure

$$
\begin{equation*}
\pi_{ \pm} \in H_{1}^{(0,1)}\left(\operatorname{gr} \Omega_{ \pm}\right) \tag{1.18}
\end{equation*}
$$

Here $\pi \nu_{\partial V} d(\partial V)$ can be understood as the pressure force which acts on a surface element perpendicular to $\nu_{\partial V}$. For every constant vector $e$, the divergence theorem yields

$$
e \cdot f_{\pi}=-\int_{\partial V} \pi e \cdot \nu_{\partial V} d(\partial V)=-\int_{V} \operatorname{div}(\pi e) d x
$$

In view of the second identity, we let $j_{\pi}(e):=-\pi e$. The tensor $J_{\pi}:=-\pi I$ yields the desired linear relations $j_{\pi}(e)=e \cdot J_{\pi}$ and $\operatorname{div} j_{\pi}(e)=e \cdot \operatorname{div} J_{\pi}$ with respect to $e$.

Due to friction on $\partial V$, there is another force acting on $V$, the stress $f_{S}=-\int_{\partial V} S \nu_{\partial V} d(\partial V)$ with the viscous stress tensor $S$. We assume that both $\Omega_{ \pm}$consist of Newtonian fluids, which means that the viscous stress tensor depends linearly on the rate-of-strain tensor

$$
D:=D(u):=\operatorname{sym}[\nabla u]=2^{-1}\left(\nabla u+[\nabla u]^{\top}\right) .
$$

Therefore we define the viscous stress tensor

$$
S:=S(u):=2 \mu D(u)=\mu\left(\nabla u+[\nabla u]^{\top}\right),
$$

where the number $\mu_{ \pm}$is the shear viscosity of the fluid $\Omega_{ \pm}$. If $e$ is constant, then

$$
e \cdot f_{S}=\int_{\partial V} e \cdot S \nu_{\partial V} d(\partial V)=\int_{V} \operatorname{div}(S e) d x .
$$

Thus we let $j_{S}(e):=S e$ and $J_{S}:=S$ and hence $\operatorname{div} j_{S}(e)=\operatorname{div}\left(e \cdot J_{S}\right)=e \cdot \operatorname{div} S$. We call

$$
\begin{equation*}
T:=T(u, \pi):=J_{S}+J_{\pi}=2 \mu D(u)-\pi I \tag{1.19}
\end{equation*}
$$

the (total) stress tensor. Since the vector $e$ is arbitrary, we obtain the differential momentum balance

$$
\begin{equation*}
\partial_{t}(\rho u)+\operatorname{div}(\rho u \otimes u-T)=0 \quad \text { in } \Omega \backslash \Gamma . \tag{1.20}
\end{equation*}
$$

1.2.4. Interface momentum balance. We recall that the momentum density $\psi=\rho u$ induces scalar densities $\psi(e)=e \cdot \psi$, whose bulk fluxes are $j(e)=e \cdot T$ for a given vector field $e$. The latter is allowed to have a possibly non-tangential restriction $\left.e\right|_{\Gamma}=e^{\alpha} \tau_{\alpha}+e_{\nu} \nu_{\Gamma}$ on $\Gamma$. Here we differ from Scriven [Scr60, p. 101] and Aris [Ari89, p. 238], who only considered vector fields with vanishing covariant derivatives, which do not cover constant vectors unless $\Gamma$ is flat. Since the interface has vanishing mass density, we have $\psi_{\Gamma}:=\left.\rho_{\Gamma} u\right|_{\Gamma}=0$ and then equations (1.16b) and (1.16c) yield

$$
\begin{equation*}
-e \cdot \llbracket T \rrbracket \nu_{\Gamma}=\operatorname{div}_{\Gamma}\left(j_{\Gamma}(e)\right) \quad \text { on } \Gamma \tag{1.21}
\end{equation*}
$$

for every vector field $e$. Here the interface flux $j_{\Gamma}(e)=j_{\Gamma, \sigma}(e)+j_{\Gamma, S}(e)$ will consist of a surface tension part $j_{\Gamma, \sigma}(e)$ and a viscous part $j_{\Gamma, S}(e)$. We first let $j_{\Gamma, \sigma}(e):=e \cdot\left(\sigma P_{\Gamma}\right)$, where $\sigma$ is the constant surface tension coefficient. If $e=e_{0} \in \mathbb{R}^{n}$ is constant, then $\operatorname{div}_{\Gamma}\left(j_{\Gamma, \sigma}(e)\right)=e \cdot \operatorname{div}_{\Gamma}\left(\sigma P_{\Gamma}\right)$.

We define the viscous flux $j_{\Gamma, S}(e):=e \cdot S_{\Gamma}$ with viscous surface stress tensor $S_{\Gamma}$. Following Scriven [Scr60], we regard $\Gamma$ as an $(n-1)$-dimensional fluid with rate-of-strain tensor

$$
D_{\Gamma}:=D_{\Gamma}(u):=2^{-1} D g_{\alpha \beta} / D t \tau^{\alpha} \otimes \tau^{\beta} .
$$

Similar to Sekomb and Skalak [SS82], we can derive the usual expression of $D_{\Gamma}$ in Euclidean coordinates. For every parametrization $y \mapsto \varphi(y)$ of $\Gamma(t)$, the map $y \mapsto \Phi(t+s, t, \varphi(y))$ is a parametrization of $\Gamma(t+s)$. Thus the tangent vectors of $\Gamma(t+s)$ are related to those of $\Gamma(t)$ by

$$
\tau_{i}(t+s, \Phi(t+s, t, x))=\partial_{x_{i}} \Phi(t+s, t, x)=\partial_{x} \Phi(t+s, t, x) \tau_{i}(t, x) .
$$

Having in mind that $\nabla_{\Gamma} u:=\tau_{\Gamma}^{j} \otimes \partial_{j} u$, we obtain

$$
(D / D t) \tau_{i}(t, x)=\left.(d / d s)\left[\partial_{x} \Phi(t+s, t, x)\right]\right|_{s=0} \tau_{i}(t, x)=\left[\nabla_{\Gamma} u(t, x)\right]^{\top} \tau_{i}(t, x) .
$$

Then the relations $g_{i j}=\tau_{i} \cdot \tau_{j}$ and $P_{\Gamma}=\tau_{i} \otimes \tau^{i}=I-\nu_{\Gamma} \otimes \nu_{\Gamma}$ yield

$$
D_{\Gamma}=\operatorname{sym}\left(P_{\Gamma}\left[\nabla_{\Gamma} u\right] P_{\Gamma}\right)=2^{-1} P_{\Gamma}\left(\nabla_{\Gamma} u+\left[\nabla_{\Gamma} u\right]^{\top}\right) P_{\Gamma} .
$$

Scriven [Scr60] proposed to consider Newtonian surface fluids for which $S_{\Gamma}$ depends linearly on $D_{\Gamma}$. Hence we define the viscous surface stress tensor

$$
S_{\Gamma}:=S_{\Gamma}(u):=\left(\lambda_{s}-\mu_{s}\right)\left(\operatorname{div}_{\Gamma} u\right) P_{\Gamma}+2 \mu_{s} D_{\Gamma},
$$

where $\lambda_{s}$ and $\mu_{s}$ are constant real numbers. The (total) surface stress tensor is defined by

$$
\begin{equation*}
T_{\Gamma}:=T_{\Gamma}(u):=\sigma P_{\Gamma}+S_{\Gamma}(u)=\sigma P_{\Gamma}+\left(\lambda_{s}-\mu_{s}\right)\left(\operatorname{div}_{\Gamma} u\right) P_{\Gamma}+2 \mu_{s} D_{\Gamma} . \tag{1.22}
\end{equation*}
$$

Then the flux $j_{\Gamma}(e)=e \cdot T_{\Gamma}$ satisfies $\operatorname{div}_{\Gamma}\left(j_{\Gamma}(e)\right)=e \cdot \operatorname{div}_{\Gamma} T_{\Gamma}+T_{\Gamma}: \nabla_{\Gamma} e$. In Section 1.3 we will see that $j_{\Gamma}(e)$ belongs to the class $H_{1}^{(0,1)}(\operatorname{gr} \Gamma ; T \Gamma)$, provided that

$$
\begin{equation*}
v=\left.P_{\Gamma} u\right|_{\Gamma} \in H_{1}^{(0,2)}(\operatorname{gr} \Gamma ; T \Gamma), \quad w=\left.\nu_{\Gamma} \cdot u\right|_{\Gamma} \in H_{1}^{(0,1)}(\operatorname{gr} \Gamma), \quad \Gamma(t) \in C^{3} . \tag{1.23}
\end{equation*}
$$

By choosing the constant vectors $e=e_{i}$ in (1.21), we obtain the interface momentum balance

$$
\begin{equation*}
-\llbracket T \rrbracket \nu_{\Gamma}=\operatorname{div}_{\Gamma} T_{\Gamma} \quad \text { on } \Gamma . \tag{1.24}
\end{equation*}
$$

By imposing the no-slip condition $\left.u\right|_{\partial \Omega}=0$, the derivation of the model $(\mathrm{N})$ is complete.

### 1.3. Properties of the model

Similar to [BP10], we will decompose the interface momentum balance (1.24) into tangential and normal parts and derive an energy identity in arbitrary control volumes; but, in contrast to [BP10], we employ covariant derivatives.

Let each $\Gamma(t)$ be of class $C^{3}$. According to Einstein's summation convention, we always sum over repeated greek indices $\alpha, \beta, \ldots \in\{1, \ldots, n-1\}$, whereas latin indices $i, j, \ldots \in\{1, \ldots, n-$ $1\}$ denote free indices. We will use the Weingarten tensor $L=l_{\alpha \beta} \tau^{\alpha} \otimes \tau^{\beta}=l^{\alpha \beta} \tau_{\alpha} \otimes \tau_{\beta}$, the mean curvature $H=g^{\alpha \beta} l_{\alpha \beta}$ and the Cristoffel symbols $\Lambda_{i j, k}=\partial_{i} \tau_{j} \cdot \tau_{k}$ and $\Lambda_{i j}^{k}=\partial_{i} \tau_{j} \cdot \tau^{k}=g^{k l} \Lambda_{i j, l}$. Then we define covariant derivatives as follows: For a tangential vector field $v \in C^{1}(\Gamma ; T \Gamma)$ and a co-vector field $\omega \in C^{1}\left(\Gamma ; T^{*} \Gamma\right)$, we let

$$
\begin{aligned}
v_{; k} & =\widetilde{\nabla}_{k} v=P_{\Gamma} \partial_{k} v=v^{\alpha}{ }_{; k} \tau_{\alpha}=\left(\partial_{k} v^{\alpha}+\Lambda_{k \beta}^{\alpha} v^{\beta}\right) \tau_{\alpha}, \\
\omega_{; k} & =\widetilde{\nabla}_{k} \omega=P_{\Gamma} \partial_{k} \omega=\omega_{\alpha ; k} \tau^{\alpha}=\left(\partial_{k} \omega_{\alpha}-\Lambda_{k \alpha}^{\beta} \omega_{\beta}\right) \tau^{\alpha} ;
\end{aligned}
$$

for a possibly non-tangential vector field $u=v+w \nu_{\Gamma} \in C^{1}\left(\Gamma ; \mathbb{R}^{n}\right)$, we let

$$
u_{; k}=\widetilde{\nabla}_{k} u=P_{\Gamma} \partial_{k} u=v_{; k}^{\alpha} \tau_{\alpha}+w \partial_{k} \nu=\left(\partial_{k} v^{\alpha}+\Lambda_{k \beta}^{\alpha} v^{\beta}-w l_{k \beta} g^{\beta \alpha}\right) \tau_{\alpha} ;
$$

and for second-order tensor fields $T \in C^{1}(\Gamma ; T \Gamma \otimes T \Gamma)$ and $D \in C^{1}\left(\Gamma ; T^{*} \Gamma \otimes T^{*} \Gamma\right)$, we let

$$
\begin{aligned}
T_{; k} & =\widetilde{\nabla}_{k} T=T_{; k}^{\alpha \beta} \tau_{\alpha} \otimes \tau_{\beta}=\left(\partial_{k} T^{\alpha \beta}+\Lambda_{k \gamma}^{\alpha} T^{\gamma \beta}+\Lambda_{k \gamma}^{\beta} T^{\alpha \gamma}\right) \tau_{\alpha} \otimes \tau_{\beta}, \\
D_{; k} & =\widetilde{\nabla}_{k} D=D_{\alpha \beta ; k} \tau^{\alpha} \otimes \tau^{\beta}=\left(\partial_{k} D_{\alpha \beta}-\Lambda_{k \alpha}^{\gamma} D_{\gamma \beta}-\Lambda_{k \beta}^{\gamma} D_{\alpha \gamma}\right) \tau^{\alpha} \otimes \tau^{\beta} .
\end{aligned}
$$

The usage of covariant derivatives (i) ensures that the derivative of a section of some bundle is again a section of that bundle, (ii) provides the simple relations

$$
\begin{equation*}
g_{i j ; k}=0, \quad g_{; k}^{i j}=0, \tag{1.25}
\end{equation*}
$$

and (iii) provides the general product rule

$$
\begin{equation*}
\left(T^{i_{1} \ldots}{ }_{j_{1} \ldots} S^{k_{1} \ldots l_{1} \ldots}\right)_{; m}=T^{i_{1} \ldots}{ }_{j_{1} \ldots ; m} S^{k_{1} \ldots l_{1} \ldots}+T^{i_{1} \ldots}{ }_{j_{1} \ldots} S^{k_{1} \ldots \ldots} l_{l_{1} \ldots m ; m} . \tag{1.26}
\end{equation*}
$$

Some relations to surface differential operators are given by

$$
\begin{align*}
\operatorname{div}_{\Gamma}\left(v^{\alpha} \tau_{\alpha}+w \nu_{\Gamma}\right) & =v_{; \alpha}^{\alpha}-w H,  \tag{1.27a}\\
D_{\Gamma}\left(v^{\alpha} \tau_{\alpha}+w \nu_{\Gamma}\right) & =2^{-1} \tau^{\alpha} \otimes \tau^{\beta}\left(v_{\alpha ; \beta}+v_{\beta ; \alpha}\right)-w L, \\
\operatorname{div}_{\Gamma}\left(T^{\alpha \beta} \tau_{\alpha} \otimes \tau_{\beta}\right) & =T^{\alpha \beta}{ }_{; \alpha} \tau_{\beta}+T^{\alpha \beta} l_{\alpha \beta} \nu_{\Gamma} \quad\left(\text { if } T^{\alpha \beta}=T^{\beta \alpha}\right), \\
l_{i j ; k} & =l_{i k ; j}=l_{j k ; i .} .
\end{align*}
$$

Second-order covariant derivatives are denoted by $\widetilde{\nabla}_{k} \widetilde{\nabla}_{l}=(\cdot)_{; l k}$. The covariant derivatives of tangential vector fields do not necessarily commute, but satisfy the relations

$$
\begin{array}{rlrl}
v^{i}{ }_{; j k}-v^{i}{ }_{; k j}=R^{i}{ }_{\alpha j k} v^{\alpha}, & v_{i ; j k}-v_{i ; k j} & =-v_{\alpha} R^{\alpha}{ }_{i j k}, \\
R^{i}{ }_{j k l} & =g^{i \alpha} R_{\alpha j k l}, & R_{i j k l} & =l_{i k} l_{j l}-l_{i l} l_{j k} . \tag{1.28b}
\end{array}
$$

The Laplace-Beltrami operators for $\psi \in C^{2}(\Gamma)$ and $u=v^{\alpha} \tau_{\alpha}+w \nu_{\Gamma} \in C^{2}(\Gamma)^{n}$ are given by

$$
\begin{aligned}
& \Delta_{\Gamma} \psi=\operatorname{div}_{\Gamma} \nabla_{\Gamma} \psi \\
& \widetilde{\Delta}_{\Gamma} u=g^{\alpha \beta}\left(\partial_{\alpha} \partial_{\beta} \psi-\Lambda_{\alpha \beta}^{\gamma} \partial_{\gamma} \psi\right), \\
& \widetilde{\nabla}_{\alpha} \widetilde{\nabla}_{\beta} u=\left(g^{\alpha \beta} v^{\gamma} ; \alpha \beta-\partial_{\alpha} w l^{\alpha \gamma}-w H_{; \delta} g^{\delta \gamma}\right) \tau_{\gamma} .
\end{aligned}
$$

We refer to Appendix A. 4 for more information on these identities.
1.3.1. Decomposition of the interface balance. We decompose (1.24) into its tangential and normal parts. From (1.22), (1.27a), and (1.27b), we infer that $T_{\Gamma}=T_{\Gamma}^{\alpha \beta} \tau_{\alpha} \otimes \tau_{\beta}$ has the components

$$
\begin{equation*}
T_{\Gamma}^{\alpha \beta}=\sigma g^{\alpha \beta}+\left(\lambda_{s}-\mu_{s}\right)\left(v^{\gamma} ; \gamma-H w\right) g^{\alpha \beta}+\mu_{s} g^{\alpha \gamma} g^{\beta \delta}\left(v_{\gamma ; \delta}+v_{\delta ; \gamma}-2 w l_{\gamma \delta}\right) . \tag{1.29}
\end{equation*}
$$

With equations (1.25), (1.26), and (1.27c), we decompose div ${ }_{\Gamma} T_{\Gamma}=T_{\Gamma}^{\alpha \beta}{ }_{; \alpha} \tau_{\beta}+T_{\Gamma}^{\alpha \beta} l_{\alpha \beta} \nu_{\Gamma}$ as

$$
\begin{aligned}
& {\left[\mu_{s} g^{\alpha \gamma} g^{\beta \delta}\left(v_{\delta ; \gamma \alpha}+v_{\gamma ; \delta \alpha}-2 w l_{\gamma \delta ; \alpha}-2 w_{; \alpha} l_{\gamma \delta}\right)+\left(\lambda_{s}-\mu_{s}\right)\left(v_{; \gamma \alpha}^{\gamma}-H_{; \alpha} w-H w_{; \alpha}\right) g^{\alpha \beta}\right] \tau_{\beta}} \\
& \quad+\left[\sigma H+\left(\lambda_{s}-\mu_{s}\right)\left(v^{\gamma} ; \gamma-H w\right) H+\mu_{s} l_{\alpha \beta} g^{\alpha \gamma} g^{\beta \delta}\left(v_{\gamma ; \delta}+v_{\delta ; \gamma}-2 w l_{\gamma \delta}\right)\right] \nu_{\Gamma} .
\end{aligned}
$$

Let us rewrite this equation in vector notation. We have $g^{\alpha \gamma} g^{\beta \delta} v_{\delta ; \gamma \alpha} \tau_{\beta}=\widetilde{\Delta}_{\Gamma} v$, and with (1.28a) and (1.28b), we obtain $g^{\alpha \gamma} g^{\beta \delta} v_{\gamma ; \delta \alpha}=\nabla_{\Gamma}$ div $_{\Gamma} v$. Identity (1.27d) yields $g^{\alpha \gamma} g^{\beta \delta} l_{\gamma \delta ; \alpha} \tau_{\beta}=\nabla_{\Gamma} H$. We proceed in a similar way with the remaining terms and obtain

$$
\begin{align*}
\operatorname{div}_{\Gamma} T_{\Gamma}= & \mu_{s} \widetilde{\Delta}_{\Gamma} v+\lambda_{s} \nabla_{\Gamma} \operatorname{div}_{\Gamma} v \\
& -\left(\lambda_{s}+\mu_{s}\right) w \nabla_{\Gamma} H+\left[\left(\mu_{s}-\lambda_{s}\right) H-2 \mu_{s} L\right] \nabla_{\Gamma} w \\
& +\left[\left(\lambda_{s}-\mu_{s}\right) \operatorname{div}_{\Gamma} v H+2 \mu_{s} L: D_{\Gamma}(v)\right] \nu_{\Gamma}  \tag{1.30}\\
& +\left[\sigma H-\left(\lambda_{s}-\mu_{s}\right) w H^{2}-2 \mu_{s} w \operatorname{tr}\left(L^{2}\right)\right] \nu_{\Gamma} .
\end{align*}
$$

We conclude that the interface momentum balance (1.24) has the tangential part

$$
\begin{align*}
-P_{\Gamma} \llbracket T \rrbracket \nu_{\Gamma} & =-\llbracket \mu \rrbracket\left[\nabla_{\Gamma} v\right] \nu_{\Gamma}-\llbracket \mu \rrbracket \nabla_{\Gamma} w-\llbracket \mu \partial_{\nu} v \rrbracket \\
& =\mu_{s} \widetilde{\Delta}_{\Gamma} v+\lambda_{s} \nabla_{\Gamma} \operatorname{div}_{\Gamma} v-\left(\lambda_{s}+\mu_{s}\right) w \nabla_{\Gamma} H+\left[\left(\mu_{s}-\lambda_{s}\right) H-2 \mu_{s} L\right] \nabla_{\Gamma} w, \tag{1.31a}
\end{align*}
$$

and the normal part

$$
\begin{align*}
-\nu_{\Gamma} \cdot \llbracket T \rrbracket \nu_{\Gamma} & =-2 \llbracket \mu \partial_{\nu} w \rrbracket+\llbracket \pi \rrbracket  \tag{1.31b}\\
& =\sigma H+\left(\lambda_{s}-\mu_{s}\right) \operatorname{div}_{\Gamma} u H+2 \mu_{s} D_{\Gamma}: L .
\end{align*}
$$

1.3.2. Energy identity. We consider the kinetic energy $\int_{V} 2^{-1} \rho|u|^{2} d x$ of a control volume $V$ in $\Omega$. By applying the transport theorem, the divergence theorem, the identity $\operatorname{div} u=0$, and the differential momentum balance (1.20), we obtain the kinetic energy balance

$$
\begin{equation*}
\frac{d}{d t} \int_{V} \frac{\rho}{2}|u|^{2} d x=-\int_{V} 2 \mu D: D d x+\int_{\partial V} T u \cdot \nu_{\partial V} d(\partial V)-\int_{V \cap \Gamma} \llbracket T u \rrbracket \cdot \nu_{\Gamma} d \Gamma . \tag{1.32}
\end{equation*}
$$

In view of the integral balance (1.10), we see that the scalar quantity $\psi=2^{-1} \rho|u|^{2}$ has the bulk source density $g=-2 \mu D: D$, the bulk flux $j=-T u$, and the interface source density $g_{\Gamma \rightarrow \Omega}=-\llbracket T u \rrbracket \cdot \nu_{\Gamma}$. The interface momentum balance (1.24) and identity (A.19) imply

$$
\begin{aligned}
-\llbracket T u \rrbracket \cdot \nu_{\Gamma} & =u \cdot \operatorname{div}_{\Gamma} T_{\Gamma}=\operatorname{div}_{\Gamma}\left(T_{\Gamma} u\right)-T_{\Gamma}: D_{\Gamma} \\
& =\operatorname{div}_{\Gamma}\left(T_{\Gamma} u\right)-\sigma \operatorname{div}_{\Gamma} u-\left(\lambda_{s}-\mu_{s}\right)\left(\operatorname{div}_{\Gamma} u\right)^{2}-2 \mu_{s} D_{\Gamma}: D_{\Gamma} .
\end{aligned}
$$

Thus, the surface transport theorem and the surface divergence theorem yield the energy identity

$$
\begin{align*}
\frac{d}{d t} & \left(\int_{V} \frac{\rho}{2}|u|^{2} d x+\int_{V \cap \Gamma} \sigma d \Gamma\right) \\
\quad= & -\int_{V} 2 \mu D: D d x-\int_{V \cap \Gamma}\left(\left(\lambda_{s}-\mu_{s}\right)\left(\operatorname{div}_{\Gamma} u\right)^{2}+2 \mu_{s} D_{\Gamma}: D_{\Gamma}\right) d \Gamma  \tag{1.33}\\
& +\int_{\partial V} T u \cdot \nu_{\partial V} d(\partial V)+\int_{C=\partial V \cap \Gamma} T_{\Gamma} u \cdot n_{C} d C .
\end{align*}
$$

In the special case $V=\Omega$ and imposing the no-slip boundary condition $\left.u\right|_{\partial \Omega}=0$, we recover the energy identity from [BP10, Theorem 3.1],

$$
\begin{align*}
& \frac{d}{d t}\left(\int_{\Omega} \frac{\rho}{2}|u|^{2} d x+\int_{\Gamma} \sigma d \Gamma\right)  \tag{1.34}\\
& \quad=-\int_{\Omega} 2 \mu D: D d x-\int_{\Gamma}\left(\left(\lambda_{s}-\mu_{s}\right)\left(\operatorname{div}_{\Gamma} u\right)^{2}+2 \mu_{s} D_{\Gamma}: D_{\Gamma}\right) d \Gamma .
\end{align*}
$$

By comparing (1.33) with the general integral balance (1.15), we see that the energy has the bulk density $\psi=2^{-1} \rho|u|^{2}$, the bulk flux $j=-T u$, the interface density $\psi_{\Gamma}=\sigma$ and the interface flux $j_{\Gamma}=-T_{\Gamma} u=-T_{\Gamma} v$. Moreover, if $\lambda_{s} \geq \mu_{s} \geq 0$, then the bulk source density $g=-2 \mu D: D$ and the interface source density $g_{\Gamma}=-\left(\lambda_{s}-\mu_{s}\right)\left(\operatorname{div}_{\Gamma} u\right)^{2}-2 \mu_{s} \operatorname{tr}\left(D_{\Gamma}^{2}\right)$ are non-positive and thus responsible for dissipation.

## CHAPTER 2

## Linear elliptic transmission problems

In this chapter we investigate the elliptic transmission problem (TP) in both a strong and a weak sense. We restate problem (TP) as the strong transmission problem

$$
\left\{\begin{align*}
-\operatorname{div}(\mu \nabla u) & =f & & \text { in } \Omega \backslash \Sigma,  \tag{2.1}\\
\mu \partial_{\nu} u & =g & & \text { on } \partial \Omega, \\
\llbracket \mu \partial_{\nu} u \rrbracket & =h_{1} & & \text { on } \Sigma, \\
\llbracket u \rrbracket & =h_{2} & & \text { on } \Sigma,
\end{align*}\right.
$$

considered in a domain $\Omega$ that contains a $C^{1}$-hypersurface $\Sigma$. Here $u: \Omega \backslash \Sigma \rightarrow \mathbb{K}$ is an unknown scalar field, $\left(f, g, h_{1}, h_{2}\right)$ are given data, $\mu: \Omega \backslash \Sigma \rightarrow(0, \infty)$ is a variable coefficient, and the jump $\llbracket . \rrbracket$ was defined on page 16. We also study the weak transmission problem

$$
\left\{\begin{align*}
\int_{\Omega} \mu \nabla u \cdot \nabla \phi d x & =\langle F \mid \phi\rangle & & \text { for all } \phi \in \mathcal{D}\left(\mathbb{R}^{n}\right),  \tag{2.2}\\
\llbracket u \rrbracket & =h_{2} & & \text { on } \Sigma,
\end{align*}\right.
$$

for given data $\left(F, h_{2}\right)$. We will see that (2.2) can be obtained from (2.1) by multiplying the first equation with $\phi$ and integrating by parts. In the case $\Sigma=\emptyset$ and $\mu_{ \pm}=1$, problem (2.2) is called the weak Neumann problem. Both problems (2.1) and (2.2) can be used to eliminate the pressure and divergence in the more complex linear problem (PL); we adopt this strategy from Köhne, Prüss, and Wilke [KPW13; Wil13]. Both problems were solved in [KPW13] for constant coefficients $\mu_{ \pm}=1 / \rho_{ \pm}$in a bounded domain $\Omega$ and the authors established optimal $H_{p}^{2}$-regularity for (2.1), optimal $\dot{H}_{p}^{1}$-regularity for (2.2), and optimal $W_{p}^{2+s}$-regularity for (2.1) under the restriction $\left(g, h_{1}, h_{2}\right)=0$. Similar transmission problems are investigated in the forthcoming monograph [PS15].

Our goal is to prove that both (2.1) and (2.2) have optimal regularity in the sense that the solution-to-data maps $u \mapsto\left(f, g, h_{1}, h_{2}\right)$ and $u \mapsto\left(F, h_{2}\right)$ are topological linear isomorphisms between suitable Banach spaces. We impose the following basic assumption on $\Omega$ and $\Sigma$.
2.1. Assumption. $\Omega \subset \mathbb{R}^{n}(n \geq 2)$ is a domain with $C^{1}$-boundary $\partial \Omega$ and $\Sigma \subset \Omega$ is a closed $C^{1}$-hypersurface such that one of the following conditions is satisfied.
(i) $\Omega$ is the whole space $\mathbb{R}^{n}$ and $\Sigma$ is empty.
(ii) $\Omega$ is a bent half-space $\mathbb{R}_{\omega}^{n}=\left\{\left(x^{\prime}, x_{n}\right) \in \mathbb{R}^{n}: x_{n}>\omega\left(x^{\prime}\right)\right\}$ with $\omega \in C_{c}^{1}\left(\mathbb{R}^{n-1}\right)$ and $\Sigma$ is empty.
(iii) $\Omega$ is the whole space $\mathbb{R}^{n}, \Sigma$ is a bent hyperplane $\Sigma_{\omega}=\left\{\left(x^{\prime}, \omega\left(x^{\prime}\right)\right): x^{\prime} \in \mathbb{R}^{n-1}\right\}$ with $\omega \in$ $C_{c}^{1}\left(\mathbb{R}^{n-1}\right)$, and $\Omega \backslash \Sigma$ consists of the bent half-spaces $\Omega_{ \pm}=\left\{\left(x^{\prime}, x_{n}\right) \in \mathbb{R}^{n}: x_{n} \gtrless \omega\left(x^{\prime}\right)\right\}$.
(iv) $\Omega$ is a bounded domain with $C^{1}$-boundary, $\Sigma$ is compact and possibly empty, and $\Omega \backslash \Sigma$ consists of disjoint open sets $\Omega_{ \pm}$with $\partial \Omega \subset \partial \Omega_{+}$and $\Sigma=\partial \Omega_{-}$.
We let $\nu_{\partial \Omega}, \nu_{\partial \Omega_{ \pm}}$, and $\nu_{\Sigma}$ denote the exterior unit normal fields on $\partial \Omega, \partial \Omega_{ \pm}$, and $\Sigma$, and we choose the orientation of $\Sigma$ such that $\nu_{\Sigma}=-\nu_{\partial \Omega_{+}}=\nu_{\partial \Omega_{-}}$on $\Sigma$.

In order to define suitable solution spaces, we recall that $u$ belongs to $H_{p}^{1}(\Omega \backslash \Sigma)$ if and only if its restrictions $u_{ \pm}=\left.u\right|_{\Omega_{ \pm}}$belong to $H_{p}^{1}\left(\Omega_{ \pm}\right)$. Other function spaces on $\Omega \backslash \Sigma$ are defined analogously. For an open subset $G \subset \mathbb{R}^{n}$ we consider the vector space

$$
\dot{\mathcal{H}}_{p}^{k}(G):=\left\{u \in H_{1, l o c}^{k}(\bar{G}): \nabla^{k} u \in L_{p}(G)\right\}, \quad \text { for } k \in \mathbb{N}_{0}, p \in[1, \infty) .
$$

This space is semi-normed with respect to $\left\|\nabla^{k} \cdot\right\|_{p}$. We call a function $u: \Omega \backslash \Sigma \rightarrow \mathbb{K}$ a strong solution to (2.1), if $u$ belongs to the space $\left(\dot{\mathcal{H}}_{p}^{2} \cap \dot{\mathcal{H}}_{p}^{1}\right)(\Omega \backslash \Sigma):=\dot{\mathcal{H}}_{p}^{2}(\Omega \backslash \Sigma) \cap \dot{\mathcal{H}}_{p}^{1}(\Omega \backslash \Sigma)$ and if (2.1) is satisfied in the sense of distributions. In particular, the first equation is understood in $\mathcal{D}^{\prime}(\Omega \backslash \Sigma)$; that is,

$$
-\int_{\Omega} \operatorname{div}(\mu \nabla u) \phi d x=\int_{\Omega} f \phi d x \quad \text { for all } \phi \in \mathcal{D}(\Omega \backslash \Sigma)
$$

where $\mathcal{D}(\Omega \backslash \Sigma)$ denotes the space of smooth functions in $\Omega$ that vanish near $\partial \Omega \cup \Sigma$. Obviously, every constant function is a strong solution of (2.1) for vanishing data; hence, we shall choose a semi-norm on $\left(\dot{\mathcal{H}}_{p}^{2} \cap \dot{\mathcal{H}}_{p}^{1}\right)(\Omega \backslash \Sigma)$ whose null-space consists of all constant functions. Such a semi-norm is given by

$$
\|u\|_{\mathbb{E}^{0}}:=\left\|\nabla^{2} u\right\|_{L_{p}(\Omega)}+\|\nabla u\|_{L_{p}(\Omega)}+\|\llbracket u \rrbracket\|_{L_{p}\left(\Sigma^{\prime}\right)}
$$

where $\Sigma^{\prime} \subset \Sigma$ is a bounded open subset with $C^{1}$-boundary that has positive measure, provided that $\Sigma \neq \emptyset$. We will prove that strong solutions are uniquely determined within the space

$$
\mathbb{E}^{0}:=\left(\left(\dot{\mathcal{H}}_{p}^{2} \cap \dot{\mathcal{H}}_{p}^{1}\right)(\Omega \backslash \Sigma),\|\cdot\|_{\mathbb{E}^{0}}\right) / \mathbb{K}
$$

We will also study strong solutions within spaces of lower or higher regularity

$$
\mathbb{E}^{k}:=\left(\bigcap_{j=1}^{k+2} \dot{\mathcal{H}}_{p}^{j}(\Omega \backslash \Sigma),\|\cdot\|_{\mathbb{E}^{k}}\right) / \mathbb{K}, \quad\|u\|_{\mathbb{E}^{k}}:=\sum_{j=1}^{k+2}\left\|\nabla^{j} u\right\|_{p}+\|\llbracket u \rrbracket\|_{L_{p}\left(\Sigma^{\prime}\right)}, \quad k \in \mathbb{N}_{0} \cup\{-1\} .
$$

Next, we derive suitable conditions on the data $\left(f, g, h_{1}, h_{2}\right)$ that are necessary for the existence of a strong solution $u \in \mathbb{E}^{0}$ of (2.1). We assume in addition that $\partial \Omega$ and $\Sigma$ are of class $C^{2-}$ and that $\mu$ belongs to $W_{\infty}^{1}(\Omega \backslash \Sigma)$; that is, $\mu_{ \pm}$are weakly differentiable in $\Omega_{ \pm}$and both $\mu_{ \pm}$ and $\nabla \mu_{ \pm}$belong to $L_{\infty}\left(\Omega_{ \pm}\right)$; thus, $\mu_{ \pm}$are Lipschitz functions. Given a strong solution $u \in \mathbb{E}^{0}$ of problem (2.1), the corresponding data ( $f, g, h_{1}, h_{2}$ ) satisfy the regularity conditions

$$
\left(f, g, h_{1}, h_{2}\right) \in L_{p}(\Omega) \times W_{p}^{1-1 / p}(\partial \Omega) \times W_{p}^{1-1 / p}(\Sigma) \times\left(\dot{\mathcal{W}}_{p}^{2-1 / p}(\Sigma) \cap \dot{\mathcal{W}}_{p}^{1-1 / p}(\Sigma) \cap L_{p}\left(\Sigma^{\prime}\right)\right) .
$$

Here the semi-normed Sobolev-Slobodeckir spaces $\left(\dot{\mathcal{W}}_{p}^{k+s}(\Sigma), \llbracket \nabla^{k} \cdot \rrbracket_{W_{p}^{s}(\Sigma)}\right)$ are defined by

$$
\dot{\mathcal{W}}_{p}^{k+s}(\Sigma):=\left\{u \in H_{p, \mathrm{loc}}^{k}(\Sigma): \llbracket \nabla^{k} u \rrbracket_{W_{p}^{s}(\Sigma)}<\infty\right\} \quad \text { for } k \in \mathbb{N}_{0}, s \in(0,1), p \in[1, \infty),
$$

and the semi-norm $\llbracket \cdot \rrbracket_{W_{p}^{s}(\Sigma)}$ is defined intrinsically by

$$
\llbracket v \rrbracket_{W_{p}^{s}(\Sigma)}:=\left(\int_{\Sigma} \int_{\Sigma} \frac{|v(x)-v(y)|^{p}}{\operatorname{dist}_{\Sigma}(x, y)^{n+s p}} d \Sigma(x) d \Sigma(y)\right)^{1 / p} .
$$

Bothe and Prüss [BP07] noticed that another joint regularity condition for $\left(f, g, h_{1}\right)$ is necessary. Indeed, let $\phi \in \mathcal{D}\left(\mathbb{R}^{n}\right)$ be a test function. Then an integration by parts yields

$$
\int_{\Omega} \mu \nabla u \cdot \nabla \phi d x=-\int_{\Omega} \operatorname{div}(\mu \nabla u) \phi d x+\int_{\partial \Omega} \mu \partial_{\nu} u \phi d(\partial \Omega)-\int_{\Sigma} \llbracket \mu \partial_{\nu} u \rrbracket \phi d \Sigma .
$$

The right-hand side can be expressed in terms of the data as a functional

$$
\begin{equation*}
\left\langle F_{\left(f, g, h_{1}\right)} \mid \phi\right\rangle:=\int_{\Omega} f \phi d x+\int_{\partial \Omega} g \phi d(\partial \Omega)-\int_{\Sigma} h_{1} \phi d \Sigma=\int_{\Omega} \mu \nabla u \cdot \nabla \phi d x . \tag{2.3}
\end{equation*}
$$

Thus the triple $\left(f, g, h_{1}\right)$ induces a continuous linear functional $\phi \mapsto\left\langle F_{\left(f, g, h_{1}\right)} \mid \phi\right\rangle$ on the normed vector space $\left(\mathcal{D}\left(\mathbb{R}^{n}\right),\|\nabla \cdot\|_{L_{p^{\prime}}(\Omega)}\right)$, where $1 / p+1 / p^{\prime}=1$. We can also define such a functional by

$$
\left\langle F_{\mu \nabla u} \mid \phi\right\rangle:=\int_{\Omega} \mu \nabla u \cdot \nabla \phi d x \quad \text { for } \phi \in \mathcal{D}\left(\mathbb{R}^{n}\right) .
$$

The completion of $\left(\mathcal{D}\left(\mathbb{R}^{n}\right),\|\nabla \cdot\|_{L_{p^{\prime}}(\Omega)}\right)$ is the homogeneous Sobolev space

$$
\dot{H}_{p^{\prime}}^{1}(\Omega):=\dot{\mathcal{H}}_{p^{\prime}}^{1}(\Omega) / \mathbb{K}, \quad\|\phi\|_{\dot{H}_{p^{\prime}}^{1}(\Omega)}:=\|\nabla \phi\|_{L_{p^{\prime}}(\Omega)},
$$

considered modulo constant functions [Gal11; Sob63]. Its topological dual space is denoted by

$$
\hat{H}_{p}^{-1}(\Omega):=\dot{H}_{p^{\prime}}^{1}(\Omega)^{*}, \quad\|F\|_{\hat{H}_{p}^{-1}(\Omega)}=\sup _{0 \neq \phi \in \dot{H}_{p^{\prime}}^{1}(\Omega)} \frac{|\langle F \mid \phi\rangle|}{\|\nabla \phi\|_{L_{p^{\prime}}(\Omega)}} .
$$

The data $\left(f, g, h_{1}, h_{2}\right)$ must therefore satisfy the joint regularity condition

$$
\begin{equation*}
F_{\left(f, g, h_{1}\right)} \in \hat{H}_{p}^{-1}(\Omega) . \tag{2.4}
\end{equation*}
$$

If $\Omega$ is bounded, then this regularity condition reduces to the compatibility condtion

$$
\begin{equation*}
\int_{\Omega} f d x+\int_{\partial \Omega} g d(\partial \Omega)-\int_{\Sigma} h_{1} d \Sigma=0 . \tag{2.5}
\end{equation*}
$$

Indeed, by choosing $\phi=1$, we see that (2.4) implies (2.5). For proving the converse implication, we consider a test function $\phi \in \mathcal{D}\left(\mathbb{R}^{n}\right)$, we let $\langle\phi\rangle_{\Omega}:=|\Omega|^{-1} \int_{\Omega} \phi d x$ denote the mean value of $\phi$, and we recall the Poincaré-Wirtinger inequality

$$
\begin{equation*}
\left\|\phi-\langle\phi\rangle_{\Omega}\right\|_{p^{\prime}} \leq C\left(\Omega, p^{\prime}\right)\|\nabla \phi\|_{p^{\prime}} \quad \text { for } \phi \in H_{p^{\prime}}^{1}(\Omega) . \tag{2.6}
\end{equation*}
$$

Then, for a given tuple $\left(f, g, h_{1}\right) \in L_{p}(\Omega) \times L_{p}(\partial \Omega) \times L_{p}(\Sigma)$ satisfying (2.5), inequality (2.6) yields

$$
\left|\left\langle F_{\left(f, g, h_{1}\right)} \mid \phi\right\rangle\right|=\left|\left\langle F_{\left(f, g, h_{1}\right)} \mid \phi-\langle\phi\rangle_{\Omega}\right\rangle\right| \leq C\|\nabla \phi\|_{p^{\prime}} ;
$$

that is, $F_{\left(f, g, h_{1}\right)}$ belongs to $\hat{H}_{p}^{-1}(\Omega)$. In this sense, (2.4) and (2.5) are equivalent, if $\Omega$ is bounded.
For a strong solution of class $\mathbb{E}^{k}\left(k \in \mathbb{N}_{0}\right)$, the corresponding data belong to the spaces

$$
\begin{aligned}
& \mathbb{F}_{\mathrm{cc}}^{k}:=\left\{\left(f, g, h_{1}, h_{2}\right) \in \mathbb{F}^{k}: F_{\left(f, g, h_{1}\right)} \in \hat{H}_{p}^{-1}(\Omega)\right\} \\
& \mathbb{F}^{k}:=H_{p}^{k}(\Omega \backslash \Sigma) \times W_{p}^{k+1-1 / p}(\partial \Omega) \times W_{p}^{k+1-1 / p}(\Sigma) \times\left(\bigcap_{j=0}^{k+1} \dot{\mathcal{W}}_{p}^{j+1-1 / p}(\Sigma) \cap L_{p}\left(\Sigma^{\prime}\right)\right) .
\end{aligned}
$$

Now we are ready to state the main result for the strong transmission problem (2.1).
2.2. Theorem (Optimal $H_{p}^{k+2}$-regularity for (2.1)). Let $\Omega$ and $\Sigma$ satisfy Assumption 2.1, let $k \in \mathbb{N}_{0}$, suppose that $\partial \Omega$ and $\Sigma$ are of class $C^{k+2-}$, and let $p \in(1, \infty)$.

If $\Omega$ is bounded, then for given $\mu \in W_{\infty}^{k+1}(\Omega \backslash \Sigma)$ with $\mu_{0} \leq \mu \leq \mu_{0}^{-1}$, the solution-to-data map

$$
\begin{equation*}
u \mapsto\left(-\operatorname{div}(\mu \nabla u), \mu \partial_{\nu} u, \llbracket \mu \partial_{\nu} u \rrbracket, \llbracket u \rrbracket\right), \quad \mathbb{E}^{k} \rightarrow \mathbb{F}_{c c}^{k} \tag{2.7}
\end{equation*}
$$

is a topological linear isomorphism.
If $\Omega$ is unbounded, then for given $\mu_{0} \in(0,1]$ there exists $\eta>0$ such that if
(i) $\omega \in C_{c}^{k+2-}\left(\mathbb{R}^{n-1}\right)$ with $\|\nabla \omega\|_{\infty} \leq \eta$ in case $\Omega=\mathbb{R}_{\omega}^{n}$ or $\Sigma=\Sigma_{\omega}$,
(ii) $\mu \in W_{\infty}^{k+1}(\Omega \backslash \Sigma)$ with $\mu_{0} \leq \mu \leq \mu_{0}^{-1}$ and $\left\|\mu_{ \pm}-\mu_{ \pm}^{*}\right\|_{\infty} \leq \eta$ for some $\mu_{ \pm}^{*} \in\left[\mu_{0}, \mu_{0}^{-1}\right]$,
then the map (2.7) is a topological linear isomorphism.
In order to prove Theorem 2.2, we first establish a corresponding result for the regularized operator $\lambda-\operatorname{div}(\mu \nabla \cdot)$ with some sufficiently large $\lambda>0$ (see Theorem 2.18) by means of a localization procedure as in [LSU68; AHS94; DHP03; KPW13]. For the case $\lambda=0$ we employ a spectral theoretic argument as in [KPW13; Wil13] and the localization procedure of Simader and Sohr [SS92]. Our main result on the weak transmission problem (2.2) is the following.
2.3. Theorem (Optimal $H_{p}^{1}$-regularity for (2.2)). Let $\Omega$ and $\Sigma$ satisfy Assumption 2.1 and let $p \in$ $(1, \infty)$.

If $\Omega$ is bounded, then for given $\mu_{ \pm} \in C\left(\bar{\Omega}_{ \pm}\right)$with $\inf \mu_{ \pm}>0$, the solution-to-data map

$$
\begin{equation*}
u \mapsto\left(F_{\mu \nabla u}, \llbracket u \rrbracket\right), \quad \mathbb{E}^{-1} \rightarrow \mathbb{F}_{c c}^{-1}=\hat{H}_{p}^{-1}(\Omega) \times\left(\dot{\mathcal{W}}_{p}^{1-1 / p}(\Sigma) \cap L_{p}\left(\Sigma^{\prime}\right)\right), \tag{2.8}
\end{equation*}
$$

is a topological linear isomorphism.
If $\Omega$ is unbounded, then for given $\mu_{0} \in(0,1]$ there exists $\eta>0$ such that if
(i) $\omega \in C_{c}^{1}\left(\mathbb{R}^{n-1}\right)$ with $\|\nabla \omega\|_{\infty} \leq \eta$ in case $\Omega=\mathbb{R}_{\omega}^{n}$ or $\Sigma=\Sigma_{\omega}$,
(ii) $\mu_{ \pm} \in L_{\infty}\left(\Omega_{ \pm}\right)$with $\mu_{0} \leq \mu_{ \pm} \leq \mu_{0}^{-1}$ and $\left\|\mu_{ \pm}-\mu_{ \pm}^{*}\right\|_{\infty} \leq \eta$ for some $\mu_{ \pm}^{*} \in\left[\mu_{0}, \mu_{0}^{-1}\right]$,
then the map (2.8) is a topological linear isomorphism.

### 2.1. The strong transmission problem for $\lambda-\operatorname{div}(\mu \nabla \cdot)$

We consider the linear operator

$$
A_{\lambda}: u \mapsto\left(\lambda u-\operatorname{div}(\mu \nabla u), \mu \partial_{\nu} u, \llbracket \mu \partial_{\nu} u \rrbracket, \llbracket u \rrbracket\right) \quad \text { for } \lambda \in \mathbb{C} \backslash \mathbb{R}_{-},
$$

which is induced by the strong transmission problem

$$
\left\{\begin{align*}
\lambda u-\operatorname{div}(\mu \nabla u) & =f & & \text { in } \Omega \backslash \Sigma,  \tag{2.9}\\
\mu \partial_{\nu} u & =g & & \text { on } \partial \Omega, \\
\llbracket \mu \partial_{\nu} u \rrbracket & =h_{1} & & \text { on } \Sigma, \\
\llbracket u \rrbracket & =h_{2} & & \text { on } \Sigma .
\end{align*}\right.
$$

Our goal is to prove that $A_{\lambda}$ is a topological linear isomorphism from the solution space

$$
\mathbb{E}^{k}=\mathbb{E}^{k}(\Omega \backslash \Sigma):=H_{p}^{k+2}(\Omega \backslash \Sigma) \quad \text { for } k \in \mathbb{N}_{0},
$$

onto the space of data

$$
\mathbb{F}^{k}=\mathbb{F}^{k}(\Omega \backslash \Sigma):=H_{p}^{k}(\Omega \backslash \Sigma) \times W_{p}^{k+1-1 / p}(\partial \Omega) \times W_{p}^{k+1-1 / p}(\Sigma) \times W_{p}^{k+2-1 / p}(\Sigma),
$$

provided that $|\lambda|$ is sufficiently large and $\partial \Omega, \Sigma$, and $\mu$ are sufficiently regular. We identify

$$
\mathbb{F}^{k}(\Omega \backslash \Sigma) \cong \begin{cases}H_{p}^{k}\left(\mathbb{R}^{n}\right) & \text { if } \Omega=\mathbb{R}^{n}, \Sigma=\emptyset \\ H_{p}^{k}(\Omega) \times W_{p}^{k+1-1 / p}(\partial \Omega) & \text { if } \Omega \neq \mathbb{R}^{n}, \Sigma=\emptyset \\ H_{p}^{k}\left(\mathbb{R}^{n} \backslash \Sigma\right) \times W_{p}^{k+1-1 / p}(\Sigma) \times W_{p}^{k+2-1 / p}(\Sigma) & \text { if } \Omega=\mathbb{R}^{n}, \Sigma \neq \emptyset\end{cases}
$$

Our strategy to solve problem (2.9) is based on solving basic model problems, perturbed model problems, and on localization. In a basic model problem, we assume that $\mu$ is constant, $\Omega$ is the whole space $\mathbb{R}^{n}$ or a half-space $\mathbb{R}_{+}^{n}$, and $\Sigma$ is a hyperplane $\mathbb{R}^{n-1} \times\{0\}$ or empty. In a perturbed model problem, we also allow for bent half-spaces $\Omega=\mathbb{R}_{\omega}^{n}=\left\{\left(x^{\prime}, x_{n}\right) \in \mathbb{R}^{n}: x_{n}>\right.$ $\left.\omega\left(x^{\prime}\right)\right\}$, bent hyperplanes $\Sigma=\Sigma_{\omega}=\left\{\left(x^{\prime}, \omega\left(x^{\prime}\right)\right): x^{\prime} \in \mathbb{R}^{n-1}\right\}$, and variable coefficients with small oscillations. In a small region of $\bar{\Omega}$, problem (2.9) looks like a perturbed model problem, after an appropriate rotation and translation. Hence, if these perturbed model problems have appropriate "local" solution operators, then we can construct a "global" solution operator for problem (2.9) in terms of the local solution operators. Such a localization technique is provided in Section 2.1.1.

During the localization procedure, we have to control leading-order and lower-order perturbations, and this can be achieved by using a smallness parameter $\eta$ and $\lambda$-dependent norms for $\mathbb{E}_{\lambda}^{k}$ and $\mathbb{F}_{\lambda}^{k}$, as defined in Section 2.1.2. These norms have useful scaling properties and allow to reduce the operator $A_{\lambda}$ to $A_{1}$ for the basic model problems. Hence, if $A_{1}$ is invertible, then $A_{\lambda}$ is uniformly invertible with respect to $\lambda$. The basic model problems for $\Sigma=\emptyset$ are well-known and we therefore turn our attention to the flat-interface model problem in Section 2.1.3. It is solved by means of the Fourier transform and with the joint $\mathcal{H}^{\infty}$ functional calculus. In Section 2.1.4, we investigate the perturbed model problem for $\Sigma=\Sigma_{\omega}$ with variable coefficient and derive the corresponding results for the remaining model problems. Here the
parameter $\eta$ bounds the oscillations of the coefficient $\mu$ and the gradient of $\omega$ and allows to control leading-order perturbations, whereas the parameter $\lambda$ is used to control lower-order perturbations. Finally, we prove optimal regularity for problem (2.9) in a bounded domain in Section 2.1.5.
2.1.1. Localization technique. We provide a localization technique that allows to invert a "global" operator $A_{\lambda}: E \rightarrow F$ having invertible "local" versions $A_{\lambda, j}: E_{j} \rightarrow F_{j}$. This technique is similar to the corresponding procedures in [LSU68; AHS94; DHP03].
2.4. Definition. Let $E$ and $E_{j}(j \in J \subset \mathbb{N})$ be Banach spaces, let $q \in[1, \infty)$, and define

$$
\mathbf{E}:=\prod_{j} E_{j}, \quad l_{q}(\mathbf{E}):=\left\{\left(x_{j}\right)_{j \in J} \in \mathbf{E}:\|x\|_{l_{q}(\mathbf{E})}<\infty\right\}, \quad\|x\|_{l_{q}(\mathbf{E})}:=\left(\sum_{j \in J}\left\|x_{j}\right\|_{E_{j}}^{q}\right)^{1 / q}
$$

Let further $\Phi_{E, j} \in \mathcal{B}\left(E ; E_{j}\right)$ and $\Psi_{E, j} \in \mathcal{B}\left(E_{j} ; E\right)$ be bounded linear operators such that

$$
\sum_{j \in J} \Psi_{E, j} \Phi_{E, j} x=x \quad \text { for all } x \in E
$$

where the series converges in $E$; and suppose that the maps

$$
\begin{aligned}
& r_{E}: l_{q}(\mathbf{E}) \rightarrow E, \quad\left(x_{j}\right)_{j \in J} \mapsto \sum_{j \in J} \Psi_{E, j} x_{j}, \\
& r_{E}^{c}: \quad E \rightarrow l_{q}(\mathbf{E}), \quad x \mapsto\left(\Phi_{E, j} x\right)_{j \in J},
\end{aligned}
$$

are linear and bounded (hence, $r_{E}$ is a retraction with co-retraction $r_{E}^{c}$ ). Then we say that the triple $\left(\mathbf{E},\left(\Phi_{E, j}\right)_{j \in J},\left(\Psi_{E, j}\right)_{j \in J}\right)$ is an $l_{q}$-approximation system for $E$.

The spaces $E$ and $E_{j}$ are related to linear operators $A_{\lambda}$ and $A_{\lambda, j}$ as follows.
2.5. Assumption. (i) $E$ and $F$ are Banach spaces over the same scalar field $\mathbb{K} \in\{\mathbb{R}, \mathbb{C}\}$, which have $l_{q}$-approximation systems $\left(\mathbf{E},\left(\Phi_{E, j}\right)_{j \in J},\left(\Psi_{E, j}\right)_{j \in J}\right)$ and $\left(\mathbf{F},\left(\Phi_{F, j}\right)_{j \in J},\left(\Psi_{F, j}\right)_{j \in J}\right)$ for some $q \in[1, \infty)$.
(ii) For some unbounded set $\Lambda \subset \mathbb{K}$, the families $\left\{\|\cdot\|_{X, \lambda}: \lambda \in \Lambda\right\}$ consist of equivalent norms on $X \in\left\{E, F, E_{j}, F_{j}: j \in J\right\}$ and we have

$$
\sup _{\lambda \in \Lambda}\left\|r_{E}\right\|_{\left.\mathcal{B}\left(l_{q}(\mathbf{E}) ; E\right)\right), \lambda}<\infty, \quad \sup _{\lambda \in \Lambda}\left\|r_{F}^{c}\right\|_{\mathcal{B}\left(F ; l_{q}(\mathbf{F})\right), \lambda}<\infty .
$$

(iii) $A_{\lambda}: E \rightarrow F(\lambda \in \Lambda)$ are bounded linear operators such that the maps $A_{\lambda}:\left(E,\|\cdot\|_{E, \lambda}\right) \rightarrow$ $\left(F,\|\cdot\|_{F, \lambda}\right)$ are uniformly bounded with respect to $\lambda \in \Lambda$.
(iv) There exist invertible operators $A_{\lambda, j} \in \mathcal{B}_{\text {isom }}\left(E_{j} ; F_{j}\right)(j \in J, \lambda \in \Lambda)$ such that

$$
\sup _{\lambda \in \Lambda}\left\|\left(f_{j}\right)_{j} \mapsto\left(A_{\lambda, j}^{-1} f_{j}\right)_{j}\right\|_{\mathcal{B}\left(l_{q}(\mathbf{F}) ; l_{q}(\mathbf{E})\right), \lambda}<\infty .
$$

(v) The operators $B_{\lambda, j}:=\Phi_{F, j} A_{\lambda}-A_{\lambda, j} \Phi_{E, j} \in \mathcal{B}\left(E ; F_{j}\right)$ satisfy

$$
\lim _{|\lambda| \rightarrow \infty}\left\|u \mapsto\left(B_{\lambda, j} u\right)_{j}\right\|_{\mathcal{B}\left(E ; l_{q}(\mathbf{F})\right), \lambda}=0 .
$$

(vi) The operators $C_{\lambda, j}:=A_{\lambda} \Psi_{E, j}-\Psi_{F, j} A_{\lambda, j} \in \mathcal{B}\left(E_{j} ; F\right)$ satisfy

$$
\lim _{|\lambda| \rightarrow \infty}\left\|\left(u_{j}\right) \mapsto \sum_{j} C_{\lambda, j} u_{j}\right\|_{\mathcal{B}\left(l_{q}(\mathbf{E}) ; F\right), \lambda}=0
$$

For later applications, it is important to establish uniform bounds for data-to-solution maps. A parameter-dependent operator $A_{\lambda} \in \mathcal{B}_{\text {isom }}(E ; F)$ is called uniformly invertible with respect to $\lambda$, if there is a number $C$ such that $\left\|A_{\lambda}^{-1}\right\|_{F \rightarrow E} \leq C$ for all $\lambda$.
2.6. Proposition (cf. [AHS94, Proposition 3.2]). If Assumption 2.5 is satisfied, then there is $\lambda_{0}>0$ such that $A_{\lambda}:\left(E,\|\cdot\|_{E, \lambda}\right) \rightarrow\left(F,\|\cdot\|_{F, \lambda}\right)$ is uniformly invertible with respect to $\lambda \in \Lambda$ with $|\lambda| \geq \lambda_{0}$.

Proof. We consider the approximate inverse

$$
R_{\lambda}: F \rightarrow E, \quad R_{\lambda} f:=\sum_{j} \Psi_{E, j} A_{\lambda, j}^{-1} \Phi_{F, j} f \quad \text { for } f \in F .
$$

Let us write

$$
\begin{aligned}
& R_{\lambda} A_{\lambda}-I_{E}=\sum_{j} \Psi_{E, j} A_{\lambda, j}^{-1}\left(\Phi_{F, j} A_{\lambda}-A_{\lambda, j} \Phi_{E, j}\right)=\sum_{j} \Psi_{E, j} A_{\lambda, j}^{-1} B_{\lambda, j}, \\
& A_{\lambda} R_{\lambda}-I_{F}=\sum_{j}\left(A_{\lambda} \Psi_{E, j}-\Psi_{F, j} A_{\lambda, j}\right) A_{\lambda, j}^{-1} \Phi_{F, j}=\sum_{j} C_{\lambda, j} A_{\lambda, j}^{-1} \Phi_{F, j} .
\end{aligned}
$$

With Assumption 2.5 we can choose an upper bound $M>0$ for the numbers

$$
\sup _{\lambda \in \Lambda}\left\|r_{E}\right\|_{\mathcal{B}\left(l_{q}(\mathbf{E}) ; E\right), \lambda}, \quad \sup _{\lambda \in \Lambda}\left\|\left(f_{j}\right)_{j} \mapsto\left(A_{\lambda, j}^{-1} f_{j}\right)_{j}\right\|_{\mathcal{B}\left(l_{q}(\mathbf{F}) ; l_{q}(\mathbf{E})\right), \lambda}, \quad \sup _{\lambda \in \Lambda}\left\|r_{F}^{c}\right\|_{\mathcal{B}\left(F ; l_{q}(\mathbf{F})\right), \lambda} .
$$

Then $R_{\lambda}$ is bounded by $M^{3}$ and we obtain the following estimates for $f \in F$ and $u \in E$ :

$$
\begin{aligned}
& \left\|R_{\lambda} A_{\lambda} u-u\right\|_{E, \lambda}=\left\|\sum_{j} \Psi_{E, j} A_{\lambda, j}^{-1} B_{\lambda, j} u\right\|_{E, \lambda} \leq M^{2}\left\|u \mapsto\left(B_{\lambda, j} u\right)_{j}\right\|_{\mathcal{B}\left(E ; l_{q}(\mathbf{F})\right), \lambda}\|u\|_{E, \lambda}, \\
& \left\|A_{\lambda} R_{\lambda} f-f\right\|_{F, \lambda}=\left\|\sum_{j} C_{\lambda, j} A_{\lambda, j}^{-1} \Phi_{F, j} f\right\|_{F, \lambda} \leq\left(\left\|\left(u_{j}\right) \mapsto \sum_{j} C_{\lambda, j} u_{j}\right\|_{\mathcal{B}\left(l_{q}(\mathbf{E}) ; F\right), \lambda}\right) M^{2}\|f\|_{F, \lambda} .
\end{aligned}
$$

Therefore we can find some $\lambda_{0} \geq 0$ such that

$$
\left\|A_{\lambda} R_{\lambda}-I_{F}\right\|_{\mathcal{B}(F), \lambda} \leq 2^{-1}, \quad\left\|R_{\lambda} A_{\lambda}-I_{E}\right\|_{\mathcal{B}(E), \lambda} \leq 2^{-1} \quad \text { for } \lambda \in \Lambda,|\lambda| \geq \lambda_{0} .
$$

Hence the operators $A_{\lambda} R_{\lambda}=I_{F}-\left(I_{F}-A_{\lambda} R_{\lambda}\right) \in \mathcal{B}(F)$ and $R_{\lambda} A_{\lambda}=I_{E}-\left(I_{E}-R_{\lambda} A_{\lambda}\right) \in \mathcal{B}(E)$ are invertible. Consequently, $R_{\lambda}\left(A_{\lambda} R_{\lambda}\right)^{-1} \in \mathcal{B}(F ; E)$ is a right-inverse and $\left(A_{\lambda} R_{\lambda}\right)^{-1} R_{\lambda} \in \mathcal{B}(F ; E)$ is a left-inverse for $A_{\lambda}$. Thus, $A_{\lambda}$ is invertible for all $\lambda \in \Lambda$ with $|\lambda| \geq \lambda_{0}$ and its inverse $A_{\lambda}^{-1}=R_{\lambda}\left(A_{\lambda} R_{\lambda}\right)^{-1}=\left(A_{\lambda} R_{\lambda}\right)^{-1} R_{\lambda}$ is bounded by $2 M^{3}$.

Next, we provide a localization set-up that can be used to construct approximation systems.
2.7. Remark. Let $\Omega$ be a bounded domain in $\mathbb{R}^{n}(n \geq 2)$ with $C^{1}$-boundary $\partial \Omega$ and let $\Sigma \subset \Omega$ be a compact $C^{1}$-hypersurface. We say that a family $\left(U_{j}\right)_{j \in J}$ of open subsets of $\mathbb{R}^{n}$ is a finite open covering for $\bar{\Omega}$ in $\mathbb{R}^{n}$, if $J$ is finite and $\bar{\Omega}$ is contained in $\bigcup_{j \in J} U_{j}$. Since $\bar{\Omega}, \partial \Omega$, and $\Sigma$ are compact, there exists $r_{0}>0$ such that for every $r \in\left(0, r_{0}\right]$ we can choose
(i) a finite open covering of balls $U_{j}=B_{r}\left(p_{j}\right)$ with $p_{j} \in \bar{\Omega}$ such that the index set can be decomposed as $J=J_{1} \cup J_{2} \cup J_{3}$ with

$$
\begin{array}{ll}
p_{j} \in \Omega \backslash \Sigma \text { and } \bar{U}_{j} \subset \Omega \backslash \Sigma & \text { if } j \in J_{1}, \\
p_{j} \in \partial \Omega \text { and } U_{j} \cap \Sigma=\emptyset & \text { if } j \in J_{2}, \\
p_{j} \in \Sigma \text { and } U_{j} \subset \Omega & \text { if } j \in J_{3},
\end{array}
$$

(ii) a family $\left(\Theta_{j}\right)_{j \in J}$ of rigid transformations

$$
\Theta_{j}: x \mapsto p_{j}+Q_{j} x, \quad B_{r}(0) \rightarrow U_{j}=B_{r}\left(p_{j}\right),
$$

with an orthogonal matrix $Q_{j}=\partial_{x} \Theta_{j} \in \mathbb{R}^{n \times n}$ such that

$$
\begin{aligned}
Q_{j} & =I & & \text { if } j \in J_{1}, \\
-Q_{j} e_{n} & =\nu_{\partial \Omega}\left(p_{j}\right) & & \text { if } j \in J_{2}, \\
Q_{j} e_{n} & =\nu_{\Sigma}\left(p_{j}\right) & & \text { if } j \in J_{3} .
\end{aligned}
$$

2.8. Definition (Localization set-up). Let $\Omega$ be a bounded domain in $\mathbb{R}^{n}(n \geq 2)$ with $C^{1}$ boundary $\partial \Omega$ and let $\Sigma \subset \Omega$ be a compact $C^{1}$-hypersurface. For $\omega: \mathbb{R}^{n-1} \rightarrow \mathbb{R}$ we put

$$
\mathbb{R}_{\omega}^{n}:=\left\{\left(x^{\prime}, x_{n}\right) \in \mathbb{R}^{n}: x_{n}>\omega\left(x^{\prime}\right)\right\}, \quad \Sigma_{\omega}:=\left\{\left(x^{\prime}, x_{n}\right) \in \mathbb{R}^{n}: x_{n}=\omega\left(x^{\prime}\right)\right\} .
$$

Let $r>0$ and $\eta>0$ be given and suppose that
(i) $\left(U_{j}\right)_{j \in J}$ is a finite open covering for $\bar{\Omega}$ in $\mathbb{R}^{n}$ with $U_{j}=B_{r}\left(p_{j}\right)$ as in Remark 2.7.(i),
(ii) $\left(\Theta_{j}\right)_{j \in J}$ is a family of rigid transformations as in Remark 2.7.(ii),
(iii) $\left(\omega_{j}\right)_{j \in J}$ is a family of functions of class $C_{c}^{1}\left(\mathbb{R}^{n-1}\right)$ which satisfy

$$
\omega_{j}(0)=\left|\nabla \omega_{j}(0)\right|=0, \quad\left\|\nabla \omega_{j}\right\|_{L_{\infty}\left(\mathbb{R}^{n-1}\right)} \leq \eta \quad \text { for all } j \in J
$$

and suppose that

$$
\begin{aligned}
\omega_{j} & =0 & & \text { if } j \in J_{1}, \\
\Theta_{j}\left(B_{r}(0) \cap \mathbb{R}_{\omega_{j}}^{n}\right) & =U_{j} \cap \Omega & & \text { if } j \in J_{2}, \\
\Theta_{j}\left(B_{r}(0) \cap \mathbb{R}^{n} \backslash \Sigma_{\omega_{j}}\right) & =U_{j} \cap \Omega \backslash \Sigma & & \text { if } j \in J_{3} .
\end{aligned}
$$

Then we call $\left(U_{j}, \Theta_{j}, \omega_{j}\right)_{j \in J}$ an $(\eta, r)$-localization set-up for $(\Omega, \Sigma)$.
2.9. Lemma. Let $\Omega \subset \mathbb{R}^{n}(n \geq 2)$ be a bounded domain with $C^{1}$-boundary $\partial \Omega$ and $\Sigma \subset \Omega$ be a compact $C^{1}$-hypersurface.
(i) If $\eta>0$ is given, then there exists $r_{0}>0$ such that for every $r \in\left(0, r_{0}\right.$ ] we can find an $(\eta, r)$ localization set-up $\left(U_{j}, \Theta_{j}, \omega_{j}\right)_{j \in J}$ for $(\Omega, \Sigma)$.
(ii) If, additionally, $\partial \Omega$ and $\Sigma$ are of class $C^{k-}$ for some $k \geq 2$, then the $\omega_{j}$ belong to $C_{c}^{k-}\left(\mathbb{R}^{n-1}\right)$ and there exists $C=C(n, p, \partial \Omega, \Sigma)>0$ such that

$$
\left\|\omega_{j}\right\|_{H_{p}^{2}\left(\mathbb{R}^{n-1}\right)} \leq C r^{(n-1) / p}\left\|\nabla^{2} \omega_{j}\right\|_{L_{\infty}\left(\mathbb{R}^{n-1}\right)} \quad \text { for all } j \in J_{2} \cup J_{3} .
$$

Proof. As in Remark 2.7, we let $U_{j}=B_{r}\left(p_{j}\right)$ form a finite open covering for $\bar{\Omega}$ and consider the rigid transformations $\Theta_{j}: x \mapsto p_{j}+Q_{j} x$. The case $j \in J_{1}$ is trivial and since the cases $j \in J_{2}$ and $j \in J_{3}$ are analogous, we concentrate on $j \in J_{3}$.

We first construct the functions $\omega_{j}$ and prove that $\left\|\nabla \omega_{j}\right\|_{\infty}$ is small. For every $p \in \Sigma$ we can find a number $r_{1}(p)>0$ and a unique height function $\omega_{p}$ on $B_{r_{1}(p)} \subset \mathbb{R}^{n-1}$, such that for $\Sigma_{\omega_{p}}:=\left\{\left(x^{\prime}, \omega_{s}\left(x^{\prime}\right)\right): x^{\prime} \in B_{r_{1}(p)}\right\}$ we have $\Theta\left(\Sigma_{\omega_{p}}\right) \subset \Sigma$ for some rigid transformation $\Theta: x \mapsto p+Q x$ with $Q e_{n}=\nu_{\Sigma}(p)$. The function $\nabla \omega_{p}$ is related to $\nu_{\Sigma}$ by (see also (A.3) on page 130)

$$
\nabla \omega_{p}=-\frac{Q P^{\prime} Q^{\top}\left(\nu_{\Sigma} \circ \Theta\right)}{Q e_{n} \cdot \nu_{\Sigma} \circ \Theta} \quad \text { on } B_{p}, \quad \text { with } P^{\prime}=I_{n}-e_{n} \otimes e_{n}
$$

Moreover, it satisfies $\omega_{p}(0)=\left|\nabla \omega_{p}(0)\right|=0$. Since $\nu_{\Sigma}$ is uniformly continuous on $\Sigma$, we obtain $\left\|\left.\nabla \omega_{p}\right|_{B_{t}}\right\|_{\infty} \rightarrow 0$ as $t \rightarrow 0$, uniformly with respect to $p \in \Sigma$. By compactness of $\Sigma$, we may choose the number $r_{1}$ uniform in $p \in \Sigma$.

Let $\chi \in \mathcal{B}\left(\mathbb{R}^{n-1}\right)$ with $0 \leq \chi \leq 1, \chi\left(x^{\prime}\right)=1$ for $\left|x^{\prime}\right| \leq 1$ and $\chi\left(x^{\prime}\right)=0$ for $\left|x^{\prime}\right| \geq 2$. For $r \in\left(0, r_{1} / 2\right]$ we define a function $\tilde{\omega}_{p, r}: \mathbb{R}^{n-1} \rightarrow \mathbb{R}$ with support in $B_{2 r}$ by

$$
\tilde{\omega}_{p, r}\left(x^{\prime}\right):= \begin{cases}\chi\left(x^{\prime} / r\right) \omega_{p}\left(x^{\prime}\right) & \text { for }\left|x^{\prime}\right|<2 r \\ 0 & \text { for }\left|x^{\prime}\right| \geq 2 r\end{cases}
$$

Then $\tilde{\omega}_{p, r}\left(x^{\prime}\right)=\omega_{p}\left(x^{\prime}\right)$ for all $x^{\prime} \in B_{r}$. From $\omega_{p}(0)=0$ and the fundamental theorem of calculus, we obtain the inequality $\left\|\left.\omega_{p}\right|_{B_{r}}\right\|_{\infty} \leq r\left\|\left.\nabla^{\prime} \omega_{p}\right|_{B_{r}}\right\|_{\infty}$. The uniform continuity of $\nu_{\Sigma}$ further implies that $\left\|\left.\nabla^{\prime} \omega_{p}\right|_{B_{r}}\right\|_{\infty} \rightarrow 0$ as $r \rightarrow 0$, uniformly in $p \in \Sigma$. Therefore

$$
\begin{aligned}
\left\|\nabla \tilde{\omega}_{p, r}\right\|_{\infty} & \leq r^{-1}\|\nabla \chi\|_{\infty}\left\|\left.\omega_{p}\right|_{B_{2 r}}\right\|_{\infty}+\left\|\left.\nabla \omega_{p}\right|_{B_{2 r}}\right\|_{\infty} \\
& \leq\left(\|\nabla \chi\|_{\infty}+1\right)\left\|\left.\nabla \omega_{p}\right|_{B_{2 r}}\right\|_{\infty} \rightarrow 0 \quad \text { as } r \rightarrow 0
\end{aligned}
$$

uniformly in $p \in \Sigma$. Thus for given $\eta>0$ we can choose a number $r_{0} \in\left(0, r_{1} / 2\right]$ such that $\left\|\nabla \tilde{\omega}_{p, r}\right\|_{\infty} \leq \eta$ for all $p \in \Sigma, r \in\left(0, r_{0}\right]$. We finally put $\omega_{j}:=\tilde{\omega}_{p_{j}, r}$ for $j \in J$ with a suitable finite index set $J(r)$. Hence assertion (i) is valid.

Having in mind that every $\omega \in W_{\infty}^{2}\left(\mathbb{R}^{n-1}\right)$ with $\omega(0)=|\nabla \omega(0)|=0$ satisfies the estimates $\left|\nabla^{k} \omega(x)\right| \leq|x|^{2-k}\left\|\nabla^{k} \omega\right\|_{\infty}$ for $k \in\{0,1,2\}$, and using the substitution $x=r y$, we obtain

$$
\left\|\nabla^{k} \tilde{\omega}_{r}\right\|_{p}=\left\|\nabla^{k}(\chi(\cdot / r) \omega)\right\|_{p} \leq C r^{(n-1) / p}\|\chi\|_{H_{p}^{k}}\left\|\nabla^{k} \omega\right\|_{\infty} \quad \text { for } k \in\{0,1,2\} .
$$

This proves assertion (ii).
2.1.2. $\lambda$-dependent norms. Let $\Omega$ and $\Sigma$ satisfy Assumption 2.1. For $p \in(1, \infty), k \in \mathbb{N}_{0}$, and $\lambda \in \mathbb{C} \backslash\{0\}$, we define the Banach spaces

$$
\begin{array}{ll}
\mathbb{E}_{\lambda}^{k}=\left(\mathbb{E}^{k},\|\cdot\|_{\mathbb{E}_{\lambda}^{k}}\right), & \mathbb{E}^{k}=H_{p}^{k+2}(\Omega \backslash \Sigma), \\
\mathbb{F}_{\lambda}^{k}=\left(\mathbb{F}^{k},\|\cdot\|_{\mathbb{F}_{\lambda}^{k}}\right), & \mathbb{F}^{k}=H_{p}^{k}(\Omega \backslash \Sigma) \times W_{p}^{k+1-1 / p}(\partial \Omega) \times W_{p}^{k+1-1 / p}(\Sigma) \cap W_{p}^{k+2-1 / p}(\Sigma),
\end{array}
$$

which are equipped with the equivalent $\lambda$-dependent norms

$$
\begin{gathered}
\|u\|_{\mathbb{E}_{\lambda}^{k}}:=\|u\|_{H_{p}^{k+2}(\Omega \backslash \Sigma), \lambda}:=\sum_{j=0}^{k+2}\left\|\lambda^{(k+2-j) / 2} \nabla^{j} u\right\|_{L_{p}(\Omega)}, \\
\left\|\left(f, g, h_{1}, h_{2}\right)\right\|_{\mathbb{F}_{\lambda}^{k}}:=\|f\|_{H_{p}^{k}(\Omega \backslash \Sigma), \lambda}+\|g\|_{W_{p}^{k+1-1 / p}(\partial \Omega), \lambda}+\left\|h_{1}\right\|_{W_{p}^{k+1-1 / p}(\Sigma), \lambda}+\left\|h_{2}\right\|_{W_{p}^{k+2-1 / p}(\Sigma), \lambda},
\end{gathered}
$$

where

$$
\begin{aligned}
\|f\|_{H_{p}^{k}(\Omega \backslash \Sigma), \lambda} & :=\sum_{j=0}^{k}\left\|\lambda^{(k-j) / 2} \nabla^{j} f\right\|_{L_{p}(\Omega)}, \\
\|g\|_{W_{p}^{k+1-1 / p}(\partial \Omega), \lambda} & :=\llbracket \nabla_{\partial \Omega}^{k} g \rrbracket_{W_{p}^{1-1 / p}(\partial \Omega)}+\sum_{j=0}^{k}|\lambda|^{1 / 2-1 / 2 p}\left\|\lambda^{(k-j) / 2} \nabla_{\partial \Omega}^{j} g\right\|_{L_{p}(\partial \Omega)}, \\
\left\|h_{1}\right\|_{W_{p}^{k+1-1 / p}(\Sigma), \lambda} & :=\llbracket \nabla_{\Sigma}^{k} h_{1} \rrbracket_{W_{p}^{1-1 / p}(\Sigma)}+\sum_{j=0}^{k}|\lambda|^{1 / 2-1 / 2 p}\left\|\lambda^{(k-j) / 2} \nabla_{\Sigma}^{j} h_{1}\right\|_{L_{p}(\Sigma)}, \\
\left\|h_{2}\right\|_{W_{p}^{k+2-1 / p}(\Sigma), \lambda} & :=\llbracket \nabla_{\Sigma}^{k+1} h_{2} \rrbracket_{W_{p}^{1-1 / p}(\Sigma)}+\sum_{j=0}^{k+1}|\lambda|^{1 / 2-1 / 2 p}\left\|\lambda^{(k+1-j) / 2} \nabla_{\Sigma}^{j} h_{2}\right\|_{L_{p}(\Sigma)} .
\end{aligned}
$$

Let us first derive these norms and have a look at its advantages. We consider the scaling

$$
u_{\lambda}(x):=\lambda^{\alpha} u\left(\lambda^{-\beta} x\right) \quad \text { for } x \in \Omega \backslash \Sigma \text { with some } \alpha, \beta \in \mathbb{R} .
$$

We only consider the cases $\Omega \in\left\{\mathbb{R}^{n}, \mathbb{R}_{+}^{n}\right\}$ and $\Sigma \in\left\{\mathbb{R}^{n-1} \times\{0\}, \emptyset\right\}$ since these are invariant under the transformation $x \mapsto \lambda^{-\beta} x$. Then

$$
\lambda u(x)-\Delta u(x)=\lambda^{-\alpha}\left(\lambda u_{\lambda}\left(\lambda^{\beta} x\right)-\lambda^{2 \beta} \Delta u_{\lambda}\left(\lambda^{\beta} x\right)\right) .
$$

Since the local operators $A_{\lambda, j}$ should be uniformly invertible in $\lambda$, we want to achieve that the equations for $u_{\lambda}$ do not depend on $\lambda$ and therefore must choose $\beta=1 / 2$. Next, the norm of the transformation $u \mapsto u_{\lambda}$ should satisfy $\|u\|_{\mathbb{E}_{\lambda}^{k}}=\left\|u_{\lambda}\right\|_{\mathbb{E}_{1}^{k}}$ and hence we require that

$$
\|u\|_{\mathbb{E}_{\lambda}^{k}}=\sum_{j=0}^{k+2}\left\|\nabla^{j} u_{\lambda}\right\|_{L_{p}(\Omega)}=\sum_{j=0}^{k+2}\left\|\lambda^{\alpha-\beta j+\beta n / p} \nabla^{j} u\right\|_{L_{p}(\Omega)}=\sum_{j=0}^{k+2}\left\|\lambda^{\alpha-j / 2+n / 2 p} \nabla^{j} u\right\|_{L_{p}(\Omega)} .
$$

Finally, we choose $\alpha=(k+2-n / p) / 2$ so that the highest order term in this norm does not depend on $\lambda$. This yields precisely the aforementioned $\mathbb{E}_{\lambda}^{k}$-norm. We keep in mind that

$$
u_{\lambda}:=\lambda^{(k+2) / 2-n / 2 p} u\left(\lambda^{-1 / 2} \cdot\right) .
$$

Similarly, we define the rescaled data

$$
\begin{aligned}
f_{\lambda} & :=\lambda^{k / 2-n / 2 p} f\left(\lambda^{-1 / 2} \cdot\right), \\
g_{\lambda} & :=\lambda^{(k+1) / 2-n / 2 p} g\left(\lambda^{-1 / 2} \cdot\right), \\
h_{1 \lambda} & :=\lambda^{(k+1) / 2-n / 2 p} h_{1}\left(\lambda^{-1 / 2} \cdot\right), \\
h_{2 \lambda} & :=\lambda^{(k+2) / 2-n / 2 p} h_{2}\left(\lambda^{-1 / 2} \cdot\right) .
\end{aligned}
$$

When we replace the functions ( $u, f, g, h_{1}, h_{2}$ ) by ( $u_{\lambda}, f_{\lambda}, g_{\lambda}, h_{1 \lambda}, h_{2 \lambda}$ ) in (2.9), we see that the system $A_{\lambda} u=\left(f, g, h_{1}, h_{2}\right)$ is equivalent to $A_{1} u_{\lambda}=\left(f_{\lambda}, g_{\lambda}, h_{1 \lambda}, h_{2 \lambda}\right)$ in case $\Omega \in\left\{\mathbb{R}^{n}, \mathbb{R}_{+}^{n}\right\}$ and $\Sigma \in\left\{\mathbb{R}^{n-1} \times\{0\}, \emptyset\right\}$. Moreover,

$$
\|u\|_{\mathbb{E}_{\lambda}^{k}}=\left\|u_{\lambda}\right\|_{\mathbb{E}_{1}^{k}}, \quad\left\|\left(f, g, h_{1}, h_{2}\right)\right\|_{\mathbb{F}_{\lambda}^{k}}=\left\|\left(f_{\lambda}, g_{\lambda}, h_{1 \lambda}, h_{2 \lambda}\right)\right\|_{\mathbb{F}_{1}^{k}} .
$$

2.1.3. Basic model problems. We first consider the system $A_{\lambda} u=\left(f, h_{1}, h_{2}\right)$ in the situation of a whole space $\Omega:=\mathbb{R}^{n}(n \geq 2)$ with flat interface $\Sigma:=\mathbb{R}^{n-1} \times\{0\} \cong \mathbb{R}^{n-1}$ and constant coefficients $\mu_{ \pm} \in(0, \infty)$; that is,

$$
\left\{\begin{array}{rlrl}
\lambda u-\mu \Delta u & =f & & \text { in } \dot{\mathbb{R}}^{n},  \tag{2.10}\\
\llbracket \mu \partial_{n} u \rrbracket & =h_{1} & & \text { on } \mathbb{R}^{n-1}, \\
\llbracket u \rrbracket=h_{2} & & \text { on } \mathbb{R}^{n-1} .
\end{array}\right.
$$

Here we have put $\dot{\Omega}:=\Omega \backslash \Sigma=\dot{\mathbb{R}}^{n}$ and $\Omega_{ \pm}:=\mathbb{R}^{n-1} \times \pm(0, \infty)$. The elements of $\Omega$ are denoted by $x=\left(x^{\prime}, x_{n}\right)$ or $\left(x^{\prime}, y\right)$ with $x^{\prime} \in \mathbb{R}^{n-1}$ and $x_{n}=y \in \mathbb{R}$, and we let $\Delta=\partial_{1}^{2}+\cdots+\partial_{n}^{2}$, $\Delta^{\prime}=\partial_{1}^{2}+\cdots+\partial_{n-1}^{2}, \nabla=\left(\partial_{1}, \ldots, \partial_{n}\right)$, and $\nabla^{\prime}=\left(\partial_{1}, \ldots, \partial_{n-1}\right)$. The parameter $\lambda$ belongs to the open sector $\Sigma_{\phi}:=\{\lambda \in \mathbb{C} \backslash\{0\}:|\arg \lambda|<\phi\}$ for $\phi \in(0, \pi)$.

We shall prove that problem (2.10) has optimal $H_{p}^{k+2}$-regularity in the following sense.
2.10. Lemma. Let $\mu_{0} \in(0,1], k \in \mathbb{N}_{0}, \phi \in(0, \pi)$, and $p \in(1, \infty)$. Then the operator

$$
A_{\lambda}: \mathbb{E}_{\lambda}^{k}\left(\dot{\mathbb{R}}^{n}\right) \rightarrow \mathbb{F}_{\lambda}^{k}\left(\dot{\mathbb{R}}^{n}\right), \quad u \mapsto\left(\lambda u-\operatorname{div}(\mu \nabla u), \llbracket \mu \partial_{n} u \rrbracket, \llbracket u \rrbracket\right)
$$

is uniformly invertible with respect to $\mu_{ \pm} \in\left[\mu_{0}, \mu_{0}^{-1}\right]$ and $\lambda \in \Sigma_{\phi}$.
Proof. (i) In order to prove uniqueness, it is sufficient to consider a solution $u \in H_{p}^{2}\left(\mathbb{R}^{n}\right)$ to (2.10) for trivial data $A_{\lambda} u=\left(f, h_{1}, h_{2}\right)=0$. When we consider $u$ as a function $y \mapsto u(\cdot, y)$ that belongs to the space $H_{p}^{2}\left(\mathbb{R} ; L_{p}\left(\mathbb{R}^{n-1}\right)\right) \cap L_{p}\left(\mathbb{R} ; H_{p}^{2}\left(\mathbb{R}^{n-1}\right)\right)$, we see that both functions $y \mapsto$ $u_{ \pm}(\cdot, y), \pm[0, \infty) \rightarrow \mathcal{S}^{\prime}\left(\mathbb{R}^{n-1}\right)$ are continuous. The functions $\omega_{ \pm}(\xi):=\left(\lambda \mu_{ \pm}^{-1}+|\xi|^{2}\right)^{1 / 2}$ satisfy $\operatorname{Re} \omega_{ \pm}(\xi)>0$ for all $\lambda \in \mathbb{C} \backslash(-\infty, 0]$ and $\xi \in \mathbb{R}^{n-1}$. Then the partially Fourier transformed equations with respect to $x \in \mathbb{R}^{n-1}$ with covariable $\xi \in \mathbb{R}^{n-1}$ are given by

$$
\left\{\begin{align*}
\omega^{2} \tilde{u}-\partial_{y}^{2} \tilde{u}=0 & \text { in } \mathcal{D}^{\prime}\left(\dot{\mathbb{R}} ; \mathcal{S}^{\prime}\left(\mathbb{R}^{n-1}\right)\right),  \tag{2.11}\\
\mu_{+} \partial_{y} \tilde{u}_{+}(\cdot, 0)-\mu_{-} \partial_{y} \tilde{u}_{-}(\cdot, 0)=0 & \text { in } \mathcal{S}^{\prime}\left(\mathbb{R}^{n-1}\right), \\
\tilde{u}_{+}(\cdot, 0)-\tilde{u}_{-}(\cdot, 0)=0 & \text { in } \mathcal{S}^{\prime}\left(\mathbb{R}^{n-1}\right)
\end{align*}\right.
$$

The first equation in (2.11) must be understood in the following sense:

$$
\begin{equation*}
\int_{-\infty}^{\infty} \tilde{u}(\cdot, y)\left(\omega^{2} \varphi(y)-\partial_{y}^{2} \varphi(y)\right) d y=0 \text { in } \mathcal{S}^{\prime}\left(\mathbb{R}^{n-1}\right) \quad \text { for } \varphi \in \mathcal{D}(\dot{\mathbb{R}}) \tag{2.12}
\end{equation*}
$$

We claim that (2.12) implies $\tilde{u}(\cdot, \pm y)=\left(\xi \mapsto e^{-\omega_{ \pm}(\xi) y}\right) c_{ \pm}$for all $y \geq 0$ and some $c_{ \pm} \in \mathcal{S}^{\prime}\left(\mathbb{R}^{n-1}\right)$. Indeed, in order to check this for $\tilde{u}_{+}(\cdot, y)$, we write an arbitrary $\varphi \in \mathcal{D}\left(\mathbb{R}_{+}\right)$as

$$
\begin{equation*}
\varphi(y)=\left(\omega_{+}^{2}-\partial_{y}^{2}\right) \psi_{\varphi}(y)+h_{+}(y)\left\langle e^{\omega_{+} \cdot} \mid \varphi\right\rangle+h_{-}(y)\left\langle e^{-\omega_{+} \cdot} \mid \varphi\right\rangle . \tag{2.13}
\end{equation*}
$$

Here $\langle\cdot \mid \cdot\rangle$ denotes bilinear integration over $\mathbb{R}_{+}$, the functions $h_{ \pm} \in \mathcal{D}\left(\mathbb{R}_{+}\right)$with $\left\langle e^{ \pm \omega_{+} \cdot} \mid h_{ \pm}\right\rangle=1$ and $\left\langle e^{ \pm \omega_{+}} \mid h_{\mp}\right\rangle=0$ are fixed (independent of $\varphi$ ), and we can calculate the solution $\psi_{\varphi} \in \mathcal{D}\left(\mathbb{R}_{+}\right)$ of (2.13) by using Green's functions (see Lemma 3.3 on page 56). Then it can be readily checked that

$$
\left\langle\tilde{u}_{+} \mid \varphi\right\rangle=\left\langle\tilde{u}_{+} \mid h_{+}\right\rangle\left\langle e^{\omega_{+} \cdot} \mid \varphi\right\rangle+\left\langle\tilde{u}_{+} \mid h_{-}\right\rangle\left\langle e^{-\omega_{+}} \mid \varphi\right\rangle \quad \text { for } \varphi \in \mathcal{D}\left(\mathbb{R}_{+}\right),
$$

with the constant distributions $c_{+, \pm}:=\left\langle\tilde{u}_{+} \mid h_{ \pm}\right\rangle$. Since $y \mapsto u(\cdot, y)$ belongs to $H_{p}^{2}\left(\mathbb{R}_{+} ; L_{p}\left(\mathbb{R}^{n-1}\right)\right)$, we must have $c_{+,+}=0$ and hence $\tilde{u}_{+}(\cdot, y)=\left(\xi \mapsto e^{-\omega_{+}(\xi) y}\right) c_{+}$for $y \geq 0$ with $c_{+}:=c_{+,-}$. Analogously, we have $\tilde{u}(\cdot,-y)=\left(\xi \mapsto e^{-\omega_{-}(\xi) y}\right) c_{-}$for $y \geq 0$ with some $c_{-} \in \mathcal{S}^{\prime}\left(\mathbb{R}^{n-1}\right)$. The remaining equations yield $c_{+}=c_{-}$and $-\mu_{+} c_{+} \omega_{+}-\mu_{-} c_{-} \omega_{-}=0$, and thus $c_{ \pm}=0$. Therefore (2.10) has at most one solution in $H_{p}^{2}\left(\dot{\mathbb{R}}^{n}\right)$.
(ii) Existence for $k=0$ and $f=0$. We construct a solution $u$ of (2.10) for given $\left(0, h_{1}, h_{2}\right) \in$ $\mathbb{F}_{\lambda}^{0}$. The partially Fourier transformed function $y \mapsto \tilde{u}(\cdot, y), \dot{\mathbb{R}} \rightarrow \mathcal{S}^{\prime}\left(\mathbb{R}^{n-1}\right)$ must satisfy the system

$$
\left\{\begin{align*}
\omega^{2} \tilde{u}^{-} \partial_{y}^{2} \tilde{u}=0 & \text { in } \mathcal{D}^{\prime}\left(\dot{\mathbb{R}} ; \mathcal{S}^{\prime}\left(\mathbb{R}^{n-1}\right)\right),  \tag{2.14}\\
\mu_{+} \partial_{y} \tilde{u}_{+}(\cdot, 0)-\mu_{-} \partial_{y} \tilde{u}_{-}(\cdot, 0)=\tilde{h}_{1} & \text { in } \mathcal{S}^{\prime}\left(\mathbb{R}^{n-1}\right), \\
\tilde{u}_{+}(\cdot, 0)-\tilde{u}_{-}(\cdot, 0)=\tilde{h}_{2} & \text { in } \mathcal{S}^{\prime}\left(\mathbb{R}^{n-1}\right) .
\end{align*}\right.
$$

Problem (2.14) has the following $\mathcal{D}^{\prime}\left(\dot{\mathbb{R}} ; \mathcal{S}^{\prime}\left(\mathbb{R}^{n-1}\right)\right)$-solution.

$$
\left[\begin{array}{c}
\tilde{u}_{+}(\cdot, y) \\
\tilde{u}_{-}(\cdot,-y)
\end{array}\right]=\frac{1}{\mu_{+}+\mu_{-}}\left[\begin{array}{cc}
-\frac{e^{-\omega_{+}+y}}{\omega_{+}} & \mu_{-} e^{-\omega_{+} y} \\
-\frac{e^{-\omega_{-}}}{\omega_{-}} & -\mu_{+} e^{-\omega_{-} y}
\end{array}\right]\left[\begin{array}{c}
\tilde{h}_{1} \\
\tilde{h}_{2}
\end{array}\right] .
$$

In order to invert the partial Fourier transform $u \mapsto \tilde{u}$, we employ the joint functional calculus for $\nabla^{\prime}$ from Theorem B. 69 on page 166. Here we consider $\nabla^{\prime}=\left(\partial_{1} \ldots, \partial_{n-1}\right)$ as an operator tuple $\mathbf{T}=\left(T_{1}, \ldots, T_{n-1}\right)$ in $X=L_{p}\left(\mathbb{R}^{n-1}\right)$ in the sense of Remark B.65. For the symbols $\omega_{ \pm, \lambda}(z)=\left(\lambda / \mu_{ \pm}-z \cdot z\right)^{1 / 2}$, we define $\omega_{ \pm}\left(\nabla^{\prime}\right)=\left(\lambda / \mu_{ \pm}-\Delta^{\prime}\right)^{1 / 2}=: L_{ \pm, \lambda}: H_{p}^{1}\left(\mathbb{R}^{n-1}\right) \rightarrow L_{p}\left(\mathbb{R}^{n-1}\right)$. With Theorem B. 25 on page 155 we define the extensions $(x, y) \mapsto\left(e^{-L_{ \pm, \lambda}} h_{2}\right)(x) \in H_{p}^{2}\left(\mathbb{R}_{+}^{n}\right)$ and $(x, y) \mapsto\left(e^{-L_{ \pm, \lambda y}} h_{1}\right)(x) \in H_{p}^{1}\left(\mathbb{R}_{+}^{n}\right)$. Then a solution to (2.10) is given by

$$
\left[\begin{array}{c}
u_{+}(\cdot, y)  \tag{2.15}\\
u_{-}(\cdot,-y)
\end{array}\right]=\frac{1}{\mu_{+}+\mu_{-}}\left[\begin{array}{cc}
-L_{+, \lambda}^{-1} e^{-y L_{+, \lambda}} & \mu_{-} e^{-y L_{+, \lambda}} \\
-L_{-, \lambda}^{-1} e^{-y L_{-, \lambda}} & -\mu_{+} e^{-y L_{-, \lambda}}
\end{array}\right]\left[\begin{array}{l}
h_{1} \\
h_{2}
\end{array}\right] .
$$

(iii) A uniform bound with respect to $\lambda$. We employ the dilations $\sigma_{t} \in \mathcal{B}_{\text {isom }}\left(L_{p}\left(\mathbb{R}^{n}\right)\right)$ and $\sigma_{t}^{\prime} \in \mathcal{B}_{\text {isom }}\left(L_{p}\left(\mathbb{R}^{n-1}\right)\right)$ with $t \in(0, \infty)$ defined by $\sigma_{t} u:=u(t \cdot)$ for $u \in L_{p}\left(\mathbb{R}^{n}\right)$ and $\sigma_{t}^{\prime} h:=h(t \cdot)$ for $h \in L_{p}\left(\mathbb{R}^{n-1}\right)$. Then $\Delta^{\prime} \sigma_{\lambda^{1 / 2}}^{\prime}=\lambda \sigma_{\lambda^{1 / 2}}^{\prime} \Delta^{\prime}$ and, with $L_{ \pm}:=L_{ \pm, 1}$, we obtain

$$
L_{ \pm, \lambda}^{2}=\lambda / \mu_{ \pm}-\Delta^{\prime}=\lambda \sigma_{\lambda^{1 / 2}}^{\prime}\left(1 / \mu_{ \pm}-\Delta^{\prime}\right) \sigma_{\lambda^{-1 / 2}}^{\prime}=\left(\lambda^{1 / 2} \sigma_{\lambda^{1 / 2}}^{\prime} L_{ \pm} \sigma_{\lambda^{-1 / 2}}^{\prime}\right)^{2} .
$$

Hence $L_{ \pm, \lambda}=\lambda^{1 / 2} \sigma_{\lambda^{1 / 2}}^{\prime} L_{ \pm} \sigma_{\lambda^{-1 / 2}}^{\prime}$ on $D\left(L_{ \pm}\right)=H_{p}^{1}\left(\mathbb{R}^{n-1}\right)$. For $h \in L_{p}\left(\mathbb{R}^{n-1}\right)$, we have

$$
\sigma_{\lambda^{-1 / 2}}\left((x, y) \mapsto e^{-y L_{ \pm, \lambda}} h\right)=\exp \left(-\lambda^{-1 / 2} y \sigma_{\lambda^{-1 / 2}}^{\prime} L_{ \pm, \lambda} \sigma_{\lambda^{1 / 2}}^{\prime}\right) \sigma_{\lambda^{-1 / 2}}^{\prime} h=\exp \left(-y L_{ \pm}\right) \sigma_{\lambda^{-1 / 2}}^{\prime} h .
$$

Then the rescaled functions $u_{ \pm, \lambda}:=\lambda^{1-n / 2 p} \sigma_{\lambda^{-1 / 2}} u_{ \pm}$and $h_{j, \lambda}:=\lambda^{j / 2-n / 2 p} \sigma_{\lambda^{-1 / 2}}^{\prime} h_{j}$ satisfy

$$
\left[\begin{array}{c}
u_{+, \lambda}(\cdot, y) \\
u_{-, \lambda}(\cdot,-y)
\end{array}\right]=\frac{1}{\mu_{+}+\mu_{-}}\left[\begin{array}{cc}
-L_{+}^{-1} e^{-y L_{+}} & \mu_{-} e^{-y L_{+}} \\
-L_{-}^{-1} e^{-y L_{-}} & -\mu_{+} e^{-y L_{-}}
\end{array}\right]\left[\begin{array}{c}
h_{1, \lambda} \\
h_{2, \lambda}
\end{array}\right] \quad \text { for } y>0
$$

For given $\mu_{0} \in(0,1)$ and $\vartheta \in(0, \pi)$, there exists $M>0$ such that $L_{ \pm}^{2}$ are operators of positive type $\mathcal{P}_{1}\left(H_{p}^{2}\left(\mathbb{R}^{n-1}\right), L_{p}\left(\mathbb{R}^{n-1}\right), M, \vartheta\right)$ for all $\mu_{ \pm} \in\left[\mu_{0}, \mu_{0}^{-1}\right]$ (see page 154) and therefore Theorem B. 25 yields the assertion for $f=0$.
(iv) If $f \in L_{p}\left(\mathbb{R}^{n}\right)$ is arbitrary, then a solution to (2.10) is given by $u+v+w$, where $u$ is defined by (2.15), $v_{ \pm}:=\left(\lambda-\mu_{ \pm} \Delta\right)^{-1} f_{ \pm} \in H_{p}^{2}\left(\mathbb{R}_{ \pm}^{n}\right)$ with $\left.v_{ \pm}\right|_{y=0}=0$ are the half-space solutions from [DHP03, Theorem 7.3], and $w$ is the solution to $\lambda w-\mu \Delta w=0, \llbracket \rho w \rrbracket=-\llbracket \rho v \rrbracket$, $\llbracket \partial_{y} w \rrbracket=-\llbracket \partial_{y} v \rrbracket$, which is defined analogously as $u$ in (2.15). Therefore the assertion for $k=0$ is proved.
(v) Existence for $k \geq 0$. Let $\left(f, h_{1}, h_{2}\right) \in \mathbb{F}_{\lambda}^{k}$ be given. We shall construct a solution to (2.10) of the form $u=v+w$, where $v, w$ are defined as follows. Let $E_{ \pm} \in \mathcal{B}\left(H_{p}^{k}\left(\mathbb{R}_{ \pm}^{n}\right) ; H_{p}^{k}\left(\mathbb{R}^{n}\right)\right)$ and $E_{ \pm, \alpha} \in \mathcal{B}\left(H_{p}^{k-|\alpha|}\left(\mathbb{R}_{ \pm}^{n}\right) ; H_{p}^{k-|\alpha|}\left(\mathbb{R}^{n}\right)\right)$ denote the extension operators from Theorem B. 6 on
page 147 with the property $\partial_{x}^{\alpha} E_{ \pm} g=E_{ \pm, \alpha} \partial_{x}^{\alpha} g$ for $g \in H_{p}^{k}\left(\mathbb{R}_{+}^{n}\right)$ and $\alpha \in \mathbb{N}_{0}^{n}$ with $|\alpha| \leq k$. Then the functions $v_{ \pm}:=\left(\lambda-\mu_{ \pm} \Delta\right)^{-1} E_{ \pm}\left(\left.f\right|_{\mathbb{R}_{ \pm}^{n}}\right)$ belong to $H_{p}^{k+2}\left(\mathbb{R}^{n}\right)$ and satisfy

$$
\left\|v_{ \pm}\right\|_{H_{p}^{2}\left(\mathbb{R}^{n}\right), \lambda} \leq C\left(n, p, \mu_{0}\right)\|f\|_{L_{p}\left(\mathbb{R}^{n}\right)} \quad \text { for } f \in H_{p}^{k}\left(\dot{\mathbb{R}}^{n}\right), \lambda \in \Sigma_{\phi}, \mu_{ \pm} \in\left[\mu_{0}, \mu_{0}^{-1}\right] .
$$

Differentiating the equation $\left(\lambda-\mu_{ \pm} \Delta\right) v_{ \pm}=E_{ \pm}\left(\left.f\right|_{\mathbb{R}_{ \pm}^{n}}\right)$ shows that $\left(\lambda-\mu_{ \pm} \Delta\right) \partial_{x}^{\alpha} v=E_{ \pm, \alpha} \partial_{x}^{\alpha} f \in$ $L_{p}\left(\mathbb{R}^{n}\right)$ for $|\alpha| \leq k$. Hence the function $v:=\chi_{\mathbb{R}_{+}^{n}} v_{+}+\chi_{\mathbb{R}_{-}^{n}} v_{-}$belongs to $H_{p}^{k+2}\left(\dot{\mathbb{R}}^{n}\right)$, solves the equation $(\lambda-\mu \Delta) v=f$ in $\dot{\mathbb{R}}^{n}$, and satisfies

$$
\|v\|_{\mathbb{E}_{\lambda}^{k}\left(\dot{\mathbb{R}}^{n}\right)} \leq C\left(n, p, \mu_{0}\right)\|f\|_{H_{p}^{k}\left(\dot{\mathbb{R}}^{n}\right), \lambda} \quad \text { for } f \in H_{p}^{k}\left(\dot{\mathbb{R}}^{n}\right), \lambda \in \Sigma_{\phi}, \mu_{ \pm} \in\left[\mu_{0}, \mu_{0}^{-1}\right] .
$$

The function $w \in H_{p}^{2}\left(\dot{\mathbb{R}}^{n}\right)$ is defined as the solution to $(\lambda-\mu \Delta) w=0, \llbracket \mu \partial_{n} w \rrbracket=h_{1}-\llbracket \mu \partial_{n} v \rrbracket$, $\llbracket w \rrbracket=h_{2}-\llbracket v \rrbracket$. From uniqueness and (2.15) we derive the representation

$$
\left[\begin{array}{c}
w_{+}(\cdot, y) \\
w_{-}(\cdot,-y)
\end{array}\right]=\frac{1}{\mu_{+}+\mu_{-}}\left[\begin{array}{cc}
-L_{+, \lambda}^{-1} e^{-y L_{+, \lambda}} & \mu_{-} e^{-y L_{+, \lambda}} \\
-L_{-, \lambda}^{-1} e^{-y L_{-, \lambda}} & -\mu_{+} e^{-y L_{-, \lambda}}
\end{array}\right]\left[\begin{array}{c}
h_{1}-\llbracket \mu \partial_{n} v \rrbracket \\
h_{2}-\llbracket v \rrbracket
\end{array}\right] .
$$

In order to verify that $w$ belongs to $H_{p}^{k+2}\left(\mathbb{R}^{n}\right)$, we let $\alpha \in \mathbb{N}_{0}^{n-1}$ with $|\alpha| \leq k$. By Theorem B.69, the operators $\nabla^{\prime|\alpha|} L_{ \pm, \lambda}^{-|\alpha|}$ are isomorphisms in $W_{p}^{s}\left(\mathbb{R}^{n-1}\right)$ for every $s \geq 0$. Hence, by using the commutativity $L_{ \pm, \lambda} e^{-y L_{ \pm, \lambda}}=e^{-y L_{ \pm, \lambda}} L_{ \pm, \lambda}$ and by applying Theorem B.25, we see that

$$
\left\|\partial_{x^{\prime}}^{\alpha} w_{ \pm}\right\|_{H_{p}^{2}\left(\mathbb{R}_{ \pm}^{n}\right)} \lesssim\left\|L_{ \pm, \lambda}^{|\alpha|} w_{ \pm}\right\|_{H_{p}^{2}\left(\mathbb{R}_{ \pm}^{n}\right)} \lesssim\left\|\nabla^{|\alpha|} g_{1}\right\|_{W_{p}^{1-1 / p}\left(\mathbb{R}^{n-1}\right)}+\left\|\nabla^{\prime \alpha \mid} g_{2}\right\|_{W_{p}^{2-1 / p}\left(\mathbb{R}^{n-1}\right)}
$$

The normal derivatives can be estimated similarly by means of $\partial_{y}^{j} e^{-y L_{ \pm, \lambda}}=e^{-y L_{ \pm, \lambda}}\left(-L_{ \pm, \lambda}\right)^{j}$. This shows that $w$ belongs to $H_{p}^{k+2}\left(\dot{\mathbb{R}}^{n}\right)$ and satisfies $(\lambda-\mu \Delta) w=0$. Hence $u=v+w$ belongs to $\mathbb{E}_{\lambda}^{k}\left(\mathbb{R}^{n}\right)$ and solves $A_{\lambda} u=\left(f, h_{1}, h_{2}\right)$. Uniform bounds for $\left\|A_{\lambda}^{-1}\right\|$ with respect to $|\arg \lambda|<\phi$ can be shown again by a scaling argument.

Lemma 2.10 includes optimal $H_{p}^{k+2}$-regularity of the whole space model problem without interface, since we can choose $\mu_{+}=\mu_{-}$and restrict the operator $A_{\lambda}$ to the case $h_{1}=h_{2}=0$.
2.11. Corollary. Let $\mu_{0} \in(0,1], k \in \mathbb{N}_{0}, \phi \in(0, \pi)$, and $p \in(1, \infty)$. Then the operator

$$
A_{\lambda}: \mathbb{E}_{\lambda}^{k}\left(\mathbb{R}^{n}\right) \rightarrow \mathbb{F}_{\lambda}^{k}\left(\mathbb{R}^{n}\right), \quad u \mapsto \lambda u-\operatorname{div}(\mu \nabla u)
$$

is uniformly invertible with respect to $\mu \in\left[\mu_{0}, \mu_{0}^{-1}\right]$ and $\lambda \in \Sigma_{\phi}$.
Finally, we consider the remaining model problem for $\Omega=\mathbb{R}_{+}^{n}$ and $\Sigma=\emptyset$.
2.12. Lemma. Let $\mu_{0} \in(0,1], k \in \mathbb{N}_{0}, \phi \in(0, \pi)$, and $p \in(1, \infty)$. Then the operator

$$
A_{\lambda}: \mathbb{E}_{\lambda}^{k}\left(\mathbb{R}_{+}^{n}\right) \rightarrow \mathbb{F}_{\lambda}^{k}\left(\mathbb{R}_{+}^{n}\right), \quad u \mapsto\left(\lambda u-\operatorname{div}(\mu \nabla u),-\mu \partial_{n} u\right)
$$

is uniformly invertible with respect to $\mu \in\left[\mu_{0}, \mu_{0}^{-1}\right]$ and $\lambda \in \Sigma_{\phi}$.
Proof. We obtain the assertion by following the lines of the proof of Lemma 2.10, except for the elimination of the boundary condition. Here a solution $u$ to $A_{\lambda} u=(0, g)$ is given by

$$
u(\cdot, y)=\frac{1}{\mu} L_{\lambda}^{-1} e^{-y L_{\lambda}} g, \quad L_{\lambda}=\sqrt{\lambda-\mu \Delta^{\prime}} .
$$

2.1.4. Perturbed model problems. We next consider the model problem $A_{\lambda} u=\left(f, h_{2}, h_{2}\right)$ for $\Omega=\mathbb{R}^{n}$, for a bent hyperplane $\Sigma_{\omega}:=\theta_{\omega}\left(\mathbb{R}^{n-1}\right)$ with $\theta_{\omega}\left(x^{\prime}\right):=\left(x^{\prime}, \omega\left(x^{\prime}\right)\right)$ for $\omega \in C_{c}^{2-}\left(\mathbb{R}^{n-1}\right)$ and for constants parameters $\mu_{ \pm}>0$. This model problem reads as follows.

$$
\left\{\begin{align*}
\lambda u-\mu \Delta u & =f & & \text { in } \mathbb{R}^{n} \backslash \Sigma_{\omega},  \tag{2.16}\\
\llbracket \mu \partial_{\nu} u \rrbracket & =h_{1} & & \text { on } \Sigma_{\omega}, \\
\llbracket u \rrbracket & =h_{2} & & \text { on } \Sigma_{\omega} .
\end{align*}\right.
$$

2.13. Remark. Problem (2.16) can be reduced to a flat interface problem with the following transformation. We only assume that $\omega$ is of class $C^{1}\left(\mathbb{R}^{n-1}\right)$ and we consider the map

$$
\Theta_{\omega}: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}, \quad\left(x^{\prime}, x_{n}\right) \mapsto\left(x^{\prime}, x_{n}+\omega\left(x^{\prime}\right)\right) .
$$

(i) It is easy to check that $\Theta_{\omega}^{-1}$ is given by $\left(x^{\prime}, x_{n}\right) \mapsto\left(x^{\prime}, x_{n}-\omega\left(x^{\prime}\right)\right)$ and that

$$
\partial \Theta_{\omega}=\left[\begin{array}{cc}
I & 0 \\
\partial \omega & 1
\end{array}\right], \quad\left[\partial \Theta_{\omega}\right]^{-1}=\left[\begin{array}{cc}
I & 0 \\
-\partial \omega & 1
\end{array}\right], \quad \operatorname{det} \partial \Theta_{\omega}=1
$$

Hence both $\Theta_{\omega}: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ and $\left.\Theta_{\omega}\right|_{\Sigma_{0}}: \Sigma_{0} \rightarrow \Sigma_{\omega}$ are $C^{1}$-diffeomorphisms.
(ii) The hypersurface $\Sigma_{\omega}$ has
(a) the tangent vectors $\tau_{j} \circ \Theta_{\omega}=e_{j}+\partial_{j} \omega e_{n}$ for $j<n$,
(b) the unit normal vector $\nu \circ \Theta_{\omega}=\beta\left(e_{n}-\nabla \omega\right)$ with $\beta:=\left(1+|\nabla \omega|^{2}\right)^{-1 / 2}$,
(c) the cotangent vectors $\tau^{j} \circ \Theta_{\omega}=e_{j}+\beta^{2} \partial_{j} \omega\left(e_{n}-\nabla \omega\right)$.

For $p \in(1, \infty), k \in \mathbb{N}_{0}, \omega \in C_{c}^{k+2-}\left(\mathbb{R}^{n-1}\right)$, and $\lambda \in \mathbb{C} \backslash\{0\}$, we employ the function spaces

$$
\begin{aligned}
\mathbb{E}_{\lambda}^{k} & =\mathbb{E}_{\lambda}^{k}\left(\mathbb{R}^{n} \backslash \Sigma_{\omega}\right)=H_{p}^{k+2}\left(\mathbb{R}^{n} \backslash \Sigma_{\omega}\right), \\
\mathbb{F}_{\lambda}^{k} & =\mathbb{F}_{\lambda}^{k}\left(\mathbb{R}^{n} \backslash \Sigma_{\omega}\right)=H_{p}^{k}\left(\mathbb{R}^{n} \backslash \Sigma_{\omega}\right) \times W_{p}^{k+1-1 / p}\left(\Sigma_{\omega}\right) \times W_{p}^{k+2-1 / p}\left(\Sigma_{\omega}\right),
\end{aligned}
$$

equipped with the $\lambda$-dependent norms from Section 2.1.2.
2.14. Lemma. Let $\mu_{0} \in(0,1], k \in \mathbb{N}_{0}, \phi \in(0, \pi)$, and $p \in(1, \infty)$. Then there exists $\eta>0$ such that for every $M>0$ we can find some $\lambda_{0} \geq 1$ such that the operator

$$
A_{\lambda}: \mathbb{E}_{\lambda}^{k}\left(\mathbb{R}^{n} \backslash \Sigma_{\omega}\right) \rightarrow \mathbb{F}_{\lambda}^{k}\left(\mathbb{R}^{n} \backslash \Sigma_{\omega}\right), \quad u \mapsto\left(\lambda u-\operatorname{div}(\mu \nabla u), \llbracket \mu \partial_{\nu} u \rrbracket, \llbracket u \rrbracket\right),
$$

is uniformly invertible with respect to

$$
\omega \in C_{c}^{k+2-}\left(\mathbb{R}^{n-1}\right),\|\nabla \omega\|_{W_{\infty}^{k+1}} \leq M,\|\nabla \omega\|_{\infty} \leq \eta, \quad \mu_{ \pm} \in\left[\mu_{0}, \mu_{0}^{-1}\right], \quad \lambda \in \Sigma_{\phi},|\lambda| \geq \lambda_{0} .
$$

Proof. (i) We study a transformation of the functions $u \in \mathbb{E}_{\lambda}^{k}$ and $\left(f, h_{1}, h_{2}\right) \in \mathbb{F}_{\lambda}^{k}$ to a flat interface situation. The map $\Theta=\Theta_{\omega}$ from Remark 2.13 is a $C^{k+2-}$-diffeomorphism from $\mathbb{R}_{ \pm}^{n}:=\mathbb{R}^{n-1} \times \pm(0, \infty)$ onto $\Omega_{ \pm}:=\left\{\left(x^{\prime}, x_{n}\right) \in \mathbb{R}^{n}: x_{n} \gtrless \omega\left(x^{\prime}\right)\right\}$ and from $\Sigma_{0}$ onto $\Sigma_{\omega}$. Both $\partial \Theta$ and $\partial \Theta^{-1}$ belong to $W_{\infty}^{k+1}\left(\mathbb{R}^{n}\right)$. We consider the pull-backs

$$
\bar{u}=u \circ \Theta, \quad \bar{f}=f \circ \Theta, \quad \bar{h}_{j}=h_{j} \circ \Theta .
$$

By means of the chain rule (B.19) and the substitution formula (A.12), it follows that

$$
\begin{aligned}
\bar{u} & \in \overline{\mathbb{E}}_{\lambda}^{k}:=H_{p}^{k+2}\left(\dot{\mathbb{R}}^{n}\right), \\
\left(\bar{f}, \bar{h}_{1}, \bar{h}_{2}\right) & \in \overline{\mathbb{F}}_{\lambda}^{k}:=H_{p}^{k}\left(\dot{\mathbb{R}}^{n}\right) \times W_{p}^{k+1-1 / p}\left(\mathbb{R}^{n-1}\right) \times W_{p}^{k+2-1 / p}\left(\mathbb{R}^{n-1}\right),
\end{aligned}
$$

and that $u \mapsto \bar{u}, \mathbb{E}_{\lambda}^{k} \rightarrow \overline{\mathbb{E}}_{\lambda}^{k}$ and $\left(f, h_{1}, h_{2}\right) \mapsto\left(\bar{f}, \bar{h}_{1}, \bar{h}_{2}\right), \mathbb{F}_{\lambda}^{k} \rightarrow \overline{\mathbb{F}}_{\lambda}^{k}$ are topological linear isomorphisms. To be more precise, let $1 \leq j \leq k+2$. Then

$$
\left\|\lambda^{(k+2-j) / 2} \nabla^{j} \bar{u}\right\|_{p} \leq \sum_{i=1}^{j} \sum_{\beta, \sigma} \frac{|\lambda|^{-(j-i) / 2}}{i!\beta!}\left\|\lambda^{(k+2-i) / 2} \nabla^{i} u \circ \Theta\right\|_{p}\left\|\partial^{\beta_{1}} \Theta\right\|_{\infty} \cdots\left\|\partial^{\beta_{i}} \Theta\right\|_{\infty},
$$

where the sum is taken over multi-indices $\beta \in \mathbb{N}^{i}$ such that $|\beta|=j$ and all $j$ ! permutations $\sigma$ of $\{1, \ldots, j\}$. From det $\partial \Theta=1$ we infer that $\left\|\nabla^{i} u \circ \Theta\right\|_{p}=\left\|\nabla^{i} u\right\|_{p}$. This shows that

$$
\begin{equation*}
C(n, k, M)^{-1}\|u\|_{\mathbb{E}_{\lambda}^{k}} \leq\|\bar{u}\|_{\mathbb{\mathbb { E }}_{\lambda}^{k}} \leq C(n, k, M)\|u\|_{\mathbb{E}_{\lambda}^{k}}, \tag{2.17}
\end{equation*}
$$

where $C(n, k, M)$ is uniform with respect to those $\omega$ and $\lambda$ that satisfy $\|\nabla \omega\|_{W_{\infty}^{k+1}} \leq M$ and $|\lambda| \geq$ 1. The relevant estimates for $\bar{f}$ in $H_{p}^{k}\left(\dot{\mathbb{R}}^{n}\right)$ follow analogously and those of $\bar{h}_{j}^{\infty}$ in $H_{p}^{k-1+j}\left(\mathbb{R}^{n-1}\right)$
follow from (A.12). Finally, since $|\partial \Theta| \leq\left(1+|\nabla \omega|^{2}\right)^{1 / 2}$ and $\left|\partial \Theta^{-1}\right| \leq\left(1+|\nabla \omega|^{2}\right)^{1 / 2}$, we infer again from (A.12) that the Slobodeckiĭ semi-norm for $s \in(0,1)$ satisfies

$$
\begin{aligned}
\llbracket g \circ \Theta \rrbracket_{W_{p}^{s}\left(\mathbb{R}^{n-1}\right)} & \leq\left(1+\|\nabla \omega\|_{\infty}^{2}\right)^{s / 2+(n-1) / 2 p} \llbracket g \rrbracket_{W_{p}^{s}\left(\Sigma_{\omega}\right)}, \\
\llbracket \bar{g} \circ \Theta^{-1} \rrbracket_{W_{p}^{s}\left(\Sigma_{\omega}\right)} & \leq\left(1+\|\nabla \omega\|_{\infty}^{2}\right)^{s / 2+(n+1) / 2 p} \llbracket \bar{g} \rrbracket_{W_{p}^{s}\left(\mathbb{R}^{n-1}\right)} .
\end{aligned}
$$

We conclude that

$$
\begin{equation*}
C(n, k, p, M)^{-1}\left\|\left(f, h_{1}, h_{2}\right)\right\|_{\mathbb{F}_{\lambda}^{k}} \leq\left\|\left(\bar{f}, \bar{h}_{1}, \bar{h}_{2}\right)\right\|_{\mathbb{F}_{\lambda}^{k}} \leq C(n, k, p, M)\left\|\left(f, h_{1}, h_{2}\right)\right\|_{\mathbb{F}_{\lambda}^{k}} \tag{2.18}
\end{equation*}
$$

(ii) We derive the transformed problem. From $(\nabla u) \circ \Theta=[\partial \Theta]^{-\top} \nabla \bar{u}$ we infer that

$$
\partial_{\nu} u \circ \Theta=\left(\nu_{\Sigma_{\omega}} \cdot \nabla u\right) \circ \Theta=-\beta \nabla^{\prime} \omega \cdot \nabla^{\prime} \bar{u}+\beta^{-1} \partial_{n} \bar{u}, \quad \beta=\left(1+\left|\nabla^{\prime} \omega\right|^{2}\right)^{-1} .
$$

With $\Theta_{m}^{-1}:=\left(\Theta^{-1}\right)_{m}$ and $\partial_{j} \Theta_{m}^{-1}=\delta_{j m}-\delta_{m n} \partial_{j} \omega$ and $\Delta \Theta_{m}^{-1}=-\delta_{m n} \Delta \omega$, we obtain

$$
\begin{aligned}
(\Delta u) \circ \Theta & =\Delta \bar{u}+\sum_{j l m} \partial_{l} \partial_{m} \bar{u}\left(\partial_{j} \Theta_{l}^{-1} \partial_{j} \Theta_{m}^{-1}-\delta_{j l} \delta_{j m}\right)+\sum_{l} \partial_{l} \bar{u} \Delta \Theta_{l}^{-1}, \\
& =\Delta \bar{u}-2 \nabla \partial_{n} \bar{u} \cdot \nabla \omega+\partial_{n}^{2} \bar{u}|\nabla \omega|^{2}+\partial_{n} \bar{u} \Delta \omega .
\end{aligned}
$$

Therefore problem (2.16) is transformed to

$$
\begin{aligned}
\lambda \bar{u}-\mu \Delta \bar{u} & =\bar{f}+F_{2} \bar{u}+F_{1} \bar{u} & & \text { in } \dot{\mathbb{R}}^{n}, \\
\llbracket \mu \partial_{n} \bar{u} \rrbracket & =\bar{h}_{1}+H \bar{u} & & \text { on } \mathbb{R}^{n-1}, \\
\llbracket \bar{u} \rrbracket & =\bar{h}_{2} & & \text { on } \mathbb{R}^{n-1},
\end{aligned}
$$

where the perturbations $F_{l}=F_{l}(\mu, \omega)$ and $H=H(\mu, \omega)$ are given by

$$
\begin{aligned}
F_{1} \bar{u} & =-\mu \Delta^{\prime} \omega \partial_{n} \bar{u}, \\
F_{2} \bar{u} & =\mu\left|\nabla^{\prime} \omega\right|^{2} \partial_{n}^{2} \bar{u}-2 \mu \partial_{n} \nabla^{\prime} \bar{u} \cdot \nabla^{\prime} \omega, \\
H \bar{u} & =\beta \nabla^{\prime} \omega \cdot \llbracket \mu \nabla^{\prime} \bar{u} \rrbracket+\left(1-\beta^{-1}\right) \llbracket \mu \partial_{n} \bar{u} \rrbracket .
\end{aligned}
$$

(iii) Let us derive suitable estimates for $F_{l}$ and $H$. Our goal is to show that

$$
\left\|F_{l}(\mu, \omega)\right\|_{\mathcal{B}\left(\overline{E_{\lambda}^{k}} ; H_{p}^{k}\left(\mathbb{R}^{n}\right)\right), \lambda}+\|H(\mu, \omega)\|_{\mathcal{B}\left(\overline{\mathbb{E}}_{\lambda}^{k} ; W_{p}^{k+1-1 / p}\left(\mathbb{R}^{n-1}\right)\right), \lambda} \rightarrow 0 \quad \text { as } \eta \rightarrow 0 \text { and } \lambda_{0} \rightarrow \infty .
$$

To be precise, we shall show that for given $\varepsilon>0$, there exist $\eta=\eta\left(n, \mu_{0}, k, \phi, p, \varepsilon\right) \in(0,1]$ and $\lambda_{0}=\lambda_{0}\left(n, \mu_{0}, k, \phi, p, M, \varepsilon\right) \geq M^{-1}$ such that the estimate

$$
\left\|F_{l}(\mu, \omega) \bar{u}\right\|_{H_{p}^{k}\left(\mathbb{R}^{n}\right), \lambda}+\|H(\mu, \omega) \bar{u}\|_{W_{p}^{k+1-1 / p}\left(\mathbb{R}^{n-1}\right), \lambda} \leq \varepsilon\|\bar{u}\|_{\mathbb{E}_{\lambda}^{k}}
$$

is valid for all $\bar{u} \in \overline{\mathbb{E}}_{\lambda}^{k}$, all $l \in\{1,2\}$, all $\omega \in C_{c}^{k+2-}\left(\mathbb{R}^{n-1}\right)$ with $\|\nabla \omega\|_{W_{\infty}^{k+1}} \leq M$ and $\|\nabla \omega\|_{\infty} \leq \eta$, all $\mu_{ \pm} \in\left[\mu_{0}, \mu_{0}^{-1}\right]$ and all $\lambda \in \Sigma_{\phi}$ with $|\lambda| \geq \lambda_{0}$.

For an estimation of $F_{l}$ we let $0 \leq j \leq k$. The product rule and Hölder's inequality yield

$$
\begin{aligned}
\left\|\lambda^{(k-j) / 2} \nabla^{j}\left(F_{1}(\mu, \omega) \bar{u}\right)\right\|_{p} & \leq C\left(n, \mu_{0}, k\right) \sum_{i=0}^{j}|\lambda|^{-1 / 2-(j-i) / 2}\left\|\nabla^{j-i+2} \omega\right\|_{\infty}\left\|\lambda^{(k+2-i) / 2} \nabla^{i} \bar{u}\right\|_{p} \\
& \leq C\left(n, \mu_{0}, k, M\right)|\lambda|^{-1 / 2}\|\bar{u}\|_{\overline{\mathbb{E}}_{\lambda}^{k}} .
\end{aligned}
$$

In the norm of $F_{2} \bar{u}$ we control the leading order terms with a factor $\eta$ as follows. For $\alpha \in \mathbb{N}_{0}^{n}$ with $|\alpha|=k$ and for $0 \leq j \leq k$ we have

$$
\begin{aligned}
\left\|\lambda^{(k-|\alpha|) / 2} F_{2}(\mu, \omega)\left(\partial_{x}^{\alpha} \bar{u}\right)\right\|_{p} \leq \eta C\left(n, \mu_{0}, k\right)\|\bar{u}\|_{\mathbb{E}_{\lambda}^{k}} \\
\left\|\lambda^{(k-j) / 2} \nabla^{j}\left(F_{2}(\mu, \omega) \bar{u}\right)\right\|_{p} \leq \eta C\left(n, \mu_{0}, k\right)\|\bar{u}\|_{\overline{\mathbb{E}}_{\lambda}^{k}}+|\lambda|^{-1 / 2} C\left(n, \mu_{0}, k, M\right)\|\bar{u}\|_{\mathbb{E}_{\lambda}^{k}} .
\end{aligned}
$$

We emphasize that the coefficient $C\left(n, \mu_{0}, k\right)$ near $\eta$ does not depend on the bound $M$ for the derivatives of $\omega$. For the estimations of $H$, we use the property $\|1-\beta(\omega)\|_{\infty} \rightarrow 0$ as $\|\nabla \omega\|_{\infty} \rightarrow 0$ and the pointwise multiplication estimate (B.7). Furthermore, a scaling argument yields

$$
\left\|\left.v\right|_{x_{n}=0}\right\|_{W_{p}^{m-1 / p}\left(\mathbb{R}^{n-1}\right), \lambda} \leq C(n, p, m)\|v\|_{H_{p}^{m}\left(\mathbb{R}_{+}^{n}\right), \lambda} \quad \text { for } v \in H_{p}^{m}\left(\mathbb{R}_{+}^{n}\right), m \in \mathbb{N}, \lambda \in \mathbb{C} \backslash\{0\} .
$$

Then the leading order terms in the norm of $H \bar{u}$ are estimated by means of

$$
\llbracket H \partial^{\alpha} \bar{u} \rrbracket_{W_{p}^{1-1 / p}} \leq \eta C\left(n, \mu_{0}, p\right) \llbracket \partial^{\alpha} \bar{u} \rrbracket_{W_{p}^{1-1 / p}}+C\left(n, \mu_{0}, p, M\right)\left\|\partial^{\alpha} \bar{u}\right\|_{p}, \quad \text { for }|\alpha|=k,
$$

where we have used $|\nabla \omega| \leq \eta \leq 1$ and $\beta \leq 1$. We therefore obtain the estimate

$$
\|H(\eta, \omega) \bar{u}\|_{W_{p}^{k+1-1 / p}\left(\mathbb{R}^{n-1}\right), \lambda} \leq \eta C\left(n, \mu_{0}, k, p\right)\|\bar{u}\|_{\overline{\mathbb{E}}_{\lambda}^{k}}+|\lambda|^{-1 / 2+1 / 2 p} C\left(n, \mu_{0}, k, p, M\right)\|\bar{u}\|_{\mathbb{\mathbb { E }}_{\lambda}^{k}} .
$$

(iv) We finally consider the operators

$$
\begin{gathered}
\bar{A}_{\lambda}: \bar{u} \mapsto\left((\lambda-\mu \Delta) \bar{u}, \llbracket \mu \partial_{n} \bar{u} \rrbracket, \llbracket \bar{u} \rrbracket\right), \\
P(\mu, \omega): \bar{u} \mapsto\left(F_{2} \bar{u}+F_{1} \bar{u}, H_{2} \bar{u}+H_{1} \bar{u}, 0\right) .
\end{gathered}
$$

In Lemma 2.10 we have proved that $\bar{A}_{\lambda}: \overline{\mathbb{E}}_{\lambda}^{k} \rightarrow \overline{\mathbb{F}}_{\lambda}^{k}$ is invertible and that

$$
\left\|\left(\bar{A}_{\lambda}\right)^{-1}\left(\bar{f}, \bar{h}_{1}, \bar{h}_{2}\right)\right\|_{\mathbb{E}_{\lambda}^{k}} \leq C\left(n, \mu_{0}, p, \phi, k\right)\left\|\left(\bar{f}, \bar{h}_{1}, \bar{h}_{2}\right)\right\|_{\mathbb{F}_{\lambda}^{k}} .
$$

With step (iii) we can choose numbers $\eta\left(n, \mu_{0}, k, \phi, p\right) \in(0,1]$ and $\lambda_{0}\left(n, \mu_{0}, k, \phi, p, M\right) \geq 1$ such that $\left\|\left(\bar{A}_{\lambda}\right)^{-1} P(\mu, \omega) \bar{u}\right\| \leq 2^{-1}\|\bar{u}\|_{\mathbb{E}_{\lambda}^{k}}$. Then a Neumann series argument and the pull-back estimates (2.17) and (2.18) imply that the desired solution $u \in \mathbb{E}_{\lambda}^{k}$ to (2.16) is given by

$$
u=A_{\lambda}^{-1}\left(f, h_{1}, h_{2}\right)=\left(\left(I-\left(\bar{A}_{\lambda}^{-1}\right) P(\mu, \omega)\right)^{-1}\left(\bar{A}_{\lambda}\right)^{-1}\left(f \circ \Theta^{-1}, h_{1} \circ \Theta^{-1}, h_{2} \circ \Theta^{-1}\right)\right) \circ \Theta .
$$

This representation also shows the uniform bounds for $A_{\lambda}^{-1}$.
Next, we consider the perturbed model problem

$$
\left\{\begin{align*}
\lambda u-\operatorname{div}(\mu \nabla u) & =f & & \text { in } \mathbb{R}^{n} \backslash \Sigma_{\omega},  \tag{2.19}\\
\llbracket \mu \partial_{\nu} u \rrbracket & =h_{1} & & \text { on } \Sigma_{\omega}, \\
\llbracket u \rrbracket & =h_{2} & & \text { on } \Sigma_{\omega},
\end{align*}\right.
$$

with variable coefficients

$$
\mu_{ \pm}: \Omega_{ \pm} \rightarrow(0, \infty), \quad \Omega_{ \pm}:=\left\{\left(x^{\prime}, x_{n}\right) \in \mathbb{R}^{n}: x_{n} \gtrless \omega\left(x^{\prime}\right)\right\} .
$$

2.15. Lemma. Let $\mu_{0} \in(0,1], k \in \mathbb{N}_{0}, \phi \in(0, \pi)$, and $p \in(1, \infty)$. Then there exists $\eta>0$ such that for every $M>0$ we can find some $\lambda_{0} \geq 1$ such that the operator

$$
A_{\lambda}: \mathbb{E}_{\lambda}^{k}\left(\mathbb{R}^{n} \backslash \Sigma_{\omega}\right) \rightarrow \mathbb{F}_{\lambda}^{k}\left(\mathbb{R}^{n} \backslash \Sigma_{\omega}\right), \quad u \mapsto\left(\lambda u-\operatorname{div}(\mu \nabla u), \llbracket \mu \partial_{\nu} u \rrbracket, \llbracket u \rrbracket\right)
$$

is uniformly invertible with respect to
(i) $\omega \in C_{c}^{k+2-}\left(\mathbb{R}^{n-1}\right)$ with $\|\nabla \omega\|_{W_{\infty}^{k+1}} \leq M$ and $\|\nabla \omega\|_{\infty} \leq \eta$,
(ii) $\mu_{ \pm} \in W_{\infty}^{k+1}\left(\Omega_{ \pm}\right)$with $\mu_{0} \leq \mu_{ \pm} \leq \mu_{0}^{-1},\left\|\mu_{ \pm}\right\|_{W_{\infty}^{k+1}} \leq M, \sup \left\{\left|\mu_{ \pm}(x)-\mu_{ \pm}(y)\right|: x, y \in \Omega_{ \pm}\right\} \leq$ $2 \eta$,
(iii) $\lambda \in \Sigma_{\phi}$ with $|\lambda| \geq \lambda_{0}$.

Proof. We may choose constants $\mu_{ \pm}^{*} \in\left[\mu_{0}, \mu_{0}^{-1}\right]$ such that $\sup \left\{\left|\mu_{ \pm}(x)-\mu_{ \pm}^{*}\right|: x \in \Omega_{ \pm}\right\} \leq \eta$, for instance $\mu_{ \pm}^{*}:=\left(\sup \mu_{ \pm}+\inf \mu_{ \pm}\right) / 2$. Then problem (2.19) can be written as

$$
\begin{aligned}
\lambda u-\mu^{*} \Delta u & =f+\operatorname{div}\left(\left(\mu-\mu^{*}\right) \nabla u\right) & & \text { in } \mathbb{R}^{n} \backslash \Sigma_{\omega}, \\
\llbracket \mu^{*} \partial_{\nu} u \rrbracket & =h_{1}+\llbracket\left(\mu^{*}-\mu\right) \partial_{\nu} u \rrbracket & & \text { on } \Sigma_{\omega}, \\
\llbracket u \rrbracket & =h_{2} & & \text { on } \Sigma_{\omega} .
\end{aligned}
$$

Let $A_{\lambda}^{*}: u \mapsto\left(\left(\lambda-\mu^{*} \Delta\right) u, \llbracket \mu^{*} \partial_{\nu} u \rrbracket, \llbracket u \rrbracket\right)$ and $P: u \mapsto\left(\operatorname{div}\left(\left(\mu-\mu^{*}\right) \nabla u\right), \llbracket\left(\mu^{*}-\mu\right) \partial_{\nu} u \rrbracket, 0\right)$. With $\left\|\mu_{ \pm}-\mu_{ \pm}^{*}\right\|_{\infty} \leq \eta$ and similar estimates as for the perturbations in Lemma 2.14, we obtain

$$
\begin{aligned}
& \left\|\operatorname{div}\left(\left(\mu-\mu^{*}\right) \nabla u\right)\right\|_{H_{p}^{k}\left(\mathbb{R}^{n} \backslash \Sigma_{\omega}\right), \lambda} \leq\left(\eta C\left(n, \mu_{0}, k\right)+|\lambda|^{-1 / 2} C\left(n, \mu_{0}, k, M\right)\right)\|u\|_{\mathbb{E}_{\lambda}^{k}}, \\
& \left\|\llbracket\left(\mu^{*}-\mu\right) \partial_{\nu} u \rrbracket\right\|_{W_{p}^{k+1-1 / p}\left(\Sigma_{\omega}\right), \lambda} \leq\left(\eta C\left(n, \mu_{0}, k, p\right)+|\lambda|^{-1 / 2+1 / 2 p} C\left(n, \mu_{0}, k, p, M\right)\right)\|u\|_{\mathbb{E}_{\lambda}^{k}} .
\end{aligned}
$$

The operators $A_{\lambda}^{*}$ are uniformly invertible by Lemma 2.14 and a Neumann series argument implies that for some $\eta>0$ and $\lambda_{0} \geq 1$, the operator $A_{\lambda}=A_{\lambda}^{*}-P$ is uniformly invertible.

Lemma 2.15 includes the case $\Omega \backslash \Sigma=\mathbb{R}^{n}$, since $\mu$ is allowed to be continuous across $\Sigma_{0}$.
2.16. Corollary. Let $\mu_{0} \in(0,1], k \in \mathbb{N}_{0}, \phi \in(0, \pi)$, and $p \in(1, \infty)$. Then there exists $\eta>0$ such that for every $M>0$ we can find some $\lambda_{0} \geq 1$ such that the operator

$$
A_{\lambda}: \mathbb{E}_{\lambda}^{k}\left(\mathbb{R}^{n}\right) \rightarrow \mathbb{F}_{\lambda}^{k}\left(\mathbb{R}^{n}\right), \quad u \mapsto \lambda u-\operatorname{div}(\mu \nabla u),
$$

is uniformly invertible with respect to
(i) $\mu \in W_{\infty}^{k+1}\left(\mathbb{R}^{n}\right)$ with $\mu_{0} \leq \mu \leq \mu_{0}^{-1},\|\mu\|_{W_{\infty}^{k+1}} \leq M$ and $\sup \left\{|\mu(x)-\mu(y)|: x, y \in \mathbb{R}^{n}\right\} \leq 2 \eta$,
(ii) $\lambda \in \Sigma_{\phi}$ with $|\lambda| \geq \lambda_{0}$.

The bent half-space problem can be solved analogously as the bent interface problem, by using the half-space result Lemma 2.12 instead of the flat interface result Lemma 2.10.
2.17. Corollary. Let $\mu_{0} \in(0,1], k \in \mathbb{N}_{0}, \phi \in(0, \pi)$, and $p \in(1, \infty)$. Then there exists $\eta>0$ such that for every $M>0$ we can find some $\lambda_{0} \geq 1$ such that the operator

$$
A_{\lambda}: \mathbb{E}_{\lambda}^{k}\left(\mathbb{R}_{\omega}^{n}\right) \rightarrow \mathbb{F}_{\lambda}^{k}\left(\mathbb{R}_{\omega}^{n}\right), \quad u \mapsto\left(\lambda u-\operatorname{div}(\mu \nabla u), \mu \partial_{\nu} u\right),
$$

is uniformly invertible with respect to
(i) $\omega \in C_{c}^{k+2-}\left(\mathbb{R}^{n-1}\right)$ with $\|\nabla \omega\|_{W_{\infty}^{k+1}} \leq M$ and $\|\nabla \omega\|_{\infty} \leq \eta$,
(ii) $\mu \in W_{\infty}^{k+1}\left(\mathbb{R}_{\omega}^{n}\right)$ with $\mu_{0} \leq \mu \leq \mu_{0}^{-1},\|\mu\|_{W_{\infty}^{k+1}} \leq M$ and $\sup \left\{|\mu(x)-\mu(y)|: x, y \in \Omega_{ \pm}\right\} \leq 2 \eta$,
(iii) $\lambda \in \Sigma_{\phi}$ with $|\lambda| \geq \lambda_{0}$.
2.1.5. Bounded domains. We solve the strong transmission problem (2.9) in a bounded domain $\Omega \subset \mathbb{R}^{n}(n \geq 2)$ with boundary $\partial \Omega \in C^{k+2-}\left(k \in \mathbb{N}_{0}\right)$ and compact interface $\Sigma \subset \Omega$ of class $C^{k+2-}$ for variable coefficients $\mu_{ \pm}: \Omega_{ \pm} \rightarrow(0, \infty)$.
2.18. Theorem. Let $\mu_{0} \in(0,1], \phi \in(0, \pi)$, and $p \in(1, \infty)$. Then there exists $\eta>0$ such that for all $M>0$ we can find some $\lambda_{0} \geq 1$ such that the operator

$$
A_{\lambda}: \mathbb{E}_{\lambda}^{k}(\Omega \backslash \Sigma) \rightarrow \mathbb{F}_{\lambda}^{k}(\Omega \backslash \Sigma), \quad u \mapsto\left(\lambda u-\operatorname{div}(\mu \nabla u), \mu \partial_{\nu} u, \llbracket \mu \partial_{\nu} u \rrbracket, \llbracket u \rrbracket\right),
$$

is uniformly invertible with respect to
(i) $\mu \in W_{\infty}^{k+1}(\Omega \backslash \Sigma)$ with $\mu_{0} \leq \mu_{ \pm} \leq \mu_{0}^{-1}$ and $\left\|\mu_{ \pm}\right\|_{W_{\infty}^{k+1}} \leq M$,
(ii) $\lambda \in \Sigma_{\phi}$ with $|\lambda| \geq \lambda_{0}$.

Proof. We apply our localization technique from Section 2.1.1.
(i) We define the global spaces

$$
E:=E_{\lambda}:=\mathbb{E}_{\lambda}^{k}(\Omega \backslash \Sigma), \quad F:=F_{\lambda}:=\mathbb{F}_{\lambda}^{k}(\Omega \backslash \Sigma),
$$

equipped with the $\lambda$-dependent norms from Section 2.1.2.
For definining local spaces we employ Lemma 2.9, which implies that for every $\eta>0$ there exist a number $r_{0}=r_{0}(\eta)>0$ and an $\left(\eta, r_{0}(\eta)\right)$-localization set-up. In particular, we can find a finite set $I=I\left(\eta, r_{0}(\eta)\right)$, an open covering for $\bar{\Omega}$ of balls $U_{j}=B_{r_{0}}\left(p_{j}\right)(j \in I)$, rigid transformations

$$
\Theta_{j}: x \mapsto p_{j}+Q_{j} x, \quad B_{r_{0}}(0) \rightarrow U_{j},
$$

and height functions $\omega_{j} \in C_{c}^{k+2-}\left(\mathbb{R}^{n-1}\right)$ with $\left\|\nabla \omega_{j}\right\|_{\infty} \leq \eta$ and $\|\nabla \omega\|_{W_{\infty}^{k+1}} \leq M(r)$. Furthermore, the index set can be decomposed into $I=I_{1} \cup I_{2} \cup I_{3}$, where $j \in I_{1}$ corresponds to the whole space case $\Omega \cap U_{j}=\Theta_{j}\left(\mathbb{R}^{n} \cap B_{r_{0}}\right), j \in I_{2}$ corresponds to the bent half-space case $\Omega \cap U_{j}=$ $\Theta_{j}\left(\mathbb{R}_{\omega_{j}}^{n} \cap B_{r_{0}}\right)$, and $j \in I_{3}$ corresponds to the bent hyperplane case $\Sigma \cap U_{j}=\Theta_{j}\left(\Sigma_{\omega_{j}} \cap B_{r_{0}}\right)$. Then we define

$$
\begin{array}{lll}
\Omega_{j}:=\mathbb{R}^{n}, & \Sigma_{j}:=\emptyset & \text { for } j \in I_{1}, \\
\Omega_{j}:=\mathbb{R}_{\omega_{\omega}}^{n}, & \Sigma_{j}:=\emptyset & \text { for } j \in I_{2}, \\
\Omega_{j}:=\mathbb{R}^{n}, & \Sigma_{j}:=\Sigma_{\omega_{j}} & \text { for } j \in I_{3} .
\end{array}
$$

Now we define the local spaces

$$
E_{j}:=E_{j, \lambda}:=\mathbb{E}_{\lambda}^{k}\left(\Omega_{j} \backslash \Sigma_{j}\right), \quad F_{j}:=F_{j, \lambda}:=\mathbb{F}_{\lambda}^{k}\left(\Omega_{j} \backslash \Sigma_{j}\right) \quad \text { for } j \in I_{1} \cup I_{2} \cup I_{3} .
$$

We keep in mind that these definitions depend on the localization set-up and this will be fixed in step (iv) during the definition of the local operators.
(ii) We next define approximation systems for $E$ and $F$. Choose a smooth partition of unity $\left(\varphi_{j}\right)_{j \in I}$ for $\bar{\Omega}$ in $\mathbb{R}^{n}$ subordinate to $\left(U_{j}\right)_{j \in I}$ and choose smooth cut-off functions $\left(\psi_{j}\right)_{j \in I}$ with $\operatorname{supp} \psi_{j} \subset B_{r_{0}}$ and $\psi_{j} \circ \Theta_{j}^{-1}=1$ on $\operatorname{supp} \varphi_{j}$. Then we have $\sum_{j} \psi_{j} \circ \Theta_{j}^{-1} \varphi_{j}=1 \mathrm{in} \bar{\Omega}$. Define

$$
\begin{aligned}
\Phi_{E, j} u & :=\left(\varphi_{j} u\right) \circ \Theta_{j}, & \Phi_{F, j}\left(f, g, h_{1}, h_{2}\right): & =\left(\varphi_{j} f, \varphi_{j} g, \varphi_{j} h_{1}, \varphi_{j} h_{2}\right) \circ \Theta_{j}, \\
\Psi_{E, j} u_{j} & :=\left(\psi_{j} u_{j}\right) \circ \Theta_{j}^{-1}, & \Psi_{F, j}\left(f_{j}, g_{j}, h_{1 j}, h_{2 j}\right) & :=\left(\psi_{j} f_{j}, \psi_{j} g_{j}, \psi_{j} h_{1 j}, \psi_{j} h_{2 j}\right) \circ \Theta_{j}^{-1} .
\end{aligned}
$$

The triples $\left(\mathbf{E},\left(\Phi_{E, j}\right),\left(\Psi_{E, j}\right)\right)$ and $\left(\mathbf{F},\left(\Phi_{F, j}\right),\left(\Psi_{F, j}\right)\right)$ are indeed $l_{q}$-approximation systems for $E$ and $F$, as can be checked by means of pointwise multiplication $W_{p}^{1-1 / p} \times W_{\infty}^{1} \rightarrow W_{p}^{1-1 / p}$ (B.7), the chain rule, the substitution formula (A.12), and the regularity condition $\omega_{j} \in C_{c}^{k+2-}\left(\mathbb{R}^{n-1}\right)$. Moreover, we may choose any $q \in[1, \infty)$, since the set $I$ is finite. Furthermore, the retractions $r_{E}$ and $r_{F}$ and the co-retractions $r_{E}^{c}$ and $r_{F}^{c}$ are defined by

$$
\begin{array}{rlrl}
r_{E}^{c} u & :=\left(\Phi_{E, j} u\right)_{j \in J}, & r_{F}^{c}\left(f, g, h_{1}, h_{2}\right) & :=\left(\Phi_{F, j}\left(f, g, h_{1}, h_{2}\right)\right)_{j \in J}, \\
r_{E}\left(u_{j}\right)_{j \in J} & :=\sum_{j \in J} \Psi_{E, j} u_{j}, & r_{F}\left(f_{j}, g_{j}, h_{1 j}, h_{2 j}\right)_{j \in J}:=\sum_{j \in J} \Psi_{F, j}\left(f_{j}, g_{j}, h_{1 j}, h_{2 j}\right) .
\end{array}
$$

These operators satisfy the estimate

$$
\left\|r_{X}\right\|_{\mathcal{B}\left(l_{q}(\mathbf{X}) ; X\right), \lambda}+\left\|r_{X}^{c}\right\|_{\mathcal{B}\left(X ; l_{q}(\mathbf{X})\right), \lambda} \leq C(n, p, k, I(\eta, r), M(r), q) \quad \text { for } X \in\{E, F\} .
$$

The numbers $\eta$ and $r$ will be fixed below for proving optimal regularity of the relevant model problems. Then the remaining perturbations will be controlled only by the largeness of $\lambda_{0}$.
(iii) In order to define the local operators $A_{\lambda, j}$, we first have to define local coefficients. As for the construction of $\omega_{j}$ in Lemma 2.9, we fix a smooth cut-off function $\chi \in \mathcal{B}\left(\mathbb{R}^{n}\right)$ with $0 \leq \chi \leq 1$, $\chi(x)=1$ for $|x| \leq 1$ and $\chi(x)=0$ for $|x| \geq 2$. For $j \in I_{3}$, we consider the transformed coefficients $\bar{\mu}_{j}:=\mu \circ \Theta_{j}$ that are defined on $B_{r_{0}} \cap \Omega_{j} \backslash \Sigma_{j}$. Given a radius $r \in\left(0, r_{0} / 2\right]$, we define

$$
\tilde{\mu}_{j, r, \pm}(x):=\bar{\mu}_{j, \pm}(0)+ \begin{cases}\chi(x / r)\left(\bar{\mu}_{j, \pm}(x)-\bar{\mu}_{j, \pm}(0)\right) & \text { for } x \in \Omega_{j, \pm},|x|<2 r, \\ 0 & \text { for } x \in \Omega_{j, \pm,},|x| \geq 2 r .\end{cases}
$$

Then $\tilde{\mu}_{j, r, \pm}(x)=\bar{\mu}_{j, \pm}(x)$ for all $x \in B_{r} \cap \Omega_{j, \pm}$ and $\left\|\tilde{\mu}_{j, r, \pm}-\bar{\mu}_{j, \pm}(0)\right\|_{\infty} \rightarrow 0$ as $r \rightarrow 0$ by uniform continuity of $\bar{\mu}_{j, \pm}$. Hence, for given $\eta>0$ we can fix a number $r=r(\eta) \in\left(0, r_{0} / 2\right]$ to ensure that the local coefficients $\mu_{j}:=\tilde{\mu}_{j, r}$ satisfy $\left|\mu_{j, \pm}(x)-\mu_{j, \pm}(y)\right| \leq 2 \eta$ for all $x, y \in \Omega_{j, \pm}$, and all $j \in I_{3}$. In the case $j \in I_{1} \cup I_{2}$, we define $\mu_{j}$ analogously.
(iv) Now we define local operators $A_{\lambda, j}$ and fix the chart radius $r$ such that these operators are invertible and satisfy Assumption 2.5.(iv). Given a function $u_{j} \in E_{j}$, we let

$$
A_{\lambda, j} u_{j}:= \begin{cases}\lambda u_{j}-\operatorname{div}\left(\mu_{j} \nabla u_{j}\right) & \text { if } j \in J_{1}, \\ \left(\lambda u_{j}-\operatorname{div}\left(\mu_{j} \nabla u_{j}\right), \mu_{j} \partial_{\nu} u_{j}\right) & \text { if } j \in J_{2}, \\ \left(\lambda u_{j}-\operatorname{div}\left(\mu_{j} \nabla u_{j}\right), \llbracket \mu_{j} \partial_{\nu} u_{j} \rrbracket, \llbracket u_{j} \rrbracket\right) & \text { if } j \in J_{3} .\end{cases}
$$

By Corollaries 2.16 and 2.17 and Lemma 2.15, we can find a number $\eta\left(n, \mu_{0}, k, \phi, p\right)>0$ such that for all $M \geq 1$, there exists $\lambda_{0}\left(n, \mu_{0}, k, \phi, p, M\right) \geq 1$ such that the operators $A_{\lambda, j}(j \in I)$ are uniformly invertible with respect to $\omega_{j} \in C_{c}^{k+2-}\left(\mathbb{R}^{n-1}\right)$ with $\left\|\nabla \omega_{j}\right\|_{W_{\infty}^{k+1}} \leq M,\left\|\nabla \omega_{j}\right\|_{\infty} \leq \eta$; and $\mu_{j} \in W_{\infty}^{k+1}\left(\Omega_{j} \backslash \Sigma_{j}\right)$ with $\mu_{0} \leq \mu_{j} \leq \mu_{0}^{-1},\left\|\mu_{j, \pm}\right\|_{W_{\infty}^{k+1}} \leq M$, and $\sup \left\{\left|\mu_{j, \pm}(x)-\mu_{j, \pm}(y)\right|:\right.$ $\left.x, y \in \Omega_{j, \pm}\right\} \leq 2 \eta$; and $\lambda \in \Sigma_{\phi}$ with $|\lambda| \geq \lambda_{0}$. In order to fulfill these conditions, we now fix a number $r \in\left(0, r_{0} / 2\right]$ and an $(\eta, r)$-localization set-up $\left(U_{j}, \Theta_{j}, \omega_{j}\right)_{j \in I(\eta, r)}$ such that $\left\|\nabla \omega_{j}\right\|_{\infty} \leq \eta$ and $\left|\mu_{ \pm}(x)-\mu_{ \pm}\left(p_{j}\right)\right| \leq \eta$ for all $x \in U_{j} \cap \Omega_{ \pm}$and all $j \in I(\eta, r)$. By compactness of $\partial \Omega$ and $\Sigma$ and since $I$ is finite, there exists $M=M(\Omega, \Sigma, r)>0$ such that $\left\|\nabla \omega_{j}\right\|_{W_{\infty}^{k+1}} \leq M$ and $\left\|\mu_{j}\right\|_{W_{\infty}^{k+1}} \leq M$ for all $j \in I(\eta, r)$. Now the aforementioned results yield suitable numbers $\lambda_{0}$ and $C$ such that

$$
\left\|A_{\lambda, j}^{-1}\right\|_{\mathcal{B}\left(F_{j} ; E_{j}\right), \lambda} \leq C \quad \text { for } j \in I, \lambda \in \Sigma_{\phi},|\lambda| \geq \lambda_{0}
$$

(v) Finally, we consider the perturbations $B_{\lambda, j}$ and $C_{\lambda, j}$. Since the mappings $\Theta_{j}$ are affine and since $\mu_{j}(x)=\mu\left(\Theta_{j}(x)\right)$ for $x \in B_{r} \cap \Omega_{j}$, we obtain

$$
\begin{aligned}
B_{\lambda, j} u & =\Phi_{F, j} A_{\lambda} u-A_{\lambda, j} \Phi_{E, j} u=\left(\varphi_{j} A_{\lambda} u\right) \circ \Theta_{j}-A_{\lambda, j}\left((\varphi u) \circ \Theta_{j}\right) \\
& =\left(\mu \nabla \varphi_{j} \cdot \nabla u+\operatorname{div}\left(\mu \nabla \varphi_{j} u\right),-\mu u \partial_{\nu} \varphi_{j},-\llbracket \mu u \rrbracket \partial_{\nu} \varphi_{j}, 0\right) \circ \Theta_{j} .
\end{aligned}
$$

This commutator is of lower order and therefore

$$
\left\|B_{\lambda, j} u\right\|_{F_{j}, \lambda} \leq|\lambda|^{-1 / 2+1 / 2 p} C\left(n, \mu_{0}, k, \phi, p, M\right)\|u\|_{E, \lambda} \quad \text { for } u \in E, \lambda \in \mathbb{C} \backslash\{0\},|\lambda| \geq 1, j \in I .
$$

Since $I$ is finite and $q \in[1, \infty)$, it follows that

$$
\sup _{0 \neq u \in E} \frac{\left\|\left(B_{\lambda, j} u\right)_{j \in I}\right\|_{l_{q}(\mathbf{F}), \lambda}}{\|u\|_{E, \lambda}} \leq|\lambda|^{-1 / 2+1 / 2 p} C\left(n, \mu_{0}, k, \phi, p, M,|I|, q\right) .
$$

For the perturbations $C_{\lambda, j}=A_{\lambda} \Psi_{E, j}-\Psi_{F, j} A_{\lambda, j}$ we obtain

$$
\sup _{0 \neq\left(u_{j}\right)_{j \in I} \in l_{q}(\mathbf{E})} \frac{\left\|\sum_{j} C_{\lambda, j} u_{j}\right\|_{F, \lambda}}{\left\|\left(u_{j}\right)_{j \in I}\right\|_{L_{q}(\mathbf{E}), \lambda}} \leq|\lambda|^{-1 / 2+1 / 2 p} C\left(n, \mu_{0}, k, \phi, p, M,|I|, q\right) .
$$

Therefore Assumption 2.5 is satisfied and Proposition 2.6 yields the assertion.

### 2.2. Transmission problems for $\operatorname{div}(\mu \nabla \cdot)$

We prove optimal $\dot{H}_{p}^{k+2}$-regularity for the strong transmission problem (2.1) and optimal $\dot{H}_{p}^{1}{ }^{-}$ regularity for the weak transmission problem (2.2). In Section 2.2 .1 we define the solution space $\mathbb{E}^{k}$ and the data space $\mathbb{F}^{k}$ that are equipped with equivalent $\lambda$-dependent norms. For the basic model problems in $\mathbb{R}^{n}, \mathbb{R}_{+}^{n}$, and $\mathbb{R}^{n}$, we prove optimal $\mathbb{E}^{k}$-regularity in Section 2.2 .2 , uniformly with respect to $\lambda$ and $\omega$. Perturbed model problems are solved in Section 2.2 .3 for sufficiently large $\lambda$. For a bounded domain $\Omega \subset \mathbb{R}^{n}$ with compact hypersurface $\Sigma \subset \Omega$, we solve the weak transmission problem in Section 2.2.4 and the strong transmission problem in Section 2.2.5.
2.2.1. $\lambda$-dependent norms for $\operatorname{div}(\mu \nabla \cdot)$. For $p \in(1, \infty), k \in \mathbb{N}_{0}$, and an open set $G \subset \mathbb{R}^{n}$, we consider the semi-normed vector space $\dot{\mathcal{H}}_{p}^{k}(G)$ from page 23. The semi-norm $\|\phi\|_{\dot{\mathcal{H}}_{p}^{k}(G)}=$ $\left\|\nabla^{k} \phi\right\|_{L_{p}(G)}$ vanishes if and only if $\phi$ belongs to the vector space $\mathcal{P}_{k-1}$ of all polynomials of degree not larger than $k-1$ [cf. Gru09, Theorem 4.19]. Therefore its quotient space

$$
\dot{H}_{p}^{k}(G):=\dot{\mathcal{H}}_{p}^{k}(G) / \mathcal{P}_{k-1}, \quad\|\phi\|_{\dot{H}_{p}^{k}(G)}:=\left\|\nabla^{k} \phi\right\|_{L_{p}(G)},
$$

the homogeneous Sobolev space, is a Banach space [cf. Gal11, Exercise III.1.2]. Alternatively, the vector space $\dot{\mathcal{H}}_{p}^{k}(G)$ becomes a Banach space when it is endowed with the norm

$$
\|\phi\|_{\dot{\mathcal{H}}_{p}^{k}(G) \cap L_{p}\left(G^{\prime}\right)}:=\left\|\nabla^{k} \phi\right\|_{L_{p}(G)}+\|\phi\|_{L_{p}\left(G^{\prime}\right)},
$$

with some non-empty bounded smooth subdomain $G^{\prime} \subset G$. The corresponding norms for different subdomains $G^{\prime}$ are equivalent [cf. Gal11, Section III.1].

Let $\Omega$ and $\Sigma$ satisfy Assumption 2.1. We consider the semi-normed vector space

$$
\dot{\mathcal{H}}_{p}^{k}(\dot{\Omega})=\dot{\mathcal{H}}_{p}^{k}(\Omega \backslash \Sigma):=\left\{u \in H_{p, \text { loc }}^{k}(\Omega \backslash \Sigma): u_{ \pm} \in H_{p, \mathrm{loc}}^{k}\left(\overline{\Omega_{ \pm}}\right), \nabla^{k} u_{ \pm} \in L_{p}\left(\Omega_{ \pm}\right)\right\},
$$

whose semi-norm $\|u\|_{\mathcal{H}_{p}^{k}(\Omega \backslash \Sigma)}:=\left\|\nabla^{k} u\right\|_{L_{p}(\Omega)}$ vanishes if and only if $u_{ \pm} \in \mathcal{P}_{k-1}$. Then, given $k \in \mathbb{N}_{0} \cup\{-1\}, \lambda \in(0, \infty)$, and $p \in(1, \infty)$, we define the solution space

$$
\begin{equation*}
\mathbb{E}_{\lambda}^{k}:=\left(\mathbb{E}^{k},\|\cdot\|_{\mathbb{E}_{\lambda}^{k}}\right), \quad \mathbb{E}^{k}:=\left(\bigcap_{j=1}^{k+2} \dot{\mathcal{H}}_{p}^{j}(\Omega \backslash \Sigma)\right) / \mathbb{K} \tag{2.20}
\end{equation*}
$$

which is a Banach space with respect to the equivalent $\lambda$-dependent norm
$\|u\|_{\mathbb{E}_{\lambda}^{k}}:=\sum_{j=1}^{k+2}\left\|\lambda^{(k+2-j) / 2} \nabla^{j} u\right\|_{L_{p}(\Omega)}+\left\|\lambda^{(k+2) / 2-1 / 2 p} \llbracket u \rrbracket\right\|_{L_{p}\left(\Sigma_{\lambda}^{\prime}\right)}+\left\|\lambda^{(k+2) / 2}\left(u-\langle u\rangle_{\Omega_{\lambda}^{\prime}}\right)\right\|_{L_{p}\left(\Omega_{\lambda}^{\prime}\right)}$.
Here $\Omega_{\lambda}^{\prime}$ bounded subdomain of $\Omega \backslash \Sigma$ with $C^{1}$-boundary and, in the case $\Sigma \neq \emptyset$, we let $\Sigma_{\lambda}^{\prime} \neq \emptyset$ be a bounded subdomain of $\Sigma$ with $C^{1}$-boundary. If $\Sigma=\emptyset$, then we let $\Sigma_{\lambda}^{\prime}=\emptyset$. If the seminorm $\|u\|_{\mathbb{E}_{\lambda}^{k}}$ vanishes, then both $u_{ \pm}$are constant in $\Omega_{ \pm}$and these constants coincide because of $\llbracket u \rrbracket=0$ on $\Sigma_{\lambda}^{\prime}$. Hence the null space of the semi-norm $\|\cdot\|_{\mathbb{E}_{\lambda}^{k}}$ consists of all constant functions and we will see that this is precisely the space of solutions with trivial data. The parameter $\lambda$ will again be useful for controlling lower-order perturbations in perturbed model problems.

In order to define the space of data, we recall from page 24 that the functionals $F_{\left(f, g, h_{1}\right)}$ and $F_{\mu \nabla u}$ are considered as elements of the dual space

$$
\hat{H}_{p}^{-1}(\Omega):=\dot{H}_{p^{\prime}}^{1}(\Omega)^{*}, \quad\|F\|_{\hat{H}_{p}^{-1}(\Omega)}:=\sup _{0 \neq \phi \in \dot{H}_{p^{\prime}}^{1}(\Omega)} \frac{|\langle F \mid \phi\rangle|}{\|\nabla \phi\|_{L_{p^{\prime}}(\Omega)}}, \quad \frac{1}{p}+\frac{1}{p^{\prime}}=1 .
$$

Then the space of data for $k=-1$ is defined by

$$
\mathbb{F}_{\mathrm{cc}, \lambda}^{-1}:=\left(\mathbb{F}_{\mathrm{cc}}^{-1},\|\cdot\|_{\mathbb{F}_{\mathrm{cc}, \lambda}^{-1}}\right), \quad \mathbb{F}_{\mathrm{cc}}^{-1}:=\hat{H}_{p}^{-1}(\Omega) \times \dot{\mathcal{W}}_{p}^{1-1 / p}(\Sigma),
$$

which is a Banach space with respect to the equivalent $\lambda$-dependent norm

$$
\left\|\left(F, h_{2}\right)\right\|_{\mathbb{F}_{\mathrm{c}, \lambda}^{-1}}:=\|F\|_{\hat{H}_{p}^{-1}(\Omega)}+\llbracket h_{2} \rrbracket_{W_{p}^{1-1 / p}(\Sigma)}+\left\|\lambda^{1 / 2-1 / 2 p} h_{2}\right\|_{L_{p}\left(\Sigma_{\lambda}^{\prime}\right)} .
$$

For $k \geq 0$, it is defined by

$$
\mathbb{F}_{\mathrm{cc}, \lambda}^{k}:=\left(\mathbb{F}_{\mathrm{cc}}^{k},\|\cdot\|_{\mathbb{F}_{\mathrm{c}, \lambda}^{k}}\right), \quad \mathbb{F}_{\mathrm{cc}}^{k}:=\left\{\left(f, g, h_{1}, h_{2}\right) \in \mathbb{F}^{k}: F_{\left(f, g, h_{1}\right)} \in \hat{H}_{p}^{-1}(\Omega)\right\},
$$

where

$$
\mathbb{F}^{k}:=H_{p}^{k}(\Omega \backslash \Sigma) \times W_{p}^{k+1-1 / p}(\partial \Omega) \times W_{p}^{k+1-1 / p}(\Sigma) \times \bigcap_{j=1}^{k+2} \dot{\mathcal{W}}_{p}^{j-1 / p}(\Sigma)
$$

which is a Banach space with respect to the equivalent $\lambda$-dependent norm

$$
\begin{aligned}
\left\|\left(f, g, h_{1}, h_{2}\right)\right\|_{\mathbb{F}_{c, \lambda}^{k}}: & =\left\|\left(f, g, h_{1}, h_{2}\right)\right\|_{\mathbb{F}_{\lambda}^{k}}+\left\|\lambda^{(k+1) / 2} F_{\left(f, g, h_{1}\right)}\right\|_{\hat{H}_{p}^{-1}(\Omega)}, \\
\left\|\left(f, g, h_{1}, h_{2}\right)\right\|_{\mathbb{F}_{\lambda}^{k}}= & \|f\|_{H_{p}^{k}(\Omega \backslash \Sigma), \lambda}+\|g\|_{W_{p}^{k+1-1 / p}(\partial \Omega), \lambda} \\
& +\left\|h_{1}\right\|_{W_{p}^{k+1-1 / p}(\Sigma), \lambda}+\left\|h_{2}\right\|_{\bigcap_{j 1}^{k+2}}^{k+2} \mathcal{W}_{p}^{j-1 / p}(\Sigma) \cap L_{p}\left(\Sigma_{\lambda}^{\prime}\right), \lambda
\end{aligned},
$$

where

$$
\begin{aligned}
&\|f\|_{H_{p}^{k}(\Omega \mid \Sigma), \lambda}= \\
&\|g\|_{j=0}^{k}\left\|\lambda^{(k-j) / 2} \nabla^{j} f\right\|_{L_{p}(\Omega)}, \\
&\left\|h_{1}\right\|_{W_{p}^{k+1-1 / p}(\partial \Omega), \lambda}=\llbracket \nabla_{\Sigma}^{k} g \rrbracket_{W_{p}^{1-1 / p}(\partial \Omega)}+\sum_{j=0}^{k}|\lambda|^{1 / 2-1 / 2 p}\left\|\lambda^{(k-j) / 2} \nabla_{\partial \Omega}^{j} g\right\|_{L_{p}(\partial \Omega)}, \\
& \| \llbracket \nabla_{\Sigma}^{k} h_{1} \rrbracket_{W_{p}^{1-1 / p}(\Sigma)}+\sum_{j=0}^{k}|\lambda|^{1 / 2-1 / 2 p}\left\|\lambda^{(k-j) / 2} \nabla_{\Sigma}^{j} h_{1}\right\|_{L_{p}(\Sigma)}, \\
&\left\|h_{2}\right\|_{\bigcap_{j=1}^{k+2} \mathcal{H}_{p}^{j-1 / p}(\Sigma) \cap L_{p}\left(\Sigma_{\lambda}^{\prime}\right), \lambda}:=\llbracket \nabla_{\Sigma}^{k+1} h_{2} \rrbracket_{W_{p}^{1-1 / p}(\Sigma)}+\sum_{j=1}^{k+1}|\lambda|^{1 / 2-1 / 2 p}\left\|\lambda^{(k+1-j) / 2} \nabla_{\Sigma}^{j} h_{2}\right\|_{L_{p}(\Sigma)} \\
&+\left\|\lambda^{(k+2) / 2-1 / 2 p} h_{2}\right\|_{L_{p}\left(\Sigma_{\lambda}^{\prime}\right)} .
\end{aligned}
$$

In the basic situations $\Omega \backslash \Sigma \in\left\{\mathbb{R}^{n}, \mathbb{R}_{+}^{n}, \dot{\mathbb{R}}^{n}\right\}, \Omega_{\lambda}^{\prime}=\lambda^{-1 / 2} \Omega_{1}^{\prime}$, and $\Sigma_{\lambda}^{\prime}=\lambda^{-1 / 2} \Sigma_{1}^{\prime}$, we obtain

$$
\|u\|_{\mathbb{E}_{\lambda}^{k}}=\left\|u_{\lambda}\right\|_{\mathbb{E}_{1}^{k}}, \quad\left\|\left(f, g, h_{1}, h_{2}\right)\right\|_{\mathbb{F}_{\lambda}^{k}}=\left\|\left(f_{\lambda}, g_{\lambda}, h_{1 \lambda}, h_{2 \lambda}\right)\right\|_{\mathbb{F}_{1}^{k}},
$$

where the rescaled functions $u_{\lambda}, f_{\lambda}, g_{\lambda}, h_{1 \lambda}$, and $h_{2 \lambda}$ were defined in Section 2.1.2.
2.2.2. Basic model problems. In order to solve problem (2.1), we have to determine the null space and range of the operator $L:=\operatorname{div}(\mu \nabla \cdot)$, considered as an unbounded operator in $L_{p}(\Omega)$. It is clear that all constant functions belong to $N(L)$ and the converse inclusion follows for $p \geq 2 n /(n+2)$ from an integration by parts. For the remaining case $p \in(1,2 n /(n+2))$, which is more involved, Simader and Sohr [SS92] obtained the following weak a priori estimate.
2.19. Theorem ([cf. SS92]). Let $n \geq 2, p \in(1, \infty)$, and let $\Omega \subset \mathbb{R}^{n}$ be either
(i) the whole space $\mathbb{R}^{n}$,
(ii) the half space $\mathbb{R}_{+}^{n}=\mathbb{R}^{n-1} \times(0, \infty)$,
(iii) a bent half space $\mathbb{R}_{\omega}^{n}=\left\{\left(x^{\prime}, x_{n}\right) \in \mathbb{R}^{n}: x_{n}>\omega\left(x^{\prime}\right)\right\}$ whose defining function $\omega \in C_{c}^{1}\left(\mathbb{R}^{n-1}\right)$ satisfies $\left\|\nabla^{\prime} \omega\right\|_{\infty} \leq \eta$ for some $\eta=\eta(n, p)>0$,
(iv) a bounded domain in $\mathbb{R}^{n}$ with $C^{1}$-boundary,
(v) or an exterior domain in $\mathbb{R}^{n}$ with $C^{1}$-boundary; that is, $\mathbb{R}^{n} \backslash \Omega$ is a bounded domain.

Then there exists a constant $C=C(n, p, \eta, \Omega)>0$ such that

$$
\begin{equation*}
\|\nabla u\|_{L_{p}(\Omega)} \leq C \sup \left\{\frac{\left|\int_{\Omega} \nabla u \cdot \nabla \phi d x\right|}{\|\nabla \phi\|_{L_{p^{\prime}}(\Omega)}}: \phi \in \dot{H}_{p^{\prime}}^{1}(\Omega) \backslash\{0\}\right\} \quad \text { for } u \in \dot{\mathcal{H}}_{p}^{1}(\Omega) \text {. } \tag{2.21}
\end{equation*}
$$

We start with the analog of Lemma 2.10 for the case $\Omega=\mathbb{R}^{n}, \Sigma=\mathbb{R}^{n-1} \times\{0\}$ and $\lambda=0$; that is, we consider the strong and the weak transmission problem

$$
\left\{\begin{align*}
-\mu \Delta u & =f & & \text { in } \dot{\mathbb{R}}^{n},  \tag{2.22}\\
\llbracket \mu \partial_{n} u \rrbracket & =h_{1} & & \text { on } \mathbb{R}^{n-1} \\
\llbracket u \rrbracket & =h_{2} & & \text { on } \mathbb{R}^{n-1} .
\end{aligned}\right\}, \quad\left\{\begin{aligned}
\int_{\mathbb{R}^{n}} \mu \nabla u \cdot \nabla \phi d x=\langle F \mid \phi\rangle & & \text { for all } \phi \in \dot{H}_{p^{\prime}}^{1}\left(\mathbb{R}^{n}\right), \\
\llbracket u \rrbracket=h_{2} & & \text { on } \mathbb{R}^{n-1},
\end{align*}\right\},
$$

with constant coefficients $\mu_{ \pm}>0$. Here the functionals $F_{\mu \nabla u}$ and $F_{\left(f, h_{1}\right)}$ on $\dot{H}_{p^{\prime}}^{1}\left(\mathbb{R}^{n}\right)$ are given

$$
\left\langle F_{\mu \nabla u} \mid \phi\right\rangle:=\int_{\mathbb{R}^{n}} \mu \nabla u \cdot \nabla \phi d x, \quad\left\langle F_{\left(f, h_{1}\right)} \mid \phi\right\rangle:=\int_{\mathbb{R}^{n}} f \phi d x-\int_{\mathbb{R}^{n-1}} h_{1} \phi d x^{\prime}, \quad \text { for } \phi \in \dot{H}_{p^{\prime}}^{1}\left(\mathbb{R}^{n}\right) .
$$

Our goal is to prove that the induced operator

$$
A: \mathbb{E}_{\lambda}^{k} \rightarrow \mathbb{F}_{\lambda}^{k}, \quad A u= \begin{cases}\left(-\operatorname{div}(\mu \nabla u), \llbracket \mu \partial_{n} u \rrbracket, \llbracket u \rrbracket\right) & \text { if } k \geq 0,  \tag{2.23}\\ \left(F_{\mu \nabla u}, \llbracket u \rrbracket\right) & \text { if } k=-1,\end{cases}
$$

is an isomorphism. In order to deal with $k=-1$, we modify the strategy of [SS92, Lemma 3.3]; thus, we first derive a variant of the Calderòn-Zygmund estimate $\left\|\nabla^{2} \phi\right\|_{p} \lesssim\|\Delta \phi\|_{p}$ for $\phi \in \mathcal{D}\left(\mathbb{R}^{n}\right)$.
2.20. Lemma. Let $n \geq 2, \Sigma:=\mathbb{R}^{n-1} \times\{0\}, \mu_{ \pm}>0$, and $p \in(1, \infty)$, and define the vector spaces

$$
\begin{aligned}
& Y:=Y_{\mu}:=\left\{\left(x^{\prime}, x_{n}\right) \mapsto a^{\prime} \cdot x^{\prime}+b \mu\left(x_{n}\right)^{-1} x_{n}+c: a^{\prime} \in \mathbb{K}^{n-1}, b, c \in \mathbb{K}\right\}, \\
& X:=X_{p, \mu}:=\left\{u \in \dot{\mathcal{H}}_{p}^{2}\left(\dot{\mathbb{R}}^{n}\right): \llbracket \mu \partial_{n} u \rrbracket=\llbracket u \rrbracket=0 \text { on } \Sigma\right\}, \quad\|u\|_{X}=\left\|\nabla^{2} u\right\|_{L_{p}\left(\mathbb{R}^{n}\right)} .
\end{aligned}
$$

Then $X / Y$ is a Banach space and the map

$$
-\mu \Delta: X / Y \rightarrow L_{p}\left(\mathbb{R}^{n}\right)
$$

is a topological linear isomorphism. In particular, there exists $C\left(n, p, \mu_{ \pm}\right)>0$ such that

$$
\begin{equation*}
C^{-1}\left\|\nabla^{2} u\right\|_{L_{p}\left(\mathbb{R}^{n}\right)} \leq\|\mu \Delta u\|_{L_{p}\left(\mathbb{R}^{n}\right)} \leq C\left\|\nabla^{2} u\right\|_{L_{p}\left(\mathbb{R}^{n}\right)} \quad \text { for all } u \in X \tag{2.24}
\end{equation*}
$$

Furthermore, the map

$$
A: u \mapsto\left(-\mu \Delta u, \llbracket \mu \partial_{n} u \rrbracket, \llbracket u \rrbracket\right), \quad \dot{\mathcal{H}}_{p}^{2}\left(\dot{\mathbb{R}}^{n}\right) / Y \rightarrow L_{p}\left(\mathbb{R}^{n}\right) \times \dot{W}_{p}^{1-1 / p}\left(\mathbb{R}^{n-1}\right) \times \dot{W}_{p}^{2-1 / p}\left(\mathbb{R}^{n-1}\right)
$$

is uniformly invertible with respect to $\mu_{ \pm} \in\left[\mu_{0}, \mu_{0}^{-1}\right]$, for every $\mu_{0} \in(0,1]$.
Proof. (i) For the injectivity of $-\mu \Delta$ modulo $Y$ we adapt an argument of Wilke [Wil13, p. 104-105]. Suppose that $u \in X$ satisfies $-\mu \Delta u=0$ in the sense of $\mathcal{D}^{\prime}\left(\dot{\mathbb{R}}^{n}\right)$. Then we even have $\Delta u=0$ in $\mathcal{D}^{\prime}\left(\mathbb{R}^{n}\right)$, but not necessarily in $\mathcal{D}^{\prime}\left(\mathbb{R}^{n}\right)$. We put $v_{+}:=u_{+}-R u_{-}$on $\mathbb{R}_{+}^{n}$ and $v_{-}:=-R v_{+}$ on $\mathbb{R}_{-}^{n}$ where $(R \phi)\left(x^{\prime}, x_{n}\right):=\phi\left(x^{\prime},-x_{n}\right)$ denotes even reflection. From $\llbracket u \rrbracket=0$ we infer that $v=0$ on $\Sigma$ and hence also $\llbracket \nabla v \rrbracket=e_{n} \llbracket \partial_{n} v \rrbracket$ on $\Sigma$. But since $\partial_{n} v_{+}=-\partial_{n}\left(R v_{-}\right)=\partial_{n} v_{-}$, we have $\llbracket \partial_{n} v \rrbracket=0$, which yields $v \in \dot{\mathcal{H}}_{p}^{2}\left(\mathbb{R}^{n}\right)$ and integrating by parts yields $-\Delta v=0$ in $\mathcal{D}^{\prime}\left(\mathbb{R}^{n}\right)$. Here the negative Laplacian represents the Riesz potential $\dot{J}_{2}=-\Delta: \dot{H}_{p}^{2}\left(\mathbb{R}^{n}\right) \rightarrow L_{p}\left(\mathbb{R}^{n}\right)$, which is a topological isomorphism (Theorem B.15). Hence $v$ must be a linear map.

In an analogous way we can check that $w_{+}:=\mu_{+} u_{+}+\mu_{-} R u_{-}$on $\mathbb{R}_{+}^{n}$ and $w_{-}:=R w_{+}$on $\mathbb{R}_{-}^{n}$ yield a function $w \in \dot{\mathcal{H}}_{p}^{2}\left(\mathbb{R}^{n}\right)$ with $\llbracket w \rrbracket=0, \partial_{n} w=0$ on $\mathbb{R}^{n-1}$, and $-\Delta w=0$. Hence also $w$ is a linear map. By using that $u_{ \pm}$are linear combinations of $v_{ \pm}$and $w_{ \pm}$, we easily see that $u \in Y$.
(ii) For the surjectivity of $-\mu \Delta$, we construct $u=v+w \in X$ where $v \in \dot{\mathcal{H}}_{p}^{2}\left(\mathbb{R}^{n}\right)$ is a representative of $(-\Delta)^{-1}\left(\mu^{-1} f\right) \in \dot{H}_{p}^{2}\left(\mathbb{R}^{n}\right)$ and $w \in \dot{\mathcal{H}}_{p}^{2}\left(\dot{\mathbb{R}}^{n}\right)$ satisfies

$$
-\Delta w=0 \text { in } \mathcal{D}^{\prime}\left(\dot{\mathbb{R}}^{n}\right), \quad \llbracket w \rrbracket=0 \text { on } \Sigma, \quad \llbracket \mu \partial_{n} w \rrbracket=-\llbracket \mu \partial_{n} v \rrbracket \text { on } \Sigma .
$$

By applying the partial Fourier transform and solving the resulting system, we obtain

$$
\tilde{w}\left(\xi^{\prime}, x_{n}\right)=\left(\left(\mu_{+}+\mu_{-}\right)\left|\xi^{\prime}\right|\right)^{-1} e^{-\left|\xi^{\prime} x_{n}\right|} \llbracket \mu \partial_{n} \tilde{v} \rrbracket\left(\xi^{\prime}\right) \quad \text { for } \xi^{\prime} \in \mathbb{R}^{n-1}, x_{n} \in \dot{\mathbb{R}} .
$$

Therefore $w$ has the following representation, which can be seen by using Jawerth's trace theorem $\left.\dot{H}_{p}^{2}\left(\mathbb{R}^{n}\right)\right|_{x_{n}=0}=\dot{W}_{p}^{2-1 / p}\left(\mathbb{R}^{n-1}\right)$ from Theorem B.31, the Riesz potential $\dot{J}_{-1}^{\prime}=\left(-\Delta^{\prime}\right)^{-1 / 2}$, and the Poisson semigroup $P\left(x_{n}\right)=e^{-x_{n}\left(-\Delta^{\prime}\right)^{1 / 2}}$ :

$$
w\left(\cdot, x_{n}\right)=\left(\mu_{+}+\mu_{-}\right)^{-1} e^{-\left|x_{n}\right|\left(-\Delta^{\prime}\right)^{1 / 2}}\left(\left(-\Delta^{\prime}\right)^{-1 / 2} \llbracket \mu \partial_{n} v \rrbracket\right) .
$$

Hence, $w$ belongs to $\dot{\mathcal{H}}_{p}^{2}\left(\dot{\mathbb{R}}^{n}\right)$ and satisfies the asserted a priori estimate. Therefore the operator $-\mu \Delta: X \rightarrow Y$ is surjective and has a bounded right-inverse.
(iii) Finally, we consider the map $A: u \mapsto\left(-\mu \Delta u, \llbracket \mu \partial_{n} u \rrbracket, \llbracket u \rrbracket\right)$, which is injective by step (i). For proving surjectivity, we let $\left(f, h_{1}, h_{2}\right) \in L_{p}\left(\mathbb{R}^{n}\right) \times \dot{W}_{p}^{1-1 / p}\left(\mathbb{R}^{n-1}\right) \times \dot{W}_{p}^{2-1 / p}\left(\mathbb{R}^{n-1}\right)$ be given. We construct $u=v+w$ with $v=(-\Delta)^{-1}\left(\mu^{-1} f\right) \in \dot{\mathcal{H}}_{p}^{2}\left(\mathbb{R}^{n}\right)$ and $-\mu \Delta w=0, \llbracket \mu \partial_{n} w \rrbracket=$ $h_{1}-\llbracket \mu \partial_{n} v \rrbracket, \llbracket w \rrbracket=h_{2}$. The function $w$ can be constructed as in step (ii) and is given by

$$
w\left(\cdot, \pm x_{n}\right)=\left(\mu_{+}+\mu_{-}\right)^{-1} e^{-\left|x_{n}\right|\left(-\Delta^{\prime}\right)^{1 / 2}}\left(-\left(-\Delta^{\prime}\right)^{-1 / 2}\left(h_{1}-\llbracket \mu \partial_{n} v \rrbracket\right) \pm \mu_{\mp} h_{2}\right) .
$$

Therefore $A$ is uniformly invertible with respect to $\mu_{ \pm} \in\left[\mu_{0}, \mu_{0}^{-1}\right]$.
2.21. Remark. The space $X_{p, \mu}=\left\{u \in \dot{\mathcal{H}}_{p}^{2}\left(\dot{\mathbb{R}}^{n}\right): \llbracket \mu \partial_{n} u \rrbracket=\llbracket u \rrbracket=0\right.$ on $\left.\Sigma\right\}$ can be identified with the standard space $\dot{\mathcal{H}}_{p}^{2}\left(\mathbb{R}^{n}\right)$ by means of the bijection

$$
T_{\mu}:=\dot{\mathcal{H}}_{p}^{2}\left(\mathbb{R}^{n}\right) \rightarrow X_{p, \mu}, \quad\left(T_{\mu} u\right)\left(x^{\prime}, x_{n}\right)=u\left(x^{\prime}, \mu\left(x_{n}\right)^{-1} x_{n}\right) .
$$

The semi-norms $\left\|\nabla^{k} T_{\mu} \cdot\right\|_{L_{p}\left(\mathbb{R}^{n}\right)}$ and $\left\|\nabla^{k} \cdot\right\|_{L_{p}\left(\mathbb{R}^{n}\right)}$ are equivalent on $\dot{\mathcal{H}}_{p}^{k}\left(\mathbb{R}^{n}\right)$ for $k \in \mathbb{N}_{0}$.
In order to deal with the case $k=-1$, we provide some density results.
2.22. Lemma. Let $n \geq 2, \Sigma:=\mathbb{R}^{n-1} \times\{0\}$, $\mu_{ \pm}>0$, and $p \in(1, \infty)$.
(i) For $u \in \dot{\mathcal{H}}_{p}^{1}\left(\mathbb{R}^{n}\right)$ and $\varepsilon>0$ there exists $u_{\varepsilon} \in X_{p, \mu} \cap \dot{\mathcal{H}}_{p}^{1}\left(\mathbb{R}^{n}\right)$ such that $\left\|\nabla\left(u_{\varepsilon}-u\right)\right\|_{L_{p}\left(\mathbb{R}^{n}\right)} \leq \varepsilon$.
(ii) For $u \in X_{p, \mu}$ and $\varepsilon>0$ there exists $u_{\varepsilon} \in X_{p, \mu} \cap \dot{\mathcal{H}}_{p}^{1}\left(\mathbb{R}^{n}\right)$ such that $\left\|\nabla^{2}\left(u_{\varepsilon}-u\right)\right\|_{L_{p}\left(\mathbb{R}^{n}\right)} \leq \varepsilon$.

Proof. (i) We shall construct $u_{\varepsilon}$ by an anisotropic mollification. Let $\varphi_{r}$ denote the Friedrichs mollifier with support $B_{r}(0) \subset \mathbb{R}^{n}$; that is, $\varphi_{r}(x)=r^{-n} \varphi(x / r)$ with some $\varphi \in \mathcal{D}\left(\mathbb{R}^{n}\right)$ such that $\varphi \geq 0, \int_{\mathbb{R}^{n}} \varphi d x=1$, and $\operatorname{supp} \varphi=B_{1}(0)$. Then we consider the function

$$
u_{r}:=T_{\mu}\left(\varphi_{r} *\left(T_{\mu}^{-1} u\right)\right) \quad \text { for } r>0 .
$$

Then $\varphi_{r} *\left(T_{\mu}^{-1} u\right)$ belongs to $C^{\infty}\left(\mathbb{R}^{n}\right)$ and hence $\llbracket u_{\varepsilon} \rrbracket=\llbracket \mu \partial_{n} u_{\varepsilon} \rrbracket=0$. Moreover,

$$
\nabla\left(T_{\mu}^{-1} u_{r}\right)=\varphi_{r} * \nabla\left(T_{\mu}^{-1} u\right) \rightarrow \nabla\left(T_{\mu}^{-1} u\right) \text { in } L_{p}\left(\mathbb{R}^{n}\right) \text { as } r \rightarrow 0
$$

and hence also $\nabla u_{r} \rightarrow \nabla u$ in $L_{p}\left(\mathbb{R}^{n}\right)$. Finally, from $T_{\mu}^{-1} u \in \dot{\mathcal{H}}_{p}^{1}\left(\mathbb{R}^{n}\right)$ and $\varphi_{r} \in \mathcal{D}\left(\mathbb{R}^{n}\right)$ we infer that $\varphi_{r} *\left(T_{\mu}^{-1} u\right)$ belongs to $\dot{\mathcal{H}}_{p}^{2}\left(\mathbb{R}^{n}\right)$. Hence, for some sufficiently small $r=r(\varepsilon)>0$, there exists some $u_{\varepsilon}:=u_{r(\varepsilon)} \in X_{p, \mu} \cap \dot{\mathcal{H}}_{p}^{1}\left(\mathbb{R}^{n}\right)$ with the desired properties.
(ii) By Remark 2.21, the function $T_{\mu}^{-1} u$ belongs to the usual homogeneous space $\dot{\mathcal{H}}_{p}^{2}\left(\mathbb{R}^{n}\right)$ and thanks to Remark B.12, there is a linear function $v_{0}: \mathbb{R}^{n} \rightarrow \mathbb{K}$ such that $\varphi_{r} *\left(\chi_{R} \cdot\left(T_{\mu}^{-1} u-v_{0}\right)\right)$ converges to $T_{\mu}^{-1} u-v_{0}$ in $\dot{\mathcal{H}}_{p}^{2}\left(\mathbb{R}^{n}\right)$ as $r \rightarrow 0$ and $R \rightarrow \infty$. Here $\chi_{R} \in \mathcal{D}\left(\mathbb{R}^{n}\right)$ denotes the radial Sobolev cut-off function with support $B_{R}(0)$. Since $\varphi_{r} *\left(\chi_{R} \cdot\left(T_{\mu}^{-1} u-v_{0}\right)\right)$ belongs to $\mathcal{D}\left(\mathbb{R}^{n}\right)$, the function $u_{\varepsilon}=T_{\mu}\left(\varphi_{r} *\left(\chi_{R} \cdot\left(T_{\mu}^{-1} u-v_{0}\right)\right)\right)$ belongs to $X_{p, \mu} \cap \dot{\mathcal{H}}_{p}^{1}\left(\mathbb{R}^{n}\right)$ and satisfies the assertion for some small $r=r(\varepsilon)>0$ and some large $R=R(\varepsilon)>0$.

We are ready to prove optimal $\mathbb{E}^{-1}$-regularity in the case $\Sigma \cong \mathbb{R}^{n-1}$ and $\llbracket u \rrbracket=0$.
2.23. Lemma. Let $n \geq 2, \Sigma:=\mathbb{R}^{n-1} \times\{0\}, \dot{\mathbb{R}}^{n}:=\mathbb{R}^{n} \backslash \Sigma, \mu_{0} \in(0,1]$, and $p \in(1, \infty)$. Then the map

$$
u \mapsto F_{\mu \nabla u}, \quad \dot{H}_{p}^{1}\left(\mathbb{R}^{n}\right) \rightarrow \hat{H}_{p}^{-1}\left(\mathbb{R}^{n}\right)
$$

is uniformly invertible with respect to $\mu_{ \pm} \in\left[\mu_{0}, \mu_{0}^{-1}\right]$.
Proof. (i) Similar as in [SS92, Lemma 3.3], we prove that $u \mapsto F_{\mu \nabla u}$ is injective and bounded from below. For $u \in\left(\dot{\mathcal{H}}_{p}^{2} \cap \dot{\mathcal{H}}_{p}^{1}\right)\left(\dot{\mathbb{R}}^{n}\right):=\dot{\mathcal{H}}_{p}^{2}\left(\dot{\mathbb{R}}^{n}\right) \cap \dot{\mathcal{H}}_{p}^{1}\left(\dot{\mathbb{R}}^{n}\right)$ and $\phi \in\left(\dot{\mathcal{H}}_{p^{\prime}}^{2} \cap \dot{\mathcal{H}}_{p^{\prime}}^{1}\right)\left(\dot{\mathbb{R}}^{n}\right)$ we obtain

$$
\int_{\mathbb{R}^{n}} \partial_{j} u \Delta \phi d x=\int_{\mathbb{R}^{n}} \mu \nabla u \cdot \nabla\left(\partial_{j}\left(\mu^{-1} \phi\right)\right) d x+\int_{\Sigma} \llbracket \delta_{j n} \nabla u \cdot \nabla \phi-\partial_{j} u \partial_{n} \phi \rrbracket d x^{\prime} .
$$

(i.a) Let $j=n$ and assume that $\llbracket u \rrbracket=0$ and $\llbracket \phi \rrbracket=0$. Then $\llbracket \nabla^{\prime} u \rrbracket=\nabla^{\prime} \llbracket u \rrbracket=0$ and $\llbracket \nabla^{\prime} \phi \rrbracket=0$. Thus, the integrand in the interface integral vanishes; that is,

$$
\llbracket \delta_{j n} \nabla u \cdot \nabla \phi-\partial_{j} u \partial_{n} \phi \rrbracket=\llbracket \nabla^{\prime} u \cdot \nabla^{\prime} \phi \rrbracket=0 .
$$

Let $Z_{n}:=\left\{\phi \in\left(\dot{\mathcal{H}}_{p^{\prime}}^{2} \cap \dot{\mathcal{H}}_{p^{\prime}}^{1}\right)\left(\dot{\mathbb{R}}^{n}\right): \llbracket \mu^{-1} \partial_{n} \phi \rrbracket=\llbracket \phi \rrbracket=0\right\}=X_{p^{\prime}, \mu^{-1}} \cap \dot{\mathcal{H}}_{p^{\prime}}^{1}\left(\mathbb{R}^{n}\right)$. Then Lemma 2.22.(ii) implies that $Z_{n}$ is dense in $X_{p^{\prime}, \mu^{-1}}$. By Lemma 2.20, the map $\mu \Delta: X_{p^{\prime}, \mu^{-1}} \rightarrow L_{p^{\prime}}\left(\mathbb{R}^{n}\right)=$ $\mu L_{p^{\prime}}\left(\mathbb{R}^{n}\right)$ is bounded and surjective and the estimate $\left\|\nabla^{2} \phi\right\|_{L_{p^{\prime}}\left(\mathbb{R}^{n}\right)} \leq C(n, p, \mu)\|\Delta \phi\|_{L_{p^{\prime}}\left(\mathbb{R}^{n}\right)}$ applies to all $\phi \in X_{p^{\prime}, \mu^{-1}}$. Therefore the space $\Delta Z_{n}$ is dense in $L_{p^{\prime}}\left(\mathbb{R}^{n}\right)$. Furthermore, $\mu^{-1} \partial_{n} Z_{n}$ is a subspace of $\dot{\mathcal{H}}_{p^{\prime}}^{1}\left(\mathbb{R}^{n}\right)$. Hence for every $u \in \dot{\mathcal{H}}_{p}^{2}\left(\dot{\mathbb{R}}^{n}\right) \cap \dot{\mathcal{H}}_{p}^{1}\left(\mathbb{R}^{n}\right)$, we obtain

$$
\begin{aligned}
\left\|\partial_{n} u\right\|_{L_{p}\left(\mathbb{R}^{n}\right)} & =\sup _{\phi \in Z_{n}, \Delta \phi \neq 0} \frac{\left|\int_{\mathbb{R}^{n}} \partial_{n} u \Delta \phi d x\right|}{\|\Delta \phi\|_{L_{p^{\prime}}\left(\mathbb{R}^{n}\right)}}=\sup _{\phi \in Z_{n}, \Delta \phi \neq 0} \frac{\left|\int_{\mathbb{R}^{n}} \mu \nabla u \cdot \nabla\left(\mu^{-1} \partial_{n} \phi\right) d x\right|}{\|\Delta \phi\|_{L_{p^{\prime}}\left(\mathbb{R}^{n}\right)}} \\
& \leq C(n, p, \mu) \sup _{\phi \in Z_{n}, \Delta \phi \neq 0} \frac{\left|\int_{\mathbb{R}^{n}} \mu \nabla u \cdot \nabla\left(\mu^{-1} \partial_{n} \phi\right) d x\right|}{\left\|\nabla^{2} \phi\right\|_{L_{p^{\prime}}\left(\mathbb{R}^{n}\right)}} \leq C^{\prime}(n, p, \mu)\left\|F_{\mu \nabla u}\right\|_{\hat{H}_{p}^{-1}\left(\mathbb{R}^{n}\right)} .
\end{aligned}
$$

By Lemma 2.22.(i), the inequality also applies to all $u \in \dot{\mathcal{H}}_{p}^{1}\left(\mathbb{R}^{n}\right)$.
(i.b) Let $j<n, \llbracket u \rrbracket=0$, and $\llbracket \mu^{-1} \phi \rrbracket=\llbracket \partial_{n} \phi \rrbracket=0$. Then the interface integral vanishes, since

$$
-\llbracket \delta_{j n} \nabla u \cdot \nabla \phi-\partial_{j} u \partial_{n} \phi \rrbracket=\llbracket \partial_{j} u \partial_{n} \phi \rrbracket=0 .
$$

We now let $Z_{j}:=\left\{\phi \in \dot{\mathcal{H}}_{p^{\prime}}^{2}\left(\mathbb{R}^{n}\right) \cap \dot{\mathcal{H}}_{p^{\prime}}^{1}\left(\dot{\mathbb{R}}^{n}\right): \llbracket \partial_{n} \phi \rrbracket=\llbracket \mu^{-1} \phi \rrbracket=0\right\}$. Then it is easy to check that $\mu^{-1} Z_{j}=X_{p^{\prime}, \mu} \cap \dot{\mathcal{H}}_{p^{\prime}}^{1}\left(\mathbb{R}^{n}\right)$ and $\mu^{-1} \partial_{j} Z_{j} \subset \dot{\mathcal{H}}_{p^{\prime}}^{1}\left(\mathbb{R}^{n}\right)$ and that $\Delta Z_{j}$ is dense in $L_{p^{\prime}}\left(\mathbb{R}^{n}\right)$. Therefore the desired inequality follows in the same way as before.
(ii) It remains to show that $u \mapsto F_{\mu \nabla u}$ is surjective. Let $F \in \hat{H}_{p}^{-1}\left(\mathbb{R}^{n}\right)=\dot{H}_{p^{\prime}}^{1}\left(\mathbb{R}^{n}\right)^{*}$. Since we may identify $\dot{H}_{p^{\prime}}^{1}\left(\mathbb{R}^{n}\right)$ with the closed subspace $\nabla \dot{H}_{p^{\prime}}^{1}\left(\mathbb{R}^{n}\right)$ of $L_{p^{\prime}}\left(\mathbb{R}^{n}\right)^{n}$, there exists some $f \in L_{p}\left(\mathbb{R}^{n}\right)^{n}$ with $\|f\|_{L_{p}\left(\mathbb{R}^{n}\right)^{n}}=\|F\|_{\hat{H}_{p}^{-1}\left(\mathbb{R}^{n}\right)}$ such that $\langle F \mid \phi\rangle=\int_{\mathbb{R}^{n}} f \cdot \nabla \phi d x$ for all $\phi \in \dot{H}_{p^{\prime}}^{1}\left(\mathbb{R}^{n}\right)$ [cf. AF03, Theorem 3.9]. Let $\left(f_{k}\right) \subset H_{p}^{1}\left(\mathbb{R}^{n}\right)^{n}$ be a sequence such that $f_{k} \rightarrow f$ in $L_{p}\left(\mathbb{R}^{n}\right)^{n}$ as $k \rightarrow \infty$ and define

$$
\left\langle F_{k} \mid \phi\right\rangle:=\int_{\mathbb{R}^{n}} f_{k} \cdot \nabla \phi d x=-\int_{\mathbb{R}^{n}} \operatorname{div} f_{k} \phi d x-\int_{\Sigma} \llbracket e_{n} \cdot f_{k} \rrbracket \phi d x^{\prime} .
$$

Hence $F_{k} \rightarrow F$ in $\hat{H}_{p}^{-1}\left(\mathbb{R}^{n}\right)$. With Lemma 2.20, we let $u_{k} \in \dot{\mathcal{H}}_{p}^{2}\left(\dot{\mathbb{R}}^{n}\right) \cap \dot{\mathcal{H}}_{p}^{1}\left(\mathbb{R}^{n}\right)$ solve the system

$$
-\mu \Delta u_{k}=-\operatorname{div} f_{k}, \quad \llbracket \mu \partial_{n} u_{k} \rrbracket=\llbracket e_{n} \cdot f_{k} \rrbracket, \quad \llbracket u_{k} \rrbracket=0 .
$$

Then we have $F_{\mu \nabla u_{k}}=F_{k}$ and $\left\|\nabla u_{k}-\nabla u_{k^{\prime}}\right\|_{L_{p}\left(\mathbb{R}^{n}\right)} \leq C\left\|F_{k}-F_{k^{\prime}}\right\|_{\hat{H}_{p}^{-1}\left(\mathbb{R}^{n}\right)}$. Therefore $\nabla u_{k} \rightarrow \nabla u$ in $L_{p}\left(\mathbb{R}^{n}\right)$ for some $u \in \dot{\mathcal{H}}_{p}^{1}\left(\mathbb{R}^{n}\right)$ and this limit satisfies $F_{\mu \nabla u}=F$.

We now prove optimal $\mathbb{E}^{k}$-regularity for the transmission problems (2.22) in the flat interface case $\Omega=\mathbb{R}^{n}$ and $\Sigma=\mathbb{R}^{n-1} \times\{0\}$. In the definition of the norms, we let $\Omega_{\lambda}^{\prime}=\lambda^{-1 / 2} \Omega^{\prime}$ and $\Sigma_{\lambda}^{\prime}=\lambda^{-1 / 2} \Sigma^{\prime}$, where $\Omega^{\prime} \neq \emptyset$ and $\Sigma^{\prime} \neq \emptyset$ are bounded open subsets of $\mathbb{R}^{n}$ and $\Sigma$ with $C^{1}$-boundaries.
2.24. Lemma. Let $n \geq 2, \mu_{0} \in(0,1], k \in \mathbb{N}_{0} \cup\{-1\}$, and $p \in(1, \infty)$. Then the map $A: \mathbb{E}_{\lambda}^{k}\left(\dot{\mathbb{R}}^{n}\right) \rightarrow$ $\mathbb{F}_{c c, \lambda}^{k}\left(\dot{\mathbb{R}}^{n}\right)$ in (2.23) is uniformly invertible with respect to $\mu_{ \pm} \in\left[\mu_{0}, \mu_{0}^{-1}\right]$ and $\lambda \in(0, \infty)$.

Proof. (i) Uniqueness. Let $k=-1$ and let $u \in \dot{\mathcal{H}}_{p}^{1}\left(\dot{\mathbb{R}}^{n}\right)$ satisfy $F_{\mu \nabla u}=0$ and $\llbracket u \rrbracket=0$. Then $u$ belongs to $\dot{\mathcal{H}}_{p}^{1}\left(\mathbb{R}^{n}\right)$ and Lemma 2.23 implies that $u$ is constant. For $k \geq 0$ we consider a function $u \in \mathbb{E}_{\lambda}^{k} \subset\left(\dot{\mathcal{H}}_{p}^{2} \cap \dot{\mathcal{H}}_{p}^{1}\right)\left(\dot{\mathbb{R}}^{n}\right)$ such that $\mu \nabla u=0$ and $\llbracket \mu \partial_{n} u \rrbracket=\llbracket u \rrbracket=0$. By Lemma 2.20, $u$ is constant.
(ii) Existence for $k=-1$. Given $F \in \hat{H}_{p}^{-1}\left(\mathbb{R}^{n}\right)$ and $h_{2} \in \dot{\mathcal{W}}_{p}^{1-1 / p}\left(\mathbb{R}^{n-1}\right)$, we construct $u=$ $v+w \in \dot{\mathcal{H}}_{p}^{1}\left(\dot{\mathbb{R}}^{n}\right) / \mathbb{K}$ as follows. Let $\mathcal{E}_{+} \in \mathcal{B}\left(\dot{W}_{p}^{1-1 / p}\left(\mathbb{R}^{n-1}\right) ; \dot{H}_{p}^{1}\left(\mathbb{R}_{+}^{n}\right)\right)$ be the extension operator from Theorem B.31. Then the equivalence class $\mathcal{E}_{+}\left(h_{2}+\mathbb{K}\right) \in \dot{H}_{p}^{1}\left(\mathbb{R}_{+}^{n}\right)$ has a representative $v_{+} \in$ $\dot{\mathcal{H}}_{p}^{1}\left(\mathbb{R}_{+}^{n}\right)$ with $\left.v_{+}\right|_{x_{n}=0}=h_{2}$. By choosing $v_{-}:=0$, the function $v$ belongs to $\dot{\mathcal{H}}_{p}^{1}\left(\mathbb{R}^{n}\right)$ and satisfies $\llbracket v \rrbracket=h_{2}$. Next, we determine $w \in \dot{\mathcal{H}}_{p}^{1}\left(\mathbb{R}^{n}\right)$ as a solution to $\int_{\mathbb{R}^{n}} \mu \nabla w \cdot \nabla \phi d x=\left\langle F-F_{\mu \nabla v} \mid \phi\right\rangle$ for $\phi \in \dot{H}_{p^{\prime}}^{1}\left(\mathbb{R}^{n}\right)$ by means of Lemma 2.23. Then $u=v+w$ belongs to $\mathbb{E}_{\lambda}^{-1}$ and solves $F_{\mu \nabla u}=F$ and $\llbracket u \rrbracket=h_{2}$.
(iii) Existence for $k \geq 0$. We construct a solution $u=u^{1}+u^{2}+u^{3}$ with

$$
\begin{aligned}
u_{ \pm}^{1} & :=\left.\mu_{ \pm}^{-1}\left((-\Delta)^{-1}\left(E_{ \pm} f_{ \pm}\right)\right)\right|_{\mathbb{R}_{ \pm}^{n}}, \\
u_{ \pm}^{2}\left(\cdot, \pm x_{n}\right) & :=-\left(\mu_{+}+\mu_{-}\right)^{-1} e^{-x_{n}\left(-\Delta^{\prime}\right)^{1 / 2}}\left(-\Delta^{\prime}\right)^{-1 / 2}\left(h_{1}-\llbracket \mu \partial_{n} u^{1} \rrbracket\right), \\
u_{ \pm}^{3}\left(\cdot, \pm x_{n}\right) & := \pm \mu_{\mp}\left(\mu_{+}+\mu_{-}\right)^{-1} e^{-x_{n}\left(-\Delta^{\prime}\right)^{1 / 2}}\left(h_{2}-\llbracket u^{1} \rrbracket\right),
\end{aligned}
$$

where $x_{n}>0$ and $E_{ \pm}: H_{p}^{k}\left(\mathbb{R}_{ \pm}^{n}\right) \rightarrow H_{p}^{k}\left(\mathbb{R}^{n}\right)$ is an extension operator. Indeed, the function $u^{1}$ belongs to $\bigcap_{j=2}^{k+2} \dot{\mathcal{H}}_{p}^{2}\left(\dot{\mathbb{R}}^{n}\right)$ by Theorem B. 15 and satisfies $-\mu \Delta u^{1}=f$. Hence $h_{1}-\llbracket \mu \partial_{n} u^{1} \rrbracket$ belongs to $\bigcap_{j=1}^{k+1} \dot{\mathcal{W}}_{p}^{j-1 / p}\left(\mathbb{R}^{n-1}\right)$ and $h_{2}-\llbracket u^{1} \rrbracket$ belongs to $\bigcap_{j=2}^{k+2} \dot{\mathcal{W}}_{p}^{j-1 / p}\left(\mathbb{R}^{n-1}\right)$. Then Theorems B. 15 and B. 28 imply $u \in \bigcap_{j=2}^{k+2} \dot{\mathcal{H}}_{p}^{j}\left(\mathbb{R}^{n}\right)$ and we have $\llbracket \mu \partial_{n} u \rrbracket=h_{1}$ and $\llbracket u \rrbracket=h_{2}$. Finally, Lemma 2.23 yields the estimate $\|\nabla u\|_{p} \lesssim\left\|F_{\left(f, h_{1}\right)}\right\|_{\hat{H}_{p}^{-1}\left(\mathbb{R}^{n}\right)}$ and therefore $u$ belongs to $\mathbb{E}_{\lambda}^{k}$.
(iv) Uniform estimates with respect to $\lambda$. We employ the rescaled functions $u_{\lambda}, f_{\lambda}, h_{1 \lambda}$, and $h_{2 \lambda}$ from page 30. Then the identity $A u=\left(f, h_{1}, h_{2}\right)$ is equivalent to $A u_{\lambda}=\left(f_{\lambda}, h_{1 \lambda}, h_{2 \lambda}\right)$ and we have

$$
\|u\|_{\mathbb{E}_{\lambda}^{k}}=\left\|u_{\lambda}\right\|_{\mathbb{E}_{1}^{k}}, \quad\left\|\left(f, h_{1}, h_{2}\right)\right\|_{\mathbb{F}_{\lambda}^{k}}=\left\|\left(f_{\lambda}, h_{1 \lambda}, h_{2 \lambda}\right)\right\|_{\mathbb{F}_{1}^{k}} .
$$

Therefore $A^{-1}$ is uniformly bounded with respect to $\lambda \in(0, \infty)$ and $\mu_{ \pm} \in\left[\mu_{0}, \mu_{0}^{-1}\right]$.
It remains to study the half-space problems

$$
\left\{\begin{array}{ll}
-\mu \Delta u=f & \text { in } \mathbb{R}_{+}^{n}, \\
-\mu \partial_{n} u=g & \text { on } \mathbb{R}^{n-1} .
\end{array}\right\}, \quad\left\{\int_{\mathbb{R}^{n}} \mu \nabla u \cdot \nabla \phi d x=\langle F \mid \phi\rangle \quad \text { for all } \phi \in \dot{H}_{p^{\prime}}^{1}\left(\mathbb{R}^{n}\right) \cdot\right\},
$$

with a constant coefficient $\mu>0$. The right one is (except for $\mu \neq 1$ ) the weak Neumann problem, which is covered by Theorem 2.19. However, we still have to verify the mapping properties with respect to the higher regularity conditions.
2.25. Lemma. Let $n \geq 2, \Omega=\mathbb{R}_{+}^{n}, \mu_{0} \in(0,1], k \in \mathbb{N}_{0} \cup\{-1\}$, and $p \in(1, \infty)$. Then the operator

$$
A: \mathbb{E}_{\lambda}^{k}\left(\mathbb{R}_{+}^{n}\right) \rightarrow \mathbb{F}_{c c, \lambda}^{k}\left(\mathbb{R}_{+}^{n}\right), \quad u \mapsto A u= \begin{cases}\left(-\operatorname{div}(\mu \nabla u),-\mu \partial_{n} u\right) & \text { if } k \geq 0 \\ F_{\mu \nabla u} & \text { if } k=-1,\end{cases}
$$

is uniformly invertible with respect to $\mu \in\left[\mu_{0}, \mu_{0}^{-1}\right]$ and $\lambda \in(0, \infty)$.
Proof. (i) Uniqueness follows from Theorem 2.19.
(ii) Existence for $k \geq 0$. We construct a solution $u=u^{1}+u^{2}$ by

$$
\begin{align*}
u^{1} & :=\left.\mu^{-1}\left((-\Delta)^{-1}\left(E_{+} f\right)\right)\right|_{\mathbb{R}_{+}^{n}}  \tag{2.25}\\
u^{2}\left(\cdot, x_{n}\right) & :=-\mu^{-1} e^{-x_{n}\left(-\Delta^{\prime}\right)^{1 / 2}}\left(-\Delta^{\prime}\right)^{-1 / 2}\left(g+\mu \partial_{n} u^{1}(\cdot, 0)\right)
\end{align*}
$$

Then $u^{1}$ belongs to $\bigcap_{j=2}^{k+2} \dot{\mathcal{H}}_{p}^{2}\left(\dot{\mathbb{R}}_{+}^{n}\right)$ by Theorem B. 15 and satisfies $-\mu \Delta u^{1}=f$. Hence $g+$ $\mu \partial_{n} u^{1}(\cdot, 0)$ belongs to $\bigcap_{j=1}^{k+1} \dot{\mathcal{W}}_{p}^{j-1 / p}\left(\mathbb{R}^{n-1}\right)$ and Theorems B. 15 and B. 28 imply $u \in \bigcap_{j=2}^{k+2} \dot{H}_{p}^{j}\left(\mathbb{R}_{+}^{n}\right)$ and we have $-\mu \Delta u=f$ and $-\mu \partial_{n} u(\cdot, 0)=g$. Finally, the weak a priori estimate implies $\|\nabla u\|_{p} \lesssim\left\|F_{(f, g)}\right\|_{\hat{H}_{p}^{-1}\left(\mathbb{R}_{+}^{n}\right)}$ and therefore $u$ belongs to $\mathbb{E}_{\lambda}^{k}$.
(iii) Existence for $k=-1$. Let $F \in \hat{H}_{p}^{-1}\left(\mathbb{R}_{+}^{n}\right)=\dot{H}_{p^{\prime}}^{1}\left(\mathbb{R}_{+}^{n}\right)^{*}$. Since we may identify $\dot{H}_{p^{\prime}}^{1}\left(\mathbb{R}_{+}^{n}\right)$ isometrically with the closed subspace $\nabla \dot{H}_{p^{\prime}}^{1}\left(\mathbb{R}_{+}^{n}\right)$ of $L_{p^{\prime}}\left(\mathbb{R}_{+}^{n}\right)^{n}$, there exists $f \in L_{p}\left(\mathbb{R}_{+}^{n}\right)^{n}$ with $\|f\|_{p}=\|F\|_{\hat{H}_{p}^{-1}\left(\mathbb{R}_{+}^{n}\right)}$ such that $\langle F \mid \phi\rangle=\int_{\mathbb{R}_{+}^{n}} f \cdot \nabla \phi d x$ for all $\phi \in \dot{H}_{p^{\prime}}^{1}\left(\mathbb{R}_{+}^{n}\right)$ [cf. AF03, Theorem 3.9]. Let $\left(f_{k}\right) \subset H_{p}^{1}\left(\dot{\mathbb{R}}_{+}^{n}\right)^{n}$ be a sequence such that $f_{k} \rightarrow f$ in $L_{p}\left(\mathbb{R}_{+}^{n}\right)^{n}$ as $k \rightarrow \infty$ and let

$$
\left\langle F_{k} \mid \phi\right\rangle:=\int_{\mathbb{R}_{+}^{n}} f_{k} \cdot \nabla \phi d x=-\int_{\mathbb{R}_{+}^{n}} \operatorname{div} f_{k} \phi d x+\int_{\mathbb{R}^{n-1}} e_{n} \cdot f_{k} \phi d x^{\prime} .
$$

Then $F_{k} \rightarrow F$ in $\hat{H}_{p}^{-1}\left(\mathbb{R}_{+}^{n}\right)$.
Next, we construct solutions $u_{k}=u_{k}^{1}+u_{k}^{2} \in \dot{\mathcal{H}}_{p}^{2}\left(\mathbb{R}_{+}^{n}\right) \cap \dot{\mathcal{H}}_{p}^{1}\left(\mathbb{R}_{+}^{n}\right)$ of the systems

$$
\mu \Delta u_{k}=\operatorname{div} f_{k} \text { in } \mathbb{R}_{+}^{n}, \quad \mu \partial_{n} u_{k}=e_{n} \cdot f_{k} \text { on } \mathbb{R}^{n-1}
$$

by using (2.25) with $f$ replaced by $-\operatorname{div} f_{k}$. Then the identity $F_{\mu \nabla u_{k}}=F_{k}$ is valid and we have $u_{k} \in \dot{\mathcal{H}}_{p}^{1}\left(\mathbb{R}_{+}^{n}\right)$ and $\left\|\nabla u_{k}-\nabla u_{k^{\prime}}\right\|_{p} \leq C\left\|F_{k}-F_{k^{\prime}}\right\|_{\hat{H}_{p}^{-1}\left(\mathbb{R}_{+}^{n}\right)}$. Therefore $\nabla u_{k} \rightarrow \nabla u$ in $L_{p}\left(\mathbb{R}_{+}^{n}\right)$ for some $u \in \dot{\mathcal{H}}_{p}^{1}\left(\mathbb{R}_{+}^{n}\right)$ and this limit satisfies $F_{\mu \nabla u}=F$.
(iv) The uniform estimates again follow from a scaling argument as on page 45.
2.2.3. Perturbed model problems. We next solve the transmission problems

$$
\left\{\begin{align*}
-\operatorname{div}(\mu \nabla u) & =f & & \text { in } \Omega \backslash \Sigma,  \tag{2.26}\\
\mu \partial_{\nu} u & =g & & \text { on } \partial \Omega, \\
\llbracket \mu \partial_{\nu} u \rrbracket & =h_{1} & & \text { on } \Sigma, \\
\llbracket u \rrbracket & =h_{2} & & \text { on } \Sigma .
\end{align*}\right\}, \quad\left\{\begin{array}{rlrl}
\int_{\Omega} \mu \nabla u \cdot \nabla \phi d x & =\langle F \mid \phi\rangle & & \text { for all } \phi \in \dot{H}_{p^{\prime}}^{1}(\Omega), \\
\llbracket u \rrbracket=h_{2} & & \text { on } \Sigma .
\end{array}\right\} .
$$

for the bent interface case $\Omega=\mathbb{R}^{n}$ and $\Sigma=\Sigma_{\omega}=\left\{\left(x^{\prime}, \omega\left(x^{\prime}\right)\right): x^{\prime} \in \mathbb{R}^{n-1}\right\}$ with variable coefficients $\mu_{ \pm}: \Omega_{ \pm} \rightarrow(0, \infty)$, where $\Omega_{ \pm}=\left\{\left(x^{\prime}, x_{n}\right) \in \mathbb{R}^{n}: x_{n} \gtrless \omega\left(x^{\prime}\right)\right\}$. In the definitions of the norms of $\mathbb{E}_{\lambda}^{k}$ and $\mathbb{F}_{\mathrm{cc}, \lambda}^{k}$ from page 40 , we let $\Omega_{\lambda}^{\prime}:=\Theta_{\omega}\left(\lambda^{-1 / 2} \Omega^{\prime}\right)$ and $\Sigma_{\lambda}^{\prime}:=\Theta_{\omega}\left(\lambda^{-1 / 2} \Sigma^{\prime}\right)$, where $\Omega^{\prime} \neq \emptyset$ and $\Sigma^{\prime} \neq \emptyset$ are bounded open subsets of $\mathbb{R}^{n}$ and $\mathbb{R}^{n-1} \times\{0\}$ with $C^{1}$-boundaries. The $C^{1}$-diffeomorphism $\Theta_{\omega}:\left(x^{\prime}, x_{n}\right) \mapsto\left(x^{\prime}, x_{n}+\omega\left(x^{\prime}\right)\right)$ was studied on page 34 .
2.26. Lemma. Let $n \geq 2, \mu_{0} \in(0,1], k \in \mathbb{N}_{0} \cup\{-1\}$, and $p \in(1, \infty)$. Then there exists $\eta>0$ such that for every $M>0$ we can find some $\lambda_{0} \geq 1$ such that the operator

$$
A: \mathbb{E}_{\lambda}^{k}\left(\mathbb{R}^{n} \backslash \Sigma_{\omega}\right) \rightarrow \mathbb{F}_{c c, \lambda}^{k}\left(\mathbb{R}^{n} \backslash \Sigma_{\omega}\right), \quad u \mapsto A u= \begin{cases}\left(-\operatorname{div}(\mu \nabla u), \llbracket \mu \partial_{\nu} u \rrbracket, \llbracket u \rrbracket\right) & \text { if } k \geq 0, \\ \left(F_{\mu \nabla u}, \llbracket u \rrbracket\right) & \text { if } k=-1,\end{cases}
$$

is uniformly invertible with respect to
(i) $\omega \in C_{c}^{1}\left(\mathbb{R}^{n-1}\right) \cap C_{c}^{k+2-}\left(\mathbb{R}^{n-1}\right)$ with $\|\nabla \omega\|_{W_{\infty}^{k+1}} \leq M$ and $\|\nabla \omega\|_{\infty} \leq \eta$,
(ii) $\mu_{ \pm} \in C\left(\overline{\Omega_{ \pm}}\right) \cap W_{\infty}^{k+1}\left(\Omega_{ \pm}\right)$with $\mu_{0} \leq \mu \leq \mu_{0}^{-1},\|\mu\|_{W_{\infty}^{k+1}} \leq M$, and $\sup \left\{\left|\mu_{ \pm}(x)-\mu_{ \pm}(y)\right|\right.$ : $\left.x, y \in \Omega_{ \pm}\right\} \leq 2 \eta$,
(iii) $\lambda \in\left[\lambda_{0}, \infty\right)$.

Proof. (i) We first consider the case of constant coefficients $\mu_{ \pm} \in\left[\mu_{0}, \mu_{0}^{-1}\right]$.
(i.a) Transformation to the flat interface. As for Lemma 2.14, we consider the pull-backs

$$
\bar{u}=u \circ \Theta, \quad \bar{h}_{2}=h_{2} \circ \Theta \quad(\text { for } k \geq-1)
$$

where the $C^{1}$-diffeomorphism $\Theta=\Theta_{\omega}:\left(x^{\prime}, x_{n}\right) \mapsto\left(x^{\prime}, x_{n}+\omega\left(x^{\prime}\right)\right)$ satisfies $\partial \Theta, \partial \Theta^{-1} \in$ $W_{\infty}^{k+1}\left(\mathbb{R}^{n}\right)$ (cf. p. 34). We further define

$$
\bar{f}=f \circ \Theta, \quad \bar{h}_{1}=\left(1+|\nabla \omega|^{2}\right)^{1 / 2} h_{1} \circ \Theta \quad(\text { for } k \geq 0) .
$$

For $\phi \in \dot{\mathcal{H}}_{p^{\prime}}^{1}\left(\mathbb{R}^{n}\right),\left(f, h_{1}, h_{2}\right) \in \mathbb{F}_{\mathrm{cc}, \lambda^{\prime}}^{0} \bar{\phi}=\phi \circ \Theta$, and $u \in \mathbb{E}_{\lambda}^{-1}$, we obtain the transformed functionals

$$
\begin{aligned}
\left\langle F_{\left(f, h_{1}\right)} \mid \phi\right\rangle & =\int_{\mathbb{R}^{n}} f \phi d x-\int_{\Sigma_{\omega}} h_{1} \phi d \sigma \\
& =\int_{\mathbb{R}^{n}} \bar{f} \bar{\phi}|\operatorname{det} \partial \Theta| d x-\int_{\mathbb{R}^{n-1}} \sqrt{1+|\nabla \omega|^{2}}\left(h_{1} \circ \Theta\right) \bar{\phi} d x^{\prime}=\left\langle F_{\left(\bar{f}, \bar{h}_{1}\right)} \mid \bar{\phi}\right\rangle, \\
\left\langle F_{\mu \nabla u} \mid \phi\right\rangle & =\int_{\mathbb{R}^{n}} \mu \nabla u \cdot \nabla \phi d x=\int_{\mathbb{R}^{n}} \bar{\mu}[\partial \Theta]^{-\top} \nabla \bar{u} \cdot[\partial \Theta]^{-\top} \nabla \bar{\phi}|\operatorname{det} \partial \Theta| d x \\
& =\left\langle F_{\bar{\mu} \nabla \bar{u}} \mid \bar{\phi}\right\rangle+\int_{\mathbb{R}^{n}} \bar{\mu} \nabla \bar{u} \cdot\left([\partial \Theta]^{-1}[\partial \Theta]^{-\top}-I\right) \nabla \bar{\phi} d x .
\end{aligned}
$$

Let $\overline{\mathbb{E}}_{\lambda}^{k}$ and $\overline{\mathbb{F}}_{\mathrm{cc}, \lambda}^{k}$ denote the corresponding spaces on $\dot{\mathbb{R}}^{n}$. Then the maps $u \mapsto \bar{u}, \mathbb{E}_{\lambda}^{k} \rightarrow \overline{\mathbb{E}}_{\lambda}^{k}$ $(k \geq-1)$ and $\left(f, h_{1}, h_{2}\right) \mapsto\left(\bar{f}, \bar{h}_{1}, \bar{h}_{2}\right), \mathbb{F}_{\mathrm{cc}, \lambda}^{k} \rightarrow \overline{\mathbb{F}}_{\mathrm{cc}, \lambda}^{k}(k \geq 0)$ are linear bijections and we obtain the estimates

$$
\begin{aligned}
C(n, k, M)^{-1}\|u\|_{\mathbb{E}_{\lambda}^{k}} & \leq\|\bar{u}\|_{\mathbb{E}_{\lambda}^{k}} \leq C(n, k, M)\|u\|_{\mathbb{E}_{\lambda}^{k}}, \\
C(n, k, p, M)^{-1}\left\|\left(f, h_{1}, h_{2}\right)\right\|_{\mathbb{F}_{c c, \lambda}^{k}} & \leq\left\|\left(\bar{f}, \bar{h}_{1}, \bar{h}_{2}\right)\right\|_{\mathbb{F}_{c c, \lambda}^{k}} \leq C(n, k, p, M)\left\|\left(f, h_{1}, h_{2}\right)\right\|_{\mathbb{F}_{c, \lambda}^{k}, \lambda}
\end{aligned}
$$

Here the numbers $C(n, k, M)$ and $C(n, k, p, M)$ are uniform with respect to $\lambda \in[1, \infty)$ and with respect to those $\omega \in C_{c}^{k+2-}\left(\mathbb{R}^{n-1}\right)$ which satisfy $\|\nabla \omega\|_{W_{\infty}^{k+1}} \leq M$. Since $2^{-1}\|\nabla \phi\|_{p^{\prime}} \leq\|\nabla \bar{\phi}\|_{p^{\prime}} \leq$ $2\|\nabla \phi\|_{p^{\prime}}$ for $\phi \in \dot{\mathcal{H}}_{p^{\prime}}^{1}\left(\mathbb{R}^{n}\right)$, the map $F \mapsto \bar{F}$, defined by $\langle\bar{F} \mid \bar{\phi}\rangle:=\langle F \mid \phi\rangle$ for $\bar{\phi} \in \dot{H}_{p^{\prime}}^{1}\left(\mathbb{R}^{n}\right)$, is an isomorphism of $\hat{H}_{p}^{-1}\left(\mathbb{R}^{n}\right)$, and we have

$$
2^{-1}\|F\|_{\hat{H}_{p}^{-1}\left(\mathbb{R}^{n}\right)} \leq\|\bar{F}\|_{\hat{H}_{p}^{-1}\left(\mathbb{R}^{n}\right)} \leq 2\|F\|_{\hat{H}_{p}^{-1}\left(\mathbb{R}^{n}\right)} \quad \text { for } F \in \hat{H}_{p}^{-1}\left(\mathbb{R}^{n}\right)
$$

(i.b) The transformed problems are given by (cf. p. 35)

$$
\left\{\begin{aligned}
\lambda \bar{u}-\mu \Delta \bar{u} & =\bar{f}+P_{2} \bar{u}+P_{1} \bar{u} & & \text { in } \dot{\mathbb{R}}^{n}, \\
\llbracket \mu \partial_{n} \bar{u} \rrbracket & =\bar{h}_{1}+H \bar{u} & & \text { on } \mathbb{R}^{n-1}, \\
\llbracket \bar{u} \rrbracket & =\bar{h}_{2} & & \text { on } \mathbb{R}^{n-1} .
\end{aligned}\right\}, \quad\left\{\begin{array}{cll}
F_{\bar{\mu} \nabla \bar{u}}=\bar{F}+P_{3} \bar{u} & & \text { in } \hat{H}_{p}^{-1}\left(\mathbb{R}^{n}\right), \\
\llbracket \bar{u} \rrbracket=\bar{h}_{2} & & \text { on } \mathbb{R}^{n-1} .
\end{array}\right\} .
$$

Here the perturbations $P_{l}=P_{l}(\mu, \omega)$ and $H=H(\mu, \omega)$ are given by

$$
\begin{aligned}
P_{1} \bar{u} & =-\mu \Delta^{\prime} \omega \partial_{n} \bar{u}, \\
P_{2} \bar{u} & =\mu\left|\nabla^{\prime} \omega\right|^{2} \partial_{n}^{2} \bar{u}-2 \mu \partial_{n} \nabla^{\prime} \bar{u} \cdot \nabla^{\prime} \omega, \\
\left\langle P_{3} \bar{u} \mid \bar{\phi}\right\rangle & =\int_{\mathbb{R}^{n}} \bar{\mu} \nabla \bar{u} \cdot\left([\partial \Theta]^{-1}[\partial \Theta]^{-\top}-I\right) \nabla \bar{\phi} d x, \\
H \bar{u} & =\nabla^{\prime} \omega \cdot \llbracket \mu \nabla^{\prime} \bar{u} \rrbracket-|\nabla \omega|^{2} \llbracket \mu \partial_{n} \bar{u} \rrbracket .
\end{aligned}
$$

For $\bar{u} \in \overline{\mathbb{E}}_{\lambda}^{k}$ and $\lambda \in\left[\lambda_{0}, \infty\right)$ we obtain the following estimates (cf. p. 35).

$$
\begin{array}{rlr}
\left\|\lambda^{(k-j) / 2} \nabla^{j}\left(P_{1} \bar{u}\right)\right\|_{p} \leq \lambda^{-1 / 2} C\left(n, \mu_{0}, k, M\right)\|\bar{u}\|_{\mathbb{E}_{\lambda}^{k}} & \text { for } 0 \leq j \leq k, \\
\left\|\lambda^{(k-j) / 2} \nabla^{j}\left(P_{2} \bar{u}\right)\right\|_{p} \leq\left(\eta C\left(n, \mu_{0}, k\right)+\lambda^{-1 / 2} C\left(n, \mu_{0}, k, M\right)\right)\|\bar{u}\|_{\overline{\mathbb{E}}_{\lambda}^{k}} & \text { for } 0 \leq j \leq k, \\
\left\|\lambda^{(k+1) / 2} P_{3} \bar{u}\right\|_{\hat{H}_{p}^{-1}\left(\mathbb{R}^{n}\right)} \leq \eta C\left(n, \mu_{0}\right)\|\bar{u}\|_{\mathbb{E}_{\lambda}^{k}}, & \\
\|H \bar{u}\|_{W_{p}^{k+1-1 / p}\left(\mathbb{R}^{n-1}\right), \lambda} \leq\left(\eta+\lambda^{-1 / 2+1 / 2 p}\right) C\left(n, \mu_{0}, k, p, M, \lambda_{0}\right)\|\bar{u}\|_{\mathbb{\mathbb { E }}_{\lambda}^{k}} & \text { for } k \geq 0 .
\end{array}
$$

Therefore a Neumann series argument as on page 36 yields the invertibility of $A$ and the uniform bounds in the case of constant coefficients $\mu_{ \pm} \in\left[\mu_{0}, \mu_{0}^{-1}\right]$.
(ii) For variable coefficients $\mu_{ \pm}$, we proceed as in the proof of Lemma 2.15. We study the perturbed problems (cf. p. 36)

$$
\left\{\begin{aligned}
\lambda u-\mu^{*} \Delta u & =f+P_{4} u & & \text { in } \mathbb{R}^{n} \backslash \Sigma_{\omega}, \\
\llbracket \mu^{*} \partial_{\nu} u \rrbracket & =h_{1}+H_{2} u & & \text { on } \Sigma_{\omega}, \\
\llbracket u \rrbracket & =\bar{h}_{2} & & \text { on } \Sigma_{\omega} .
\end{aligned}\right\}, \quad\left\{\begin{aligned}
F_{\mu^{*} \nabla u}=F+P_{5} u & & \text { in } \hat{H}_{p}^{-1}\left(\mathbb{R}^{n}\right), \\
\llbracket \bar{u} \rrbracket=\bar{h}_{2} & & \text { on } \Sigma_{\omega} .
\end{aligned}\right\},
$$

where the perturbations $P_{l}=P_{l}(\mu, \omega)$ and $H_{2}=H_{2}(\mu, \omega)$ are given by

$$
\begin{aligned}
P_{4} u & =\operatorname{div}\left(\left(\mu-\mu^{*}\right) \nabla u\right), \\
\left\langle P_{5} u \mid \phi\right\rangle & =\int_{\mathbb{R}^{n}}\left(\mu-\mu^{*}\right) \nabla u \cdot \nabla \phi d x, \\
H_{2} u & =\llbracket\left(\mu^{*}-\mu\right) \partial_{\nu} u \rrbracket .
\end{aligned}
$$

These perturbations can be estimated as follows.

$$
\begin{array}{rlrl}
\left\|P_{4} u\right\|_{H_{p}^{k}\left(\mathbb{R}^{n} \backslash \Sigma_{\omega}\right), \lambda} & \leq\left(\eta C\left(n, \mu_{0}, k\right)+\lambda^{-1 / 2} C\left(n, \mu_{0}, k, M\right)\right)\|u\|_{\mathbb{E}_{\lambda}^{k}} & & \text { if } k \geq 0, \\
\left\|\lambda^{(k+1) / 2} P_{5} u\right\|_{\hat{H}_{p}^{-1}\left(\mathbb{R}^{n}\right)} \leq \eta\|u\|_{\mathbb{E}_{\lambda}^{k}}, & & \\
\left\|H_{2} u\right\|_{W_{p}^{k+1-1 / p}\left(\Sigma_{\omega}\right), \lambda} & \leq\left(\eta C\left(n, \mu_{0}, k, p\right)+\lambda^{-1 / 2+1 / 2 p} C\left(n, \mu_{0}, k, p, M\right)\right)\|u\|_{\mathbb{E}_{\lambda}^{k}} & \text { if } k \geq 0 .
\end{array}
$$

Again, a Neumann series argument yields the uniform invertibility of $A: \mathbb{E}_{\lambda}^{k} \rightarrow \mathbb{F}_{\mathrm{cc}, \lambda}^{k}$.
The solvability of the perturbed model problem (2.26) in case $\Omega=\mathbb{R}^{n}$ and $\Omega_{\lambda}^{\prime}=\lambda^{-1 / 2} \Omega^{\prime}$ for $\Omega^{\prime} \subset \mathbb{R}^{n}$ and $\Sigma=\Sigma^{\prime}=\emptyset$ follows again by considering continuous coefficient functions.
2.27. Corollary. Let $n \geq 2, \mu_{0} \in(0,1], k \in \mathbb{N}_{0} \cup\{-1\}$, and $p \in(1, \infty)$. Then there exists $\eta>0$ such that for every $M>0$ we can find some $\lambda_{0} \geq 1$ such that the operator

$$
A: \mathbb{E}_{\lambda}^{k}\left(\mathbb{R}^{n}\right) \rightarrow \mathbb{F}_{c c, \lambda}^{k}\left(\mathbb{R}^{n}\right), \quad u \mapsto A u= \begin{cases}-\operatorname{div}(\mu \nabla u) & \text { if } k \geq 0, \\ F_{\mu \nabla u} & \text { if } k=-1,\end{cases}
$$

is uniformly invertible with respect to $\mu \in C\left(\mathbb{R}^{n}\right) \cap W_{\infty}^{k+1}\left(\mathbb{R}^{n}\right)$ with $\mu_{0} \leq \mu \leq \mu_{0}^{-1},\|\mu\|_{W_{\infty}^{k+1}} \leq M$, and $\sup \left\{|\mu(x)-\mu(y)|: x, y \in \mathbb{R}^{n}\right\} \leq 2 \eta$, and $\lambda \in\left[\lambda_{0}, \infty\right)$.

The bent half-space problems (2.26) for $\Omega=\mathbb{R}_{\omega}^{n}$ and $\Omega_{\lambda}^{\prime}=\Theta_{\omega}\left(\lambda^{-1 / 2} \Omega^{\prime}\right)$ with $\Omega^{\prime} \subset \mathbb{R}_{+}^{n}$ and $\Sigma=\Sigma^{\prime}=\emptyset$ can be solved analogously as the bent interface problem, by following the lines of the proof of Lemma 2.26 and by using the half-space result Lemma 2.25.
2.28. Corollary. Let $n \geq 2, \mu_{0} \in(0,1], k \in \mathbb{N}_{0} \cup\{-1\}$, and $p \in(1, \infty)$. Then there exists $\eta>0$ such that for every $M>0$ we can find some $\lambda_{0} \geq 1$ such that the operator

$$
A: \mathbb{E}_{\lambda}^{k}\left(\mathbb{R}_{\omega}^{n}\right) \rightarrow \mathbb{F}_{c c, \lambda}^{k}\left(\mathbb{R}_{\omega}^{n}\right), \quad u \mapsto A u= \begin{cases}\left(-\operatorname{div}(\mu \nabla u), \mu \partial_{\nu} u\right) & \text { if } k \geq 0, \\ F_{\mu \nabla u} & \text { if } k=-1,\end{cases}
$$

is uniformly invertible with respect to
(i) $\omega \in C_{c}^{1}\left(\mathbb{R}^{n-1}\right) \cap C_{c}^{k+2-}\left(\mathbb{R}^{n-1}\right)$ with $\|\nabla \omega\|_{W_{\infty}^{k+1}} \leq M$ and $\|\nabla \omega\|_{\infty} \leq \eta$,
(ii) $\mu \in C\left(\overline{\mathbb{R}_{\omega}^{n}}\right) \cap W_{\infty}^{k+1}\left(\mathbb{R}_{\omega}^{n}\right)$ with $\mu_{0} \leq \mu \leq \mu_{0}^{-1},\|\mu\|_{W_{\infty}^{k+1}} \leq M$, and $\sup \{|\mu(x)-\mu(y)|: x, y \in$ $\left.\mathbb{R}_{\omega}^{n}\right\} \leq 2 \eta$,
(iii) $\lambda \in\left[\lambda_{0}, \infty\right)$.
2.2.4. The weak transmission problem in bounded domains. We next consider the problems (2.26) in bounded domain $\Omega \subset \mathbb{R}^{n}$ with $C^{1}$-boundary $\partial \Omega$ and $C^{1}$-interface $\Sigma \subset \Omega$ and variable coefficients $\mu_{ \pm} \in C\left(\bar{\Omega}_{ \pm}\right)$. We first study uniqueness of weak solutions.
2.29. Lemma. Let $\Omega$ and $\Sigma$ be bounded, let $\mu_{ \pm} \in C\left(\overline{\Omega_{ \pm}}\right)$with $\inf \mu_{ \pm}>0$ and $p \in(1, \infty)$. Then every solution $u \in H_{p}^{1}(\Omega)$ of the problem

$$
\int_{\Omega} \mu \nabla u \cdot \nabla \phi d x=0 \quad \text { for all } \phi \in \mathcal{D}\left(\mathbb{R}^{n}\right)
$$

is a constant function.
Proof. The proof of this lemma is easy for $p \in[2, \infty)$, since we can choose $\phi=\bar{u}$, the complex conjugate of $u$. For $p \in(1,2)$ we employ the localization procedure of [SS92, Lemma 3.9].
(i) First, we assume that $\nabla u$ belongs to $L_{2}(\Omega)$. If $p \geq 2$, then $u$ belongs to $H_{2}^{1}(\Omega)$ by the Poincaré-Wirtinger inequality. By choosing $\phi=\bar{u} \in H_{2}^{1}(\Omega)$ we obtain $\int_{\Omega} \mu \nabla u \cdot \nabla \bar{u} d x=0$ and hence $\nabla u=0$ in $\Omega$, which yields the assertion. In the case $p \in(1,2)$, we let

$$
\begin{equation*}
1 / q_{j}:=1 / p-j / n, \quad j \in\{0,1, \ldots, k\} \tag{2.27}
\end{equation*}
$$

where $k \in \mathbb{N}_{0}$ is chosen such that $1 / q_{k} \leq 1 / 2<1 / q_{k-1}$. From the Sobolev embedding theorem we obtain the embedding $H_{q_{j}}^{1}(\Omega) \hookrightarrow L_{q_{j+1}}(\Omega)$. For $j \leq k$ we have $\nabla u \in L_{2}(\Omega) \hookrightarrow L_{q_{j}}(\Omega)$ and therefore induction yields $u \in H_{q_{k}}^{1}(\Omega) \hookrightarrow L_{q_{k+1}}(\Omega) \hookrightarrow L_{2}(\Omega)$. Hence $u$ belongs to $H_{2}^{1}(\Omega)$ and we again obtain $\nabla u=0$ in $\Omega$. It remains to prove that $\nabla u \in L_{2}(\Omega)$ for all $u \in H_{p}^{1}(\Omega)$ with $F_{\mu \nabla u}=0$.
(ii) Localization set-up. Lemma 2.9 implies that for every given $\eta>0$ there exists $r_{0}(\eta)>0$ such that for all $r \in\left(0, r_{0}(\eta)\right]$ we can find an $(\eta, r)$-localization set-up $\left(U_{j}, \Theta_{j}, \omega_{j}\right)_{j \in I(\eta, r)}$ for $\Omega \backslash \Sigma$. Let $I=I_{1} \cup I_{2} \cup I_{3}, \Theta_{j}: x \mapsto p_{j}+Q_{j} x, \Omega_{j}$, and $\Sigma_{j}$ have the same meaning as in the proof of Theorem 2.18 on page 37. We may also assume that the sets $\left(\Theta_{j}\left(B_{r 2^{-k-1}}\right)\right)_{j \in I}$ cover $\bar{\Omega}$ with $k$ from step (i).

There exists $\mu_{0} \in(0,1]$ such that $\mu_{0} \leq \mu_{ \pm} \leq \mu_{0}^{-1}$ in $\Omega \backslash \Sigma$. We now choose the number $\eta\left(n, \mu_{0}, p\right)>0$ such that Lemma 2.26 and Corollaries 2.27 and 2.28 are applicable. Then there exists $r_{\mu}(\eta)>0$ such that $\left|\mu_{ \pm}(x)-\mu_{ \pm}(y)\right| \leq 2 \eta$ for all $x, y \in \Omega_{ \pm}$with $|x-y| \leq 2 r_{\mu}(\eta)$. We define local coefficient functions $\mu_{j}$ as on page 38 and obtain $\left\|\mu_{j, \pm}-\mu_{j, \pm}^{*}\right\|_{\infty} \leq \eta$ for some constants $\mu_{j, \pm}^{*}$ and all $j \in I(\eta, r)$, provided that $r \in\left(0, r_{0}(\eta) / 2\right] \cap\left(0, r_{\mu}(\eta) / 2\right]$. Now the aforementioned results are applicable and yield suitable numbers $\lambda_{0} \geq 1$ and $C \geq 1$ such that

$$
\left\|\nabla\left(A_{j}^{-1} F_{j}\right)\right\|_{L_{p}\left(\Omega_{j}\right)} \leq C\left\|F_{j}\right\|_{\hat{H}_{p}^{-1}\left(\Omega_{j}\right)} \quad \text { for } F_{j} \in \hat{H}_{p}^{-1}\left(\Omega_{j}\right), j \in I, \lambda \in\left[\lambda_{0}, \infty\right)
$$

(iii) We now show that $\nabla u$ belongs to $L_{2}(\Omega)$ by refining the argument in step (i). Let $j \in I_{3}$ be fixed. We define the numbers $q_{l}(l \in\{0,1, \ldots, k\})$ by (2.27) and let $r_{l}:=r 2^{-l}(l \in\{0,1, \ldots, k+$ $1\}$ ). We further choose $\psi_{l} \in \mathcal{D}\left(B_{r_{l}}\right)$ such that $0 \leq \psi_{l} \leq 1$ and $\psi_{l}=1$ on $B_{r_{l+1}} \subset \mathbb{R}^{n}$. For every $v \in \dot{\mathcal{H}}_{p^{\prime}}^{1}\left(\mathbb{R}^{n}\right)$, we let $v_{l}:=\left.v\right|_{B_{r_{l}}}-\langle v\rangle_{B_{r_{l}}}$. Since $\left(\psi_{l} v_{l}\right) \circ \Theta_{l}^{-1}$ belongs to $\dot{\mathcal{H}}_{p^{\prime}}^{1}(\Omega)$ and since $\partial_{x} \Theta_{j}$ is orthogonal, we obtain

$$
\int_{\mathbb{R}^{n}} \mu_{j} \nabla\left(u \circ \Theta_{j}\right) \cdot \nabla\left(\psi_{l} v_{l}\right) d x=\int_{\Omega} \mu \nabla u \cdot \nabla\left(\left(\psi_{l} v_{l}\right) \circ \Theta_{j}^{-1}\right) d x=0
$$

With $\bar{u}:=u \circ \Theta_{j}$, this yields

$$
\begin{equation*}
\int_{\mathbb{R}^{n}} \mu_{j} \nabla\left(\psi_{l} \bar{u}\right) \cdot \nabla v_{l} d x=\int_{B_{r_{l}}} \mu_{j} \bar{u} \nabla \psi_{l} \cdot \nabla v_{l} d x-\int_{B_{r_{l}}} \mu_{j} \nabla \bar{u} \cdot\left(v_{l} \nabla \psi_{l}\right) d x \tag{2.28}
\end{equation*}
$$

From (2.28) we shall deduce that $\psi_{l} \bar{u} \in H_{q_{l}}^{1}\left(\mathbb{R}^{n}\right)$ for $l \in\{0,1, \ldots, k\}$ by induction. For $l=0$ the assertion is valid since $q_{0}=p$. Next, suppose that $\psi_{l-1} \bar{u}$ belongs to $H_{q_{l-1}}^{1}\left(\mathbb{R}^{n}\right)$ for some $l \in\{1, \ldots, k\}$. With $\psi_{l-1}=1$ on $B_{r_{l}}=B_{r_{l-1} / 2}$, this implies $\bar{u} \in H_{q_{l-1}}^{1}\left(B_{r_{l}}\right) \hookrightarrow L_{q_{l}}\left(B_{r_{l}}\right)$ by the Sobolev embedding theorem. With the dual exponents $q_{l}^{\prime}$, defined by $1 / q_{l}^{\prime}:=1-1 / q_{l}=$
$1 / p^{\prime}+l / n(l \in\{0,1, \ldots, k\})$, we obtain $v \in \dot{\mathcal{H}}_{p^{\prime}}^{1}\left(\mathbb{R}^{n}\right) \subset \dot{\mathcal{H}}_{q_{l}^{\prime}, \text { loc }}^{1}\left(\mathbb{R}^{n}\right) \subset L_{q_{l-1}^{\prime}, \text { loc }}\left(\mathbb{R}^{n}\right)$ from the Poincaré-Wirtinger inequality and the Sobolev embedding theorem. Hence

$$
\begin{aligned}
& \left|\int_{B_{r_{l}}} \mu_{j} \bar{u} \nabla \psi_{l} \cdot \nabla v_{l} d x\right| \leq\left\|\mu_{j} \bar{u}\right\|_{L_{q_{l}}\left(B_{r_{l}}\right)}\left\|\nabla \psi_{l}\right\|_{\infty}\left\|\nabla v_{l}\right\|_{L_{q_{l}^{\prime}}\left(B_{r_{l}}\right)} \leq C_{1}\left(\mu, u, \psi_{l}\right)\left\|\nabla v_{l}\right\|_{L_{p^{\prime}}\left(\mathbb{R}^{n}\right)}, \\
& \left|\int_{B_{r_{l}}} \mu_{j} \nabla \bar{u} \cdot\left(v_{l} \nabla \psi_{l}\right) d x\right| \leq\left\|\mu_{j} \nabla \bar{u}\right\|_{L_{q_{l-1}}\left(B_{r_{l}}\right)}\|v\|_{L_{q_{l-1}^{\prime}}\left(B_{r_{l}}\right)}\left\|\nabla \psi_{l}\right\|_{\infty} \leq C_{2}\left(\mu, u, \psi_{l}\right)\left\|\nabla v_{l}\right\|_{L_{p^{\prime}}\left(\mathbb{R}^{n}\right)} .
\end{aligned}
$$

Since the map $v \mapsto v_{l}+\mathbb{K}, \dot{\mathcal{H}}_{p^{\prime}}^{1}\left(\mathbb{R}^{n}\right) \rightarrow \dot{H}_{p^{\prime}}^{1}\left(B_{r_{l}}\right)$ is surjective, the identity (2.28) and Lemma 2.26 imply $\nabla\left(\psi_{l} \bar{u}\right) \in L_{q_{l}}\left(\mathbb{R}^{n}\right)$ and hence $\psi_{l} \bar{u} \in H_{q_{l}}^{1}\left(\mathbb{R}^{n}\right)$. Induction therefore yields $\psi_{k} \bar{u} \in$ $H_{q_{k}}^{1}\left(B_{r_{k}}\right) \hookrightarrow H_{2}^{1}\left(B_{r_{k}}\right)$ and hence $\left.\bar{u}\right|_{B_{r 2}-k-1} \in H_{2}^{1}\left(B_{r_{j} 2^{-k-1}}\right)$. In the case $j \in I_{1} \cup I_{2}$ we proceed analogously, by using Corollaries 2.27 and 2.28 instead of Lemma 2.26. Since the open sets $\Theta_{j}\left(B_{r_{j} 2^{-k-1}}\right)$ cover $\bar{\Omega}$, we obtain $u \in H_{2}^{1}(\Omega \backslash \Sigma)$. Then step (i) yields the assertion.

We are ready to prove that the weak transmission problem (2.2) has optimal $\dot{H}_{p}^{1}$-regularity.
Proof of Theorem 2.3. The cases $\Omega \backslash \Sigma \in\left\{\mathbb{R}^{n}, \mathbb{R}_{\omega}^{n}, \mathbb{R}^{n} \backslash \Sigma_{\omega}\right\}$ were solved in Lemma 2.26 and Corollaries 2.27 and 2.28. For the remaining case we follow the proof of [SS92, Theorem 1.3].
(i) We prove the weak a priori estimate

$$
\begin{equation*}
\|\mu \nabla u\|_{L_{p}(\Omega)} \leq C\left\|F_{\mu \nabla u}\right\|_{\hat{H}_{p}^{-1}(\Omega)} \quad \text { for } u \in \dot{\mathcal{H}}_{p}^{1}(\Omega) \tag{2.29}
\end{equation*}
$$

Assume that it is not true. Then we find a sequence $\left(u_{k}\right) \subset \dot{\mathcal{H}}_{p}^{1}(\Omega)$ such that

$$
1=\left\|\mu \nabla u_{k}\right\|_{L_{p}(\Omega)} \geq k\left\|F_{\mu \nabla u_{k}}\right\|_{\hat{H}_{p}^{-1}(\Omega)} \quad \text { for all } k \in \mathbb{N}
$$

We may assume that $\int_{\Omega} u_{k} d x=0$, so that the sequence $\left\|u_{k}\right\|_{L_{p}(\Omega)}$ is bounded by the PoincaréWirtinger inequality. Since $H_{p}^{1}(\Omega)$ is compactly embedded into $L_{p}(\Omega)$, we may also assume that the sequence $\left(u_{k}\right)$ converges in $L_{p}(\Omega)$ to some limit $u \in L_{p}(\Omega)$. Furthermore, the space $Z:=\left\{v \in \dot{\mathcal{H}}_{p}^{1}(\Omega): \int_{\Omega} v d x=0\right\}$ with norm $\|\mu \nabla \cdot\|_{L_{p}(\Omega)}$ is isomorphic to the closed subspace $\mu \nabla Z$ of $L_{p}(\Omega)^{n}$ and therefore $Z$ is reflexive. Hence we may even assume that $u$ belongs to $Z$ and that $\left(u_{k}\right)$ converges weakly to $u$; that is, on the one hand $F_{\mu \nabla u_{k}} \rightarrow F_{\mu \nabla u}$ in $\hat{H}_{p}^{-1}(\Omega)$, but also

$$
\begin{equation*}
\lim _{k \rightarrow \infty} \int_{\Omega} \mu \nabla u_{k} \cdot f d x=\int_{\Omega} \mu \nabla u \cdot f d x \quad \text { for all } f \in L_{p^{\prime}}(\Omega)^{n} \tag{2.30}
\end{equation*}
$$

Thus, $F_{\mu \nabla u}=0$ and hence Lemma 2.29 implies that $\nabla u=0$.
Next, as in step (ii) in the proof of Lemma 2.29, we consider an open covering $\left(U_{j}\right)$ for $\bar{\Omega}$ with $j \in I=I_{1} \cup I_{2} \cup I_{3}$, and rigid transformations $\Theta_{j}: B_{r}(0) \subset \mathbb{R}^{n} \rightarrow U_{j} \subset \mathbb{R}^{n}$. We assume that the smaller sets $\Theta_{j}\left(B_{r_{j} / 2}\right)$ form an open covering for $\bar{\Omega}$ and choose functions $\psi_{j} \in \mathcal{D}\left(B_{r}\right)$ with $0 \leq \psi_{j} \leq 1$ and $\psi_{j}=1$ on $B_{r_{j} / 2}$. The weak a priori estimates for the model problems in Lemma 2.26 and Corollaries 2.27 and 2.28 imply that there is a number $C(n, \mu, p, \eta)>0$ such that

$$
\left\|\nabla\left(\psi_{j} u_{k} \circ \Theta_{j}\right)\right\|_{L_{p}\left(\Omega_{j}\right)} \leq C\left\|F_{\mu_{j} \nabla\left(\psi_{j} u_{k} \circ \Theta_{j}\right)}\right\|_{\hat{H}_{p}^{-1}\left(\Omega_{j}\right)} \quad \text { for } j \in I, k \in \mathbb{N}
$$

where the height functions $\omega_{j} \in C_{c}^{1}\left(\mathbb{R}^{n-1}\right), j \in I_{2} \cup I_{3}$, satisfy $\left\|\nabla^{\prime} \omega_{j}\right\|_{\infty} \leq \eta$ and the local coefficients $\mu_{j}$ satisfy $\left\|\mu_{j}-\mu_{j}^{*}\right\|_{\infty} \leq \eta$ for some locally constant functions $\mu_{j}^{*}$.

Let $j$ be fixed and put $B:=B_{r}, \bar{u}_{k}:=u_{k} \circ \Theta_{j}$. We choose a sequence $\left(v_{k}\right) \subset \dot{\mathcal{H}}_{p^{\prime}}^{1}(B)$ with $\int_{B} v_{k} d x=0$ and $\left\|\nabla v_{k}\right\|_{L_{p^{\prime}}(B)}=1$ which converges strongly to some $v$ in $L_{p^{\prime}}(B)$ and satisfies

$$
\left|\int_{B} \mu_{j} \nabla\left(\psi_{j} \bar{u}_{k}\right) \cdot \nabla v_{k} d x\right| \geq d_{k}-\frac{1}{k}, \quad \text { with } d_{k}:=\left\|F_{\mu_{j} \nabla\left(\psi_{j} \bar{u}_{k}\right)}\right\|_{\hat{H}_{p}^{-1}\left(\Omega_{j}\right)} .
$$

In order to show that $d_{k} \rightarrow 0$, we compute

$$
\begin{aligned}
& \int_{B} \mu_{j} \nabla\left(\psi_{j} \bar{u}_{k}\right) \cdot \nabla v_{k} d x \\
& \quad=\int_{B} \mu_{j} \nabla \bar{u}_{k} \cdot \nabla\left(\psi_{j} v_{k}\right) d x+\int_{B} \mu_{j} \bar{u}_{k} \nabla \psi_{j} \cdot \nabla v_{k} d x-\int_{B} \mu_{j} \nabla \bar{u}_{k} \cdot\left(v_{k} \nabla \psi_{j}\right) d x .
\end{aligned}
$$

Here the first summand on the right-hand side vanishes for $k \rightarrow \infty$, which can be seen by transforming the integral from $B$ to $\Theta_{j}(B)$ with the orthogonality of $\partial_{x} \Theta_{j}$ and by using that $F_{\mu \nabla u_{k}} \rightarrow 0$ in $\hat{H}_{p}^{-1}(\Omega)$. The second integral vanishes, since $u_{k} \rightarrow 0$ in $L_{p}(\Omega)$ and since $\left\|\mu_{j}\right\|_{\infty}$, $\left\|\nabla \psi_{j}\right\|_{\infty},\left\|\nabla v_{k}\right\|_{p^{\prime}},\left\|\nabla^{\prime} \omega_{j}\right\|_{\infty}$ are bounded. Finally, since $\left\|v_{k}\right\|_{p^{\prime}}$ is bounded, we may use (2.30) and $\nabla u=0$ to conclude that also the third integral vanishes.

We have shown that $\lim _{k \rightarrow \infty} F_{\mu_{j} \nabla\left(\psi_{j} u_{k} \circ \Theta_{j}\right)}=0$ in $\hat{H}_{p}^{-1}\left(\Omega_{j}\right)$ for each $j$. The weak a priori estimates for the model problems therefore imply that $\lim _{k \rightarrow \infty} \nabla\left(\psi_{j} u_{k} \circ \Theta_{j}\right)=0$ in $L_{p}\left(B_{r}\right)^{n}$ for every $j$. With $\psi_{j}=1$ on $B_{r_{j} / 2}$ and since the sets $\Theta_{j}\left(B_{r_{j} / 2}\right)$ cover $\bar{\Omega}$, we conclude that $\nabla u_{k} \rightarrow 0$ in $L_{p}(\Omega)$. This is a contradiction to $\left\|\mu \nabla u_{k}\right\|_{L_{p}(\Omega)}=1$. Therefore estimate (2.29) is valid.
(ii) Existence for given $F \in \hat{H}_{p}^{-1}(\Omega)$ and $h_{2}=0$. We employ the strategy from [SS92, Lemma 3.1]. Since the space $\nabla \dot{\mathcal{H}}_{p}^{1}(\Omega)$ is closed in $L_{p}(\Omega)^{n}$, it follows from step (i) that $X:=\left\{F_{\mu \nabla u}: u \in\right.$ $\left.\dot{\mathcal{H}}_{p}^{1}(\Omega)\right\}$ is a closed subspace of $\hat{H}_{p}^{-1}(\Omega)$. We assume that $X \neq \hat{H}_{p}^{-1}(\Omega)$ and seek a contradiction. The Hahn-Banach theorem yields a non-trivial functional $J \in\left(\hat{H}_{p}^{-1}(\Omega)\right)^{*} \backslash\{0\}$ such that $\left.J\right|_{X}=$ 0 . Since closed subspaces and quotient spaces of reflexive spaces are again reflexive, we may identify $\left(\hat{H}_{p}^{-1}(\Omega)\right)^{*}=\left(\dot{H}_{p^{\prime}}^{1}(\Omega)\right)^{* *} \cong \dot{H}_{p^{\prime}}^{1}(\Omega)$. Hence there exists a unique $\phi \in \dot{H}_{p^{\prime}}^{1}(\Omega)$ with $\|\nabla \phi\|_{L_{p^{\prime}}(\Omega)^{n}}=\|J\|_{\hat{H}_{p}^{-1}(\Omega)^{*}} \neq 0$ such that $\langle J \mid F\rangle=\langle F \mid \phi\rangle$ for every $F \in \hat{H}_{p}^{-1}(\Omega)$. Using assertion (i) for $p^{\prime}$ instead of $p$ and considering only the functionals $F_{\mu \nabla u}$ for $u \in \dot{\mathcal{H}}_{p}^{1}(\Omega)$, we see that

$$
\|\nabla \phi\|_{L_{p^{\prime}}(\Omega)} \lesssim \sup _{0 \neq u \in \dot{H}_{p}^{1}(\Omega)} \frac{\left|\int_{\Omega} \mu \nabla \phi \cdot \nabla u d x\right|}{\|\nabla u\|_{L_{p}(\Omega)}}=\sup _{0 \neq u \in \dot{H}_{p}^{1}(\Omega)} \frac{\left|\left\langle\left. J\right|_{X} \mid F_{\mu \nabla u}\right\rangle\right|}{\|\nabla u\|_{L_{p}(\Omega)}}=0
$$

This is a contradiction to $\nabla \phi \neq 0$. Therefore the map $u \mapsto F_{\mu \nabla u}, \dot{\mathcal{H}}_{p}^{1}(\Omega) \rightarrow \hat{H}_{p}^{-1}(\Omega)$ is surjective.
(iii) Existence for given $F \in \hat{H}_{p}^{-1}(\Omega)$ and $h_{2} \in W_{p}^{1-1 / p}(\Sigma)$. We construct $u=v+w$ as follows. Let $\mathcal{E}_{+} \in \mathcal{B}\left(W_{p}^{1-1 / p}(\Sigma) ; H_{p}^{1}\left(\Omega_{+}\right)\right)$denote a co-retraction for the trace operator $H_{p}^{1}\left(\Omega_{+}\right) \rightarrow$ $W_{p}^{1-1 / p}(\Sigma)$. Then we define $v_{+}:=\mathcal{E}_{+} h_{2}$ and $v_{-}:=0$, so that $v \in H_{p}^{1}(\Omega \backslash \Sigma)$ with $\llbracket v \rrbracket=h_{2}$. Finally, we determine $w \in H_{p}^{1}(\Omega)$ as a solution to $\int_{\Omega} \mu \nabla w \cdot \nabla \phi d x=\left\langle F-F_{\mu \nabla v} \mid \phi\right\rangle$ for $\phi \in H_{p^{\prime}}^{1}(\Omega)$. Then $u=v+w$ solves (2.2) and hence $u \mapsto\left(F_{\mu \nabla u}, \llbracket u \rrbracket\right)$ is surjective. We conclude that the map $u \mapsto\left(F_{\mu \nabla u}, \llbracket u \rrbracket\right), \mathbb{E}^{-1} \rightarrow \mathbb{F}^{-1}$ induced by the weak transmission problem (2.2) is invertible.
2.2.5. The strong transmission problem in bounded domains. In order to solve the strong transmission problem (2.1) in the case $\lambda=0$, we employ the following fact.
2.30. Proposition (cf. [EN00, Corollary IV.1.19] and [Lun95, Remark A.2.4]). Let $A: D(A) \subset$ $X \rightarrow X$ be a densely defined linear operator in a Banach space $X$ with compact resolvent. Then $\sigma(A)$ consists only of poles of $\lambda \mapsto(\lambda-A)^{-1}$ with finite algebraic multiplicity. If $\lambda \in \sigma(A)$ satisfies $N\left(\lambda_{0}-A\right)=N\left(\left(\lambda_{0}-A\right)^{2}\right)$, then $X=N(\lambda-A) \oplus R(\lambda-A)$ as a topological direct sum.

Proof of Theorem 2.2. The result for the cases $\Omega \backslash \Sigma \in\left\{\mathbb{R}^{n} \backslash \Sigma_{\omega}, \mathbb{R}_{\omega}^{n}, \mathbb{R}^{n}\right\}$ was proved in Lemma 2.26 and Corollaries 2.27 and 2.28 and it remains to consider a bounded domain.
(i) Homogenous boundary conditions. We define $L_{p, 0}(\Omega):=\left\{f \in L_{p}(\Omega): \int_{\Omega} f d x=0\right\}$ and, for $k \geq 0$, we consider the operator

$$
L=-\operatorname{div}(\mu \nabla \cdot), \quad D(L)=\left\{u \in H_{p}^{k+2}(\Omega \backslash \Sigma): \mu \partial_{\nu} u=0 \text { on } \partial \Omega, \llbracket \mu \partial_{\nu} u \rrbracket=\llbracket u \rrbracket=0 \text { on } \Sigma\right\} .
$$

Theorem 2.18 implies that $\lambda-L: D(L) \rightarrow H_{p}^{k}(\Omega \backslash \Sigma)$ is invertible for $|\lambda| \geq \lambda_{0}$ and hence the resolvent set of $L$ is not empty. The resolvent is also compact. From Lemma 2.29 we infer that $N(L)=\mathbb{K}$ and an integration by parts shows that $R(L) \subset H_{p}^{k}(\Omega \backslash \Sigma) \cap L_{p, 0}(\Omega)$. We also have the topological direct sum $L_{p}(\Omega)=L_{p, 0}(\Omega) \oplus \mathbb{K}$ where the projection onto $L_{p, 0}(\Omega)$ is given by $u \mapsto u-\langle u\rangle_{\Omega}$ where $\langle u\rangle_{\Omega}:=|\Omega|^{-1} \int_{\Omega} u d x$ denotes the mean value of $u$ in $\Omega$. Hence also

$$
H_{p}^{k}(\Omega \backslash \Sigma)=\left(H_{p}^{k}(\Omega \backslash \Sigma) \cap L_{p, 0}(\Omega)\right) \oplus \mathbb{K}
$$

In order to apply Proposition 2.30, we let $u \in N\left(L^{2}\right)$. Then $L u \in R(L) \cap N(L) \subset L_{p, 0}(\Omega) \cap \mathbb{K}$ which yields $L u=0$ and hence $u \in N(L)$. Therefore we also have $H_{p}^{k}(\Omega \backslash \Sigma)=R(L) \oplus \mathbb{K}$ which yields $R(L)=H_{p}^{k}(\Omega \backslash \Sigma) \cap L_{p, 0}(\Omega)$. Thus, the operator $L: D(L) \cap L_{p, 0}(\Omega) \rightarrow H_{p}^{k}(\Omega \backslash \Sigma) \cap L_{p, 0}(\Omega)$ is therefore bijective and bounded and therefore invertible by the closed graph theorem. As a consequence, the strong transmission problem admits at most one solution within $H_{p}^{k+2}(\Omega \backslash$ इ) $\cap L_{p, 0}(\Omega)$.
(ii) Existence. For given data ( $f, g, h_{1}, h_{2}$ ), we construct a solution $u=u^{1}+u^{2}$ to (2.1), by solving the subproblems

$$
\left\{\begin{aligned}
\lambda u_{1}-\operatorname{div}\left(\mu \nabla u_{1}\right) & =\langle f\rangle_{\Omega} & & \text { in } \Omega, \\
\mu \partial_{\nu} u_{1} & =g & & \text { on } \partial \Omega, \\
\llbracket \mu \partial_{\nu} u_{1} \rrbracket & =h_{1} & & \text { on } \Sigma, \\
\llbracket u_{1} \rrbracket & =h_{2} & & \text { on } \Sigma .
\end{aligned}\right\}, \quad\left\{\begin{aligned}
&-\operatorname{div}\left(\mu \nabla u_{2}\right)=\lambda u_{1}+f-\langle f\rangle_{\Omega} \\
& \text { in } \Omega, \\
& \mu \partial_{\nu} u_{2}=0 \\
& \llbracket \mu \partial_{\nu} u_{2} \rrbracket=0 \text { on } \partial \Omega, \\
& \llbracket u_{2} \rrbracket=0 \\
& \text { on } \Sigma, \\
& \text { an } \Sigma .
\end{aligned}\right\} .
$$

The first problem is solvable for some sufficiently large $\lambda \in[1, \infty)$ by Theorem 2.18. Then the compatibility condition on ( $f, g, h_{1}$ ) implies $\left\langle\lambda u_{1}\right\rangle_{\Omega}=0$ and therefore $\lambda u_{1}+f-\langle f\rangle_{\Omega}$ belongs to $H_{p}^{k}(\Omega \backslash \Sigma) \cap L_{p, 0}(\Omega)$. Hence the problem for $u^{2}$ is solvable by step (i). The proof of Theorem 2.2 is complete.

## CHAPTER 3

## The linearized problem

We investigate the linear problem (PL), which we restate as

$$
\left\{\begin{align*}
\rho \partial_{t} u-\mu \Delta u+\nabla \pi & =f_{u} & & \text { in } J \times \Omega \backslash \Sigma,  \tag{3.1}\\
\operatorname{div} u & =f_{d} & & \text { in } J \times \Omega \backslash \Sigma, \\
\llbracket u \rrbracket & =0 & & \text { on } J \times \Sigma, \\
L_{v}\left(u, h ; u_{*}\right) & =g_{v} & & \text { on } J \times \Sigma, \\
L_{w}\left(u, \pi, h ; u_{*}\right) & =g_{w} & & \text { on } J \times \Sigma, \\
\partial_{t} h-u \cdot \nu_{\Sigma} & =g_{h} & & \text { on } J \times \Sigma, \\
\left.u\right|_{\partial \Omega} & =0 & & \text { on } J \times \partial \Omega, \\
h_{t=0} & =0 & & \text { on } \Sigma, \\
\left.u\right|_{t=0} & =0 & & \text { in } \Omega \backslash \Sigma,
\end{align*}\right.
$$

Here we consider a bounded domain $\Omega \subset \mathbb{R}^{n}(n \geq 2)$ with smooth boundary $\partial \Omega$ and compact smooth interface $\Sigma \subset \Omega$ such that $\Omega \backslash \Sigma$ consists of disjoint open sets $\Omega_{+}$and $\Omega_{-}$with $\partial \Omega_{+} \cap$ $\partial \Omega_{-}=\Sigma$. We choose the unit normal vector field $\nu_{\Sigma}=\nu_{\partial \Omega_{-}}=-\nu_{\partial \Omega_{+}}$that points into $\Omega_{+}$. Given two functions $\psi_{ \pm}$on $\Omega_{ \pm}$, we put $\psi:=\psi_{+} \chi_{+}+\psi_{-} \chi_{-}$with the characteristic functions $\chi_{ \pm}$of $\Omega_{ \pm}$, and we define the jump $\llbracket \psi \rrbracket:=\left.\psi_{+}\right|_{\Sigma}-\left.\psi_{-}\right|_{\Sigma}$. In this way we define the density $\rho=\rho_{+} \chi_{+}+\rho_{-} \chi_{-}$and viscosity $\mu=\mu_{+} \chi_{+}+\mu_{-} \chi_{-}$with positive constants $\rho_{ \pm}$and $\mu_{ \pm}$. Moreover, $J=(0, T)$ is a bounded interval with $T \in(0, \infty)$ and $u_{*}: J \times \Sigma \rightarrow \mathbb{R}^{n}$ is a possibly nontangential vector field. In a tubular neighborhood $B_{r}(\Sigma) \subset \Omega$ of $\Sigma$, there exists a nonlinear projection $\Pi: B_{r}(\Sigma) \rightarrow \Sigma$ and we decompose the velocity field $u$ into

$$
u=v+w \nu_{\Sigma} \circ \Pi, \quad v:=\left[P_{\Sigma} \circ \Pi\right] u, \quad w:=\left(\nu_{\Sigma} \circ \Pi \mid u\right) .
$$

Analogously, we let $u_{*}=v_{*}+w_{*} \nu_{\Sigma}$ on $\Sigma$. Then the operators $L_{v}$ and $L_{w}$ are defined by

$$
\begin{aligned}
L_{v}\left(u, h ; u_{*}\right):= & -\mu_{s} \widetilde{\Delta}_{\Sigma} v-\lambda_{s} \nabla_{\Sigma} \operatorname{div}_{\Sigma} v-\llbracket \mu \partial_{\nu} v \rrbracket-\llbracket \mu \rrbracket \nabla_{\Sigma} w+\left(\lambda_{s}+\mu_{s}\right) w_{*} \nabla_{\Sigma} \Delta_{\Sigma} h, \\
L_{w}\left(u, \pi, h ; u_{*}\right):= & -\operatorname{tr}\left(\left[\left(\lambda_{s}-\mu_{s}\right) H_{\Sigma}+2 \mu_{s} L_{\Sigma}\right] \nabla_{\Sigma} v\right)-2 \llbracket \mu \partial_{\nu} w \rrbracket+\llbracket \pi \rrbracket \\
& -\operatorname{tr}\left(\left[\sigma+\left(\lambda_{s}-\mu_{s}\right)\left(\operatorname{div}_{\Sigma} v_{*}-2 H_{\Sigma} w_{*}\right)+2 \mu_{s}\left(D_{\Sigma}\left(v_{*}\right)-2 w_{*} L_{\Sigma}\right)\right] \nabla_{\Sigma}^{2} h\right) .
\end{aligned}
$$

Here the surface shear viscosity $\mu_{s}$ is a positive constant and $\lambda_{s}$ (the surface dilational visocosity if $n=3$ ) is a real number. Moreover, we employ the surface gradient $\nabla_{\Sigma} w=\tau^{j} \partial_{j} w$, the surface divergence $\operatorname{div}_{\Sigma} u=\tau^{j} \cdot \partial_{j} u$, the scalar Laplace-Beltrami operator $\Delta_{\Sigma} h=\operatorname{div}_{\Sigma} \nabla_{\Sigma} h$, the tangential Laplace-Beltrami operator $\widetilde{\Delta}_{\Sigma} v=g^{j k} \widetilde{\nabla}_{j} \widetilde{\nabla}_{k} v$, the Weingarten tensor $L_{\Sigma}=-\nabla_{\Sigma} \nu_{\Sigma}$, and the $(n-1)$-fold mean curvature $H_{\Sigma}=\operatorname{tr} L_{\Sigma}$. More information on the differential geometric quantities is given in Appendix A.

In this chapter we prove that problem (3.1) has optimal regularity in the sense that, for suitable Banach spaces ${ }_{0} \mathbb{E}$ and ${ }_{0} \mathbb{F}$, the solution-to-data map of problem (3.1), given by $(u, \pi, h) \mapsto$ $\left(f_{u}, f_{d}, g_{v}, g_{w}, g_{h}\right),{ }_{0} \mathbb{E} \rightarrow{ }_{0} \mathbb{F}$, is a topological linear isomorphism. To this end, it is crucial to understand the situation of a flat interface $\Sigma=\mathbb{R}^{n} \times\{0\} \cong \mathbb{R}^{n}$ in the whole space $\Omega=\mathbb{R}^{n+1}$. For the corresponding model problem (MP) we prove optimal regularity in Section 3.1 (see Theorems 3.1 and 3.14). Next, in Section 3.2, we prove optimal regularity for a perturbed model
problem with a bent hyperplane and variable coefficients (see Theorem 3.16). Finally, Section 3.3 contains the main result on optimal regularity for problem (3.1) in a bounded configuration (see Theorem 3.21).

### 3.1. The interface conditions

In this section we prove optimal regularity for the model problem (MP) that corresponds to problem (3.1) in the situation of a flat interface $\Sigma=\mathbb{R}^{n} \times\{0\} \cong \mathbb{R}^{n}$ in the whole space $\Omega=\mathbb{R}^{n+1}$ $(n \in \mathbb{N})$ with $\left(f_{u}, f_{d}\right)=0$. We restate this problem as

$$
\left\{\begin{array}{rlrl}
\rho\left(\tau+\partial_{t}\right) u-\mu \Delta u+\nabla \pi=0 & & \text { in } J \times \dot{\mathbb{R}}^{n+1},  \tag{3.2}\\
\operatorname{div} u=0 & & \text { in } J \times \dot{\mathbb{R}}^{n+1}, \\
\llbracket u \rrbracket & =0 & & \text { on } J \times \mathbb{R}^{n}, \\
-\mu_{s} \Delta_{x} v-\lambda_{s} \nabla_{x} \operatorname{div}_{x} v-c_{5} \llbracket \mu \nabla_{x} w \rrbracket-c_{6} \llbracket \mu \partial_{y} v \rrbracket+c_{1} \nabla_{x} \Delta_{x} h=g_{v} & & \text { on } J \times \mathbb{R}^{n}, \\
-\operatorname{tr}\left(\left(c_{2}+2 C_{3}\right) \nabla_{x} v\right)-2 \llbracket \mu \partial_{y} w \rrbracket+\llbracket \pi \rrbracket-\operatorname{tr}\left(\left(c_{\sigma}+C_{4}\right) \nabla_{x}^{2} h\right)=g_{w} & & \text { on } J \times \mathbb{R}^{n}, \\
\left(\tau+\partial_{t}\right) h-w & =g_{h} & & \text { on } J \times \mathbb{R}^{n}, \\
\left.h\right|_{t=0}=0 & & \text { on } \mathbb{R}^{n}, \\
\left.u\right|_{t=0}=0 & & \text { in } \mathbb{R}^{n+1} .
\end{array}\right.
$$

In this section we let $\rho_{ \pm}, \mu_{ \pm}, \sigma$, and $\mu_{s}$ be positive constants, $\lambda_{s}$ be a real number, $\tau \in[0, \infty)$ be a constant, $J=(0, T)$ or $J=(0, \infty)$, and $\mathbb{R}^{n+1}=\mathbb{R}^{n} \times(\mathbb{R} \backslash\{0\})$. The elements of $\mathbb{R}^{n+1}$ are denoted by $(x, y)$ with $x \in \mathbb{R}^{n}$ and $y \in \mathbb{R}$. The parameters $c_{1}, c_{2}, C_{3}, C_{4}, c_{5}, c_{6}$, and $c_{\sigma}$ are defined by

$$
\left\{\begin{align*}
c_{1} & :=\left(\lambda_{s}+\mu_{s}\right) \vartheta_{w}, & c_{2} & :=\left(\lambda_{s}-\mu_{s}\right) \operatorname{tr} \vartheta_{L},  \tag{3.3}\\
C_{3} & :=\mu_{s} \vartheta_{L}, & C_{4} & :=2 \mu_{s}\left(\vartheta_{D v}-2 \vartheta_{w} \vartheta_{L}\right), \\
c_{5}, c_{6} & \in\{0,1\}, & c_{\sigma} & :=\sigma+\left(\lambda_{s}-\mu_{s}\right) \operatorname{tr}\left(\vartheta_{D v}-2 \vartheta_{w} \vartheta_{L}\right),
\end{align*}\right.
$$

and depend on

$$
\vartheta=\left(\vartheta_{w}, \vartheta_{L}, \vartheta_{D v}\right) \quad \text { for } \vartheta_{w} \in \mathbb{R}, \vartheta_{L} \in \mathbb{R}^{n \times n}, \vartheta_{D v} \in \mathbb{R}^{n \times n} .
$$

In Section 3.3 we will relate these parameters to the normal reference velocity $w_{*}$, the Weingarten map $L_{\Sigma}$, and the tangential rate-of-strain tensor $D_{\Sigma}\left(v_{*}\right)$ in problem (3.1). We further abbreviate

$$
\vartheta_{H}:=\operatorname{tr}\left(\vartheta_{L}\right), \quad \vartheta_{d v}:=\operatorname{tr}\left(\vartheta_{D v}\right), \quad \vartheta_{D u}:=\vartheta_{D v}-\vartheta_{w} \vartheta_{L}, \quad \vartheta_{d u}:=\operatorname{tr}\left(\vartheta_{D u}\right)=\vartheta_{d v}-\vartheta_{w} \vartheta_{H},
$$

and we define

$$
d_{0}\left(\vartheta_{D u}\right):=\sigma+\left(\lambda_{s}-\mu_{s}\right) \operatorname{tr} \vartheta_{D u}+2 \mu_{s} \min _{\xi \in \mathbb{R}^{n} \backslash\{0\}}|\xi|^{-2} \xi^{\top}\left[\vartheta_{D u}\right] \xi \quad \text { for } \vartheta_{D u} \in \mathbb{R}^{n \times n}
$$

Then we define the following parameter set for problem (3.2) and a given number $M>0$ :

$$
\begin{equation*}
\mathcal{P}_{M}:=\left\{\vartheta=\left(\vartheta_{w}, \vartheta_{L}, \vartheta_{D v}\right) \in \mathbb{R} \times \mathbb{R}^{n \times n} \times \mathbb{R}^{n \times n}:|\vartheta| \leq M, d_{0}\left(\vartheta_{D u}\right) \geq 1 / M\right\} . \tag{3.4}
\end{equation*}
$$

Our main result on problem (3.2) reads as follows, where the solution space is denoted by

$$
\begin{align*}
{ }_{0} \mathbb{E}={ }_{0} \mathbb{E}(J, \tau):= & \left\{(u, \pi, h) \in{ }_{0} \mathbb{E}_{u, v, w}(J) \times{ }_{0} \mathbb{E}_{\pi,[\pi \pi]}(J) \times{ }_{0} \mathbb{E}_{h}(J):\right. \\
& \left.\rho\left(\tau+\partial_{t}\right) u-\mu \Delta u+\nabla \pi=0, \operatorname{div} u=0\right\} . \tag{3.5}
\end{align*}
$$

Here the relevant function spaces are defined in Figure 3.1 on the next page.
3.1. Theorem. Let $\lambda_{s}+\mu_{s}>0, c_{5} \in\{0,1\}, c_{6}=1, J=(0, \infty), p \in(1, \infty)$, and $M>0$.

Then there exists $\tau \in(0, \infty)$ such that the solution-to-data map $(u, \pi, h) \mapsto\left(g_{v}, g_{w}, g_{h}\right), 0 \mathbb{E} \rightarrow$ ${ }_{0} \mathbb{G}_{v} \times{ }_{0} \mathbb{G}_{w} \times{ }_{0} \mathbb{G}_{h}$ of problem (3.2) is uniformly invertible with respect to $\vartheta \in \mathcal{P}_{M}$.

$$
\begin{aligned}
{ }_{0} \mathbb{E}_{u} & :=\left\{u \in{ }_{0} H_{p}^{1}\left(J ; L_{p}\left(\mathbb{R}^{n+1}\right)^{n+1}\right) \cap L_{p}\left(J ; H_{p}^{2}\left(\dot{\mathbb{R}}^{n+1}\right)^{n+1}\right): \llbracket u \rrbracket=0\right\}, \\
{ }_{0} \mathbb{E}_{v} & :={ }_{0} W_{p}^{1-1 / 2 p}\left(J ; L_{p}\left(\mathbb{R}^{n}\right)^{n}\right) \cap{ }_{0} W_{p}^{1 / 2-1 / 2 p}\left(J ; H_{p}^{2}\left(\mathbb{R}^{n}\right)^{n}\right) \cap L_{p}\left(J ; W_{p}^{3-1 / p}\left(\mathbb{R}^{n}\right)^{n}\right), \\
{ }_{0} \mathbb{E}_{w} & :={ }_{0} W_{p}^{1-1 / 2 p}\left(J ; H_{p}^{1}\left(\mathbb{R}^{n}\right)\right) \cap L_{p}\left(J ; W_{p}^{3-1 / p}\left(\mathbb{R}^{n}\right)\right), \\
{ }_{0} \mathbb{E}_{u, v, w} & :=\left\{u=(v, w) \in{ }_{0} \mathbb{E}_{u}:\left.v\right|_{y=0} \in_{0} \mathbb{E}_{v},\left.w\right|_{y=0} \in{ }_{0} \mathbb{E}_{w}\right\}, \\
\mathbb{E}_{\pi} & :=L_{p}\left(J ; \dot{H}_{p}^{1}\left(\dot{\mathbb{R}}^{n+1}\right)\right), \\
{ }_{0} \mathbb{E}_{\pi, \llbracket \pi \rrbracket} & :=\left\{\pi \in \mathbb{E}_{\pi}: \llbracket \pi \rrbracket \in_{0} \mathbb{G}_{w}\right\}, \\
{ }_{0} \mathbb{E}_{h} & :={ }_{0} W_{p}^{2-1 / 2 p}\left(J ; H_{p}^{1}\left(\mathbb{R}^{n}\right)\right) \cap{ }_{0} H_{p}^{1}\left(J ; W_{p}^{3-1 / p}\left(\mathbb{R}^{n}\right)\right) \cap L_{p}\left(J ; W_{p}^{4-1 / p}\left(\mathbb{R}^{n}\right)\right), \\
{ }_{0} \mathbb{G}_{v} & :={ }_{0} W_{p}^{1 / 2-1 / 2 p}\left(J ; L_{p}\left(\mathbb{R}^{n}\right)^{n}\right) \cap L_{p}\left(J ; W_{p}^{1-1 / p}\left(\mathbb{R}^{n}\right)^{n}\right), \\
{ }_{0} \mathbb{G}_{w} & :={ }_{0} W_{p}^{1 / 2-1 / 2 p}\left(J ; H_{p}^{1}\left(\mathbb{R}^{n}\right)\right) \cap L_{p}\left(J ; W_{p}^{2-1 / p}\left(\mathbb{R}^{n}\right)\right), \\
{ }_{0} \mathbb{G}_{h} & :={ }_{0} W_{p}^{1-1 / 2 p}\left(J ; H_{p}^{1}\left(\mathbb{R}^{n}\right)\right) \cap L_{p}\left(J ; W_{p}^{3-1 / p}\left(\mathbb{R}^{n}\right)\right) .
\end{aligned}
$$

Figure 3.1. Function spaces ${ }_{0} \mathbb{E} \ldots$ and ${ }_{0} \mathbb{G} . .$. for problem (MP).
This theorem is a central result of this thesis, as it provides the basic functional analytic framework for proving that the linear problem (3.1) has optimal regularity, and these function spaces are also appropriate for proving that problem (T) is locally well-posed. We can easily conclude the following result on bounded time intervals.
3.2. Corollary. Let $\lambda_{s}+\mu_{s}>0, c_{5} \in\{0,1\}, c_{6}=1, \tau=0, p \in(1, \infty), T_{0} \in(0, \infty)$, and $M>0$.

Then the solution-to-data map $(u, \pi, h) \mapsto\left(g_{v}, g_{w}, g_{h}\right),{ }_{0} \mathbb{E}(J, 0) \rightarrow{ }_{0} \mathbb{G}_{v}(J) \times{ }_{0} \mathbb{G}_{w}(J) \times{ }_{0} \mathbb{G}_{h}(J)$ of problem (3.2) is uniformly invertible with respect to $\vartheta \in \mathcal{P}_{M}$ and $J=(0, T)$ with $T \in\left(0, T_{0}\right]$.
Proof. As in [PSS07, p. 720] and [DK13, Remark 1.70], we consider the multiplication operator

$$
\left(\mathcal{M}_{\tau} u\right)(t):=e^{\tau t} u(t) \quad \text { for } u \in L_{1, \mathrm{loc}}\left(\mathbb{R}_{+} ; X\right), \tau \in \mathbb{R},
$$

with exponential weight $t \mapsto e^{\tau t}$. Then it is easy to verify the operator identities

$$
\mathcal{M}_{\tau}^{-1}=\mathcal{M}_{-\tau}, \quad \partial_{t} \mathcal{M}_{\tau}=\mathcal{M}_{\tau}\left(\tau+\partial_{t}\right) .
$$

Hence we have $\partial_{t}=\mathcal{M}_{\tau}\left(\tau+\partial_{t}\right) \mathcal{M}_{-\tau}$.
Theorem 3.1 yields a number $\tau \geq 0$ such that the solution-to-data map

$$
S_{\infty, \tau}:(u, \pi, h) \mapsto\left(g_{v}, g_{w}, g_{h}\right), \quad{ }_{0} \mathbb{E}\left(\mathbb{R}_{+}, \tau\right) \rightarrow_{0} \mathbb{G}_{v}\left(\mathbb{R}_{+}\right) \times{ }_{0} \mathbb{G}_{w}\left(\mathbb{R}_{+}\right) \times{ }_{0} \mathbb{G}_{h}\left(\mathbb{R}_{+}\right)
$$

of problem (3.2) is uniformly invertible with respect to $\vartheta \in \mathcal{P}_{M}$. By Lemma B. 9 on page 148, there exist linear extension operators $\mathcal{E}_{J, j}:{ }_{0} \mathbb{G}_{j}(J) \rightarrow{ }_{0} \mathbb{G}_{j}\left(\mathbb{R}_{+}\right)(j \in\{v, w, h\})$ that are uniformly bounded with respect to $T \in(0, \infty)$. The data-to-solution map for (3.2) on $J$ is therefore given by

$$
\left(g_{v}, g_{w}, g_{h}\right) \mapsto(u, \pi, h)=\left.\left(\mathcal{M}_{\tau} S_{\infty, \tau}^{-1} \mathcal{M}_{-\tau}\left(\mathcal{E}_{J, v} g_{v}, \mathcal{E}_{J, w} g_{w}, \mathcal{E}_{J, h} g_{h}\right)\right)\right|_{[0, T]},
$$

and its asserted mapping properties can be easily checked.
The proof of Theorem 3.1 is prepared in the following subsections and given on page 68. In Section 3.1.1, we apply the Fourier-Laplace transformation to problem (3.2) and express the transformed solution ( $\hat{u}, \llbracket \hat{\pi} \rrbracket, \hat{h})$ by means of Green's functions and the values of $\left(\hat{u}, \partial_{y} \hat{u}_{ \pm}, \hat{\pi}, \hat{h}\right)$ at $y=0$. The latter satisfy a linear system (3.12) for given unknowns ( $\hat{g}_{v}, \hat{g}_{w}, \hat{g}_{h}$ ) whose determinant (3.13), which we call the interface symbol of problem (3.2), does not vanish. Moreover, the system (3.12) fits into the theory of Denk and Kaip [DK13] on $N$-parabolic mixed order systems, as will be shown in Section 3.1.2. By applying their theory in Section 3.1.3, we obtain suitable function spaces on $J \times \Sigma$ such that the map $\left(\left.u\right|_{\Sigma}, \partial_{y} u_{ \pm} \mid \Sigma, \llbracket \pi \rrbracket, h\right) \mapsto\left(g_{v}, g_{w}, g_{h}\right)$ is
uniformly invertible with respect to the parameter $\vartheta \in \mathcal{P}_{M}$. In Section 3.1.4, we employ an appropriate extension technique as in [DHP03; DHP07; DK13] for proving that $(u, \pi)$ satisfy the desired interior regularity conditions. Finally, inhomogeneous bulk data $\left(f_{u}, f_{d}\right)$ are resolved in Section 3.1 .5 by using optimal regularity of elliptic transmission problems from Chapter 2 and of the Stokes problem in a half space from [DHP01].
3.1.1. The interface symbol. We first adapt the computations of Denk and Kaip [DK13, Section 4.7] and derive the linear system (3.12) for the transformed interface values of $(u, \pi, h)$. Assume that $(u, \pi, h)$ is a solution of problem (3.2), which can be transformed with the Fourier transformation $x \rightsquigarrow \xi, \nabla_{x} \rightsquigarrow i \xi$ and the Laplace transformation $t \rightsquigarrow \lambda$. The transformed functions are denoted by $\hat{u}(\lambda, \xi, y), \hat{\pi}(\lambda, \xi, y)$, and $\hat{h}(\lambda, \xi)$. For $j \in\{1,2\}$, we define

$$
\hat{u}_{j}(\lambda, \xi, y):=\hat{u}\left(\lambda, \xi,(-1)^{j} y\right), \quad \hat{\pi}_{j}(\lambda, \xi, y):=\hat{\pi}\left(\lambda, \xi,(-1)^{j} y\right) \quad \text { for } y>0 .
$$

The transformed tangential and normal velocities $\hat{v}_{j}$ and $\hat{w}_{j}$ are defined analogously and we let $\rho_{2}:=\rho_{+}, \rho_{1}:=\rho_{-}$, and so on. We consider the parabolic case $\lambda \in \Sigma_{\phi}=\{\lambda \in \mathbb{C}:|\arg \lambda|<\phi\}$ with $\phi \in(\pi / 2, \pi)$. Since $\tau+\Sigma_{\phi}$ is a subset of $\Sigma_{\phi}$ for $\tau \geq 0$, we may replace $\tau+\lambda \in \tau+\Sigma_{\phi}$ by $\lambda \in \Sigma_{\phi}$ in the following computations.

The Fourier-Laplace transformed equation of $\llbracket u \rrbracket=0$ is $\llbracket \hat{u} \rrbracket=0$, and hence $\hat{u}_{2}=\hat{u}_{1}, \hat{v}_{2}=$ $\hat{v}_{1}=: \hat{v}$, and $\hat{w}_{2}=\hat{w}_{1}=: \hat{w}$ at $y=0$. For $j \in\{1,2\}, \lambda \in \Sigma_{\phi}, \xi \in \mathbb{R}^{n}$, and $k \in\{3,4\}$, we define

$$
\omega_{j}(\lambda, \xi):=\left(\rho_{j} \mu_{j}^{-1} \lambda+|\xi|^{2}\right)^{1 / 2}, \quad c_{k}(\xi):=|\xi|^{-2} \xi^{\top} C_{k} \xi
$$

Then (3.2) is transformed to the following system.

$$
\begin{align*}
\mu_{j} \omega_{j}^{2} \hat{u}_{j}-\mu_{j} \partial_{y}^{2} \hat{u}_{j}+\left(i \xi,(-1)^{j} \partial_{y}\right)^{\top} \hat{\pi}_{j} & =0,  \tag{3.6a}\\
i \xi \cdot \hat{v}_{j}+(-1)^{j} \partial_{y} \hat{w}_{j} & =0,  \tag{3.6b}\\
\llbracket \hat{u} \rrbracket & =0,  \tag{3.6c}\\
\left(\mu_{s}|\xi|^{2}+\lambda_{s} \xi \otimes \xi\right) \hat{v}-c_{6}\left(\mu_{2} \partial_{y} \hat{v}_{2}+\mu_{1} \partial_{y} \hat{v}_{1}\right)-c_{5} \llbracket \mu \rrbracket i \xi \hat{w}-c_{1} i \xi|\xi|^{2} \hat{h} & =\hat{g}_{v},  \tag{3.6d}\\
-\left(c_{2} i \xi+2 C_{3} i \xi\right) \cdot \hat{v}-2\left(\mu_{2} \partial_{y} \hat{w}_{2}+\mu_{1} \partial_{y} \hat{w}_{1}\right)+\llbracket \hat{\pi} \rrbracket+\left(c_{\sigma}+c_{4}(\xi)\right)|\xi|^{2} \hat{h} & =\hat{g}_{w},  \tag{3.6e}\\
\lambda \hat{h}-\hat{w} & =\hat{g}_{h} . \tag{3.6f}
\end{align*}
$$

Equation (3.6a) can be eliminated with the following result on Green's functions

$$
k_{ \pm}(y, s):=k_{ \pm}(y, s ; \tau):=\left(e^{-\tau|y-s|} \pm e^{-\tau(y+s)}\right) / 2 \tau \quad \text { for } y, s \geq 0, \tau \in \mathbb{C} \backslash\{0\} .
$$

3.3. Lemma. For $\mu \in \mathbb{C}$ and $f \in C([0, \infty))$ with $\left(s \mapsto e^{-\tau s} f(s)\right) \in L_{1}(0, \infty)$, the functions

$$
v_{ \pm}(y):=\mu e^{-\tau y}-\int_{0}^{\infty} k_{ \pm}(y, s) f(s) d s \quad \text { for } y \geq 0
$$

solve the initial value problems

$$
\begin{array}{lll}
\partial_{y}^{2} v_{+}-\tau^{2} v_{+}=f, & v_{+}(0)=\mu-\frac{1}{\tau} \int_{0}^{\infty} e^{-\tau s} f(s) d s, & \partial_{y} v_{+}(0)=-\tau \mu \\
\partial_{y}^{2} v_{-}-\tau^{2} v_{-}=f, & v_{-}(0)=\mu, & \partial_{y} v_{-}(0)=-\tau \mu-\int_{0}^{\infty} e^{-\tau s} f(s) d s
\end{array}
$$

Proof. These assertions can be verified easily.
Consequently, equation (3.6a) can be eliminated when we represent the transformed functions $(\hat{v}, \hat{w}, \hat{\pi})$ in terms of unknown transformed functions $\hat{p}_{j}(\lambda, \xi), \hat{\Phi}_{v}^{j}(\lambda, \xi)$, and $\hat{\Phi}_{w}^{j}(\lambda, \xi)$ as
follows:

$$
\begin{align*}
& \hat{\pi}_{j}(\lambda, \xi, y):=\hat{p}_{j}(\lambda, \xi) e^{-|\xi| y}  \tag{3.7a}\\
& \hat{v}_{j}(\lambda, \xi, y):=\hat{\Phi}_{v}^{j}(\lambda, \xi) e^{-\omega_{j}(\lambda, \xi) y}-\int_{0}^{\infty} k_{-}\left(y, s ; \omega_{j}(\lambda, \xi)\right) \frac{i \xi \hat{\pi}_{j}(\lambda, \xi, s)}{\mu_{j}} d s  \tag{3.7b}\\
& \hat{w}_{j}(\lambda, \xi, y):=\hat{\Phi}_{w}^{j}(\lambda, \xi) e^{-\omega_{j}(\lambda, \xi) y}-\int_{0}^{\infty} k_{+}\left(y, s ; \omega_{j}(\lambda, \xi)\right) \frac{(-1)^{j} \partial_{s} \hat{\pi}_{j}(\lambda, \xi, s)}{\mu_{j}} d s \tag{3.7c}
\end{align*}
$$

These functions satisfy the following interface conditions.

$$
\begin{align*}
\left.\hat{\pi}_{j}\right|_{y=0} & =\hat{p}_{j}, & \llbracket \hat{\pi} \rrbracket & =\hat{p}_{2}-\hat{p}_{1}  \tag{3.8a}\\
\left.\hat{v}_{j}\right|_{y=0} & =\hat{\Phi}_{v}^{1}=\hat{\Phi}_{v}^{2}=: \hat{\Phi}_{v}, & \left.\partial_{y} \hat{v}_{j}\right|_{y=0} & =-\omega_{j} \hat{\Phi}_{v}-\frac{i \xi \hat{p}_{j}}{\mu_{j}\left(\omega_{j}+|\xi|\right)} \\
\left.\hat{w}_{j}\right|_{y=0} & =\hat{\Phi}_{w}^{j}+\frac{(-1)^{j}|\xi| \hat{p}_{j}}{\mu_{j} \omega_{j}\left(\omega_{j}+|\xi|\right)}, & \left.\partial_{y} \hat{w}_{j}\right|_{y=0} & =-\omega_{j} \hat{\Phi}_{w}^{j} \tag{3.8b}
\end{align*}
$$

We employ the abbreviations $\alpha_{j}:=\mu_{j} \omega_{j}\left(\omega_{j}+|\xi|\right)$ and $\Omega_{+}:=\alpha_{1}+\alpha_{2}$. Then, with $\hat{p}_{1}=\hat{p}_{2}-\llbracket \hat{\pi} \rrbracket$, $\left.\hat{w}_{2}\right|_{y=0}=\left.\hat{w}_{1}\right|_{y=0}$, and (3.8c), we represent $\hat{p}_{1}$ and $\hat{p}_{2}$ as

$$
\begin{equation*}
\hat{p}_{j}=\frac{\alpha_{1} \alpha_{2}}{|\xi| \Omega_{+}}\left(\hat{\Phi}_{w}^{1}-\hat{\Phi}_{w}^{2}\right)+\frac{(-1)^{j} \alpha_{j}}{\Omega_{+}} \llbracket \hat{\pi} \rrbracket . \tag{3.9}
\end{equation*}
$$

Hence the transformed functions $\hat{w}_{j}$ and $\partial_{y} \hat{v}_{j}$ are given by

$$
\begin{align*}
\hat{w}_{1} & =\hat{w}_{2}=\frac{\alpha_{2}}{\Omega_{+}} \hat{\Phi}_{w}^{2}+\frac{\alpha_{1}}{\Omega_{+}} \hat{\Phi}_{w}^{1}+\frac{|\xi|}{\Omega_{+}} \llbracket \hat{\pi} \rrbracket  \tag{3.10a}\\
\partial_{y} \hat{v}_{1} & =\frac{\alpha_{2} \omega_{1} i \xi}{\Omega_{+}|\xi|} \hat{\Phi}_{w}^{2}-\frac{\alpha_{2} \omega_{1} i \xi}{\Omega_{+}|\xi|} \hat{\Phi}_{w}^{1}-\omega_{1} \hat{\Phi}_{v}+\frac{\omega_{1} i \xi \llbracket \hat{\pi} \rrbracket}{\Omega_{+}} \\
\partial_{y} \hat{v}_{2} & =\frac{\alpha_{1} \omega_{2} i \xi}{\Omega_{+}|\xi|} \hat{\Phi}_{w}^{2}-\frac{\alpha_{1} \omega_{2} i \xi}{\Omega_{+}|\xi|} \hat{\Phi}_{w}^{1}-\omega_{2} \hat{\Phi}_{v}-\frac{\omega_{2} i \xi \llbracket \hat{\pi} \rrbracket}{\Omega_{+}}
\end{align*}
$$

It remains to formulate a linear system for the unknowns $\hat{\Phi}_{v}, \hat{\Phi}_{w}^{2}, \hat{\Phi}_{w}^{1}, \hat{h}$, and $\llbracket \hat{\pi} \rrbracket$. We abbreviate

$$
\begin{aligned}
\Omega^{\prime} & :=c_{6} \mu_{1} \omega_{1}+c_{6} \mu_{2} \omega_{2}+\mu_{s}|\xi|^{2}, \\
L_{w}^{1} & :=c_{6} \mu_{1} \omega_{1} \alpha_{2}+c_{6} \mu_{2} \omega_{2} \alpha_{1}-c_{5} \llbracket \mu \rrbracket|\xi| \alpha_{1}, \\
L_{w}^{2} & :=-c_{6} \mu_{1} \omega_{1} \alpha_{2}-c_{6} \mu_{2} \omega_{2} \alpha_{1}-c_{5} \llbracket \mu \rrbracket|\xi| \alpha_{2}, \\
L_{q} & :=c_{6} \mu_{2} \omega_{2}-c_{6} \mu_{1} \omega_{1}-c_{5} \llbracket \mu \rrbracket|\xi| .
\end{aligned}
$$

Then equations (3.6d) to (3.6f), (3.9) and (3.10) yield

$$
\begin{equation*}
\left(\Omega^{\prime} \mathrm{id}_{n}+\lambda_{s} \xi \otimes \xi\right) \hat{\Phi}_{v}+\frac{i \xi L_{w}^{2}}{|\xi| \Omega_{+}} \hat{\Phi}_{w}^{2}+\frac{i \xi L_{w}^{1}}{|\xi| \Omega_{+}} \hat{\Phi}_{w}^{1}+\frac{i \xi L_{q}}{\Omega_{+}} \llbracket \hat{\pi} \rrbracket-c_{1} i \xi|\xi|^{2} \hat{h}=\hat{g}_{v} \tag{3.11a}
\end{equation*}
$$

$$
\begin{equation*}
-\left(c_{2} i \xi+2 C_{3} i \xi\right) \cdot \hat{\Phi}_{v}+2 \mu_{2} \omega_{2} \hat{\Phi}_{w}^{2}+2 \mu_{1} \omega_{1} \hat{\Phi}_{w}^{1}+\llbracket \hat{\pi} \rrbracket+\left(c_{\sigma}+c_{4}\right)|\xi|^{2} \hat{h}=\hat{g}_{w} \tag{3.11b}
\end{equation*}
$$

$$
\begin{equation*}
\lambda \hat{h}-\frac{\alpha_{2}}{\Omega_{+}} \hat{\Phi}_{w}^{2}-\frac{\alpha_{1}}{\Omega_{+}} \hat{\Phi}_{w}^{1}-\frac{|\xi|}{\Omega_{+}} \llbracket \hat{\pi} \rrbracket=\hat{g}_{h} \tag{3.11c}
\end{equation*}
$$

Consequently, system (3.6) becomes

$$
\underbrace{\left[\begin{array}{ccccc}
\Omega^{\prime} I_{n}-\lambda_{s} i \xi \otimes i \xi & \frac{L_{w}^{2} i \xi}{\Omega_{+}|\xi|} & \frac{L_{w}^{1} i \xi}{\Omega_{+}|\xi|} & -c_{1} i \xi|\xi|^{2} & L_{q} \frac{i \xi}{\Omega_{+}}  \tag{3.12}\\
i \xi^{\top} & -\omega_{2} & 0 & 0 & 0 \\
i \xi^{\top} & 0 & \omega_{1} & 0 & 0 \\
0 & -\frac{\alpha_{2}}{\Omega_{+}} & -\frac{\alpha_{1}}{\Omega_{+}} & \lambda & -\frac{|\xi|}{\Omega_{+}} \\
-c_{2} i \xi^{\top}-2 i \xi^{\top} C_{3} & 2 \mu_{2} \omega_{2} & 2 \mu_{1} \omega_{1} & \left(c_{\sigma}+c_{4}\right)|\xi|^{2} & 1
\end{array}\right]}_{\hat{\mathcal{L}}(\lambda, \xi)}\left[\begin{array}{c}
\hat{\Phi}_{v} \\
\hat{\Phi}_{w}^{2} \\
\hat{\Phi}_{w}^{1} \\
\hat{h} \\
\llbracket \hat{\pi} \rrbracket
\end{array}\right]=\left[\begin{array}{c}
\hat{g}_{v} \\
0 \\
0 \\
\hat{g}_{h} \\
\hat{g}_{w}
\end{array}\right] .
$$

In order to compute the interface symbol $\operatorname{det} \mathcal{L}(\lambda, \xi)$, we
(i) subtract row $n+1$ from row $n+2$,
(ii) add $c_{2} \cdot($ row $n+1$ ) to row $n+4$,
(iii) add $\lambda_{s} i \xi \otimes($ row $n+1)$ to rows $1, \ldots, n$,
(iv) add $-\Omega^{\prime-1} i \xi^{\top} \cdot($ rows $1, \ldots, n)$ to row $n+1$, and
(v) add $2 \Omega^{\prime-1} i \xi^{\top} C_{3} \cdot($ rows $1, \ldots, n$ ) to row $n+4$.

In this way we calculcate
$\operatorname{det} \hat{\mathcal{L}}(\lambda, \xi)$

$$
\left.\begin{array}{l}
=\operatorname{det}\left[\begin{array}{ccccc}
\Omega^{\prime} I_{n} & * & * & * & * \\
0 & -\omega_{2}+\frac{|\xi|^{2}}{\Omega^{\prime}}\left(\frac{L_{w}^{2}}{\Omega_{+}|\xi|}-\lambda_{s} \omega_{2}\right) & \frac{|\xi|^{2}}{\Omega^{\prime}} \frac{L_{w}^{1}}{\Omega_{+}|\xi|} & -c_{1} \frac{|\xi|^{4}}{\Omega^{\prime}} & \frac{L_{q}}{\Omega^{\prime} \Omega_{+}}|\xi|^{2} \\
0 & \omega_{1} & 0 & 0 \\
0 & -\frac{\alpha_{2}}{\Omega_{+}} & -\frac{\alpha_{1}}{\Omega_{+}} & \lambda & -\frac{|\xi|}{\Omega_{+}} \\
0 & -\frac{2|\xi|^{2} c_{3}}{\Omega^{\prime}}\left(\frac{L_{2}^{2}}{\Omega_{+}+\xi \mid}-\lambda_{s} \omega_{2}\right) & -\frac{2|\xi|^{2} c_{3} \omega_{1}}{\Omega^{\prime}} \frac{L_{w}^{1}}{\Omega_{+}|\xi|} & \left(c_{\sigma}+c_{4}\right)|\xi|^{2} & -\frac{2 c_{1} c_{3}|\xi|^{4}}{\Omega^{\prime}}
\end{array}\right] 1-\frac{2 c_{3} L_{q}|\xi|^{2}}{\Omega^{\prime} \Omega_{+}}
\end{array}\right] .
$$

Here an asterisk $*$ denotes a non-specified entry. The remaining ( $4 \times 4$ )-determinant can be calculated with the software Maxima [Max] and we refer to page 173 in Appendix B. 5 for the source code. Therefore the interface symbol can be written as

$$
\begin{equation*}
\operatorname{det} \hat{\mathcal{L}}(\lambda, \xi)=-\omega_{1}(\lambda, \xi) \omega_{2}(\lambda, \xi) \Omega_{+}(\lambda, \xi)^{-1} \Omega^{\prime}(\lambda, \xi)^{n-1} P(\lambda, \xi) \tag{3.13}
\end{equation*}
$$

where the symbol $P(\lambda, \xi)$ is defined as follows. Define the mean value $\langle\psi\rangle:=\left(\psi_{1}+\psi_{2}\right) / 2$, the jump $\llbracket \psi \rrbracket:=\psi_{2}-\psi_{1}$, and let

$$
\begin{aligned}
d(\xi): & c_{\sigma}+2 \vartheta_{w} c_{3}(\xi)+c_{4}(\xi)+\vartheta_{w} c_{2} \\
= & \sigma+\left(\lambda_{s}-\mu_{s}\right)\left(\vartheta_{d v}-2 \vartheta_{H} \vartheta_{w}\right) \\
& -2 \vartheta_{w}|\xi|^{-2} i \xi^{\top}\left[\mu_{s} \vartheta_{L}\right] i \xi-|\xi|^{-2} i \xi^{\top}\left[2 \mu_{s}\left(\vartheta_{D v}-2 \vartheta_{w} \vartheta_{L}\right)\right] i \xi+\left(\lambda_{s}-\mu_{s}\right) \vartheta_{w} \vartheta_{H} \\
= & \sigma+\left(\lambda_{s}-\mu_{s}\right)\left(\vartheta_{d v}-\vartheta_{H} \vartheta_{w}\right)+|\xi|^{-2} \xi^{\top}\left[\vartheta_{w} \vartheta_{L}+2 \vartheta_{w} 2 \mu_{s}\left(\vartheta_{D v}-\vartheta_{w} \vartheta_{L}\right)\right] \xi \\
= & \sigma+\left(\lambda_{s}-\mu_{s}\right) \vartheta_{d u}+2 \mu_{s}|\xi|^{-2} \xi^{\top}\left[\vartheta_{D u}\right] \xi .
\end{aligned}
$$

Then, with $\beta_{s}:=\lambda_{s}+\mu_{s}$, the symbol $P(\lambda, \xi)$ in (3.13) is given by

$$
\begin{aligned}
P(\lambda, \xi)= & \underline{\beta_{s} d(\xi)|\xi|^{5}}+2\langle\mu \mu\rangle\left(c_{6} c_{\sigma}+c_{4} c_{6}\right)|\xi|^{4} \\
& +\frac{c_{1}(\llbracket \mu \omega \rrbracket-\llbracket \mu \rrbracket|\xi|)|\xi|^{4}}{} \\
& +\left(2 c_{6} c_{\sigma}\left\langle\langle\mu \omega\rangle+\left(2 \vartheta_{3}+c_{2}\right) c_{5} \llbracket \mu \rrbracket \lambda-\llbracket \mu \rrbracket^{2} c_{5} \lambda+2 c_{4} c_{6}\langle\mu \omega\rangle\right\rangle\right)|\xi|^{3} \\
& \left.+\left(4 c_{6}\langle\mu \omega\rangle\right)^{2}+2 c_{6} \mu_{1} \mu_{2}\left\langle\left\langle\omega^{2}\right\rangle+4 c_{6}\left\langle\mu \mu^{2} \omega\right\rangle\right| \xi\left|+\left(c_{5} \llbracket \mu \rrbracket-c_{2} c_{6}-2 c_{3} c_{6}\right) \llbracket \mu \omega \rrbracket\right| \xi \mid\right) \lambda|\xi| \\
+ & \underline{2 \beta_{s}\left(\left\langle\mu \omega^{2}\right\rangle+\langle\mu \mu \omega\rangle|\xi|\right) \lambda|\xi|^{2}} \\
+ & \underline{4 c_{6}\left\langle\mu \mu \omega^{2}\right\rangle\langle\mu \omega \omega\rangle \lambda .}
\end{aligned}
$$

Here the underlined terms are the principal parts as we will see in the next section.
3.1.2. Invertibility of the interface symbol. Our goal is to show that $\hat{\mathcal{L}}$ is an $N$-parabolic mixed-order system in the sense of Definition B. 77 on page 168. Moreover, since our localization procedure will require uniform invertibility with respect to the reference velocity $u_{*}=v_{*}+$ $w_{*} \nu_{\Sigma}$ and $L_{\Sigma}$, we will also study the dependence on the related parameters $\vartheta=\left(\vartheta_{w}, \vartheta_{L}, \vartheta_{D v}\right) \in$ $\mathcal{P}_{M}$. First, we show that the interface symbol det $\hat{\mathcal{L}}$ of problem (3.2) is an $N$-parabolic symbol. To this end, we replace

$$
i \xi \rightsquigarrow z \in{\overline{B \Sigma_{\delta}}}_{\delta}^{n}, \quad|\xi| \rightsquigarrow|z|_{-}=\sqrt{-z \cdot z},
$$

and we define the complex $(n+4) \times(n+4)$-matrix

$$
\hat{\mathcal{L}}(\lambda, z ; \vartheta):=\left[\begin{array}{ccccc}
\Omega^{\prime} I_{n}-\lambda_{S} z \otimes z & \frac{L_{w}^{2} z}{\Omega_{+}|z|_{-}} & \frac{L_{w}^{1} z}{\Omega_{+}|z|_{-}} & -c_{1}(\vartheta) z|z|_{-}^{2} & L_{q} \frac{z}{\Omega_{+}}  \tag{3.14}\\
z^{\top} & -\omega_{2} & 0 & 0 & 0 \\
z^{\top} & 0 & \omega_{1} & 0 & 0 \\
0 & -\frac{\alpha_{2}}{\Omega_{+}} & -\frac{\alpha_{1}}{\Omega_{+}} & \lambda & -\frac{|z|_{-}}{\Omega_{+}} \\
-c_{2}(\vartheta) z^{\top}-2 z^{\top} C_{3}(\vartheta) & 2 \mu_{2} \omega_{2} & 2 \mu_{1} \omega_{1} & c_{\sigma}(\vartheta)|z|_{-}^{2}-z^{\top} C_{4}(\vartheta) z & 1
\end{array}\right] .
$$

We replace the functions $d(\xi)$ and $P(\lambda, \xi)$ as

$$
\begin{aligned}
d(z ; \vartheta):= & \sigma+\left(\lambda_{s}-\mu_{s}\right) \operatorname{tr} \vartheta_{D u}+2 \mu_{s}|z|_{-}^{-2} i z^{\top}\left[\vartheta_{D u}\right] i z \quad \text { with } c_{j}(z ; \vartheta):=-|z|_{-}^{-2} z^{\top} C_{j}(\vartheta) z, \\
P(\lambda, z ; \vartheta):= & \beta_{s} d(z ; \vartheta)|z|_{-}^{5}+2\langle\mu \mu\rangle\left(c_{6} c_{\sigma}(\vartheta)+c_{4}(z ; \vartheta) c_{6}\right)|z|_{-}^{4} \\
& +c_{1}(\vartheta)\left(\llbracket \mu \omega \rrbracket-\llbracket \mu \rrbracket|z|_{-}\right)|z|_{-}^{4} \\
& \left.+\left(2 c_{6} c_{\sigma}(\vartheta)\langle\mu \omega\rangle\right\rangle+\left(2 c_{3}(z ; \vartheta)+c_{2}(\vartheta)\right) c_{5} \llbracket \mu \rrbracket \lambda-\llbracket \mu \rrbracket^{2} c_{5} \lambda+2 c_{4}(z ; \vartheta) c_{6}\langle\mu \omega\rangle\right)|z|_{-}^{3} \\
& \left.+\left(4 c_{6}\langle\mu \omega\rangle\right\rangle^{2}+2 c_{6} \mu_{1} \mu_{2}\left\langle\omega^{2}\right\rangle+4 c_{6}\left\langle\mu^{2} \omega\right\rangle|z|-\right) \lambda|z|_{-} \\
& +\left(c_{5} \llbracket \mu \rrbracket-c_{2}(\vartheta) c_{6}-2 c_{3}(z ; \vartheta) c_{6}\right) \llbracket \mu \omega \rrbracket \lambda|z|_{-}^{2} \\
+ & \left.\left.2 \beta_{s}\left(《 \mu \omega^{2}\right\rangle+\langle\mu \omega\rangle\right\rangle|z|_{-}\right) \lambda|z|_{-}^{2} \\
= & \left.4 c_{6}\left\langle\mu \omega^{2}\right\rangle\langle\mu \omega\rangle\right\rangle .
\end{aligned}
$$

It is straightforward to check that $P$ belongs to the symbol class in Definition B. 72 on page 166.
Next, we employ the $\gamma$-orders and $\gamma$-principal parts of the symbols $\omega_{j}, \Omega^{\prime}, \Omega_{+}$, and $P$, which are defined in Definition B. 73 and given in Figure 3.2 on the following page. Due to Theorem B.75, it is sufficient to show that the principal parts of $\omega_{j}, \Omega_{+}, \Omega^{\prime}$, and $P$ do not vanish, and therefore the function $d(z ; \vartheta)$ should not vanish. Let us derive a condition on the parameter tuple $\vartheta \in \mathcal{P}_{M}$ which ensures that

$$
\begin{equation*}
\text { there is } \delta>0 \text { such that } \operatorname{Re} d(z ; \vartheta)>0 \text { for all } z \in B \Sigma_{\delta}^{n} \text {. } \tag{3.15}
\end{equation*}
$$

$$
\begin{aligned}
& d_{\gamma}\left(\omega_{j}\right)= \begin{cases}1 & : \gamma \in(0,2], \\
\gamma / 2 & : \gamma \in[2, \infty] .\end{cases} \\
& d_{\gamma}\left(\Omega_{+}\right)= \begin{cases}2 & : \gamma \in(0,2], \\
\gamma & : \gamma \in[2, \infty] .\end{cases} \\
& d_{\gamma}\left(\Omega^{\prime}\right)= \begin{cases}2 & : \gamma \in(0, \infty], c_{6}=0 \\
2 & : \gamma \in(0,4], c_{6}=1 \\
\gamma / 2 & : \gamma \in[4, \infty], c_{6}=1\end{cases} \\
& d_{\gamma}(P)= \begin{cases}5 & : \gamma \in(0,1] \\
4+\gamma & : \gamma \in[0,2] \\
2+2 \gamma & : \gamma \in[2, \infty], c_{6}=0, \\
2+2 \gamma & : \gamma \in[2,4], c_{6}=1, \\
5 \gamma / 2 & : \gamma \in[4, \infty], c_{6}=1 .\end{cases} \\
& \pi_{\gamma} \omega_{j}(\lambda, z)= \begin{cases}|z|_{-} & : \gamma \in(0,2) \\
\left(\rho_{j} \mu_{j}^{-1} \lambda+|z|_{-}^{2}\right)^{1 / 2} & : \gamma=2, \\
\left(\rho_{j} \mu_{j}^{-1}\right)^{1 / 2} \lambda^{1 / 2} & : \gamma \in(2, \infty]\end{cases} \\
& \pi_{\gamma} \Omega_{+}(\lambda, z)= \begin{cases}4\langle\mu\rangle|z|_{-}^{2} & : \gamma \in(0,2), \\
\Omega_{+}(\lambda, z) & : \gamma=2, \\
2\langle\rho\rangle \lambda & : \gamma \in(2, \infty] .\end{cases} \\
& \pi_{\gamma} \Omega^{\prime}(\lambda, z)= \begin{cases}\mu_{s}|z|_{-}^{2} & : \gamma \in(0, \infty], c_{6}=0, \\
\mu_{s}|z|_{-}^{2} & : \gamma \in(0,4), c_{6}=1, \\
2\langle\sqrt{\rho \mu}\rangle \lambda^{1 / 2}+\mu_{s}|z|_{-}^{2} & : \gamma=4, c_{6}=1, \\
2\langle\sqrt{\rho \mu}\rangle \lambda^{1 / 2} & : \gamma \in(4, \infty], c_{6}=1 .\end{cases} \\
& \pi_{\gamma} P(\lambda, z ; \vartheta)= \begin{cases}\beta_{s} d(z ; \vartheta)|z|_{-}^{5} & : \gamma \in(0,1), \\
\beta_{s} d(z ; \vartheta)|z|_{-}^{5}+4 \beta_{s}\langle\mu \mu\rangle \lambda|z|_{-}^{4} & : \gamma=1, \\
4 \beta_{s}\langle\mu \mu\rangle \lambda|z|_{-}^{4} & : \gamma \in(1,2), \\
2 \beta_{s}\left(\left\langle\mu \omega^{2}\right\rangle+\langle\mu \omega\rangle|z|_{-}\right) \lambda|z|_{-}^{2} & : \gamma=2, \\
2 \beta_{s}\langle\rho\rangle \lambda^{2}|z|_{-}^{2} & : \gamma \in(2,4), \\
\left.\left.2 \beta_{s}\langle\rho\rangle\right\rangle \lambda^{2}|z|_{-}^{2}+4 c_{6}\langle\rho\rangle\right\rangle\langle\sqrt{\rho \mu}\rangle \lambda^{5 / 2} & : \gamma=4, \\
\left.2 \beta_{s}\left\langle\langle\rho\rangle \lambda^{2}\right| z\right|_{-} ^{2} & : \gamma \in(4, \infty], c_{6}=0, \\
4\langle\rho\rangle\langle\sqrt{\rho \mu}\rangle \lambda^{5 / 2} & : \gamma \in(4, \infty], c_{6}=1 .\end{cases}
\end{aligned}
$$

FIgURE 3.2. The $\gamma$-orders and $\gamma$-principal parts of the symbols $\omega_{j}, \Omega_{+}, \Omega^{\prime}$, and $P$.
It is shown in Lemma B. 55 that an estimate $C^{-1}|z| \leq\left||z|_{-}\right| \leq C|z|$ applies for $z \in \overline{B \Sigma}_{\delta}^{n} \backslash\{0\}$ and $\delta \in(0, \pi / 4)$. Moreover, Lemma B. 55 yields the following estimates for $j \in\{3,4\}$.
$\left|c_{j}(z ; \vartheta)\right| \leq n^{1 / 2}\left|C_{j}(\vartheta)\right|, \quad\left|\operatorname{Im} c_{j}(z ; \vartheta)\right| \leq \sin (4 \delta)\left|c_{j}(z ; \vartheta)\right| \quad$ for $z \in \overline{B \Sigma}_{\delta}^{n} \backslash\{0\}, \delta \in(0, \pi / 8]$.
Hence a sufficient condition for (3.15) is

$$
\begin{equation*}
d_{0}\left(\vartheta_{D u}\right)=\sigma+\left(\lambda_{s}-\mu_{s}\right) \operatorname{tr} \vartheta_{D u}+2 \mu_{s} \min _{\xi \in \mathbb{R}^{n} \backslash\{0\}}|\xi|^{-2} \xi^{\top}\left[\vartheta_{D u}\right] \xi>0 . \tag{3.16}
\end{equation*}
$$

Indeed, suppose that $d_{0}\left(\vartheta_{D u}\right) \geq 1 / M$ and $\left|\vartheta_{D u}\right| \leq M$ for some $M>0$. Then

$$
\operatorname{Re} d(z ; \vartheta) \geq d_{0}\left(\vartheta_{D u}\right)+2 \mu_{s}\left(\min _{z \in \overline{B \Sigma_{\delta}^{n} \backslash\{0\}}} \operatorname{Re} \frac{i z^{\top}\left[\vartheta_{D u}\right] i z}{|z|_{-}^{2}}-\min _{\xi \in \mathbb{R}^{n} \backslash\{0\}} \frac{\xi^{\top}\left[\vartheta_{D u}\right] \xi}{|\xi|^{2}}\right) \rightarrow d_{0}\left(\vartheta_{D u}\right)
$$

as $\delta \rightarrow 0$. Hence there exists $\delta=\delta(R) \in(0, \pi / 8)$ such that $\operatorname{Re} d(z ; \vartheta) \geq 1 /(2 M)$ for all $z \in B \Sigma_{\delta}^{n}$. In view of the inclusion $i \mathbb{R}^{n} \subset \overline{B \Sigma}_{\delta}^{n}$, we see that condition (3.16) is also necessary for (3.15).
3.4. Lemma. Let $\rho_{j}, \mu_{j}, \sigma$, and $\mu_{s}$ be positive constants and let $\lambda_{s}$ be a real number.
(i) If $\beta_{s}=\lambda_{s}+\mu_{s}>0$, then for given $M>0$ and $\phi \in(\pi / 2, \pi)$ there exists $\delta \in(0, \pi / 8]$ such that $P: \bar{\Sigma}_{\phi} \times \overline{B \Sigma_{\delta}^{n}} \times \mathcal{P}_{M} \rightarrow \mathbb{C}$ is $N$-parabolic.
(ii) Conversely, if $P(\cdot, \cdot ; \vartheta): \bar{\Sigma}_{\phi} \times \overline{B \Sigma}_{\delta}^{n} \rightarrow \mathbb{C}$ is $N$-parabolic for some $\phi \in(\pi / 2, \pi), \delta \in(0, \pi / 8]$, and $\vartheta \in \mathbb{R} \times \mathbb{R}^{n \times n} \times \mathbb{R}^{n \times n}$, then $\lambda_{s}+\mu_{s}>0$ and $\vartheta$ satisfies (3.16).

Proof. (i) In view of Theorem B.75, it is sufficient to show that the principal parts of the symbol $P$ do not vanish, in the sense that

$$
\pi_{\gamma} P(\lambda, z ; \vartheta) \neq 0 \quad \text { for all } \gamma \in(0, \infty], \lambda \in \bar{\Sigma}_{\phi} \backslash\{0\}, z \in \overline{B \Sigma}_{\delta}^{n} \backslash\{0\}, \vartheta \in \mathcal{P}_{M}
$$

First, we choose $\delta(M) \in(0, \pi / 8]$ such that $\operatorname{Re} d(z ; \vartheta) \geq 1 /(2 M)$ for $z \in \overline{B \Sigma}_{\delta}^{n}$ with $\delta \in(0, \delta(M)]$. Then Lemma B. 55 implies that $\pi_{\gamma} P(\lambda, z ; \vartheta)$ does not vanish for all $\gamma \in(0,1)$. Next, let $\gamma \in$ $[1,2)$. Since $\arg \left(d(z ; \vartheta)|z|_{-}\right) \leq 5 \delta$, there exists $\delta_{1}(M) \in(0, \delta(M)]$ such that $\left.|d(z ; \vartheta)| z\right|_{-} \mid \geq$ $n^{-1 / 4}|z| /(2 M)$ for all $\delta \in\left(0, \delta_{1}(M)\right]$ and $z \in \overline{B \Sigma}_{\delta}^{n}$. Hence for some $\delta=\delta(M, \phi) \leq \delta_{1}(M)$ with $5 \delta+\phi<\pi$ the number $\beta_{s} d(z ; \vartheta)|z|_{-}+4 \beta_{s}\langle\mu\rangle \lambda \lambda$ belongs to $\bar{\Sigma}_{\phi} \backslash\{0\}$, which implies that $\pi_{\gamma} P(\lambda, z ; \vartheta)$ does not vanish for $\gamma \in[1,2)$. In the case $\gamma=2$, we write

$$
\begin{equation*}
\pi_{2} P(\lambda, z ; \vartheta)=2 \beta_{s}\langle\mu \omega\rangle\left(\frac{\left\langle\mu \omega^{2}\right\rangle / \lambda^{1 / 2}}{\langle\mu \omega\rangle / \lambda^{1 / 2}}+|z|_{-}\right) \lambda|z|_{-}^{2} . \tag{3.17}
\end{equation*}
$$

Recall that $\omega_{j}(\lambda, z)^{2}=\rho_{j} \mu_{j}^{-1} \lambda+|z|_{-}^{2}$ belongs to $\bar{\Sigma}_{\phi} \backslash\{0\}$. Hence we obtain $\mid \arg \left(\left\langle\left\langle\mu \omega^{2}\right\rangle / \lambda^{1 / 2}\right) \mid \leq\right.$ $\phi / 2+2 \delta$ and $\left|\arg \left(\langle\mu \omega\rangle / \lambda^{1 / 2}\right)\right| \leq \phi / 2+\delta$ and therefore $\left.\mid \arg \left(\left\langle\mu \omega^{2}\right\rangle\right\rangle /\langle\mu \omega\rangle\right) \mid \leq \phi+3 \delta$. By choosing $\delta<(\pi-\phi) / 4$, it follows that $\left.\left\langle\left\langle\mu \omega^{2}\right\rangle\right\rangle /\langle\mu \omega\rangle\right\rangle+|z|_{-} \neq 0$ and then $\pi_{2} P(\lambda, z ; \vartheta) \neq 0$. The remaining cases $\gamma \in(2, \infty]$ can be treated similary. Therefore $P$ has non-vanishing principal parts and hence, by Theorem B.75, it is $N$-parabolic.
(ii) To prove the converse assertion, let $P(\cdot, \cdot ; \vartheta)$ be $N$-parabolic. Then $\pi_{\gamma} P(\cdot, \cdot ; \vartheta)$ does not vanish for all $\gamma \in(0,1)$ and we conclude that $\beta_{s} \neq 0$ and $d(i \xi ; \vartheta) \neq 0$ for all $\xi \in \mathbb{R}^{n} \backslash\{0\}$. Next,

$$
\left.\left.\left.\pi_{4} P(\lambda, i \xi ; \vartheta)=\left(2 \beta_{s}\langle\rho\rangle\right\rangle|\xi|^{2}+4\langle\rho\rangle\right\rangle\langle\rho \mu\rangle\right\rangle \lambda^{1 / 2}\right) \lambda^{2} \neq 0 \quad \text { for all } \lambda>0, \xi \in \mathbb{R}^{n} \backslash\{0\} .
$$

This yields $\beta_{s}>0$ by using $\left.\langle\rho\rangle\right\rangle 0$ and $\left.\langle\rho \mu\rangle\right\rangle>0$. Finally,

$$
\left.\pi_{1} P(\lambda, i \xi ; \vartheta)=\left(\left(\lambda_{s}+\mu_{s}\right) d(i \xi ; \vartheta)|\xi|+2 \beta_{s}\langle\mu\rangle\right\rangle \lambda\right)|\xi|^{4} \neq 0 \quad \text { for all } \lambda>0, \xi \in \mathbb{R}^{n} \backslash\{0\},
$$

and this implies $d(i \xi ; \vartheta)>0$ for all $\xi \in \mathbb{R}^{n} \backslash\{0\}$, since $\beta_{s}>0$ and $\langle\mu\rangle>0$. This in turn yields $d_{0}\left(\vartheta_{D u}\right)=\min \left\{d(i \xi ; \vartheta): \xi \in \mathbb{R}^{n},|\xi|=1\right\}>0$.
3.5. Corollary. Let $\rho_{j}, \mu_{j}, \sigma$, and $\mu_{s}$ be positive constants and let $\lambda_{s}$ be a real number.
(i) If $\lambda_{s}+\mu_{s}>0$, then for given $M>0$ and $\phi \in(\pi / 2, \pi)$ there exists $\delta \in(0, \pi / 8]$ such that $\operatorname{det} \hat{\mathcal{L}}: \bar{\Sigma}_{\phi} \times \overline{B \Sigma_{\delta}^{n}} \times \mathcal{P}_{M} \rightarrow \mathbb{C}$ is $N$-parabolic.
 and $\vartheta \in \mathbb{R} \times \mathbb{R}^{n \times n} \times \mathbb{R}^{n \times n}$, then $\lambda_{s}+\mu_{s}>0$ and $\vartheta$ satisfies (3.16).
Proof. (i) It is easy to check that $\omega_{1} \omega_{2} \Omega_{+}^{-1}$ satisfies the homogeneity property

$$
\left(\omega_{1} \omega_{2} \Omega_{+}^{-1}\right)\left(\eta^{2} \lambda, \eta z\right)=\left(\omega_{1} \omega_{2} \Omega_{+}^{-1}\right)(\lambda, z) \quad \text { for all } \eta>0,
$$

and therefore belongs to the symbol class $S_{N}\left(\bar{\Sigma}_{\phi} \times \overline{B \Sigma_{\delta}}\right)$ with $\phi \in(\pi / 2, \pi)$ and $\delta \in(0, \pi / 8)$. We further have $\Omega^{\prime}=c_{6} \mu_{1} \omega_{1}+c_{6} \mu_{2} \omega_{2}+\mu_{s}|z|_{-}^{2} \in S_{N}\left(\bar{\Sigma}_{\phi} \times \overline{B \Sigma}_{\delta}^{n}\right)$ and $P \in S_{N}\left(\bar{\Sigma}_{\phi} \times \overline{B \Sigma}_{\delta}^{n} \times \mathcal{P}_{M}\right)$ if $\phi \in(\pi / 2, \pi)$ and if $\delta=\delta(M, \phi)$ is chosen as in Lemma 3.4. These symbols have strictly positive order functions and therefore [DK13, Lemma 3.33] yields

$$
\operatorname{det} \hat{\mathcal{L}}=-\omega_{1} \omega_{2} \Omega_{+}^{-1} \cdot \Omega^{\prime n-1} \cdot P \in S_{N}\left(\bar{\Sigma}_{\phi} \times \overline{B \Sigma}_{\delta}^{n} \times \mathcal{P}_{M}\right)
$$

(ii) Let $\operatorname{det} \hat{\mathcal{L}}(\cdot, \cdot ; \vartheta)$ be $N$-parabolic for some $\vartheta \in \mathcal{P}_{M}$ and recall that $\omega_{j}, \Omega_{+}^{-1}$, and $\Omega^{\prime}$ are $N$ parabolic. Hence their principal parts do not vanish and the above representation of $\operatorname{det} \hat{\mathcal{L}}$ and the identities in Figure 3.2 show that $\pi_{\gamma} P(\lambda, i \xi ; \vartheta) \neq 0$ for all $\lambda>0$ and $\xi \in \mathbb{R}^{n} \backslash\{0\}$. Therefore Lemma 3.4 implies that $\lambda_{s}+\mu_{s}>0$ and $\vartheta$ satisfies (3.16).
3.1.3. Function spaces. Next, we construct spaces $\mathbb{H}$ and $\mathbb{F}$ such that the interface operator

$$
\hat{\mathcal{L}}\left(\tau+\mathcal{D}_{t}, \mathcal{D}_{x} ; \vartheta\right): \mathbb{H} \rightarrow \mathbb{F}, \quad\left(\Phi_{v}, \Phi_{w}^{2}, \Phi_{w}^{1}, h, \llbracket \pi \rrbracket\right) \mapsto\left(g_{v}, 0,0, g_{h}, g_{w}\right)
$$

is uniformly invertible with respect to $\vartheta \in \mathcal{P}_{M}$, for every $M>0$. Here the operator $\hat{\mathcal{L}}(\tau+$ $\left.\mathcal{D}_{t}, \mathcal{D}_{x} ; \vartheta\right)$ is defined by the joint functional calculus of $\left(\mathcal{D}_{t}, \mathcal{D}_{x}\right)$ from Theorem B.70 on page 166. We note that every component of $\hat{\mathcal{L}}(\lambda, z ; \vartheta)$ belongs to the symbol class $S\left(\bar{\Sigma}_{\phi} \times \overline{B \Sigma}_{\delta}^{n} \times \mathcal{P}_{M}\right)$ from page 166 for $\phi \in(\pi / 2, \pi)$ and some $\delta \in(0, \pi / 8]$. From now on we restict our considerations to the case

$$
c_{5} \in\{0,1\}, \quad c_{6}=1
$$

In order to apply Theorem B.79, we first estimate the $\gamma$-orders of the components of $\hat{\mathcal{L}}$.

$$
d_{\gamma}(\hat{\mathcal{L}}) \leq\left[\begin{array}{ccccc}
\max \{2, \gamma / 2\} & \max \{1, \gamma / 2\} & \max \{1, \gamma / 2\} & 3 & 1-\max \{1, \gamma / 2\} \\
1 & \max \{1, \gamma / 2\} & -\infty & -\infty & -\infty \\
1 & -\infty & \max \{1, \gamma / 2\} & -\infty & -\infty \\
-\infty & 0 & 0 & \gamma & 1-\max \{2, \gamma\} \\
1 & \max \{1, \gamma / 2\} & \max \{1, \gamma / 2\} & 2 & 0
\end{array}\right]
$$

Here the relation $\leq$ is considered component-wise and an entry $-\infty$ corresponds to a vanishing component of $\hat{\mathcal{L}}$.

We define the row-wise order functions $s_{j}$ and the column-wise order functions $t_{i}$ by

$$
\begin{aligned}
s_{1}(\gamma)=\cdots=s_{n}(\gamma) & :=1, & t_{1}(\gamma)=\cdots=t_{n}(\gamma) & :=\max \{2, \gamma / 2\}-1, \\
s_{n+1}(\gamma)=s_{n+2}(\gamma) & :=0, & t_{n+1}(\gamma)=t_{n+2}(\gamma) & :=\max \{1, \gamma / 2\} \\
s_{n+3}(\gamma) & :=-1, & t_{n+3}(\gamma) & :=\max \{1, \gamma\}+1, \\
s_{n+4}(\gamma) & :=0, & & t_{n+4}(\gamma)
\end{aligned}:=0 .
$$

Then it follows that

$$
\begin{aligned}
\sum_{j} s_{j}(\gamma)+\sum_{i} t_{i}(\gamma) & =n \max \{2, \gamma / 2\}+2 \max \{1, \gamma / 2\}+\max \{1, \gamma\} \\
& =\max \{2 n+3,2 n+2+\gamma, 2 n+2 \gamma,(n+4) \gamma / 2\}=d_{\gamma}(\operatorname{det} \hat{\mathcal{L}})
\end{aligned}
$$

Moreover, for all $i, j \in\{1, \ldots, n+4\}$, the function $s_{j}+t_{i}$ is an upper order function for $\hat{\mathcal{L}}_{j i}$.
3.6. Corollary. Let $\rho_{j}, \mu_{j}, \sigma$, and $\mu_{s}$ be positive constants and let $\lambda_{s}$ be a real number such that $\lambda_{s}+$ $\mu_{s}>0$. Then for given numbers $M>0$ and $\phi \in(\pi / 2, \pi)$, there exists $\delta \in(0, \pi / 8]$ such that the symbol $\hat{\mathcal{L}}: \bar{\Sigma}_{\phi} \times \overline{B \Sigma}_{\delta}^{n} \times \mathcal{P}_{M} \rightarrow \mathbb{C}^{(n+4) \times(n+4)}$ is an $N$-parabolic mixed-order system.
3.7. Remark. The preceding choice of the order functions differs from [DK13, p. 223]. In particular, we take care of the additional entries $\hat{\mathcal{L}}_{i, n+3}(i \leq n)$ with $\gamma$-order lesser or equal to 3 and we avoid the difference $\max \{2, \gamma / 2\}-\max \{1, \gamma / 2\}$, which is neither convex nor concave and hence not an order function.

|  | $l=0, \gamma \in(0,1]$ | $l=1, \gamma \in(1,2]$ | $l=2, \gamma \in(2,4]$ | $l=3, \gamma \in(4, \infty)$ |
| :---: | :---: | :---: | :---: | :---: |
| $t_{i}$ | $b_{0}\left(t_{i}\right)+m_{0}\left(t_{i}\right) \gamma$ | $b_{1}\left(t_{i}\right)+m_{1}\left(t_{i}\right) \gamma$ | $b_{2}\left(t_{i}\right)+m_{2}\left(t_{i}\right) \gamma$ | $b_{3}\left(t_{i}\right)+m_{3}\left(t_{i}\right) \gamma$ |
| $t_{1}=\cdots=t_{n}$ | $1+0 \gamma$ | $1+0 \gamma$ | $1+0 \gamma$ | $-1+\frac{1}{2} \gamma$ |
| $t_{n+1}=t_{n+2}$ | $1+0 \gamma$ | $1+0 \gamma$ | $0+\frac{1}{2} \gamma$ | $0+\frac{1}{2} \gamma$ |
| $t_{n+3}$ | $2+0 \gamma$ | $1+1 \gamma$ | $1+1 \gamma$ | $1+1 \gamma$ |
| $t_{n+4}$ | $0+0 \gamma$ | $0+0 \gamma$ | $0+0 \gamma$ | $0+0 \gamma$ |
| $s_{j}$ | $b_{0}\left(s_{j}\right)+m_{0}\left(s_{j}\right) \gamma$ | $b_{1}\left(s_{j}\right)+m_{1}\left(s_{j}\right) \gamma$ | $b_{2}\left(s_{j}\right)+m_{2}\left(s_{j}\right) \gamma$ | $b_{3}\left(s_{j}\right)+m_{3}\left(s_{j}\right) \gamma$ |
| $s_{1}=\cdots=s_{n}$ | $1+0 \gamma$ | $1+0 \gamma$ | $1+0 \gamma$ | $1+0 \gamma$ |
| $s_{n+1}=s_{n+2}$ | $0+0 \gamma$ | $0+0 \gamma$ | $0+0 \gamma$ | $0+0 \gamma$ |
| $s_{n+3}$ | $-1+0 \gamma$ | $-1+0 \gamma$ | $-1+0 \gamma$ | $-1+0 \gamma$ |
| $s_{n+4}$ | $0+0 \gamma$ | $0+0 \gamma$ | $0+0 \gamma$ | $0+0 \gamma$ |

Figure 3.3. Upper order functions for the symbol $\hat{\mathcal{L}}$.

$$
\begin{array}{rlll}
\mathbb{H}_{1}=\cdots=\mathbb{H}_{n}:=L_{p}\left(W_{p}^{3-\frac{1}{p}}\right) \cap L_{p}\left(W_{p}^{3-\frac{1}{p}}\right) & \cap_{0} W_{p}^{\frac{1}{2}-\frac{1}{2 p}}\left(H_{p}^{2}\right) \cap_{0} W_{p}^{1-\frac{1}{2 p}}\left(L_{p}\right) & \left(\text { for } \Phi_{v}\right), \\
\mathbb{H}_{n+1}= & \mathbb{H}_{n+2}:=L_{p}\left(W_{p}^{3-\frac{1}{p}}\right) \cap L_{p}\left(W_{p}^{3-\frac{1}{p}}\right) & \cap_{0} W_{p}^{1-\frac{1}{2 p}}\left(H_{p}^{1}\right) \cap_{0} W_{p}^{1-\frac{1}{2 p}}\left(H_{p}^{1}\right) & \left(\text { for } \Phi_{w}^{j}\right), \\
\mathbb{H}_{n+3}:=L_{p}\left(W_{p}^{4-\frac{1}{p}}\right) \cap H_{p}^{1}\left(W_{p}^{3-\frac{1}{p}}\right) \cap_{0} W_{p}^{\frac{3}{2}-\frac{1}{2 p}}\left(H_{p}^{2}\right) \cap_{0} W_{p}^{\frac{3}{2}-\frac{1}{2 p}}\left(H_{p}^{2}\right) & (\text { for } h), \\
\mathbb{H}_{n+4}:=L_{p}\left(W_{p}^{2-\frac{1}{p}}\right) \cap L_{p}\left(W_{p}^{2-\frac{1}{p}}\right) & \cap_{0} W_{p}^{\frac{1}{2}-\frac{1}{2 p}}\left(H_{p}^{1}\right) \cap_{0} W_{p}^{\frac{1}{2}-\frac{1}{2 p}}\left(H_{p}^{1}\right) & (\text { for } \llbracket \pi \rrbracket), \\
\mathbb{F}_{1}=\cdots=\mathbb{F}_{n}:=L_{p}\left(W_{p}^{1-\frac{1}{p}}\right) \cap L_{p}\left(W_{p}^{1-\frac{1}{p}}\right) & \cap_{0} W_{p}^{\frac{1}{2}-\frac{1}{2 p}}\left(L_{p}\right) \cap_{0} W_{p}^{\frac{1}{2}-\frac{1}{2 p}}\left(L_{p}\right) & \left(\text { for } g_{v}\right), \\
\mathbb{F}_{n+1}= & \mathbb{F}_{n+2}:=L_{p}\left(W_{p}^{2-\frac{1}{p}}\right) \cap L_{p}\left(W_{p}^{2-\frac{1}{p}}\right) & \cap_{0} W_{p}^{\frac{1}{2}-\frac{1}{2 p}}\left(H_{p}^{1}\right) \cap_{0} W_{p}^{\frac{1}{2}-\frac{1}{2 p}}\left(H_{p}^{1}\right), \\
& \mathbb{F}_{n+3}:=L_{p}\left(W_{p}^{3-\frac{1}{p}}\right) \cap L_{p}\left(W_{p}^{3-\frac{1}{p}}\right) & \cap_{0} W_{p}^{\frac{1}{2}-\frac{1}{2 p}}\left(H_{p}^{2}\right) \cap_{0} W_{p}^{\frac{1}{2}-\frac{1}{2 p}}\left(H_{p}^{2}\right) & \left(\text { for } g_{h}\right), \\
& \mathbb{F}_{n+4}:=L_{p}\left(W_{p}^{2-\frac{1}{p}}\right) \cap L_{p}\left(W_{p}^{2-\frac{1}{p}}\right) & \cap_{0} W_{p}^{\frac{1}{2}-\frac{1}{2 p}}\left(H_{p}^{1}\right) \cap_{0} W_{p}^{\frac{1}{2}-\frac{1}{2 p}}\left(H_{p}^{1}\right) & \left(\text { for } g_{w}\right) .
\end{array}
$$

Here we abbreviate $H_{p}^{\alpha}\left(W_{p}^{\beta}\right):=H_{p}^{\alpha}\left(\mathbb{R}_{+} ; W_{p}^{\beta}\left(\mathbb{R}^{n}\right)\right)$ and $W_{p}^{\alpha}\left(H_{p}^{\beta}\right):=W_{p}^{\alpha}\left(\mathbb{R}_{+} ; H_{p}^{\beta}\left(\mathbb{R}^{n}\right)\right)$.
FIGURE 3.4. Function spaces for the operator $\hat{\mathcal{L}}\left(\tau+\mathcal{D}_{t}, \mathcal{D}_{x} ; \vartheta\right)$.

We choose the following parameters and scales for the construction of the spaces $\mathbb{H}_{i}$ and $\mathbb{F}_{j}$.

$$
\left.\begin{array}{rlrlrl}
\left(\gamma_{0}, \gamma_{1}\right] & =(0,1], & \left(\gamma_{1}, \gamma_{2}\right] & =(1,2], & \left(\gamma_{2}, \gamma_{3}\right] & =(2,4], \\
s_{0}^{\prime} & =0, & s_{1}^{\prime} & =0, & s_{2}^{\prime} & =1 / 2-1 / 2 p, \\
r_{0}^{\prime} & =2-1 / p, & r_{1}^{\prime} & =2-1 / p, & r_{2}^{\prime} & =1, \\
\mathcal{F}_{0}\left(\mathcal{K}_{0}\right) & =H_{p}\left(B_{p, p}\right), & \mathcal{F}_{1}\left(\mathcal{K}_{1}\right) & =H_{p}\left(B_{p, p}\right), & \mathcal{F}_{2}\left(\mathcal{K}_{2}\right) & =B_{p, p}\left(H_{p}\right),
\end{array}\right) r_{3}^{\prime}=1 / 2-1 / 2 p, \mathcal{F}_{3}\left(\mathcal{K}_{3}\right)=B_{p, p}\left(H_{p}\right) .
$$

By using the definitions of $t_{i}$ and $s_{j}$, we obtain the representations in Figure 3.3 on this page. Then we define the spaces $\mathbb{H}_{i}$ and $\mathbb{F}_{j}$ by

$$
\mathbb{H}_{i}:=\bigcap_{l=0}^{3}{ }_{0} \mathcal{F}_{l}^{s_{l}^{\prime}+m_{l}\left(t_{i}\right)}\left(\mathcal{K}_{l}^{r_{l}^{\prime}+b_{l}\left(t_{i}\right)}\right), \quad \mathbb{F}_{j}:=\bigcap_{l=0}^{3} 0 \mathcal{F}_{l}^{s_{l}^{\prime}-m_{l}\left(s_{j}\right)}\left(\mathcal{K}_{l}^{r_{l}^{\prime}-b_{l}\left(s_{j}\right)}\right) .
$$

In our situation this yields the spaces in Figure 3.4 on the current page. These spaces are ad-
missible in the sense of Definition B. 78 and therefore Theorem B. 79 yields the following result.
3.8. Lemma. Let $\rho_{j}, \mu_{j}, \sigma$, and $\mu_{s}$ be positive constants and let $\lambda_{s}$ be a real number such that $\lambda_{s}+\mu_{s}>$ 0 . Then, given $p \in(1, \infty), M>0$, and $\phi \in(\pi / 2, \pi)$, there exist $\delta \in(0, \pi / 8)$ and $\tau \in[0, \infty)$ such that

$$
\hat{\mathcal{L}}\left(\tau+\mathcal{D}_{t}, \mathcal{D}_{x} ; \vartheta\right): \prod_{i=1}^{n+4} \mathbb{H}_{i} \rightarrow \prod_{j=1}^{n+4} \mathbb{F}_{j}
$$

is uniformly invertible with respect to $\vartheta \in \mathcal{P}_{M}$.
By restricting the inverse of $\hat{\mathcal{L}}\left(\tau+\mathcal{D}_{t}, \mathcal{D}_{x} ; \vartheta\right)$ to tuples of the form $\left(g_{v}, 0,0, g_{h}, g_{w}\right)$, and by using that $\partial_{t} h=g_{h}+w$ belongs to the subspace ${ }_{0} \mathbb{G}_{h}$ of $\mathbb{F}_{n+3}$ for $g_{h} \in{ }_{0} \mathbb{G}_{h}$ and $w \in{ }_{0} \mathbb{E}_{w}$, we obtain the following result for the spaces from Figure 3.1.
3.9. Corollary. In the situation of Lemma 3.8, the map

$$
\begin{aligned}
\quad\left(g_{v}, g_{w}, g_{h}\right) & \mapsto\left(\Phi_{v}, \Phi_{w}^{2}, \Phi_{w}^{1}, h, \llbracket \pi \rrbracket\right)^{\top}=\left[\hat{\mathcal{L}}\left(\tau+\mathcal{D}_{t}, \mathcal{D}_{x} ; \vartheta\right)\right]^{-1}\left(g_{v}, 0,0, g_{h}, g_{w}\right)^{\top}, \\
{ }_{0} \mathbb{G}_{v} \times{ }_{0} \mathbb{G}_{w} \times{ }_{0} \mathbb{G}_{h} & \rightarrow{ }_{0} \mathbb{E}_{v} \times{ }_{0} \mathbb{E}_{w} \times{ }_{0} \mathbb{E}_{w} \times{ }_{0} \mathbb{E}_{h} \times{ }_{0} \mathbb{G}_{w}
\end{aligned}
$$

is uniformly bounded with respect to $\vartheta \in \mathcal{P}_{M}$.
3.1.4. Interior regularity. Our next goal is to verify the interior regularity conditions

$$
u \in H_{p}^{1,2}\left(\mathbb{R}_{+} \times \dot{\mathbb{R}}^{n+1}\right):=H_{p}^{1}\left(\mathbb{R}_{+} ; L_{p}\left(\mathbb{R}^{n+1}\right)\right) \cap L_{p}\left(\mathbb{R}_{+} ; H_{p}^{2}\left(\mathbb{R}^{n+1}\right)\right), \quad \pi \in L_{p}\left(\mathbb{R}_{+} ; \dot{H}_{p}^{1}\left(\dot{\mathbb{R}}^{n+1}\right)\right)
$$

for the functions $u$ and $\pi$ from (3.7) and Corollary 3.9. We can easily obtain the pressure regularity from the properties of the Poisson semigroup, but for the velocity we need to study more involved extension operators. The Fourier-Laplace transformed functions $\hat{\pi}_{j}, \hat{v}_{j}$, and $\hat{w}_{j}$ of $\pi$ and $\left.u\right|_{\Sigma}=(v, w)$ were given in (3.7), where we employed Green's functions $k_{ \pm}$from Lemma 3.3, the Poisson extension symbol $e^{-|\xi| y}$ and the parabolic extension symbol $e^{-\omega_{j}(\lambda, \xi) y}$ with

$$
\omega_{j}(\lambda, \xi)=\left(\rho_{j} \mu_{j}^{-1}(\tau+\lambda)+|\xi|^{2}\right)^{1 / 2} \quad \text { for } \lambda \in \bar{\Sigma}_{\phi}, \xi \in \mathbb{R}^{n} .
$$

Here $\tau>0$ is chosen as in Lemma 3.8. For computing the integrals in $\hat{v}$ and $\hat{w}$, we let $y>0$, $\omega \in \mathbb{C}$, and $\eta:=|z|_{-} \in \mathbb{C}$ with $\operatorname{Re} \omega, \operatorname{Re} \eta$, and $\operatorname{Re}(\omega-\eta)>0$. Then

$$
\begin{aligned}
\int_{0}^{\infty} k_{ \pm}(y, s ; \omega) e^{-\eta s} d s & =\int_{0}^{\infty} \frac{e^{-\omega|y-s|} \pm e^{-\omega(y+s)}}{2 \omega} e^{-\eta s} d s \\
& =\int_{0}^{y} \frac{e^{-\omega y}}{2 \omega}\left(e^{\omega s} \pm e^{-\omega s}\right) e^{-\eta s} d s+\int_{y}^{\infty} \frac{e^{\omega y} \pm e^{-\omega y}}{2 \omega} e^{-\omega s-\eta s} d s \\
& =\frac{e^{-\omega y}}{2 \omega}\left(\frac{e^{(\omega-\eta) y}-1}{\omega-\eta} \mp \frac{e^{-(\omega+\eta) y}-1}{\omega+\eta}\right)+\frac{e^{\omega y} \pm e^{-\omega y}}{2 \omega} \frac{e^{-(\omega+\eta) y}}{\omega+\eta} \\
& =\frac{e^{-\eta y}-e^{-\omega y}}{2 \omega(\omega-\eta)}+\frac{e^{-\eta y} \pm e^{-\omega y}}{2 \omega(\omega+\eta)} .
\end{aligned}
$$

From the identites $g_{h}=\left(\tau+\partial_{t}\right) h-w$ and (3.8c) we infer that

$$
\hat{p}_{j}(\lambda, \xi)=\frac{(-1)^{j} \mu_{j} \omega_{j}\left(\omega_{j}+|\xi|\right)}{|\xi|}\left((\tau+\lambda) \hat{h}-\hat{g}_{h}-\hat{\Phi}_{w}^{j}\right) .
$$

Plugging in these identities into (3.7) yields the representations in Figure 3.5 on the facing page.
In order to prove the interior regularity of the velocity, we employ general extension operators $E[s]$ induced by an extension symbol $s(\lambda, \xi, y)$ like the parabolic extension symbol $e^{-\omega(\lambda, \xi) y}$ or the Stokes extension symbol $\omega \frac{e^{-\omega y}-e^{-|\xi| y}}{\omega-|\xi|}$ and apply these to the extension symbols $V_{w, j}, V_{h, j}$, $W_{w, j}$, and $W_{h, j}$. To this purpose, we first prove the boundedness of certain integral operators by comparing their kernels with the Hilbert transform. Similar results were established by Denk, Hieber and Prüss [DHP03, Section 6.4], [DHP07, Section 4], and by Denk and Kaip [DK13, Section 3.5].

Let $\omega_{j}(\lambda, \xi)=\left(\rho_{j} \mu_{j}^{-1}(\tau+\lambda)+|\xi|^{2}\right)^{1 / 2}$ and define the extension symbols

$$
\begin{aligned}
V_{w, j}(\lambda, \xi, y) & :=\frac{(-1)^{j+1} i \xi \omega_{j}}{|\xi|} \frac{e^{-\omega_{j} y}-e^{-|\xi| y}}{\omega_{j}-|\xi|}, & V_{h, j}(\lambda, \xi, y):=\frac{(-1)^{j} i \xi \omega_{j}}{|\xi|} \frac{e^{-\omega_{j} y}-e^{-|\xi| y}}{\omega_{j}-|\xi|}, \\
W_{w, j}(\lambda, \xi, y) & :=\omega_{j} \frac{e^{-\omega_{j} y}-e^{-|\xi| y}}{\omega_{j}-|\xi|}, & W_{h, j}(\lambda, \xi, y):=-\frac{|\xi| e^{-\omega_{j} y}-\omega_{j} e^{-|\xi| y}}{\omega_{j}-|\xi|} .
\end{aligned}
$$

The Fourier-Laplace transforms $\hat{u}_{j}(\lambda, \xi, y) \quad:=\hat{u}\left(\lambda, \xi,(-1)^{j} y\right)$ and $\hat{\pi}_{j}(\lambda, \xi, y) \quad:=$ $\hat{\pi}\left(\lambda, \xi,(-1)^{j} y\right)$ of $u_{j}=\left(v_{j}, w_{j}\right)$ and $\pi_{j}$ are given by $\hat{v}_{j}(\lambda, \xi, y)=e^{-\omega_{j} y} \hat{\Phi}_{v}(\lambda, \xi)+V_{w, j}(\lambda, \xi, y) \hat{\Phi}_{w}^{j}(\lambda, \xi)+V_{h, j}(\lambda, \xi, y)\left((\tau+\lambda) \hat{h}(\lambda, \xi)-\hat{g}_{h}(\lambda, \xi)\right)$,
$\hat{w}_{j}(\lambda, \xi, y)=W_{w, j}(\lambda, \xi, y) \hat{\Phi}_{w}^{j}(\lambda, \xi)+W_{h, j}(\lambda, \xi, y)\left((\tau+\lambda) \hat{h}(\lambda, \xi)-\hat{g}_{h}(\lambda, \xi)\right)$, $\hat{\pi}_{j}(\lambda, \xi, y)=e^{-|\xi| y}(-1)^{j} \mu_{j} \omega_{j}\left(\omega_{j}+|\xi|\right)|\xi|^{-1}\left((\tau+\lambda) \hat{h}(\lambda, \xi)-\hat{g}_{h}(\lambda, \xi)-\hat{\Phi}_{w}^{j}(\lambda, \xi)\right)$.

Figure 3.5. The interior Fourier-Laplace transformed velocity and pressure.
3.10. Lemma. Let $E$ be a Banach space of class $\mathcal{H T}$ with property $(\alpha)$ and let $\phi \in(\pi / 2, \pi)$ and $\delta \in(0, \pi / 2)$ such that $\phi+2 \delta<\pi$. Suppose that the mapping

$$
k:(\lambda, z, y, \bar{y}) \mapsto k(\lambda, z, y, \bar{y}), \quad \Sigma_{\phi} \times B \Sigma_{\delta}^{n} \times \mathbb{R}_{+} \times \mathbb{R}_{+} \rightarrow \mathcal{B}(E)
$$

is holomorphic with respect to $(\lambda, z)$ for every $(y, \bar{y})$ and that

$$
M(k):=\sup \left\{|(y+\bar{y}) k(\lambda, z, y, \bar{y})|_{\mathcal{B}(E)}: \lambda \in \Sigma_{\phi}, z \in B \Sigma_{\delta}^{n}, y, \bar{y} \in \mathbb{R}_{+}\right\}<\infty .
$$

With the joint functional calculus for $\left(\mathcal{D}_{t}, \mathcal{D}_{x}\right)$ from Theorem B. 70 we define

$$
k\left(\mathcal{D}_{t}, \mathcal{D}_{x}, y, \bar{y}\right) \in \mathcal{B}\left(L_{p}\left(\mathbb{R}_{+} \times \mathbb{R}^{n} ; E\right)\right) \quad \text { for } y, \bar{y} \in \mathbb{R}_{+}, p \in(1, \infty) .
$$

Define an integral operator $G[k]$ by

$$
(G[k] u)(y)=\int_{0}^{\infty} k\left(\mathcal{D}_{t}, \mathcal{D}_{x}, y, \bar{y}\right) u(\cdot, \cdot, \bar{y}) d \bar{y}, \quad \text { for } y \in(0, \infty), u \in C_{c}^{\infty}\left(\mathbb{R}_{+} \times \mathbb{R}^{n} \times \mathbb{R}_{+} ; E\right) .
$$

Then $G[k]$ can be extended uniquely to a bounded operator in $L_{p}\left(\mathbb{R}_{+} ; L_{p}\left(\mathbb{R}_{+}^{n+1} ; E\right)\right)$ such that

$$
\|G[k]\|_{\mathcal{B}\left(L_{p}\left(\mathbb{R}_{+} ; L_{p}\left(\mathbb{R}_{+}^{n+1} ; E\right)\right)\right)} \leq C M(k)\|\mathcal{H} \mathcal{T}\|_{\mathcal{B}\left(L_{p}\left(\mathbb{R}_{+}\right)\right)} .
$$

Here $\mathcal{H} \mathcal{T}$ is the one-sided Hilbert transform on $L_{p}\left(\mathbb{R}_{+}\right)$and the number $C$ satisfies

$$
\left\|f\left(\mathcal{D}_{t}, \mathcal{D}_{x}\right)\right\|_{\mathcal{B}\left(L_{p}\left(\mathbb{R}_{+} \times \mathbb{R}^{n} ; E\right)\right)} \leq C\|f\|_{\infty} \quad \text { for all } f \in \mathcal{H}^{\infty}\left(\Sigma_{\phi} \times B \Sigma_{\delta}^{n}\right)
$$

Proof. The one-sided scalar Hilbert transform

$$
(\mathcal{H} \mathcal{T} u)(y)=\int_{0}^{\infty} \frac{u(\bar{y})}{y+\bar{y}} d \bar{y} \quad \text { for } y \in \mathbb{R}_{+}, u \in L_{p}\left(\mathbb{R}_{+}\right)
$$

is bounded in $L_{p}\left(\mathbb{R}_{+}\right)$with norm $\tan (\pi / 2 p)$ if $p \in(1,2]$ and $\cot (\pi / 2 p)$ if $p \in[2, \infty)$ [see ME88]. By applying Theorem B. 70 to the family of bounded holomorphic functions $\{(y+\bar{y}) k(\cdot, \cdot, y, \bar{y})$ : $\left.y, \bar{y} \in \mathbb{R}_{+}\right\} \subset \mathcal{H}^{\infty}\left(\Sigma_{\phi} \times B \Sigma_{\delta}^{n}\right)$, we obtain

$$
\begin{aligned}
\|(G[k] u)(t)\|_{L_{p}\left(\mathbb{R}_{+}^{n+1} ; E\right)} & \leq C M(k)\left\|y \mapsto \int_{0}^{\infty} \frac{1}{y+\bar{y}}|u(t, \cdot, \bar{y})|_{E} d \bar{y}\right\|_{L_{p}\left(\mathbb{R}_{+}^{n+1}\right)} \\
& \leq C M(k)\|\mathcal{H} \mathcal{T}\|_{L_{p}(\mathbb{R})}\|u(t, \cdot, \cdot)\|_{L_{p}\left(\mathbb{R}_{+}^{n+1} ; E\right)}
\end{aligned}
$$

Next, we may rearrange the $t$ - and $y$-variables using Fubini's theorem:

$$
\|t \mapsto u(t, \cdot, \cdot)\|_{L_{p}\left(\mathbb{R}_{+} ; L_{p}\left(\mathbb{R}_{+}^{n+1} ; E\right)\right)}=\|y \mapsto u(\cdot, \cdot, y)\|_{L_{p}\left(\mathbb{R}_{+} ; L_{p}\left(\mathbb{R}_{+}^{n+1} ; E\right)\right)} .
$$

Hence the assertion follows.

Next we define the aforementioned extension operators with the Volevich trick (3.18). We employ the anisotropic Sobolev-Slobodeckiĭ spaces

$$
{ }_{0} W_{p}^{t, s}(J \times \Omega ; E):={ }_{0} W_{p}^{t}\left(J ; L_{p}(\Omega ; E)\right) \cap L_{p}\left(J ; W_{p}^{s}(\Omega ; E)\right) .
$$

3.11. Lemma (Extension operators from $y=0$ to $y \in \mathbb{R}_{+}$). Let $E$ be a Banach space of class $\mathcal{H} \mathcal{T}$ with property $(\alpha)$ and let $\phi \in(\pi / 2, \pi)$ and $\delta \in(0, \pi / 8]$ such that $\phi+2 \delta<\pi$. Suppose that the mapping

$$
s:(\lambda, z, y) \mapsto s(\lambda, z, y), \quad \Sigma_{\phi} \times B \Sigma_{\delta}^{n} \times \mathbb{R}_{+} \rightarrow \mathcal{B}(E)
$$

is holomorphic and bounded with respect to $(\lambda, z)$ for every $y$. Then we define $S(y):=s\left(\mathcal{D}_{t}, \mathcal{D}_{x}, y\right) \in$ $\mathcal{B}\left(L_{p}\left(\mathbb{R}_{+} \times \mathbb{R}^{n} ; E\right)\right)$ by Theorem B. 70 and we consider the operator

$$
\begin{equation*}
(E[s] f)(y)=S(y)\left(\left.f\right|_{y=0}\right)=-\int_{0}^{\infty}\left(\partial_{\bar{y}} S(y+\bar{y}) f(\bar{y})+S(y+\bar{y}) \partial_{\bar{y}} f(\bar{y})\right) d \bar{y}, \quad y>0 . \tag{3.18}
\end{equation*}
$$

acting on appropriate functions $f: \mathbb{R}_{+} \times \mathbb{R}_{+}^{n+1} \rightarrow E$, which are specified below. Since $E[s] f$ only depends on $\left.f\right|_{y=0}$, we may also consider $E[s]$ as an extension operator which maps functions $f: \mathbb{R}_{+} \times$ $\mathbb{R}^{n} \rightarrow E$ to functions $E[s] f: \mathbb{R}_{+} \times \mathbb{R}_{+}^{n+1} \rightarrow E$.
(i) Let $\omega(\lambda, z):=(\tau+\lambda-z \cdot z)^{1 / 2}$ with some $\tau \geq 0$ and suppose that $s$ satisfies

$$
\begin{equation*}
\sup \left\{\left|y \omega(\lambda, z)^{1-j} \partial_{y}^{j} s(\lambda, z, y)\right|: y>0, \lambda \in \Sigma_{\phi}, z \in B \Sigma_{\delta}^{n}, j \in\{0,1,2,3\}\right\}<\infty . \tag{3.19}
\end{equation*}
$$

Then the operator $E[s]$ is bounded in ${ }_{0} H_{p}^{1,2}\left(\mathbb{R}_{+} \times \mathbb{R}_{+}^{n+1} ; E\right)$. Hence, considered as an extension operator, $E[s]$ is bounded as

$$
E[s]:{ }_{0} W_{p}^{1-1 / 2 p, 2-1 / p}\left(\mathbb{R}_{+} \times \mathbb{R}^{n} ; E\right) \rightarrow_{0} H_{p}^{1,2}\left(\mathbb{R}_{+} \times \mathbb{R}_{+}^{n+1} ; E\right) .
$$

(ii) Suppose that s satisfies the weaker condition

$$
\sup \left\{|y z s(\lambda, z, y)|,\left|y \omega(\lambda, z)^{1-j} \partial_{y}^{j} s(\lambda, z, y)\right|: y>0, \lambda \in \Sigma_{\phi}, z \in B \Sigma_{\delta}^{n}, j \in\{1,2,3\}\right\}<\infty .
$$

Let $P(y)=e^{-\sqrt{-\Delta_{x}} y}$ denote the Poisson semigroup. Then $E[s]$ is a bounded operator as a map

$$
\left\{f \in{ }_{0} H_{p}^{1,2}\left(\mathbb{R}_{+} \times \mathbb{R}_{+}^{n+1} ; E\right): f(\cdot, \cdot, y)=P(y) f(\cdot, \cdot, 0) \text { for } y>0\right\} \rightarrow_{0} H_{p}^{1,2}\left(\mathbb{R}_{+} \times \mathbb{R}_{+}^{n+1} ; E\right) .
$$

Hence, considered as an extension operator, $E[s]$ is bounded as [cf. PS10, Proposition 3.3]

$$
E[s]:{ }_{0} H_{p}^{1}\left(\mathbb{R}_{+} ; \dot{W}_{p}^{-1 / p}\left(\mathbb{R}^{n} ; E\right)\right) \cap W_{p}^{1-1 / 2 p, 2-1 / p}\left(\mathbb{R}_{+} \times \mathbb{R}^{n} ; E\right) \rightarrow_{0} H_{p}^{1,2}\left(\mathbb{R}_{+} \times \mathbb{R}_{+}^{n+1} ; E\right) .
$$

Proof. In order to apply Lemma 3.10 we consider the kernels

$$
(\lambda, z, y, \bar{y}) \mapsto s(\lambda, z, y+\bar{y}), \quad(\lambda, z, y, \bar{y}) \mapsto \partial_{\bar{y}} s(\lambda, z, y+\bar{y}),
$$

which are again denoted by $s$ end $\partial_{\bar{y}} s$, respectively. Then

$$
E[s]=-G\left[\partial_{\bar{y}} s\right]-G[s] \partial_{\bar{y}} .
$$

(i) By means of Theorem B. 70 and Remarks B. 65 we define the operator $L:=\omega\left(\mathcal{D}_{t}, \mathcal{D}_{x}\right)$. By the Kalton-Weis-Theorem B. 47 and by using Theorem B. 34 and Corollary B.37, the operators

$$
\begin{aligned}
L^{2}: \quad{ }_{0} H_{p}^{1,2}\left(\mathbb{R}_{+} \times \mathbb{R}^{n} ; E\right) & \rightarrow L_{p}\left(\mathbb{R}_{+} \times \mathbb{R}^{n} ; E\right), \\
L=\left(L^{2}\right)^{1 / 2}:{ }_{0} H_{p}^{1 / 2,1}\left(\mathbb{R}_{+} \times \mathbb{R}^{n} ; E\right) & \rightarrow L_{p}\left(\mathbb{R}_{+} \times \mathbb{R}^{n} ; E\right)
\end{aligned}
$$

are invertible with $\mathcal{R}$-bounded $\mathcal{H}^{\infty}$-calculi of angle not larger than $\pi / 2$ and $\pi / 4$, respectively.
For given numbers $\phi \in(\pi / 2, \pi)$ and $\delta \in(0, \pi / 8]$ with $\phi+2 \delta<\pi$, the functions $z_{j} \omega(\lambda, z)^{-1}$, $z_{j} z_{k} \omega(\lambda, z)^{-2}$, and $\lambda \omega(\lambda, z)^{-2}$ are bounded with respect to $\lambda \in \bar{\Sigma}_{\phi}$ and $z \in \overline{B \Sigma}_{\delta}^{n}$ by Example B.56. Therefore Theorem B. 70 implies that the operators $\partial_{x_{j}} L^{-1}, \partial_{x_{j}} \partial_{x_{k}} L^{-2}$, and $\partial_{t} L^{-2}$ are bounded in $L_{p}\left(\mathbb{R}_{+} \times \mathbb{R}_{+}^{n+1} ; E\right)$. In order to prove that $E[s]$ maps ${ }_{0} H_{p}^{1,2}\left(\mathbb{R}_{+} \times \mathbb{R}_{+}^{n+1} ; E\right)$ into itself, we use the multiplicative property $(\omega s)\left(\mathcal{D}_{t}, \mathcal{D}_{x}, y\right)=\operatorname{Ls}\left(\mathcal{D}_{t}, \mathcal{D}_{x}, y\right)$ of the joint functional
calculus and deduce the following identity for $f \in{ }_{0} H_{p}^{1,2}\left(\mathbb{R}_{+} \times \mathbb{R}_{+}^{n+1} ; E\right)$ and $l, m \in\{0,1,2\}$ with $0 \leq l+m \leq 2$ :

$$
\begin{aligned}
L^{l} \partial_{y}^{m} E[s] f & =L^{l} \partial_{y}^{m}\left(-G\left[\partial_{y} s\right] f-G[\omega s] L^{-1} \partial_{y} f\right) \\
& =-G\left[\omega^{-m} \partial_{y}^{1+m} s\right] L^{l+m} f-G\left[\omega^{1-m} \partial_{y}^{m} s\right] L^{l+m-1} \partial_{y} f .
\end{aligned}
$$

Since $s$ satisfies (3.19), it follows from Lemma 3.10 that the operators $G\left[\omega^{1-j} \partial_{y}^{j} s\right](j \in\{0,1,2,3\})$ are bounded in $L_{p}\left(\mathbb{R}_{+} \times \mathbb{R}_{+}^{n+1} ; E\right)$. Moreover, the functions $L^{l+m} f$ and $L^{l+m-1} \partial_{y} f$ belong to $L_{p}\left(\mathbb{R}_{+} \times \mathbb{R}_{+}^{n+1} ; E\right)$ and depend continuously on $f \in{ }_{0} H_{p}^{1,2}\left(\mathbb{R}_{+} \times \mathbb{R}_{+}^{n+1} ; E\right)$. Therefore assertion (i) is valid.
(ii) We represent the relevant derivatives of $E[s] f$ as

$$
\begin{aligned}
\nabla_{x}^{j} E[s] f & =-G\left[\partial_{y} s\right] \nabla_{x}^{j} f-G[|z|-s] \nabla_{x}^{j}{\sqrt{-\Delta_{x}}}^{-1} \partial_{y} f & \text { for } j \in\{0,1,2\}, \\
\partial_{t} E[s] f & =-G\left[\partial_{y} s\right] \partial_{t} f-G[|z|-s] \partial_{t}{\sqrt{-\Delta_{x}}}^{-1} \partial_{y} f, & \\
\nabla_{x}^{j} \partial_{y} E[s] f & =-G\left[\omega^{-1} \partial_{y}^{2} s\right] \nabla_{x} L f-G\left[\partial_{y} s\right] \nabla_{x}^{j} \partial_{y} f & \text { for } j \in\{0,1\}, \\
\partial_{y}^{2} E[s] f & =-G\left[\omega^{-2} \partial_{y}^{3} s\right] L^{2} f-G\left[\omega^{-1} \partial_{y}^{2} s\right] L \partial_{y} f . &
\end{aligned}
$$

The operators $G\left[|z|_{-} s\right]$ and $G\left[\omega^{1-j} \partial_{y}^{j} s\right](j \in\{1,2,3\})$ are bounded in $L_{p}\left(\mathbb{R}_{+} \times \mathbb{R}_{+}^{n+1} ; E\right)$ by Lemma 3.10. For functions $f \in{ }_{0} H_{p}^{1,2}\left(\mathbb{R}_{+} \times \mathbb{R}_{+}^{n+1} ; E\right)$ of the form $f(y)=P(y) f(0)$, we have $\partial_{y} f=-\sqrt{-\Delta_{x}} f$ and hence $\sqrt{-\Delta_{x}}{ }^{-1} \partial_{y} f=-f$. For a given function

$$
\begin{aligned}
g \in & { }_{0} H_{p}^{1}\left(\mathbb{R}_{+} ; \dot{W}_{p}^{-1 / p}\left(\mathbb{R}^{n} ; E\right)\right) \cap W_{p}^{1-1 / 2 p, 2-1 / p}\left(\mathbb{R}_{+} \times \mathbb{R}^{n} ; E\right) \\
& \hookrightarrow{ }_{0} H_{p}^{1}\left(\mathbb{R}_{+} ; \dot{W}_{p}^{-1 / p}\left(\mathbb{R}^{n} ; E\right)\right) \cap L_{p}\left(\mathbb{R}_{+} ;\left(\dot{W}_{p}^{-1 / p} \cap \dot{W}_{p}^{2-1 / p}\right)\left(\mathbb{R}^{n} ; E\right)\right)
\end{aligned}
$$

the Poisson extension $f(\cdot, \cdot, y)=P(y) g$ belongs to ${ }_{0} H_{p}^{1,2}\left(\mathbb{R}_{+} \times \mathbb{R}_{+}^{n+1} ; E\right)$ by Theorem B.28. Therefore $E[s]$ satisfies assertion (ii).

In order to deal with the symbols $V_{w, j}$ and $W_{w, j}$, we study the Stokes extension symbol

$$
\begin{equation*}
s(\lambda, z, y):=\omega(\lambda, z) \frac{e^{-\omega(\lambda, z) y}-e^{-\eta(z) y}}{\omega(\lambda, z)-\eta(z)}, \quad \omega(\lambda, z):=\left(\tau+\lambda+|z|_{-}^{2}\right)^{1 / 2}, \eta(z):=|z|_{-} \tag{3.20}
\end{equation*}
$$

for $\lambda \in \Sigma_{\phi}, z \in B \Sigma_{\delta}^{n}$, and $y \in(0, \infty)$.
3.12. Lemma. Let $n \in \mathbb{N}, \phi \in(\pi / 2, \pi), \delta \in(0, \pi / 8]$ with $\phi+2 \delta<\pi$, and $\tau>0$ and define $s$ by (3.20). Then for every $j \in \mathbb{N}$, there exists $C>0$ such that

$$
\sup \left\{|y z s(\lambda, z, y)|,\left|y \omega(\lambda, z)^{1-j} \partial_{y}^{j} s(\lambda, z, y)\right|: \lambda \in \Sigma_{\phi}, z \in B \Sigma_{\delta}^{n}, y \in(0, \infty)\right\} \leq C
$$

Proof. It is useful [cf. SS08, p. 186] to represent the difference quotient as

$$
\frac{e^{-\omega y}-e^{-\eta y}}{\omega-\eta}=\int_{0}^{1} \frac{d}{d s} \frac{e^{-g(s) y}}{\omega-\eta} d s=-y \int_{0}^{1} e^{-g(s) y} d s, \quad \text { where } g(s):=(1-s) \eta+s \omega \text {. }
$$

Lemma B. 55 and Example B. 56 imply $|\eta| \sim \operatorname{Re} \eta$ and $|\omega| \sim \operatorname{Re} \omega$ and hence $\operatorname{Re} g(s) \gtrsim|z|$. For $\alpha \in(0, \infty)$ and $x \in(0, \infty)$, the inequality $e^{-x} \leq(\alpha / e x)^{\alpha}$ is valid. Therefore

$$
|y z s(\lambda, z, y)| \leq y^{2}|z \omega| \int_{0}^{1} e^{-\operatorname{Re} g(s) y} d s \leq y^{2}|z \omega| \int_{0}^{1} \frac{4 d s}{e^{2} y^{2}(\operatorname{Re} g(s))^{2}}=\frac{4|z \omega|}{e^{2} \operatorname{Re} \eta \operatorname{Re} \omega} \lesssim 1
$$

Let us show that $\left|y \omega^{1-j} \partial_{y}^{j} s\right| \lesssim 1$ for $j \in \mathbb{N}$. Since $\phi+2 \delta<\pi$, the inequality (B.14) yields $|\eta| \lesssim|\omega|$. In the case $|\eta| \leq 2^{-1}|\omega|$, we have $|\omega-\eta| \geq 2^{-1}|\omega|$ and hence

$$
\left|y \omega^{1-j} \partial_{y}^{j} s\right| \leq y\left|\omega^{2-j} \frac{\omega^{j} e^{-\omega y}-\eta^{j} e^{-\eta y}}{\omega-\eta}\right| \leq y|\omega|^{2-j} \frac{|\omega|^{j} e^{-\operatorname{Re} \omega y}+|\eta|^{j} e^{-\operatorname{Re} \eta y}}{|\omega-\eta|} \lesssim 1 .
$$

Next, the Leibniz rule yields

$$
\begin{aligned}
y \partial_{y}^{j} s & =-y \partial_{y}^{j}\left(y \omega \int_{0}^{1} e^{-g(s) y} d s\right) \\
& =(-1)^{j+1} y^{2} \omega \int_{0}^{1} g(s)^{j} e^{-g(s) y} d s+j(-1)^{j} y \omega \int_{0}^{1} g(s)^{j-1} e^{-g(s) y} d s .
\end{aligned}
$$

Hence, in the remaining case $2^{-1}|\omega| \leq|\eta| \lesssim|\omega|$, we have $\operatorname{Re} g(s) \sim|\omega|$ and thus

$$
\begin{aligned}
\left|y \omega^{1-j} \partial_{y}^{j} s\right| & \leq y^{2}|\omega|^{2-j} \int_{0}^{1}|\omega|^{j} e^{-\operatorname{Re} g(s) y} d s+j y|\omega|^{2-j} \int_{0}^{1}|\omega|^{j-1} e^{-\operatorname{Re} g(s) y} d s \\
& \leq|\omega|^{2} \int_{0}^{1} \frac{4}{e^{2}(\operatorname{Re} g(s))^{2}} d s+j|\omega| \int_{0}^{1} \frac{1}{e(\operatorname{Re} g(s))^{2}} d s \lesssim 1 .
\end{aligned}
$$

We are ready to prove Theorem 3.1.
Proof of Theorem 3.1. The boundedness of the solution-to-data map $(u, \pi, h) \mapsto\left(g_{v}, g_{w}, g_{h}\right)$ follows from the mixed derivative embeddings on page 159 and the spatial trace theorem on page 156. Moreover, the functions $u, \pi$, and $h$ are Fourier-Laplace transformable in the sense of distributions and their transforms have the representations in (3.7). Hence the uniqueness in Lemma 3.8 and Corollary 3.9 imply that problem (3.2) has at most one solution.

In order to construct a solution, we let $\left(g_{v}, g_{w}, g_{h}\right) \in{ }_{0} \mathbb{G}_{v} \times{ }_{0} \mathbb{G}_{w} \times{ }_{0} \mathbb{G}_{h}$ be given and define the functions $u, \pi$, and $h$ as in Figure 3.5, Lemma 3.8, and Corollary 3.9. Then ( $u, \pi, h$ ) solves problem (3.2), which follows from the injectivity of the Fourier-Laplace transformation. It remains to prove that the data-to-solution $\operatorname{map}\left(g_{v}, g_{w}, g_{h}\right) \mapsto(u, \pi, h)$ is uniformly bounded with respect to $\vartheta \in \mathcal{P}_{M}$.
(i) Corollary 3.9 implies that $\left(g_{v}, g_{w}, g_{h}\right) \mapsto h$ is uniformly bounded with respect to $\vartheta \in \mathcal{P}_{M}$.
(ii) The pressure $\pi$ has the symbol $\hat{\pi}_{j}(\lambda, \xi, y)=e^{-|\xi| y} \hat{p}_{j}(\lambda, \xi)$ where $e^{-|\xi| y}$ is the symbol of the Poisson semigroup $P(y)$. Therefore Theorem B.28.(iv) and $\partial_{y} P(y)=\sqrt{-\Delta_{x}} P(y)$ yield

$$
\left\|\nabla_{(x, y)} P(y) p_{j}\right\|_{L_{p}\left(\mathbb{R}_{+} \times \mathbb{R}_{+}^{n+1}\right)} \lesssim\left\|p_{j}\right\|_{L_{p}\left(\mathbb{R}_{+} ; \dot{W}_{p}^{1-1 / p}\left(\mathbb{R}^{n}\right)\right)}
$$

In view of the divergence conditions (3.6b) and the identity (3.9) for $\hat{p}_{j}$, we obtain

$$
\hat{p}_{j}=-\frac{\alpha_{1} \alpha_{2}}{\Omega_{+}\left(\omega_{1}+\omega_{2}\right)} \frac{i \xi}{|\xi|} \cdot \hat{\Phi}_{v}+\frac{(-1)^{j} \alpha_{j}}{\Omega_{+}} \llbracket \hat{\pi} \rrbracket .
$$

Hence, by Corollary 3.9, the interface pressures $p_{j}=\left.\pi_{j}\right|_{y=0}$ satisfy

$$
p_{j} \in{ }_{0} W_{p}^{1 / 2-1 / 2 p}\left(\mathbb{R}_{+} ; L_{p}\left(\mathbb{R}^{n}\right)\right) \cap L_{p}\left(\mathbb{R}_{+} ; W_{p}^{2-2 / p}\left(\mathbb{R}^{n}\right)\right) \hookrightarrow L_{p}\left(\mathbb{R}_{+} ; \dot{W}_{p}^{1-1 / p}\left(\mathbb{R}^{n}\right)\right)
$$

and therefore $\pi_{j}$ belongs to $\mathbb{E}_{\pi}=L_{p}\left(\mathbb{R}_{+} ; \dot{H}_{p}^{1}\left(\mathbb{R}_{+}^{n+1}\right)\right)$ and satisfies $\llbracket \pi \rrbracket \in{ }_{0} \mathbb{G}_{w} ;$ thus, $\pi \in{ }_{0} \mathbb{E}_{\pi, \llbracket \pi \rrbracket}$. Moreover, the map $\left(g_{v}, g_{w}, g_{h}\right) \mapsto \pi$ is uniformly bounded with respect to $\vartheta \in \mathcal{P}_{M}$.
(iii) Corollary 3.9 yields $p_{j} \in{ }_{0} \mathbb{G}_{w}, \Phi_{v} \in{ }_{0} \mathbb{E}_{v}$, and $\Phi_{w}^{j} \in{ }_{0} \mathbb{E}_{w}$. Therefore the identities (3.8b) and (3.8c) yield $\left.v\right|_{y=0}=\Phi_{v} \in{ }_{0} \mathbb{E}_{v}$ and $\left.w\right|_{y=0} \in{ }_{0} \mathbb{E}_{w}$. Since $\Phi_{v}$ belongs to the Dirichlet trace space ${ }_{0} W_{p}^{1-1 / 2 p, 2-1 / p}\left(\mathbb{R}_{+} \times \mathbb{R}^{n}\right)$ of ${ }_{0} H_{p}^{1,2}\left(\mathbb{R}_{+} \times \mathbb{R}_{+}^{n+1}\right)$, we conclude from Theorem B. 25 that the parabolic extension $\left[y \mapsto e^{-L_{j} y}\right] \Phi_{v}$ with $L_{j}:=\omega_{j}\left(\mathcal{D}_{t}, \mathcal{D}_{x}\right)^{-1}$ belongs to ${ }_{0} H_{p}^{1,2}\left(\mathbb{R}_{+} \times \mathbb{R}_{+}^{n+1}\right)$. Next, from (3.6b) we infer that $\hat{\Phi}_{w}^{j}=(-1)^{j} i \xi \omega_{j}^{-1} \hat{\Phi}_{v}^{j}$, and with ${ }_{0} W_{p}^{1-1 / 2 p, 2-1 / p}\left(\mathbb{R}_{+} \times \mathbb{R}^{n}\right) \hookrightarrow$ ${ }_{0} H_{p}^{1 / 2}\left(\mathbb{R}_{+} ; W_{p}^{1-1 / p}\left(\mathbb{R}^{n}\right)\right.$ ) (see Proposition B.44) we obtain

$$
\Phi_{w}^{j}=(-1)^{j} \dot{J}_{1} L_{j}^{-1} \Phi_{v}^{j} \in \dot{J}_{1} L_{j}^{-1}{ }_{0} H_{p}^{1 / 2}\left(\mathbb{R}_{+} ; W_{p}^{1-1 / p}\left(\mathbb{R}^{n}\right)\right) \hookrightarrow{ }_{0} H_{p}^{1}\left(\mathbb{R}_{+} ; \dot{W}_{p}^{-1 / p}\left(\mathbb{R}^{n}\right)\right)
$$

Hence the Poisson extension $(t, x, y) \mapsto\left(P(y) \Phi_{w}^{j}\right)(t, x)$ belongs to ${ }_{0} H_{p}^{1}\left(\mathbb{R}_{+} ; L_{p}\left(\mathbb{R}_{+}^{n+1}\right)\right)$. By Lemma 3.12 and Example B.56, the Stokes extension symbols $W_{w, j}$ and $V_{w, j}$ satisfy the assumption of Lemma 3.11.(ii). Since $V_{h, j}=-V_{w, j}$ and since $\left(\tau+\partial_{t}\right) h-g_{h}$ belongs to the space
${ }_{0} H_{p}^{1}\left(\mathbb{R}_{+} ; \dot{W}_{p}^{-1 / p}\left(\mathbb{R}^{n}\right)\right) \cap_{0} W_{p}^{1-1 / 2 p, 2-1 / p}\left(\mathbb{R}_{+} \times \mathbb{R}^{n}\right)$, we conclude that $v$ belongs to ${ }_{0} H_{p}^{1,2}\left(\mathbb{R}_{+} \times\right.$ $\dot{\mathbb{R}}^{n+1}$ ) and that $\left(g_{v}, g_{w}, g_{h}\right) \mapsto v$ is uniformly bounded with respect to $\vartheta \in \mathcal{P}_{M}$. Finally, the symbol $W_{h, j}$ also satisfies the assumption of Lemma 3.11. (ii) and therefore $w$ belongs to ${ }_{0} H_{p}^{1,2}\left(\mathbb{R}_{+} \times\right.$ $\left.\dot{\mathbb{R}}^{n+1}\right)^{n}$. We conclude that $u=(v, w)$ belongs to ${ }_{0} \mathbb{E}_{u, v, w}$ and that $\left(g_{v}, g_{w}, g_{h}\right) \mapsto u$ is uniformly bounded with respect to $\vartheta \in \mathcal{P}_{M}$. The proof of Theorem 3.1 is complete.
3.1.5. Inhomogeneous bulk equations. The next step towards optimal regularity of problem (3.1) is to allow for additional data $\left(f_{u}, f_{d}\right)$; that is, we consider the problem

$$
\left\{\begin{array}{rlrl}
\rho \partial_{t} u-\mu \Delta u+\nabla \pi=f_{u} & & \text { in } J \times \dot{\mathbb{R}}^{n+1},  \tag{3.21}\\
\operatorname{div} u=f_{d} & & \text { in } J \times \dot{\mathbb{R}}^{n+1}, \\
\llbracket u \rrbracket & =0 & & \text { on } J \times \mathbb{R}^{n}, \\
-\mu_{s} \Delta_{x} v-\lambda_{s} \nabla_{x} \operatorname{div}_{x} v-\llbracket \mu \partial_{y} v \rrbracket-c_{5} \llbracket \mu \nabla_{x} w \rrbracket+c_{1} \nabla_{x} \Delta_{x} h=g_{v} & & \text { on } J \times \mathbb{R}^{n}, \\
-\operatorname{tr}\left(\left(c_{2}+2 C_{3}\right) \nabla_{x} v\right)-2 \llbracket \mu \partial_{y} w \rrbracket+\llbracket \pi \rrbracket-\operatorname{tr}\left(\left(c_{\sigma}+C_{4}\right) \nabla_{x}^{2} h\right)=g_{w} & & \text { on } J \times \mathbb{R}^{n}, \\
\left(\tau+\partial_{t}\right) h-w & =g_{h} & & \text { on } J \times \mathbb{R}^{n}, \\
\left.h\right|_{t=0} & =0 & & \text { on } \mathbb{R}^{n}, \\
\left.u\right|_{t=0}=0 & & \text { in } \mathbb{R}^{n+1} .
\end{array}\right.
$$

Here we still consider a flat interface $\Sigma=\mathbb{R}^{n} \times\{0\} \cong \mathbb{R}^{n}$ in the whole space $\Omega=\mathbb{R}^{n+1}$, but restrict our investigation to a bounded time interval $J=(0, T)$ with $T \in(0, \infty)$ and $\tau=0$. The physical parameters $\rho_{1}, \rho_{2}, \mu_{1}, \mu_{2}, \sigma, \lambda_{s}$, and $\mu_{s}$, and the abbreviations $c_{1}, c_{2}, C_{3}, C_{4}$, and $c_{\sigma}$ are the same as on page 54 . For the additional data $\left(f_{u}, f_{d}\right)$ we consider the conditions

$$
\begin{aligned}
& f_{u} \in \mathbb{F}_{u}:=L_{p}\left(J ; L_{p}\left(\mathbb{R}^{n+1}\right)^{n+1}\right), \\
& f_{d} \in{ }_{0} \mathbb{F}_{d}:={ }_{0} H_{p}^{1}\left(J ; \dot{H}_{p}^{-1}\left(\mathbb{R}^{n+1}\right)\right) \cap L_{p}\left(J ; H_{p}^{1}\left(\dot{\mathbb{R}}^{n+1}\right)\right) .
\end{aligned}
$$

In the previously considered case $\operatorname{div} u=f_{d}=0$, the term $\partial_{y} w$ was of class ${ }_{0} \mathbb{G}_{w}$, since

$$
\partial_{y} w=\operatorname{div} u-\operatorname{div}_{x} v=-\operatorname{div}_{x} v \in_{0} \mathbb{G}_{w} \quad \text { for } u \in{ }_{0} \mathbb{E}_{u, v, w} \text { with } \operatorname{div} u=0 .
$$

In order to maintain this property for $f_{d} \neq 0$, we consider the additional conditions

$$
\left.\partial_{y} w_{ \pm}\right|_{\Sigma} \in{ }_{0} \mathbb{G}_{w},\left.\quad f_{d \pm}\right|_{\Sigma} \in_{0} \mathbb{G}_{w} .
$$

Then the space of suitable divergence data can be characterized as follows.
3.13. Lemma. The divergence operator

$$
\begin{align*}
\operatorname{div}:{ }_{0} \mathbb{E}_{u, v, w, \partial_{y} w} & :=\left\{u \in{ }_{0} \mathbb{E}_{u, v, w}:\left.\partial_{y} w_{ \pm}\right|_{y=0} \in{ }_{0} \mathbb{G}_{w}\right\} \\
\rightarrow{ }_{0} \mathbb{F}_{d, \Sigma} & :=\left\{f_{d} \in{ }_{0} \mathbb{F}_{d}:\left.f_{d \pm}\right|_{y=0} \in{ }_{0} \mathbb{G}_{w}\right\} \tag{3.22}
\end{align*}
$$

is a retraction.
Proof. We have to show that div: ${ }_{0} \mathbb{E}_{u, v, w, \partial_{y} w} \rightarrow{ }_{0} \mathbb{F}_{d, \Sigma}$ is bounded and surjective and has a bounded right-inverse. The divergence theorem with interface implies that the map (3.22) is bounded. In order to construct a bulk velocity field $u \in \mathbb{E}_{u, v, w, \partial_{\nu} w}$ for given divergence data $f_{d} \in{ }_{0} \mathbb{F}_{d, \Sigma}$, we employ the data-to-solution operators $S_{ \pm}: f_{d \pm} \mapsto\left(u_{ \pm}, \pi_{ \pm}\right)$for the one-phase Stokes problems

$$
\left(\partial_{t}-\Delta\right) u_{ \pm}+\nabla \pi_{ \pm}=0 \text { in } J \times \Omega_{ \pm}, \operatorname{div} u_{ \pm}=f_{d \pm} \text { in } J \times \Omega_{ \pm},\left.u_{ \pm}\right|_{y=0}=0 \text { on } J \times \mathbb{R}^{n}
$$

for $\Omega_{ \pm}= \pm \mathbb{R}_{+}^{n+1}$ from [BP07, Theorem 6.1]. The function $u$ from $\left(u_{ \pm}, \pi_{ \pm}\right)=S_{ \pm}\left(f_{d \pm}\right)$ belongs to ${ }_{0} \mathbb{E}_{u}$, the traces $\left.v\right|_{y=0}$ and $\left.w\right|_{y=0}$ vanish, and hence belong to ${ }_{0} \mathbb{E}_{v}$ and ${ }_{0} \mathbb{E}_{w}$, respectively. Moreover, $\left.\partial_{y} w_{ \pm}\right|_{y=0}=\left.f_{d \pm}\right|_{y=0}$ belong to ${ }_{0} \mathbb{G}_{w}$. Therefore $f_{d} \mapsto u$ is a bounded right-inverse for (3.22).

$$
\begin{aligned}
{ }_{0} \mathbb{E}_{u} & :=\left\{u \in{ }_{0} H_{p}^{1}\left(J ; L_{p}\left(\mathbb{R}^{n+1}\right)^{n+1}\right) \cap L_{p}\left(J ; H_{p}^{2}\left(\mathbb{R}^{n+1}\right)^{n+1}\right): \llbracket u \rrbracket=0\right\}, \\
{ }_{0} \mathbb{E}_{v} & :={ }_{0} W_{p}^{1-1 / 2 p}\left(J ; L_{p}\left(\mathbb{R}^{n}\right)^{n}\right) \cap{ }_{0} W_{p}^{1 / 2-1 / 2 p}\left(J ; H_{p}^{2}\left(\mathbb{R}^{n}\right)^{n}\right) \cap L_{p}\left(J ; W_{p}^{3-1 / p}\left(\mathbb{R}^{n}\right)^{n}\right), \\
{ }_{0} \mathbb{E}_{w} & :={ }_{0} W_{p}^{1-1 / 2 p}\left(J ; H_{p}^{1}\left(\mathbb{R}^{n}\right)\right) \cap L_{p}\left(J ; W_{p}^{3-1 / p}\left(\mathbb{R}^{n}\right)\right), \\
{ }_{0}{ }_{u, v, w, \partial_{y} w} & :=\left\{u=(v, w) \in{ }_{0} \mathbb{E}_{u}:\left.v\right|_{y=0} \in{ }_{0} \mathbb{E}_{v},\left.w\right|_{y=0} \in{ }_{0} \mathbb{E}_{w},\left.\partial_{y} w_{ \pm}\right|_{y=0} \in{ }_{0} \mathbb{G}_{w}\right\}, \\
\mathbb{E}_{\pi} & :=L_{p}\left(J ; \dot{H}_{p}^{1}\left(\dot{R}^{n+1}\right)\right), \\
{ }_{0} \mathbb{E}_{\pi, \llbracket \pi \rrbracket} & :=\left\{\pi \in \mathbb{E}_{\pi}: \llbracket \pi \rrbracket \in{ }_{0} \mathbb{G}_{w}\right\}, \\
{ }_{0} \mathbb{E}_{h} & :={ }_{0} W_{p}^{2-1 / 2 p}\left(J ; H_{p}^{1}\left(\mathbb{R}^{n}\right)\right) \cap{ }_{0} H_{p}^{1}\left(J ; W_{p}^{3-1 / p}\left(\mathbb{R}^{n}\right)\right) \cap L_{p}\left(J ; W_{p}^{4-1 / p}\left(\mathbb{R}^{n}\right)\right), \\
\mathbb{F}_{u} & :=L_{p}\left(J ; L_{p}\left(\mathbb{R}^{n+1}\right)^{n+1}\right), \\
{ }_{0} \mathbb{F}_{d} & :={ }_{0} H_{p}^{1}\left(J ; \dot{H}_{p}^{-1}\left(\mathbb{R}^{n+1}\right)\right) \cap L_{p}\left(J ; H_{p}^{1}\left(\dot{R}^{n+1}\right)\right), \\
{ }_{0} \mathbb{F}_{d, \Sigma} & :=\left\{f_{d} \in{ }_{0} \mathbb{F}_{d}: f_{d \pm} \mid{ }_{y=0} \in{ }_{0} \mathbb{G}_{w}\right\}, \\
{ }_{0} \mathbb{G}_{v} & :={ }_{0} W_{p}^{1 / 2-1 / 2 p}\left(J ; L_{p}\left(\mathbb{R}^{n}\right)^{n}\right) \cap L_{p}\left(J ; W_{p}^{1-1 / p}\left(\mathbb{R}^{n}\right)^{n}\right), \\
{ }_{0} \mathbb{G}_{w} & :={ }_{0} W_{p}^{1 / 2-1 / 2 p}\left(J ; H_{p}^{1}\left(\mathbb{R}^{n}\right)\right) \cap L_{p}\left(J ; W_{p}^{2-1 / p}\left(\mathbb{R}^{n}\right)\right), \\
{ }_{0} \mathbb{G}_{h} & :={ }_{0} W_{p}^{1-1 / 2 p}\left(J, H_{p}^{1}\left(\mathbb{R}^{n}\right)\right) \cap L_{p}\left(J, W_{p}^{3-1 / p}\left(\mathbb{R}^{n}\right)\right) .
\end{aligned}
$$

Figure 3.6. Function spaces ${ }_{0} \mathbb{E} \ldots,{ }_{0} \mathbb{F} \ldots$, and ${ }_{0} \mathbb{G} \ldots$ on $\left(J, \mathbb{R}^{n+1}, \mathbb{R}^{n}\right)$.

We are ready to prove optimal regularity for problem (3.21). The relevant function spaces are summarized in Figure 3.6 on this page.
3.14. Theorem. Let $\lambda_{s}+\mu_{s}>0, c_{5} \in\{0,1\}, c_{6}=1, p \in(1, \infty), T_{0} \in(0, \infty)$, and $M>0$. Then the solution-to-data map

$$
\begin{aligned}
(u, \pi, h) & \mapsto\left(f_{u}, f_{d}, g_{v}, g_{w}, g_{h}\right), \\
{ }_{0} \mathbb{E}_{u, v, w, \partial_{y} w} \times{ }_{0} \mathbb{E}_{\pi, \llbracket \pi \rrbracket} \times{ }_{0} \mathbb{E}_{h} & \rightarrow \mathbb{F}_{u} \times{ }_{0} \mathbb{F}_{d, \Sigma} \times{ }_{0} \mathbb{G}_{v} \times{ }_{0} \mathbb{G}_{w} \times{ }_{0} \mathbb{G}_{h}
\end{aligned}
$$

of problem (3.21) is uniformly invertible with respect to $T \in\left(0, T_{0}\right]$ and $\vartheta \in \mathcal{P}_{M}$.
Proof. Boundedness of the solution-to-data map follows from the mixed derivative embeddings on page 159, the divergence theorem with interface, and the spatial trace theorem on page 156. Injectivity follows from Corollary 3.2. For proving surjectivity, we construct a solution

$$
(u, \pi, h)=\left(u_{1}, 0,0\right)+\left(u_{2}, \pi_{2}, 0\right)+\left(u_{3}, \pi_{3}, h_{3}\right)
$$

First, with the co-retraction $\operatorname{div}^{c}:{ }_{0} \mathbb{F}_{d, \Sigma} \rightarrow{ }_{0} \mathbb{E}_{u, v, w, \partial_{y} w}$ from Lemma 3.13, we choose $u_{1}=\operatorname{div}^{c} f_{d}$. Second, let $f_{u, 2}:=f_{u}-\left(\rho \partial_{t}-\mu \Delta\right) u_{1}$ and let $P=I-\nabla \Delta^{-1}$ div denote the Helmholtz projection in $L_{p}\left(\mathbb{R}^{n+1}\right)^{n+1}$. Then $P f_{u, 2}$ belongs to $\mathbb{F}_{u}$ and we seek a solution $u_{2} \in{ }_{0} \mathbb{E}_{u}$ of the Stokes problem

$$
\begin{equation*}
\rho \partial_{t} u_{2}-\mu \Delta u_{2}=P f_{u, 2}, \quad \operatorname{div} u_{2}=0,\left.\quad u_{2}\right|_{\Sigma}=0 \tag{3.23}
\end{equation*}
$$

Since (3.23) consists of two separated one-phase Stokes problems in $J \times \mathbb{R}_{ \pm}^{n+1}$, we obtain the desired solution map $P f_{u, 2} \mapsto u_{2}$ from [DHP01, Theorem 7.6]. We trivially have $0=\left.v_{2}\right|_{\Sigma} \in{ }_{0} \mathbb{E}_{v}$, $0=\left.w_{2}\right|_{\Sigma} \in{ }_{0} \mathbb{E}_{w}$, and $0=\left.\partial_{y} w_{2}\right|_{\Sigma}=\left.\operatorname{div} u_{2}\right|_{\Sigma}-\left.\operatorname{div}_{x} v_{2}\right|_{\Sigma} \in{ }_{0} \mathbb{G}_{w}$. Therefore $u_{2}$ belongs to ${ }_{0} \mathbb{E}_{u, v, w, \partial_{y} w}$. With Lemma 2.23, we define $\pi_{2}$ as the solution to the weak Neumann transmission problem

$$
\left\langle\nabla \pi_{2}(t, \cdot), \nabla \varphi\right\rangle_{\mathbb{R}^{n+1}}=\left\langle(I-P) f_{u, 2}(t, \cdot), \nabla \varphi\right\rangle_{\mathbb{R}^{n+1}} \quad \text { for all } \varphi \in \dot{H}_{p^{\prime}}^{1}\left(\mathbb{R}^{n+1}\right), \quad \llbracket \pi_{2} \rrbracket=0
$$

Hence $\pi_{2}$ belongs to ${ }_{0} \mathbb{E}_{\pi, \llbracket \pi]}$ and satisfies $\nabla \pi_{2}=(I-P) f_{u, 2}$ in $\mathbb{F}_{u}$. The uniform boundedness of $\left(f_{u}, f_{d}\right) \mapsto\left(u_{1}+u_{2}, \pi_{2}, 0\right)$ with respect to $T \in\left(0, T_{0}\right]$ follows by extension of the data from $J=(0, T)$ to $(0, \infty)$ with Lemma B. 9 and by restriction to $\left(0, T_{0}\right)$.

Finally, Corollary 3.2 yields a unique solution $\left(u_{3}, \pi_{3}, h_{3}\right) \in{ }_{0} \mathbb{E}$ of

$$
\begin{aligned}
\rho \partial_{t} u_{3}-\mu \Delta u_{3}+\nabla \pi_{3} & =0 & & \text { in } J \times \dot{\mathbb{R}}^{n+1}, \\
\operatorname{div} u_{3} & =0 & & \text { in } J \times \dot{\mathbb{R}}^{n+1}, \\
L_{u}\left(u_{3}, \pi_{3}, h_{3} ; \vartheta\right) & =\left(g_{v}, g_{w}\right)-L_{u}\left(u_{1}+u_{2}, \pi_{2}, 0 ; \vartheta\right) & & \text { on } J \times \mathbb{R}^{n}, \\
\partial_{t} h_{3}-w_{3} & =g_{h}+w_{1}+w_{2} & & \text { on } J \times \mathbb{R}^{n},
\end{aligned}
$$

and the map $\left(f_{u}, f_{d}, g_{v}, g_{w}, g_{h}\right) \mapsto\left(u_{3}, \pi_{3}, h_{3}\right)$ is uniformly bounded with respect to $T \in\left(0, T_{0}\right.$ ] and $\vartheta \in \mathcal{P}_{M}$. Hence the proof of Theorem 3.14 is complete.

### 3.2. Bent hyperplanes and variable coefficients

We generalize Theorem 3.14 to a situation where the interface is a bent hyperplane

$$
\begin{equation*}
\Sigma=\Sigma_{\omega}:=\left\{\left(x^{\prime}, \omega\left(x^{\prime}\right)\right): x^{\prime} \in \mathbb{R}^{n-1}\right\} \quad \text { with } \omega \in B C^{4}\left(\mathbb{R}^{n-1}\right) \tag{3.24}
\end{equation*}
$$

in $\Omega=\mathbb{R}^{n}(n \geq 2)$ and the coefficients on the interface may depend on $\left(t, x^{\prime}\right)$. In a tubular neighborhood $B_{r}(\Sigma)$ with projection $\Pi_{\Sigma}: B_{r}(\Sigma) \rightarrow \Sigma$, we decompose $u=v+w \nu_{\Sigma} \circ \Pi$ with $v:=\left[P_{\Sigma} \circ \Pi_{\Sigma}\right] u$ and $w:=\left(\nu_{\Sigma} \circ \Pi_{\Sigma} \mid u\right)$. We consider the perturbed model problem

$$
\left\{\begin{align*}
\rho \partial_{t} u-\mu \Delta u+\nabla \pi & =f_{u} & & \text { in } J \times\left(\mathbb{R}^{n} \backslash \Sigma_{\omega}\right),  \tag{3.25}\\
\operatorname{div} u & =f_{d} & & \text { in } J \times\left(\mathbb{R}^{n} \backslash \Sigma_{\omega}\right), \\
\llbracket u \rrbracket & =0 & & \text { on } J \times \Sigma_{\omega}, \\
L_{v}(u, h ; \vartheta, \omega) & =g_{v} & & \text { on } J \times \Sigma_{\omega}, \\
L_{w}(u, \pi, h ; \vartheta, \omega) & =g_{w} & & \text { on } J \times \Sigma_{\omega}, \\
\partial_{t} h-w & =g_{h} & & \text { on } J \times \Sigma_{\omega}, \\
\left.h\right|_{t=0} & =0 & & \text { on } \Sigma_{\omega}, \\
\left.u\right|_{t=0} & =0 & & \text { in } \mathbb{R}^{n} .
\end{align*}\right.
$$

Here $J=(0, T)$ is bounded, the parameter triple $\vartheta=\left(\vartheta_{L}, \vartheta_{w}, \vartheta_{D v}\right)$ consists of fixed functions

$$
\vartheta_{L}: \Sigma_{\omega} \rightarrow \mathbb{R}^{n \times n}, \quad\left(\vartheta_{w}, \vartheta_{D v}\right): J \times \Sigma_{\omega} \rightarrow \mathbb{K} \times \mathbb{K}^{n \times n},
$$

and, similar to (3.3) on page 54 , we define further parameters

$$
\begin{cases}\vartheta_{1}:=\left(\lambda_{s}+\mu_{s}\right) \vartheta_{w}, & \vartheta_{2}:=\left(\lambda_{s}-\mu_{s}\right) \operatorname{tr} \vartheta_{L},  \tag{3.26}\\ \vartheta_{3}:=\mu_{s} \vartheta_{L}, & \vartheta_{4}:=2 \mu_{s}\left[\vartheta_{D v}-2 \vartheta_{w} \vartheta_{L}\right], \\ c_{5} \in\{0,1\}, & \vartheta_{\sigma}:=\sigma+\left(\lambda_{s}-\mu_{s}\right) \operatorname{tr}\left[\vartheta_{D v}-2 \vartheta_{w} \vartheta_{L}\right] .\end{cases}
$$

Then the operators $L_{v}$ and $L_{w}$ are given by

$$
\begin{aligned}
L_{v}(u, h ; \omega, \vartheta) & =-\mu_{s} \widetilde{\Delta}_{\Sigma_{\omega}} v-\lambda_{s} \nabla_{\Sigma_{\omega}} \operatorname{div}_{\Sigma_{\omega}} v-\llbracket \mu \partial_{\nu_{\Sigma_{\omega}}} v \rrbracket-c_{5} \llbracket \mu \nabla_{\Sigma_{\omega}} w \rrbracket+\vartheta_{1} \nabla_{\Sigma_{\omega}} \Delta_{\Sigma_{\omega}} h, \\
L_{w}(u, \pi, h ; \omega, \vartheta) & =-\operatorname{tr}\left(\left[\vartheta_{2}+2 \vartheta_{3}\right] \nabla_{\Sigma_{\omega}} v\right)-2 \llbracket \mu \partial_{\nu_{\Sigma_{\omega}}} w \rrbracket+\llbracket \pi \rrbracket-\operatorname{tr}\left(\left[\vartheta_{\sigma}+\vartheta_{4}\right] \nabla_{\Sigma_{\omega}}^{2} h\right) .
\end{aligned}
$$

More details on these operators will be given in Figure 3.8 on page 73 and Section 4.3.
We will prove optimal regularity for problem (3.25) for the following class of parameters and by using the function spaces from Figure 3.7 on the next page.
3.15. Definition. Given $M, T, \eta, R \in(0, \infty)$, the set $\mathcal{P}_{M, T, \eta, R}$ consists of all $\left(\vartheta^{*}, \omega, \vartheta\right)$ such that
(i) the constant tuple $\vartheta^{*}=\left(\vartheta_{L}^{*}, \vartheta_{w}^{*}, \vartheta_{D v}^{*}\right)$ belongs to the parameter set $\mathcal{P}_{M}$ from page 54,
(ii) the map $\omega \in B C^{4}\left(\mathbb{R}^{n-1}\right)$ satisfies $\|\omega\|_{B C^{1} \cap H_{p}^{2}} \leq \eta,\|\omega\|_{B C^{4}} \leq R$, and $\omega(0)=|\nabla \omega(0)|=0$,

$$
\begin{aligned}
&{ }_{0} \mathbb{E}_{u}:=\left\{u \in{ }_{0} H_{p}^{1}\left(J ; L_{p}(\Omega)^{n}\right) \cap L_{p}\left(J ; H_{p}^{2}(\Omega \backslash \Sigma)^{n}\right):\left.u\right|_{\partial \Omega}=0\right\}, \\
&{ }_{0} \mathbb{E}_{v}:={ }_{0} W_{p}^{1-1 / 2 p}\left(J ; L_{p}(\Sigma ; T \Sigma)\right) \\
& \cap{ }_{0} W_{p}^{1 / 2-1 / 2 p}\left(J ; H_{p}^{2}(\Sigma ; T \Sigma)\right) \cap L_{p}\left(J ; W_{p}^{3-1 / p}(\Sigma ; T \Sigma)\right), \\
&{ }_{0} \mathbb{E}_{w}:={ }_{0} W_{p}^{1-1 / 2 p}\left(J ; H_{p}^{1}(\Sigma)\right) \cap L_{p}\left(J ; W_{p}^{3-1 / p}(\Sigma)\right), \\
&{ }_{0} \mathbb{E}_{u, v, w, \partial_{\nu} w}:=\left\{u \in{ }_{0} \mathbb{E}_{u}: \llbracket u \rrbracket=0,\left.v\right|_{\Sigma} \in{ }_{0} \mathbb{E}_{v},\left.w\right|_{\Sigma} \in{ }_{0} \mathbb{E}_{w}, \partial_{\nu} w_{ \pm} \in{ }_{0} \mathbb{G}_{w}\right\}, \\
& \mathbb{E}_{\pi}:=L_{p}\left(J ; \dot{H}_{p}^{1}(\Omega \backslash \Sigma)\right), \\
&{ }_{0} \mathbb{E}_{\pi, \llbracket \pi \rrbracket}:=\left\{\pi \in \mathbb{E}_{\pi}: \llbracket \pi \rrbracket \in{ }_{0} \mathbb{G}_{w}\right\}, \\
&{ }_{0} \mathbb{E}_{h}:={ }_{0} W_{p}^{2-1 / 2 p}\left(J ; H_{p}^{1}(\Sigma)\right) \cap{ }_{0} H_{p}^{1}\left(J ; W_{p}^{3-1 / p}(\Sigma)\right) \cap L_{p}\left(J ; W_{p}^{4-1 / p}(\Sigma)\right), \\
& \mathbb{F}_{u}:=L_{p}(J \times \Omega)^{n}, \\
&{ }_{0} \mathbb{F}_{d}:={ }_{0} H_{p}^{1}\left(J ; \hat{H}_{p}^{-1}(\Omega)\right) \cap L_{p}\left(J ; H_{p}^{1}(\Omega \backslash \Sigma)\right), \\
&{ }_{0} \mathbb{F}_{d, \Sigma}:=\left\{f_{d} \in{ }_{0} \mathbb{F}_{d}: f_{d \pm} \mid \Sigma \in{ }_{0} \mathbb{G}_{w}\right\}, \\
&{ }_{0} \mathbb{G}_{v}:={ }_{0} W_{p}^{1 / 2-1 / 2 p}\left(J ; L_{p}(\Sigma ; T \Sigma)\right) \cap L_{p}\left(J ; W_{p}^{1-1 / p}(\Sigma ; T \Sigma)\right), \\
&{ }_{0} \mathbb{G}_{w}:={ }_{0} W_{p}^{1 / 2-1 / 2 p}\left(J ; H_{p}^{1}(\Sigma)\right) \cap L_{p}\left(J ; W_{p}^{2-1 / p}(\Sigma)\right), \\
&{ }_{0} \mathbb{G}_{h}:={ }_{0} W_{p}^{1-1 / 2 p}\left(J ; H_{p}^{1}(\Sigma)\right) \cap L_{p}\left(J ; W_{p}^{3-1 / p}(\Sigma)\right) .
\end{aligned}
$$

The spaces $L_{p}, H_{p}^{k}$, and $W_{p}^{s}\left(p \in(1, \infty), k \in \mathbb{N}_{0}, s \in[0, \infty)\right)$ are endowed with the intrinsic norm (B.5). The corresponding spaces $\mathbb{E} \ldots, \mathbb{F} \ldots$, and $\mathbb{G} \ldots$ are defined by replacing ${ }_{0} W_{p}^{s}$ by $W_{p}^{s}$ and ${ }_{0} H_{p}^{k}$ by $H_{p}^{k}$. The scalar-valued versions of ${ }_{0} \mathbb{E}_{v}$ and ${ }_{0} \mathbb{G}_{v}$ are denoted by the same symbol.

Figure 3.7. Function spaces ${ }_{0} \mathbb{E} \ldots,{ }_{0} \mathbb{F} \ldots$, and ${ }_{0} \mathbb{G} \ldots$ on $(J, \Omega, \Sigma)$.
(iii) the triple $\vartheta=\left(\vartheta_{L}, \vartheta_{w}, \vartheta_{D v}\right)$ consists of functions $\vartheta_{L}: \Sigma_{\omega} \rightarrow \mathbb{R}^{n \times n}$ and $\left(\vartheta_{w}, \vartheta_{D v}\right):(0, T) \times$ $\Sigma_{\omega} \rightarrow \mathbb{R} \times \mathbb{R}^{n \times n}$ that satisfy the inequalities

$$
\begin{aligned}
\left\|\vartheta_{L}-\vartheta_{L}^{*}\right\|_{B C\left(\Sigma_{\omega}\right) \cap H_{p}^{1}\left(\Sigma_{\omega}\right)} \leq \eta, \quad\left\|\vartheta_{L}-\vartheta_{L}^{*}\right\|_{B C^{2}\left(\Sigma_{\omega}\right)} \leq R, \\
\left\|\vartheta_{w}-\vartheta_{w}^{*}\right\|_{C\left([0, T] ; B C\left(\Sigma_{\omega}\right) \cap H_{p}^{1}\left(\Sigma_{\omega}\right)\right)} \leq \eta, \quad\left\|\vartheta_{w}-\vartheta_{w}^{*}\right\|_{\mathbb{G}_{w}(T)} \leq R \\
\left\|\vartheta_{D v}-\vartheta_{D v}^{*}\right\|_{C\left([0, T] ; B C\left(\Sigma_{\omega}\right) \cap H_{p}^{1}\left(\Sigma_{\omega}\right)\right)} \leq \eta, \quad\left\|\vartheta_{D v}-\vartheta_{D v}^{*}\right\|_{\mathbb{G}_{w}(T)} \leq R .
\end{aligned}
$$

3.16. Theorem. Let $\rho_{ \pm,} \mu_{ \pm}, \sigma, \mu_{s}, \lambda_{s}+\mu_{s}>0$ and let $p \in(n+2, \infty), M>0$, and $T_{1}>0$ be fixed. Then there exists $\eta>0$ such that for given $R>0$ we can find a number $T_{0} \in\left(0, T_{1}\right]$ such that the solution-to-data map

$$
\begin{aligned}
(u, \pi, h) & \mapsto\left(f_{u}, f_{d}, g_{v}, g_{w}, g_{h}\right), \\
{ }_{0} \mathbb{E}:={ }_{0} \mathbb{E}_{u, v, w, \partial_{\nu} w} \times{ }_{0} \mathbb{E}_{\pi, \llbracket \pi \rrbracket} \times{ }_{0} \mathbb{E}_{h} & \rightarrow{ }_{0} \mathbb{F}:=\mathbb{F}_{u} \times{ }_{0} \mathbb{F}_{d, \Sigma} \times{ }_{0} \mathbb{G}_{v} \times{ }_{0} \mathbb{G}_{w} \times{ }_{0} \mathbb{G}_{h}
\end{aligned}
$$

of problem (3.25) is uniformly invertible with respect to $T \in\left(0, T_{0}\right]$ and $\left(\vartheta^{*}, \omega, \vartheta\right) \in \mathcal{P}_{M, T_{1}, \eta, R}$.
We point out that the number $\eta$ depends on the bound $M$ for $\vartheta^{*}$ but not on the bound $R$ for $\omega$ and $\vartheta$. This will be important for the localization procedure for a bounded domain. The proof of Theorem 3.16 will be reduced to an application of Theorem 3.14 for the flat interface problem (3.21). To this end, we consider the usual defining diffeomorphism

$$
\begin{equation*}
\Theta_{\omega}\left(x^{\prime}, x_{n}\right)=\left(x^{\prime}, x_{n}+\omega\left(x^{\prime}\right)\right) \quad \text { for } x^{\prime} \in \mathbb{R}^{n-1}, x_{n} \in \mathbb{R} \tag{3.27}
\end{equation*}
$$

Let $\partial=\left(\partial^{\prime}, \partial_{n}\right), \partial^{\prime}=\left(\partial_{1}, \ldots, \partial_{n-1}\right), \nabla=\partial^{\top}, \nabla^{\prime}=\partial^{\prime \top}$ and $\Sigma_{\omega}=\left\{\left(x^{\prime}, \omega\left(x^{\prime}\right)\right): x^{\prime} \in \mathbb{R}^{n-1}\right\}$. The defining diffeomorphism $\Theta_{\omega}\left(x^{\prime}, x_{n}\right)=\left(x^{\prime}, x_{n}+\omega\left(x^{\prime}\right)\right)$ for $\Sigma_{\omega}$ satisfies

$$
\partial \Theta_{\omega}=I+e_{n} \otimes \nabla \omega=\left[\begin{array}{cc}
I^{\prime} & 0 \\
\partial^{\prime} \omega & 1
\end{array}\right], \quad\left[\partial \Theta_{\omega}\right]^{-1}=I-e_{n} \otimes \nabla \omega=\left[\begin{array}{cc}
I^{\prime} & 0 \\
-\partial^{\prime} \omega & 1
\end{array}\right]
$$

The tangent vectors $\tau_{j}$, cotangent vectors $\tau^{j}$, unit normal vector $\nu$, metric components $g_{j k}$, $g^{j k}$ and Christoffel symbols $\Lambda_{j k, l}, \Lambda_{k j}^{l}$ of $\Sigma_{\omega}$ are given by

$$
\begin{aligned}
\nu \circ \Theta_{\omega} & =\beta\left(e_{n}-\nabla \omega\right), & \beta & =\left(1+|\nabla \omega|^{2}\right)^{-1 / 2}, \\
\tau_{j} \circ \Theta_{\omega} & =e_{j}+\partial_{j} \omega e_{n}, & \tau^{j} \circ \Theta_{\omega} & =e_{j}+\beta^{2} \partial_{j} \omega\left(e_{n}-\nabla \omega\right), \\
g_{j k} \circ \Theta_{\omega} & =\delta_{j k}+\partial_{j} \omega \partial_{k} \omega, & g^{j k} \circ \Theta_{\omega} & =\delta_{j k}-\beta^{2} \partial_{j} \omega \partial_{k} \omega, \\
\Lambda_{j k, l} \circ \Theta_{\omega} & =\partial_{j} \partial_{k} \omega \partial_{l} \omega, & \Lambda_{j k}^{l} \circ \Theta_{\omega} & =\beta^{2} \partial_{j} \partial_{k} \omega \partial_{l} \omega .
\end{aligned}
$$

The projections $P^{\prime}=I-e_{n} \otimes e_{n}$ and $P_{\Sigma_{\omega}}=I-\nu_{\Sigma_{\omega}} \otimes \nu_{\Sigma_{\omega}}$ satisfy

$$
P^{\prime}=\left[\begin{array}{ll}
I^{\prime} & 0 \\
0 & 0
\end{array}\right], \quad P_{\Sigma_{\omega}} \circ \Theta_{\omega}=\left[\begin{array}{cc}
I^{\prime}-\beta^{2} \nabla^{\prime} \omega \otimes \nabla^{\prime} \omega & \beta^{2} \nabla^{\prime} \omega \\
\beta^{2} \partial^{\prime} \omega & 1-\beta^{2}
\end{array}\right]
$$

For a scalar field $\varphi$, a tangential vector field $v$ and a not necessarily tangential vector field $u$, the gradient $\nabla_{\Sigma_{\omega}} \varphi, \nabla_{\Sigma_{\omega}} u$, divergence $\operatorname{div}_{\Sigma_{\omega}} u$, scalar Laplace-Beltrami operator $\Delta_{\Sigma_{\omega}} \varphi=$ $\operatorname{div}_{\Sigma_{\omega}} \nabla_{\Sigma_{\omega}} \varphi$ and tangential Laplace-Beltrami operator $\widetilde{\Delta}_{\Sigma_{\omega}} v=g^{j k} \widetilde{\nabla}_{j} \widetilde{\nabla}_{k} v$ are given by

$$
\begin{array}{rlrl}
\nabla_{\Sigma_{\omega}} \varphi & =\tau^{j} \partial_{j} \varphi, & \nabla_{\Sigma_{\omega}} u & =\tau^{j} \otimes \partial_{j} u \\
\Delta_{\Sigma_{\omega}} \varphi=g^{j k}\left(\partial_{j} \partial_{k} \varphi-\Lambda_{j k}^{l} \partial_{l} \varphi\right), & \operatorname{div}_{\Sigma_{\omega}} u & =\tau^{j} \cdot \partial_{j} u \\
\widetilde{\Delta}_{\Sigma_{\omega}} v & =g^{j k} P_{\Sigma_{\omega}} \partial_{j}\left(P_{\Sigma_{\omega}} \partial_{k} v\right)
\end{array}
$$

FIGURE 3.8. Differential geometric identities for bent hyperplanes.

Since $x^{\prime} \mapsto \Theta_{\omega}\left(x^{\prime}, 0\right)$ is a global parametrization for $\Sigma_{\omega}$, we can compute the relevant differential geometric quantities of $\Sigma_{\omega}$ by a straightforward application of Appendix A. The relevant identities are collected in Figure 3.8 on this page.

In Lemma 3.17 we will prove that the induced transformations for solutions and data induce isomorphisms between the function spaces in $\mathbb{R}^{n} \backslash \Sigma_{\omega}$ and the corresponding spaces in $\dot{\mathbb{R}}^{n}=\mathbb{R}^{n} \backslash \Sigma_{0}$, which are uniformly bounded and uniformly invertible with respect to $\omega$. We also derive transformation identities for the velocity components $v$ and $w$ which are collected in Figure 3.9 on the next page. These will be employed for deriving the transformed version (3.48) of problem (3.25). This transformed problem corresponds to the basic model problem (3.21) with additional perturbations. We will control those pertubations by means of appropriate interval-dependent estimates and estimates for pointwise multiplication and continuous embeddings (see Lemmas 3.18 and 3.19). For proving Theorem 3.16, we require smallness of $\eta$ in order to control the leading-order perturbations and smallness of $T$ for controlling the lower-order perturbations.
3.17. Lemma. Let $\omega \in B C^{4}\left(\mathbb{R}^{n-1}\right)$ and $J=(0, T)$ with $T \in(0, \infty]$. Consider the transformations

$$
\begin{aligned}
& (u, \pi, h) \circ \Theta_{\omega}=\left(\left[\partial \Theta_{\omega}\right] \bar{u}, \bar{\pi}, \bar{h}\right), \\
& \left(f_{u}, f_{d}, g_{u}, g_{h}\right) \circ \Theta_{\omega}=\left(\left[\partial \Theta_{\omega}\right] \bar{f}_{u}, \bar{f}_{d},\left[\partial \Theta_{\omega}\right] \bar{g}_{u}, \bar{g}_{h}\right),
\end{aligned}
$$

and the decompositions $u=v+w \nu_{\Sigma_{\omega}} \bar{u}=\bar{v}+\bar{w} e_{n}, g_{u}=g_{v}+g_{w} \nu_{\Sigma_{\omega}}$, and $\bar{g}_{u}=\bar{g}_{v}+\bar{g}_{w} e_{n}$.

Define $\Theta_{\omega}$ as in Figure 3.8 and let

$$
u \circ \Theta_{\omega}=\left[\partial \Theta_{\omega}\right] \bar{u}, \quad u=v+w \nu_{\Sigma_{\omega}}, \quad \bar{u}=\bar{v}+\bar{w} e_{n} .
$$

Then the following identities are valid.
$\bar{v}=P^{\prime}\left(v \circ \Theta_{\omega}\right)+Q_{v}(\omega) w \circ \Theta_{\omega}$,
$Q_{v}(\omega)=-\beta \nabla \omega$,
(3.29b)
$\bar{w}=w \circ \Theta_{\omega}+Q_{w}(\omega) w \circ \Theta_{w}$,
$Q_{w}(\omega)=\beta^{-1}-1$,

$$
\begin{equation*}
\partial_{n} \bar{w}=\beta^{2}\left(\partial_{\nu} w\right) \circ \Theta_{\omega}+\beta^{3} \partial_{j} \omega\left(\partial_{j} w\right) \circ \Theta_{\omega}, \tag{3.29c}
\end{equation*}
$$ $v \circ \Theta_{\omega}=\left[I+e_{n} \otimes \nabla \omega\right] \bar{v}+\left\{\left(1-\beta^{2}\right) e_{n}-\beta^{2} \nabla \omega\right\} \bar{w}$,

$w \circ \Theta_{\omega}=\bar{w}+(\beta-1) \bar{w}$,

Their derivations are given in the proof of Lemma 3.17.
Figure 3.9. Identities for the transformed velocity field.

Then, given $R>0$, the operators

| (3.28a) | $\bar{u} \mapsto u$, | $\mathbb{E}_{u}\left(\mathbb{R}^{n} \backslash \Sigma_{0}\right) \rightarrow \mathbb{E}_{u}\left(\mathbb{R}^{n} \backslash \Sigma_{\omega}\right)$, |
| :---: | :---: | :---: |
| (3.28b) | $(\bar{v}, \bar{w}) \mapsto(v, w)$, | $\mathbb{E}_{v}\left(\Sigma_{0}\right) \times \mathbb{E}_{w}\left(\Sigma_{0}\right) \rightarrow \mathbb{E}_{v}\left(\Sigma_{\omega}\right) \times \mathbb{E}_{w}\left(\Sigma_{\omega}\right)$, |
| (3.28c) | $\bar{u} \mapsto u$, | $\mathbb{E}_{u, v, w, \partial_{n} w}\left(\mathbb{R}^{n} \backslash \Sigma_{0}\right) \rightarrow \mathbb{E}_{u, v, w, \partial_{\nu} w}\left(\mathbb{R}^{n} \backslash \Sigma_{\omega}\right)$, |
| (3.28d) | $\bar{\pi} \mapsto \pi$, | $\mathbb{E}_{\pi}\left(\mathbb{R}^{n} \backslash \Sigma_{0}\right) \rightarrow \mathbb{E}_{\pi}\left(\mathbb{R}^{n} \backslash \Sigma_{\omega}\right)$, |
| (3.28e) | $\bar{\pi} \mapsto \pi$, | $\mathbb{E}_{\pi,[\pi]]}\left(\mathbb{R}^{n} \backslash \Sigma_{0}\right) \rightarrow \mathbb{E}_{\pi,[\pi]]}\left(\mathbb{R}^{n} \backslash \Sigma_{\omega}\right)$, |
| (3.28f) | $\bar{h} \mapsto h$, | $\mathbb{E}_{h}\left(\Sigma_{0}\right) \rightarrow \mathbb{E}_{h}\left(\Sigma_{\omega}\right)$, |
| (3.28g) | $\bar{f}_{u} \mapsto f_{u}$, | $\mathbb{F}_{u}\left(\mathbb{R}^{n} \backslash \Sigma_{0}\right) \rightarrow \mathbb{F}_{u}\left(\mathbb{R}^{n} \backslash \Sigma_{\omega}\right)$, |
| (3.28h) | $\bar{f}_{d} \mapsto f_{d}$, | $\mathbb{F}_{d}\left(\mathbb{R}^{n} \backslash \Sigma_{0}\right) \rightarrow \mathbb{F}_{d}\left(\mathbb{R}^{n} \backslash \Sigma_{\omega}\right)$, |
| (3.28i) | $\bar{f}_{d} \mapsto f_{d}$, | $\mathbb{F}_{d, \Sigma}\left(\mathbb{R}^{n} \backslash \Sigma_{0}\right) \rightarrow \mathbb{F}_{d, \Sigma}\left(\mathbb{R}^{n} \backslash \Sigma_{\omega}\right)$, |
| (3.28j) | $\left(\bar{g}_{v}, \bar{g}_{w}\right) \mapsto\left(g_{v}, g_{w}\right)$, | $\mathbb{G}_{v}\left(\Sigma_{0}\right) \times \mathbb{G}_{w}\left(\Sigma_{0}\right) \rightarrow \mathbb{G}_{v}\left(\Sigma_{\omega}\right) \times \mathbb{G}_{w}\left(\Sigma_{\omega}\right)$, |
| (3.28k) | $\bar{g}_{h} \mapsto g_{h}$, | $\mathbb{G}_{h}\left(\Sigma_{0}\right) \rightarrow \mathbb{G}_{h}\left(\Sigma_{\omega}\right)$ |

are uniformly bounded and uniformly invertible with respect to $\|\nabla \omega\|_{B C^{3}} \leq R$ and $T \in(0, \infty]$.
Proof. (i) In order to estimate the norms of transformed functions, we employ the chain rule from Remark B. 85 on page 171 for representing their derivatives. Define $\Theta=\Theta_{\omega}$ as in (3.27).
(i.a) For $\bar{x} \in \mathbb{R}^{n}$ and $\bar{\alpha} \in \mathbb{R}^{n}$ we put $x=\Theta(\bar{x})$ and $\alpha=[\partial \Theta(\bar{x})] \bar{\alpha}$. Then the derivatives of $\Theta$ and $\Theta^{-1}$ read as follows.

$$
\begin{aligned}
{[\partial \Theta(\bar{x})] \bar{\alpha} } & =\bar{\alpha}+e_{n}\left(\nabla \omega\left(\bar{x}^{\prime}\right) \mid \bar{\alpha}\right), & {\left[\partial^{j} \Theta(\bar{x})\right]\left(\bar{\alpha}_{1}, \ldots, \bar{\alpha}_{j}\right) } & =e_{n}\left[\partial^{j} \omega\left(\bar{x}^{\prime}\right)\right]\left(\bar{\alpha}_{1}, \ldots, \bar{\alpha}_{j}\right) \text { for } j \geq 2, \\
{\left[\partial \Theta^{-1}(x)\right] \alpha } & =\alpha-e_{n}\left(\nabla \omega\left(\bar{x}^{\prime}\right) \mid \bar{\alpha}\right), & & \left.\partial^{j} \Theta^{-1}(x)\right]\left(\alpha_{1}, \ldots, \alpha_{j}\right)
\end{aligned}=-e_{n}\left[\partial^{j} \omega\left(\bar{x}^{\prime}\right)\right]\left(\bar{\alpha}_{1}, \ldots, \bar{\alpha}_{j}\right) \text { for } j \geq 2 .
$$

(i.b) For a sufficiently smooth function $\varphi: \mathbb{R}^{n} \rightarrow \mathbb{R}$ we put $\bar{\varphi}:=\varphi \circ \Theta$ and obtain

$$
\begin{aligned}
{[\partial \varphi(x)] \alpha } & =[\partial \bar{\varphi}(\bar{x})] \bar{\alpha}, \\
{\left[\partial^{2} \varphi(x)\right]\left(\alpha_{1}, \alpha_{2}\right) } & =\left[\partial^{2} \bar{\varphi}(\bar{x})\right]\left(\bar{\alpha}_{1}, \bar{\alpha}_{2}\right)-[\partial \bar{\varphi}(\bar{x})][\partial \Theta(\bar{x})]^{-1}\left[\partial^{2} \Theta(\bar{x})\right]\left(\bar{\alpha}_{1}, \bar{\alpha}_{2}\right) .
\end{aligned}
$$

Omitting the arguments $x$ and $\bar{x}$ and the square brackets around derivatives, we have

$$
\begin{aligned}
& \partial^{3} \varphi\left(\alpha_{1}, \alpha_{2}, \alpha_{3}\right) \\
& =\partial^{3} \bar{\varphi}\left(\bar{\alpha}_{1}, \bar{\alpha}_{2}, \bar{\alpha}_{3}\right) \\
& \quad-\partial^{2} \bar{\varphi}\left\{\left(\partial \Theta^{-1} \partial^{2} \Theta\left(\bar{\alpha}_{1}, \bar{\alpha}_{2}\right), \bar{\alpha}_{3}\right)+\left(\partial \Theta^{-1} \partial^{2} \Theta\left(\bar{\alpha}_{1}, \bar{\alpha}_{3}\right), \bar{\alpha}_{2}\right)+\left(\partial \Theta^{-1} \partial^{2} \Theta\left(\bar{\alpha}_{2}, \bar{\alpha}_{3}\right), \bar{\alpha}_{1}\right)\right\} \\
& \quad+\partial \bar{\varphi} \partial \Theta^{-1} \partial^{2} \Theta\left\{\left(\partial \Theta^{-1} \partial^{2} \Theta\left(\bar{\alpha}_{1}, \bar{\alpha}_{2}\right), \bar{\alpha}_{3}\right)+\left(\partial \Theta^{-1} \partial^{2} \Theta\left(\bar{\alpha}_{1}, \bar{\alpha}_{3}\right), \bar{\alpha}_{2}\right)+\left(\partial \Theta^{-1} \partial^{2} \Theta\left(\bar{\alpha}_{2}, \bar{\alpha}_{3}\right), \bar{\alpha}_{1}\right)\right\} \\
& \quad-\partial \bar{\varphi} \partial \Theta^{-1} \partial^{3} \Theta\left(\bar{\alpha}_{1}, \bar{\alpha}_{2}, \bar{\alpha}_{3}\right) .
\end{aligned}
$$

(i.c) For $\bar{u}: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ we let $u(x):=[\partial \Theta(\bar{x})] \bar{u}(\bar{x})$ and write $u=\partial \Theta \bar{u}$, to be short. Then

$$
\begin{aligned}
\partial u \alpha_{1}= & \partial \Theta \partial \bar{u} \bar{\alpha}+\partial^{2} \Theta(\bar{u}, \bar{\alpha}) \\
\partial^{2} u\left(\alpha_{1}, \alpha_{2}\right)= & \partial \Theta \partial^{2} \bar{u}\left(\bar{\alpha}_{1}, \bar{\alpha}_{2}\right)+\partial^{2} \Theta\left\{\left(\partial \bar{u} \bar{\alpha}_{1}, \bar{\alpha}_{2}\right)+\left(\bar{\alpha}_{1}, \partial \bar{u}^{2}\right)\right\} \\
& -\partial \Theta \partial \bar{u} \partial \Theta^{-1} \partial^{2} \Theta\left(\bar{\alpha}_{1}, \bar{\alpha}_{2}\right)-\partial^{2} \Theta\left(\bar{u}, \partial \Theta^{-1} \partial^{2} \Theta\left(\bar{\alpha}_{1}, \bar{\alpha}_{2}\right)\right)+\partial^{3} \Theta\left(\bar{u}, \bar{\alpha}_{1}, \bar{\alpha}_{2}\right) .
\end{aligned}
$$

(ii) Lemma B. 10 on page 148 yields the pointwise multiplication estimate

$$
\begin{equation*}
\llbracket u v \rrbracket_{W_{p}^{\sigma}\left(\Sigma_{\omega}\right)} \leq\|u\|_{L_{\infty}\left(\Sigma_{\omega}\right)} \llbracket v \rrbracket_{W_{p}^{\sigma}\left(\Sigma_{\omega}\right)}+C\left(n, s, p,\|\nabla \omega\|_{B C^{1}}\right)\|u\|_{W_{\infty}^{1}\left(\Sigma_{\omega}\right)}\|v\|_{L_{p}(\Sigma)} \tag{3.30}
\end{equation*}
$$

for $u \in W_{\infty}^{1}\left(\Sigma_{\omega}\right), v \in W_{p}^{\sigma}\left(\Sigma_{\omega}\right), \sigma \in(0,1), p \in[1, \infty)$.
(iii) Now we are prepared for proving that the operators (3.28) are uniformly invertible. We will frequently employ the differential geometric identities in Figure 3.8.
(iii.a) The invertibility of $\bar{u} \mapsto u$ in (3.28a), $\bar{\pi} \mapsto \pi$ in (3.28d) and $\bar{f}_{u} \mapsto f_{u}$ in (3.28g) easily follows from (i). These operators are uniformly invertible with respect to $\|\nabla \omega\|_{B C^{2}} \leq M$.
(iii.b) We recall that $W_{p}^{s}\left(\Sigma_{\omega}\right)$ is equipped with the intrinsic Sobolev-Slobodeckiŭ norm (B.5). Hence, from $\nabla_{\Sigma_{\omega}}^{k+1} h=\left(\nabla_{\Sigma_{\omega}} \otimes\right)^{k} \nabla_{\Sigma_{\omega}} h(k \leq 2)$ and (3.30) we infer that the operator $\bar{h} \mapsto h$ in (3.28f) is invertible, uniformly with respect to $\|\nabla \omega\|_{B C^{3}} \leq M$. In a similar way we see that $\bar{g}_{h} \mapsto g_{h}$ in (3.28k) is uniformly invertible with respect to $\|\nabla \omega\|_{B C^{2}} \leq M$.
(iii.c) The transformed normal velocity satisfies

$$
\begin{aligned}
\bar{w}=e_{n} \cdot \bar{u} & =e_{n} \cdot[\partial \Theta]^{-1}\left(v+\nu_{\Sigma_{\omega}} w\right) \circ \Theta_{\omega} \\
& =\left(1+|\nabla \omega|^{2}\right)^{1 / 2} w \circ \Theta_{\omega} \\
& =w \circ \Theta_{\omega}+Q_{w}(\omega) w \circ \Theta_{\omega} \quad \text { with } Q_{w}(\omega):=\left(1+|\nabla \omega|^{2}\right)^{1 / 2}-1=\beta^{-1}-1 .
\end{aligned}
$$

Hence the identities (3.29b) and (3.29e) are valid and the operator $w \mapsto \bar{w}, \mathbb{E}_{w}\left(\Sigma_{\omega}\right) \rightarrow \mathbb{E}_{w}\left(\Sigma_{0}\right)$ is uniformly invertible with respect to $\|\nabla \omega\|_{B C^{3}} \leq M$.
(iii.d) The projection $P^{\prime}=I-e_{n} \otimes e_{n}$ satisfies

$$
P^{\prime}\left[\partial \Theta_{\omega}\right]^{-1}\left[P_{\Sigma_{\omega}} \circ \Theta_{\omega}\right]=P^{\prime}+\beta \nabla \omega \otimes\left(\nu_{\Sigma_{\omega}} \circ \Theta_{\omega}\right)
$$

Therefore $v$ is related to $(\bar{v}, \bar{w})$ by

$$
\begin{aligned}
\bar{v}=P^{\prime} \bar{u} & =P^{\prime}\left[\partial \Theta_{\omega}\right]^{-1}\left(P_{\Sigma_{\omega}} v+w \nu_{\Sigma_{\omega}}\right) \circ \Theta_{\omega} \\
& =P^{\prime}\left(v \circ \Theta_{\omega}\right)+Q_{v}(\omega) w \circ \Theta_{\omega} \quad \text { with } Q_{v}(\omega):=-\nabla \omega\left(1+|\nabla \omega|^{2}\right)^{-1 / 2}=-\beta \nabla \omega .
\end{aligned}
$$

This yields (3.29a). With $\mathbb{E}_{w} \hookrightarrow \mathbb{E}_{v}$, it follows that $(v, w) \mapsto \bar{v}, \mathbb{E}_{v}\left(\Sigma_{\omega}\right) \times{ }_{0} \mathbb{E}_{w}\left(\Sigma_{\omega}\right) \rightarrow{ }_{0} \mathbb{E}_{v}\left(\Sigma_{0}\right)$ is uniformly bounded with respect to $\|\nabla \omega\|_{B C^{3}} \leq M$. The inverse representation is given by

$$
\begin{aligned}
v \circ \Theta_{\omega} & =\left[P_{\Sigma_{\omega}} \circ \Theta_{\omega}\right]\left[I+e_{n} \otimes \nabla \omega\right]\left(\bar{v}+\bar{w} e_{n}\right) \\
& =\left[I+e_{n} \otimes \nabla \omega\right] \bar{v}+\left\{\left(1-\beta^{2}\right) e_{n}+\beta^{2} \nabla \omega\right\} \bar{w} .
\end{aligned}
$$

Therefore identity (3.29d) is true and the operator $(\bar{v}, \bar{w}) \mapsto(v, w)$ in (3.28b) is uniformly invertible with respect to $\|\nabla \omega\|_{B C^{3}} \leq M$.
(iii.e) With $w \circ \Theta_{\omega}=\beta \bar{w}$ we obtain

$$
\begin{aligned}
\left(\partial_{\nu} w\right) \circ \Theta_{\omega} & =\beta\left(e_{n}-\nabla \omega\right) \cdot \nabla\left((\beta \bar{w}) \circ \Theta_{\omega}^{-1}\right) \circ \Theta_{\omega} \\
& =\beta\left[I-e_{n} \otimes \nabla \omega\right]\left(e_{n}-\nabla \omega\right) \cdot \nabla(\beta \bar{w}) \\
& =\beta\left(e_{n}-\nabla \omega+e_{n}|\nabla \omega|^{2}\right) \cdot \nabla(\beta \bar{w}) \\
& =\partial_{n} \bar{w}-\beta \nabla^{\prime} \omega \cdot \nabla^{\prime}(\beta \bar{w}), \\
\partial_{n} \bar{w} & =\beta e_{n} \cdot \nabla\left(w \circ \Theta_{\omega}\right) \\
& =\beta\left[I+e_{n} \otimes \nabla \omega\right] e_{n} \cdot\left(\nu \partial_{\nu} w+\tau^{j} \partial_{j} w\right) \circ \Theta_{\omega} \\
& =\beta^{2}\left(\partial_{\nu} w\right) \circ \Theta_{\omega}+\beta^{3} \partial_{j} \omega\left(\partial_{j} w\right) \circ \Theta_{\omega} .
\end{aligned}
$$

Thus, equations (3.29c) and (3.29f) are valid and the operator $u \mapsto \bar{u}$ in (3.28c) is uniformly invertible with respect to $\|\nabla \omega\|_{B C^{3}} \leq M$.
(iii.f) The relation between $g_{u} \circ \Theta_{\omega}$ and $\bar{g}_{u}$ is analogous to that of $\left.u\right|_{\Sigma} \circ \Theta_{\omega}$ and $\left.\bar{u}\right|_{\Sigma_{0}}$. Hence equations (3.29a), (3.29b), (3.29d) and (3.29e) yield

$$
\begin{align*}
\bar{g}_{v} & =P^{\prime}\left(g_{v} \circ \Theta_{\omega}\right)+Q_{v}(\omega) g_{w} \circ \Theta_{\omega}, \\
g_{v} \circ \Theta_{\omega} & =\left[I+e_{n} \otimes \nabla \omega\right] \bar{g}_{v}+\beta^{2}\left(\nabla \omega-e_{n}|\nabla \omega|^{2}\right) \bar{g}_{w}, \quad g_{w} \circ \Theta_{\omega}=\beta \bar{g}_{w} . \tag{3.31}
\end{align*}
$$

Therefore (3.28j) and (3.28e) are uniformly invertible with respect to $\|\nabla \omega\|_{B C^{2}} \leq M$.
(iii.g) For given $\bar{f}_{d} \in \mathbb{F}_{d}\left(\mathbb{R}^{n} \backslash \Sigma_{0}\right)$ and $\varphi \in \dot{H}_{p^{\prime}}^{1}\left(\mathbb{R}^{n}\right)$, we obtain

$$
\int_{\mathbb{R}^{n}} f_{d} \varphi d x=\int_{\mathbb{R}^{n}} \bar{f}_{d} \varphi \circ \Theta_{\omega} d x, \quad \text { with } \operatorname{det} \partial \Theta_{\omega}=1
$$

The map $\varphi \mapsto \varphi \circ \Theta_{\omega}, \dot{H}_{p^{\prime}}^{1}\left(\mathbb{R}^{n}\right) \rightarrow \dot{H}_{p^{\prime}}^{1}\left(\mathbb{R}^{n}\right)$ is uniformly invertible and therefore

$$
\bar{f}_{d} \mapsto f_{d}, \quad H_{p}^{1}\left(J ; \hat{H}_{p}^{-1}\left(\mathbb{R}^{n}\right)\right) \rightarrow H_{p}^{1}\left(J ; \hat{H}_{p}^{-1}\left(\mathbb{R}^{n}\right)\right)
$$

is uniformly invertible with respect to $\|\nabla \omega\|_{\infty} \leq M$. The estimates for the remaining norms follow similarly as above and therefore (3.28h) and (3.28i) are uniformly invertible with respect to $\|\nabla \omega\|_{B C^{2}} \leq M$. The proof of Lemma 3.17 is complete.

In order to take advantage of short time intervals we will frequently employ the following interval-dependent estimates, where we study the time-dependence of certain embedding constants.
3.18. Lemma. Let $X$ be a Banach space, $J=(0, T)$ with $T \in(0, \infty)$, and $p \in[1, \infty)$. Then

$$
\begin{array}{ll}
\|u\|_{L_{p}(J)} \leq T^{1 / p-1 / q}\|u\|_{L_{q}(J)}, & \text { for } u \in L_{q}(J ; X), q \in[p, \infty] \\
\|v\|_{L_{p}(J)} \leq T \frac{1}{1-1 / p}\left\|\partial_{t} v\right\|_{L_{p}(J)}, & \\
\text { for } v \in{ }_{0} W_{p}^{1}(J ; X), p>1,
\end{array}
$$

Proof. (3.32a) follows from Hölder's inequality. To prove (3.32b), we apply Hardy's inequality (B.4) to $\partial_{t} v$. (3.32c) follows from Lemma B.5. (3.32d) can be verified directly:

$$
\llbracket u \rrbracket_{W_{p}^{\alpha}(0, T)}=\left(\int_{0}^{T} \int_{0}^{T}|t-s|^{(\beta-\alpha) p} \frac{|u(t)-u(s)|_{X}^{p}}{|t-s|^{1+\beta p}} d s d t\right)^{1 / p} \leq T^{\beta-\alpha} \llbracket u \rrbracket_{W_{p}^{\beta}(0, T)} .
$$

Estimate (3.32e) follows from Hardy's inequality as in [PS11, Proposition 5.1].
Next we provide appropriate estimates for controlling perturbations in ${ }_{0} \mathbb{G}_{v, 0} \mathbb{G}_{w}$, and ${ }_{0} \mathbb{G}_{h}$. Basically, such estimates were used in [PSS07] and [PS10].
3.19. Lemma. Let $\Omega$ be a domain in $\mathbb{R}^{n}(n \geq 2)$ containing a smooth (possibly empty) hypersurface $\Sigma$ such that Assumption 2.1 on page 23 is satisfied and let $p \in(1, \infty)$. Then the following assertions are valid.
(i) If $p>2$, then for all $\delta \in(0,1 / 2)$ and $T_{0} \in(0, \infty)$ there exists $C\left(\delta, T_{0}\right)>0$ such that

$$
\begin{equation*}
\left\|\left(T^{-1} u, T^{-\delta} \nabla u\right)\right\|_{\mathbb{F}_{u}(T)} \leq C\left(\delta, T_{0}\right)\|u\|_{0 \mathbb{E}_{u}(T)} \tag{3.33}
\end{equation*}
$$

for all $u \in{ }_{0} \mathbb{E}_{u}(T)$ and $T \in\left(0, T_{0}\right]$.
(ii) If $p>2$, then for all $T_{0} \in(0, \infty)$ there exists $C\left(T_{0}\right)>0$ such that

$$
\begin{equation*}
\left\|\left(T^{-1 / 2} v, T^{-1 / 4} \nabla^{\prime} v\right)\right\|_{0 \mathbb{G}_{v}(T)}+\left\|T^{-1 / 4} v\right\|_{0 \mathbb{G}_{w}(T)} \leq C\left(T_{0}\right)\|v\|_{0 \mathbb{E}_{v}(T)} \tag{3.34}
\end{equation*}
$$

for all $v \in{ }_{0} \mathbb{E}_{v}(T)$ and $T \in\left(0, T_{0}\right]$.
(iii) If $p>2$, then for all $\delta \in(0,1 / 2)$ and $T_{0} \in(0, \infty)$ there exists $C\left(\delta, T_{0}\right)>0$ such that

$$
\begin{equation*}
\left\|\left(T^{-1 / 2} w, T^{-\delta} \nabla^{\prime} w\right)\right\|_{0} \mathbb{G}_{v}(T)+\left\|T^{-\delta} w\right\|_{0} \mathbb{G}_{w}(T) \leq C\left(\delta, T_{0}\right)\|w\|_{0 \mathbb{E}_{w}(T)} \tag{3.35}
\end{equation*}
$$

for all $w \in{ }_{0} \mathbb{E}_{w}(T)$ and $T \in\left(0, T_{0}\right]$.
(iv) If $p>3$, then for all $\delta \in(0,3 / 2)$ and $T_{0} \in(0, \infty)$ there exists $C\left(\delta, T_{0}\right)>0$ such that

$$
\begin{equation*}
\left\|\left(T^{-3 / 2} h, T^{-\delta} \nabla^{\prime} h, T^{-1} \nabla^{\prime 2} h\right)\right\|_{0 \mathbb{G}_{v}(T)}+\left\|\left(T^{-3 / 2} h, T^{-1} \nabla^{\prime} h\right)\right\|_{0 \mathbb{G}_{w}(T)} \leq C\left(\delta, T_{0}\right)\|h\|_{0 \mathbb{E}_{h}(T)} \tag{3.36}
\end{equation*}
$$

for all $h \in{ }_{0} \mathbb{E}_{h}(T)$ and $T \in\left(0, T_{0}\right]$.
(v) There exists $C>0$ such that for all $T \in(0, \infty), \varphi \in B C^{1}(\Sigma)$, and $g_{v} \in \mathbb{G}_{v}(T)$ we have

$$
\begin{equation*}
\left\|\varphi g_{v}\right\|_{\mathbb{G}_{v}(T)} \leq C\|\varphi\|_{\infty}\left\|g_{v}\right\|_{\mathbb{G}_{v}(T)}+C\|\varphi\|_{\infty}^{1 / p}\left\|\nabla^{\prime} \varphi\right\|_{\infty}^{1-1 / p}\left\|g_{v}\right\|_{L_{p}\left(0, T ; L_{p}(\Sigma)\right)} \tag{3.37}
\end{equation*}
$$

and if $p>3$, then for all $\delta \in(0,1 / 2-1 / 2 p)$ and $T_{0} \in(0, \infty)$ there exists $C\left(\delta, T_{0}\right)>0$ such that

$$
\begin{equation*}
\left\|g_{v}\right\|_{L_{p}\left(0, T ; L_{p}(\Sigma)\right)} \leq T^{\delta} C\left(\delta, T_{0}\right)\left\|g_{v}\right\|_{0} \mathbb{G}_{v}(T) \quad \text { for all } g_{v} \in{ }_{0} \mathbb{G}_{v}(T), T \in\left(0, T_{0}\right] \tag{3.38}
\end{equation*}
$$

There exists $C>0$ such that for all $T \in(0, \infty)$ we have
(3.39) $\left\|\vartheta g_{v}\right\|_{\mathbb{G}_{v}(T)} \leq C\|\vartheta\|_{\infty}\left\|g_{v}\right\|_{\mathbb{G}_{v}(T)}+C\|\vartheta\|_{\mathbb{G}_{v}(T)}\left\|g_{v}\right\|_{\infty} \quad$ for all $\vartheta, g_{v} \in \mathbb{G}_{v}(T) \cap L_{\infty}(J \times \Sigma)$, and if $p>n+2$, then we have the continuous embedding

$$
\mathbb{G}_{v}(T) \hookrightarrow C([0, T] ; B C(\Sigma))
$$

and for all $\delta \in(0,1 / 2-(n+2) / 2 p)$ and $T_{0} \in(0, \infty)$ there exists $C\left(\delta, T_{0}\right)>0$ such that

$$
\begin{equation*}
\left\|g_{v}\right\|_{C([0, T] ; B C(\Sigma))} \leq T^{\delta} C\left(\delta, T_{0}\right)\left\|g_{v}\right\|_{0 \mathbb{G}_{v}(T)} \quad \text { for all } g_{v} \in{ }_{0} \mathbb{G}_{v}(T), T \in(0, \infty) \tag{3.40}
\end{equation*}
$$

(vi) There exists $C>0$ such that for all $T \in(0, \infty), \varphi \in B C^{2}\left(\mathbb{R}^{n-1}\right)$, and $g_{w} \in \mathbb{G}_{w}(T)$ we have

$$
\begin{equation*}
\left\|\varphi g_{w}\right\|_{\mathbb{G}_{w}(T)} \leq C\|\varphi\|_{L_{\infty} \cap H_{p}^{1}}\left\|g_{w}\right\|_{\mathbb{G}_{w}(T)}+C\|\varphi\|_{B C^{2}}\left\|g_{w}\right\|_{L_{p}\left(0, T ; H_{p}^{1}(\Sigma)\right)} \tag{3.41}
\end{equation*}
$$

and if $p>3$, then for all $\delta \in(0,1 / 2-1 / 2 p)$ and $T_{0} \in(0, \infty)$ there exists $C\left(\delta, T_{0}\right)>0$ such that

$$
\begin{equation*}
\left\|g_{w}\right\|_{L_{p}\left(0, T ; H_{p}^{1}(\Sigma)\right)} \leq T^{\delta} C\left(\delta, T_{0}\right)\left\|g_{w}\right\|_{0 \mathbb{G}_{w}(T)} \quad \text { for } g_{w} \in{ }_{0} \mathbb{G}_{w}(T), T \in\left(0, T_{0}\right] \tag{3.42}
\end{equation*}
$$

If $p \in(n+2, \infty)$, then we have the continuous embedding

$$
\mathbb{G}_{w}(T) \hookrightarrow \tilde{\mathbb{G}}_{w}(T):=C\left([0, T] ; H_{p}^{1}(\Sigma)\right) \cap L_{p}\left(0, T ; B C^{1}(\Sigma)\right)
$$

there exists $C>0$ such that for all $T \in(0, \infty), \vartheta \in \mathbb{G}_{w}(T)$, and $g_{w} \in \mathbb{G}_{w}(T)$ we have

$$
\begin{equation*}
\left\|\vartheta g_{w}\right\|_{\mathbb{G}_{w}(T)} \leq C\|\vartheta\|_{C\left([0, T] ; H_{p}^{1}(\Sigma)\right)}\left\|g_{w}\right\|_{\mathbb{G}_{w}(T)}+C\|\vartheta\|_{\mathbb{G}_{w}(T)}\left\|g_{w}\right\|_{\tilde{\mathbb{G}}_{w}(T)} \tag{3.43}
\end{equation*}
$$

and for all $\delta \in(0,1 / 2-3 / 2 p)$ and $T_{0} \in(0, \infty)$ there exists $C\left(\delta, T_{0}\right)>0$ such that

$$
\begin{equation*}
\left\|g_{w}\right\|_{0 \tilde{\mathbb{G}}_{w}(T)} \leq T^{\delta} C\left(\delta, T_{0}\right)\left\|g_{w}\right\|_{0 \mathbb{G}_{w}(T)} \quad \text { for all } g_{w} \in{ }_{0} \mathbb{G}_{w}(T), T \in(0, \infty) \tag{3.44}
\end{equation*}
$$

(vii) There is $C>0$ such that for all $T \in(0, \infty), \varphi \in B C^{3}\left(\mathbb{R}^{n-1}\right)$, and $g_{h} \in \mathbb{G}_{h}(T)$ we have

$$
\begin{equation*}
\left\|\varphi g_{h}\right\|_{\mathbb{G}_{h}(T)} \leq C\|\varphi\|_{L_{\infty} \cap H_{p}^{1}}\left\|g_{h}\right\|_{\mathbb{G}_{h}(T)}+C\|\varphi\|_{B C^{3}}\left\|g_{h}\right\|_{L_{p}\left(0, T ; H_{p}^{2}(\Sigma)\right)} \tag{3.45}
\end{equation*}
$$

and if $p>3$, then for all $T_{0} \in(0, \infty)$ there exists $C\left(T_{0}\right)>0$ such that

$$
\begin{equation*}
\left\|g_{h}\right\|_{L_{p}\left(0, T ; H_{p}^{2}(\Sigma)\right)} \leq T^{1 / 2-1 / 2 p} C\left(T_{0}\right)\left\|g_{h}\right\|_{0 \mathbb{G}_{h}(T)} \quad \text { for all } g_{h} \in{ }_{0} \mathbb{G}_{h}(T), T \in\left(0, T_{0}\right] \tag{3.46}
\end{equation*}
$$

Proof. We will frequently employ the embeddings (B.1), (B.2), (B.3) page 145, and the mixed derivative embeddings from Proposition B. 44 on page 159. From Lemma B. 9 on page 148 we infer that the embedding constants for the relevant subspaces with vanishing initial values are uniformly bounded with respect to $T \in\left(0, T_{0}\right]$, by extension to $(0, \infty)$ and restriction to $\left(0, T_{0}\right)$. Moreover, Lemma 3.18 on page 76 yields a factor $T^{\delta} C\left(\delta, T_{0}\right)$ for the norm bound of an embedding into a space with lower temporal regularity.

For $\tau, \sigma \in(0,1), p \in[1, \infty)$, and $q \in[1, \infty]$, we abbreviate

We may assume that the norms of $\mathbb{G}_{v}, \mathbb{G}_{w}$, and $\mathbb{G}_{h}$ are given by

$$
\begin{aligned}
\|v\|_{\mathbb{G}_{v}} & =\llbracket v \rrbracket_{1 / 2-1 / 2 p, p ; p}+\|v\|_{p ; p}+\llbracket v \rrbracket_{p ; 1-1 / p, p}, \\
\|w\|_{\mathbb{G}_{w}} & =\llbracket(w, \nabla w) \rrbracket_{1 / 2-1 / 2 p, p ; p}+\|(w, \nabla w)\|_{p ; p}+\llbracket \nabla w \rrbracket_{p ; 1-1 / p, p}, \\
\|h\|_{\mathbb{G}_{h}} & =\llbracket(h, \nabla h) \rrbracket_{1-1 / 2 p, p ; p}+\left\|\left(h, \nabla h, \nabla^{2} h\right)\right\|_{p ; p}+\llbracket \nabla^{2} h \rrbracket_{p ; 1-1 / p, p},
\end{aligned}
$$

since these norms are equivalent to the usual ones and the corresponding constants only depend on $p$ and $n$ but not on $T$. Lemma B. 10 yields the estimate

$$
\begin{equation*}
\llbracket u v \rrbracket_{W_{p}^{\sigma}} \leq\|u\|_{\infty} \llbracket v \rrbracket_{W_{p}^{\sigma}}+C(n, p, \sigma)\|u\|_{\infty}^{1-\sigma}\|\nabla u\|_{\infty}^{\sigma}\|v\|_{p} \tag{3.47}
\end{equation*}
$$

for $u \in W_{\infty}^{1}(\Omega)$ and $v \in W_{p}^{\sigma}(\Omega)$.
The inequality (3.32c) and the mixed derivative embeddings yield

$$
\|\nabla u\|_{L_{p}\left(0, T ; L_{p}\left(\mathbb{R}^{n}\right)\right)} \leq T^{\delta} C\left(\delta, T_{0}\right)\|u\|_{0 W_{p}^{\delta}\left(0, T ; H_{p}^{1}\left(\mathbb{R}^{n}\right)\right)} \leq T^{\delta} C\left(\delta, T_{0}\right)\|u\|_{0 \mathbb{E}_{u}}
$$

for $\delta \in(0,1 / 2)$, provided $1 / 2>1 / p$ which is true for $p>2$. Together with (3.32b), this proves assertion (i). With (3.32c) and (3.32d) we obtain

$$
\begin{aligned}
\|v\|_{0} \mathbb{G}_{v}(T) & =\llbracket v \rrbracket_{1 / 2-1 / 2 p, p ; p}+\|v\|_{p ; p}+\llbracket v \rrbracket_{p ; 1-1 / p, p} \\
& \leq C\left\{T^{1 / 2}+T^{1-1 / 2 p}\right\} \llbracket v \rrbracket_{1-1 / 2 p, p ; p}+C T^{\delta}\|v\|_{0 W_{p}^{\delta}\left(0, T ; W_{p}^{1-1 / p}(\Sigma)\right)}
\end{aligned}
$$

for $v \in{ }_{0} \mathbb{E}_{v}(T)$ and $\delta \in(1 / p, 1)$. Moreover, for $\rho \in(0,1 / 2]$ with $1 / p<\delta<1 / 2-1 / 2 p+\rho \leq$ $1-1 / 2 p$ and $2-4 \rho>1-1 / p$ the mixed derivative embeddings yield

$$
\|v\|_{W_{p}^{\delta}\left(0, T ; W_{p}^{1-1 / p}(\Sigma)\right)} \leq C\left(T_{0}\right)\|v\|_{H_{p}^{1 / 2-1 / 2 p+\rho}\left(0, T ; W_{p}^{2-4 \rho}(\Sigma)\right)} \leq C\left(T_{0}\right)\|v\|_{0 \mathbb{E}_{v}(T)} .
$$

The number $\rho$ must belong to $(0,1 / 2] \cap(\delta-1 / 2+1 / 2 p, 1 / 4+1 / 4 p)$ and this interval is nonempty if $\delta<3 / 4-1 / 4 p$, which is true for $\delta \leq 1 / 2$. The embedding estimates (3.34), (3.35), (3.36), (3.38), (3.40), (3.42), (3.44), and (3.46) follow similarly and hence assertions (ii), (iii), and (iv) are valid.

The bilinear estimates (3.37), (3.41), and (3.45) can be verified by means of the spatial pointwise multiplication inequality (3.47), since the factor $\varphi$ does not depend on time. Hence (vii) is valid.

Finally, the bilinear estimates (3.39) and (3.43) follow from (3.47), Sobolev's embedding, and the pointwise multiplication estimate in Lemma B.81. Therefore (v) and (vi) are also true.

Proof of Theorem 3.16. For a given parameter tuple $\left(\vartheta^{*}, \omega, \vartheta\right) \in \mathcal{P}_{M, T_{1}, \eta, R}$, we define $\vartheta_{1}^{*}, \vartheta_{2}^{*}, \vartheta_{3}^{*}$, $\vartheta_{4}^{*}$, and $\vartheta_{\sigma}^{*}$ according to (3.26). Let $z=(u, \pi, h), \bar{z}=(\bar{u}, \bar{\pi}, \bar{h}), f=\left(f_{u}, f_{d}, g_{v}, g_{w}, g_{h}\right)$, and $\bar{f}=\left(\bar{f}_{u}, \bar{f}_{d}, \bar{g}_{v}, \bar{g}_{w}, \bar{g}_{h}\right)$ be related as in Lemma 3.17. We introduce the transformed operators

$$
\begin{aligned}
\bar{L}_{v}\left(\bar{u}, \bar{h} ; \vartheta^{*}\right) & :=-\mu_{s} \Delta^{\prime} \bar{v}-\lambda_{s} \nabla^{\prime} \operatorname{div}^{\prime} \bar{v}-\llbracket \mu \partial_{n} \bar{v} \rrbracket-c_{5} \llbracket \mu \nabla^{\prime} \bar{w} \rrbracket+\vartheta_{1}^{*} \nabla^{\prime} \Delta^{\prime} \bar{h}, \\
\bar{L}_{w}\left(\bar{u}, \bar{\pi}, \bar{h} ; \vartheta^{*}\right) & :=-\operatorname{tr}\left(\left[\vartheta_{2}^{*}+2 \vartheta_{3}^{*}\right] \nabla^{\prime} \bar{v}\right)-2 \llbracket \mu \partial_{n} \bar{w} \rrbracket+\llbracket \bar{\pi} \rrbracket-\operatorname{tr}\left(\left[\vartheta_{\sigma}^{*}+\vartheta_{4}^{*} \nabla^{\prime 2} \bar{h}\right),\right. \\
F_{u}(\bar{u}, \bar{\pi} ; \omega) & :=\bar{\mu}\left(\left[\partial \Theta_{\omega}\right]^{-1}(\Delta u) \circ \Theta_{\omega}-\Delta \bar{u}\right)+\left(I-\left[\partial \Theta_{\omega}\right]^{-1}\left[\partial \Theta_{\omega}\right]^{-\top}\right) \nabla \bar{\pi}, \\
G_{v}\left(\bar{u}, \bar{\pi}, \bar{h} ; \vartheta^{*}, \omega, \vartheta\right) & :=\bar{L}_{v}\left(\bar{u}, \bar{h} ; \vartheta^{*}\right)-P^{\prime} L_{v}(u, h ; \omega, \vartheta) \circ \Theta_{\omega}-Q_{v}(\omega) L_{w}(u, \pi, h ; \omega, \vartheta) \circ \Theta_{\omega}, \\
G_{w}\left(\bar{u}, \bar{\pi}, \bar{h} ; \vartheta^{*}, \omega, \vartheta\right) & :=\bar{L}_{w}\left(\bar{u}, \bar{\pi}, \bar{h} ; \vartheta^{*}\right)-L_{w}(u, \pi, h ; \omega, \vartheta) \circ \Theta_{\omega}-Q_{w}(\omega) L_{w}(u, \pi, h ; \omega, \vartheta) \circ \Theta_{\omega}, \\
G_{h}(\bar{w} ; \omega) & :=\left(\left(1+|\nabla \omega|^{2}\right)^{-1 / 2}-1\right) \bar{w} .
\end{aligned}
$$

Here actually $G_{v}$ and $G_{w}$ do not depend on $\bar{\pi}$ and we will therefore write $G_{j}\left(\bar{u}, \bar{h} ; \vartheta^{*}, \omega, \vartheta\right)$ for $j \in\{v, w\}$. More details on these operators will be given below and we will show that problem (3.25) is equivalent to the following problem for $\Sigma_{0}=\mathbb{R}^{n-1} \times\{0\}$.

$$
\left\{\begin{align*}
\bar{\rho} \partial_{t} \bar{u}-\bar{\mu} \Delta \bar{u}+\nabla \bar{\pi} & =\bar{f}_{u}+F_{u}(\bar{u}, \bar{\pi} ; \omega) & & \text { in } J \times \dot{\mathbb{R}}^{n},  \tag{3.48}\\
\operatorname{div} \bar{u} & =\bar{f}_{d} & & \text { in } J \times \dot{\mathbb{R}}^{n}, \\
\llbracket \bar{u} \rrbracket & =0 & & \text { on } J \times \Sigma_{0}, \\
\bar{L}_{v}\left(\bar{u}, \bar{h} ; \vartheta^{*}\right) & =\bar{g}_{v}+G_{v}\left(\bar{u}, \bar{h} ; \vartheta^{*}, \omega, \vartheta\right) & & \text { on } J \times \Sigma_{0}, \\
\bar{L}_{w}\left(\bar{u}, \bar{\pi}, \bar{h} ; \vartheta^{*}\right) & =\bar{g}_{w}+G_{w}\left(\bar{u}, \bar{h} ; \vartheta^{*}, \omega, \vartheta\right) & & \text { on } J \times \Sigma_{0}, \\
\partial_{t} \bar{h}-\bar{w} & =\bar{g}_{h}+G_{h}(\bar{w} ; \omega) & & \text { on } J \times \Sigma_{0}, \\
\left.\bar{h}\right|_{t=0} & =0 & & \text { on } \Sigma_{0}, \\
\left.\bar{u}\right|_{t=0} & =0 & & \text { in } \mathbb{R}^{n} .
\end{align*}\right.
$$

Let us abbreviate

$$
S\left(\vartheta^{*}\right) \bar{z}:=\left[\begin{array}{c}
\bar{\rho} \partial_{t} \bar{u}-\bar{\mu} \Delta \bar{u}+\nabla \bar{\pi} \\
\operatorname{div} \bar{u} \\
\bar{L}_{v}\left(\bar{u}, \bar{h} ; \vartheta^{*}\right) \\
\bar{L}_{w}\left(\bar{u}, \bar{\pi}, \bar{h} ; \vartheta^{*}\right) \\
\partial_{t} \bar{h}-\bar{w}
\end{array}\right], \quad F\left(\vartheta^{*}, \omega, \vartheta\right) \bar{z}:=\left[\begin{array}{c}
F_{u}(\bar{u}, \bar{\pi} ; \omega) \\
0 \\
G_{v}\left(\bar{u}, \bar{h} ; \vartheta^{*}, \omega, \vartheta\right) \\
G_{w}\left(\bar{u}, \bar{h} ; \vartheta^{*}, \omega, \vartheta\right) \\
G_{h}(\bar{w} ; \omega)
\end{array}\right] .
$$

Analogously as in Figure 3.7 on page 72 we let

$$
{ }_{0} \overline{\mathbb{E}}:={ }_{0} \overline{\mathbb{E}}_{u, v, w, \partial_{\nu} w} \times{ }_{0} \overline{\mathbb{E}}_{\pi, \llbracket \pi \rrbracket} \times{ }_{0} \overline{\mathbb{E}}_{h}, \quad{ }_{0} \overline{\mathbb{F}}:=\overline{\mathbb{F}}_{u} \times{ }_{0} \overline{\mathbb{F}}_{d, \Sigma} \times{ }_{0} \overline{\mathbb{G}}_{v} \times{ }_{0} \overline{\mathbb{G}}_{w} \times{ }_{0} \overline{\mathbb{G}}_{h}
$$

denote the corresponding spaces defined with $\Sigma_{0}$ instead of $\Sigma_{\omega}$. Our goal is to prove that

$$
\begin{equation*}
\bar{z} \mapsto \bar{f}=\left[S\left(\vartheta^{*}\right)-F\left(\vartheta^{*}, \omega, \vartheta\right)\right] \bar{z}, \quad 0 \overline{\mathbb{E}}(T) \rightarrow_{0} \overline{\mathbb{F}}(T) \tag{3.49}
\end{equation*}
$$

is uniformly invertible with respect to $T \in\left(0, T_{1}\right]$ and $\left(\vartheta^{*}, \omega, \vartheta\right) \in \mathcal{P}_{M, T_{1}, \eta, R}$.
Theorem 3.14 shows that $S\left(\vartheta^{*}\right):{ }_{0} \overline{\mathbb{E}}(T) \rightarrow{ }_{0} \overline{\mathbb{F}}(T)$ is uniformly invertible with respect to $T \in\left(0, T_{1}\right]$ and $\vartheta^{*} \in \mathcal{P}_{M}$, for every $T_{1} \in(0, \infty)$ and $M \in(0, \infty)$, where $\mathcal{P}_{M}$ is defined in equation (3.4) on page 54. In order to apply a Neumann series argument, it remains to ensure that

$$
\left\|\left[S\left(\vartheta^{*}\right)\right]^{-1} F\left(\vartheta^{*}, \omega, \vartheta\right)\right\|_{\mathcal{B}\left({ }_{0} \overline{\mathbb{E}}(T)\right)}<1 .
$$

(i) We first compute the perturbations in more detail and we abbreviate

$$
X(\bar{x}):=\Theta_{\omega}(\bar{x}), \quad \bar{X}(x):=\Theta_{\omega}^{-1}(x), \quad \text { for } x, \bar{x} \in \mathbb{R}^{n} .
$$

By using the summation convention, we have

$$
\partial_{l} X_{k}=\delta_{l k}+\delta_{k n} \partial_{l} \omega, \quad \partial_{j} \bar{X}_{m}=\delta_{j m}-\delta_{m n} \partial_{j} \omega .
$$

Then the following identity is valid, where the values of $u$ and $\bar{X}$ are taken at $(t, x) \in J \times\left(\mathbb{R}^{n} \backslash\right.$ $\left.\Sigma_{\omega}\right)$ and those of $\bar{u}$ and $X$ at $(t, \bar{x}) \in J \times\left(\mathbb{R}^{n} \backslash \Sigma_{0}\right)$ with $x=X(\bar{x})$.

$$
\begin{align*}
\Delta u_{k}= & \Delta \bar{u}_{k}+\left(\partial_{l} X_{k} \partial_{j} \bar{X}_{m} \partial_{j} \bar{X}_{p}-\delta_{k l} \delta_{j m} \delta_{j p}\right) \partial_{m} \partial_{p} \bar{u}_{l} \\
& +\left(\partial_{l} \partial_{p} X_{k} \partial_{j} \bar{X}_{m} \partial_{j} \bar{X}_{p}-\partial_{p} \bar{X}_{m} \partial_{j} \partial_{r} X_{p} \partial_{j} \bar{X}_{r} \partial_{l} X_{k}\right) \partial_{m} \bar{u}_{l}  \tag{3.50}\\
& +\left(\partial_{l} \partial_{m} \partial_{p} X_{k} \partial_{j} \bar{X}_{m} \partial_{j} \bar{X}_{p}-\partial_{p} \bar{X}_{m} \partial_{j} \partial_{r} X_{p} \partial_{j} \bar{X}_{r} \partial_{l} \partial_{m} X_{k}\right) \bar{u}_{l} .
\end{align*}
$$

Moreover, a straightforward computation shows that the divergence satisfies

$$
\operatorname{div} \bar{u}=\partial_{k}\left(\left[\partial \Theta_{\omega}\right]_{k l}^{-1} u_{l} \circ \Theta_{\omega}\right)=\operatorname{div} u \circ \Theta_{\omega},
$$

and therefore no perturbations of the divergence equation appear.
The representations of $G_{v}$ and $G_{w}$ and the corresponding equations in (3.48) follow from (3.31).
(ii) In order to control the perturbation $F_{u}(\bar{u}, \bar{\pi} ; \omega)$ in $\overline{\mathbb{F}}_{u}(T)=L_{p}\left(0, T ; L_{p}\left(\mathbb{R}^{n}\right)\right)$ we note that $\left\|I-\partial \Theta_{\omega}\right\|_{\infty} \rightarrow 0$ as $\|\nabla \omega\|_{\infty} \rightarrow 0$. Hence there exists $\eta>0$ such that

$$
\left\|\left(I-\left[\partial \Theta_{\omega}\right]^{-1}\left[\partial \Theta_{\omega}\right]^{-\top}\right) \nabla \bar{\pi}\right\|_{L_{p}\left(0, T ; L_{p}\left(\mathbb{R}^{n}\right)\right)} \leq \varepsilon\|\nabla \bar{\pi}\|_{L_{p}\left(0, T ; L_{p}\left(\mathbb{R}^{n}\right)\right)} \quad \text { for } T \in(0, \infty),\|\nabla \omega\|_{\infty} \leq \eta .
$$

Next, we rewrite the transformation formula (3.50) as

$$
(\Delta u) \circ \Theta_{\omega}-\Delta \bar{u}=a^{j k}(\omega) \partial_{j} \partial_{k} \bar{u}+b^{j}(\omega) \partial_{j} \bar{u}+c(\omega) \bar{u},
$$

where the coefficients $a^{j k}(\omega), b^{j}(\omega)$, and $c(\omega)$ are functions on $\mathbb{R}^{n}$ which satisfy the following estimate. For given $\varepsilon>0$ and $R>0$ there exist $\eta>0, R_{b} \geq 0$, and $R_{c} \geq 0$ such that

$$
\left\|a^{j k}(\omega)\right\|_{\infty} \leq \varepsilon, \quad\left\|b^{j}(\omega)\right\|_{\infty} \leq R_{b}, \quad\|c(\omega)\|_{\infty} \leq R_{c},
$$

for all $\omega \in B C^{3}\left(\mathbb{R}^{n-1}\right)$ with $\|\nabla \omega\|_{\infty} \leq \eta$ and $\|\nabla \omega\|_{B C^{2}} \leq R$. By controlling the lower-order terms $\bar{u}$ and $\nabla \bar{u}$ with estimate (3.33), we conclude that for given $\varepsilon>0$ there exists $\eta>0$ such that

$$
\begin{aligned}
\left\|(\Delta u) \circ \Theta_{\omega}-\Delta \bar{u}\right\|_{L_{p}\left(0, T ; L_{p}\left(\mathbb{R}^{n}\right)\right)} & \leq \varepsilon\|u\|_{\overline{\mathbb{E}}_{u}(T)}+C(R)\|\bar{u}\|_{L_{p}\left(0, T ; H_{p}^{1}\left(\mathbb{R}^{n}\right)\right)} \\
& \leq \varepsilon\|u\|_{0 \overline{\mathbb{E}}_{u}(T)}+C(R) T^{\delta} C\left(\delta, T_{1}\right)\|\bar{u}\|_{0 \mathbb{\mathbb { E }}_{u}(T)},
\end{aligned}
$$

provided that $u \in{ }_{0} \overline{\mathbb{E}}_{u}(T), T \in\left(0, T_{1}\right], \delta \in(1 / p, 1 / 2), T_{1} \in(0, \infty),\|\nabla \omega\|_{\infty} \leq \eta$, and $\|\nabla \omega\|_{B C^{2}} \leq$ $R$. We conclude that for $M, T_{1}, \varepsilon$, and $R>0$ there are $\eta\left(M, T_{1}, \varepsilon\right)>0$ and $T_{0}\left(M, T_{1}, \varepsilon, R\right) \in$ $\left(0, T_{1}\right]$ such that

$$
\left.\left\|\bar{z} \mapsto F_{u}(\bar{u}, \bar{\pi} ; \omega)\right\|_{0} \overline{\mathbb{E}}(T) \rightarrow \overline{\mathbb{F}}_{u}(T)\right) \leq \varepsilon \quad \text { for all }\left(\vartheta^{*}, \omega, \vartheta\right) \in \mathcal{P}_{M, T_{1}, \eta, R} .
$$

(iii) We next control the perturbation $G_{v}$ in ${ }_{0} \overline{\mathbb{G}}_{v}(T)$. First, the estimates (3.37) and (3.38) and Lemma 3.17 yield an estimate

$$
\begin{aligned}
\left\|Q_{v}(\omega) L_{w}(z ; \omega, \vartheta) \circ \Theta_{\omega}\right\|_{0_{0}(T)} & \leq C \eta\|\bar{z}\|_{0 \overline{\mathbb{E}}(T)}+C(R)\left\|L_{w}(z ; \omega, \vartheta) \circ \Theta_{\omega}\right\|_{L_{p}\left(0, T ; L_{p}(\Sigma)\right)} \\
& \leq C \eta\|\bar{z}\|_{0 \overline{\mathbb{E}}(T)}+C(R) T^{\delta} C\left(\delta, T_{1}\right)\|\bar{z}\|_{0 \overline{\mathbb{E}}(T)},
\end{aligned}
$$

uniformly with respect to $\bar{z} \in_{0} \overline{\mathbb{E}}(T), T \in\left(0, T_{1}\right],\|\nabla \omega\|_{\infty} \leq \eta$, and $\|\nabla \omega\|_{B C^{3}} \leq R$.
It remains to estimate the difference $\bar{L}_{v}\left(\bar{u}, \bar{h} ; \vartheta^{*}\right)-P^{\prime} L_{v}(u, h ; \omega, \vartheta) \circ \Theta_{\omega}$ in ${ }_{0} \overline{\mathbb{G}}_{v}(T)$. In view of

$$
\left\|P^{\prime}-P_{\Sigma_{\omega}} \circ \Theta_{\omega}\right\|_{\infty} \rightarrow 0 \quad \text { as }\|\nabla \omega\|_{\infty} \rightarrow 0
$$

and estimate (3.38), we may omit the projection $P^{\prime}$ in the above difference and therefore it remains to estimate the following differences in ${ }_{0} \overline{\mathbb{G}}_{v}(T)$.

$$
\begin{align*}
& \Delta^{\prime} \bar{v}-\left(\widetilde{\Delta}_{\Sigma_{\omega}} v\right) \circ \Theta_{\omega},  \tag{3.51a}\\
& \nabla^{\prime} \operatorname{div}^{\prime} \bar{v}-\left(\nabla_{\Sigma_{\omega}} \operatorname{div}_{\Sigma_{\omega}} v\right) \circ \Theta_{\omega},  \tag{3.51b}\\
& \llbracket \bar{\mu} \partial_{v} \bar{v} \rrbracket-\llbracket \mu \partial_{\nu_{\Sigma_{\omega}}} v \rrbracket \circ \Theta_{\omega},  \tag{3.51c}\\
& \llbracket \bar{\mu} \nabla^{\prime} \bar{w} \rrbracket-\llbracket \mu \nabla_{\Sigma_{\omega}} w \rrbracket \circ \Theta_{\omega},  \tag{3.51d}\\
& \vartheta_{w}^{*} \nabla^{\prime} \Delta^{\prime} \bar{h}-\left(\vartheta_{w} \nabla_{\Sigma_{\omega}} \Delta_{\Sigma_{\omega}} h\right) \circ \Theta_{\omega} . \tag{3.51e}
\end{align*}
$$

The differences (3.51a), (3.51b), and (3.51d) can be controlled by applying the identities in Figures 3.8 and 3.9, Lemma 3.17, and the estimates (3.34), (3.35), (3.37) and (3.38) and we obtain

$$
\begin{aligned}
& \left\|\left(\Delta^{\prime} \bar{v}-\left(\widetilde{\Delta}_{\Sigma_{\omega}} v\right) \circ \Theta_{\omega}, \nabla^{\prime} \operatorname{div}^{\prime} \bar{v}-\left(\nabla_{\Sigma_{\omega}} \operatorname{div}_{\Sigma_{\omega}} v\right) \circ \Theta_{\omega}, \llbracket \bar{\mu} \nabla^{\prime} \bar{w} \rrbracket-\llbracket \mu \nabla_{\Sigma_{\omega}} w \rrbracket \circ \Theta_{\omega}\right)\right\|_{0 \overline{\mathbb{G}}_{v}(T)} \\
& \leq\left\{C \eta+C(R) T^{1 / 4} C\left(T_{1}\right)\right\}\|(\bar{v}, \bar{w})\|_{0 \mathbb{E}_{v}(T) \times_{0} \mathbb{E}_{w}(T)},
\end{aligned}
$$

uniformly with respect to $\bar{v} \in{ }_{0} \overline{\mathbb{E}}_{v}(T), \bar{w} \in{ }_{0} \overline{\mathbb{E}}_{w}(T), T \in\left(0, T_{1}\right], T_{1} \in(0, \infty),\|\nabla \omega\|_{\infty} \leq \eta$, and $\|\nabla \omega\|_{B C^{3}} \leq R$. For (3.51e) we employ the estimates (3.36), (3.39) and (3.40) and obtain

$$
\left\|\vartheta_{w}^{*} \nabla^{\prime} \Delta^{\prime} \bar{h}-\left(\vartheta_{w} \nabla_{\Sigma_{\omega}} \Delta_{\Sigma_{\omega}} h\right) \circ \Theta_{\omega}\right\|_{0} \overline{\mathbb{G}}_{v}(T) \leq\left\{C \eta+C(R) T^{\delta} C\left(\delta, T_{1}\right)\right\}\|\bar{h}\|_{o \mathbb{E}_{h}(T)} .
$$

In order to deal with $\partial_{\nu} v$, we note that $v=\left[P_{\Sigma_{\omega}} \circ \Pi_{\Sigma_{\omega}}\right] u$ near $\Sigma_{\omega}$ with the nonlinear projection $\Pi_{\Sigma_{\omega}}$ onto $\Sigma_{\omega}$ from on page 138. Hence we obtain

$$
\begin{aligned}
\left\|\llbracket \bar{\mu} \partial_{n} \bar{v} \rrbracket-\llbracket \mu \partial_{\nu_{\Sigma_{\omega}}} v \rrbracket \circ \Theta_{\omega}\right\|_{0 \overline{\mathbb{G}}_{v}(T)} & \leq C \eta\|\bar{u}\|_{\overline{\mathbb{E}}_{u}(T)}+C(R)\left\|\left(\bar{u}, \nabla^{\prime} \bar{u}\right)\right\|_{0} \overline{\mathbb{G}}_{v}(T) \\
& \leq\left\{C \eta+C(R) T^{-1 / 4} C\left(T_{1}\right)\right\}\|\bar{u}\|_{0 \overline{\mathbb{E}}_{u, v, w, \partial_{\nu} w}(T)} .
\end{aligned}
$$

We conclude that for $\varepsilon>0$ there is $\eta>0$ such that for $R, T_{1} \in(0, \infty)$ there is $T_{0} \in\left(0, T_{1}\right]$ such that

$$
\left\|\bar{z} \mapsto G_{v}\left(\bar{u}, \bar{h} ; \vartheta^{*}, \omega, \vartheta\right)\right\|_{0 \overline{\mathbb{E}}(T) \rightarrow 0} \overline{\mathbb{G}}_{v}(T) \leq \varepsilon \quad \text { for all }\left(\vartheta^{*}, \omega, \vartheta\right) \in \mathcal{P}_{M, T_{1}, \eta, R}, T \in\left(0, T_{0}\right] .
$$

(iv) We next control $G_{w}$ in $\overline{\mathbb{G}}_{w}(T)$. The estimates (3.41) and (3.42) and Lemma 3.17 yield

$$
\begin{aligned}
\left\|Q_{w}(\omega) L_{w}(z ; \omega, \vartheta) \circ \Theta_{\omega}\right\|_{0} \bar{G}_{w}(T) & \leq C \eta\|\bar{z}\|_{0 \overline{\mathrm{E}}(T)}+C(R)\left\|L_{w}(z ; \omega, \vartheta) \circ \Theta_{\omega}\right\|_{L_{p}\left(0, T ; H_{p}^{1}(\Sigma)\right)} \\
& \leq C \eta\|\bar{z}\|_{0 \overline{\mathrm{E}}(T)}+C(R) T^{\delta} C\left(\delta, T_{1}\right)\|\bar{z}\|_{0 \overline{\mathrm{E}}(T)},
\end{aligned}
$$

uniformly with respect to $\bar{z} \in_{0} \overline{\mathbb{E}}(T), T \in\left(0, T_{1}\right],\|\nabla \omega\|_{\infty} \leq \eta$, and $\|\nabla \omega\|_{B C^{3}} \leq R$. It remains to estimate the difference $\bar{L}_{w}\left(\bar{u}, \bar{\pi}, \bar{h} ; \vartheta^{*}\right)-L_{w}(u, \pi, h ; \omega, \vartheta) \circ \Theta_{\omega}$ which consists of
( 3.52 g )

$$
\begin{align*}
& \vartheta_{L}^{*} \nabla^{\prime} \bar{v}-\left(\vartheta_{L} \nabla_{\Sigma_{\omega}} v\right) \circ \Theta_{\omega},  \tag{3.52a}\\
& \llbracket \mu \partial_{n} \bar{w}-\llbracket \mu \partial_{\nu_{\Sigma_{\omega}}} w \rrbracket \circ \Theta_{\omega},  \tag{3.52b}\\
& \operatorname{tr}\left(\nabla^{\prime 2} \bar{h}\right)-\operatorname{tr}\left(\nabla_{\Sigma_{\omega}}^{2} h\right) \circ \Theta_{\omega},  \tag{3.52c}\\
& \operatorname{tr}\left(\operatorname{tr} \vartheta_{D v}^{*} \nabla^{\prime 2} \bar{h}\right)-\operatorname{tr}\left(\operatorname{tr} \vartheta_{D v} \nabla_{\Sigma_{\omega}}^{2} h\right) \circ \Theta_{\omega},  \tag{3.52d}\\
& \operatorname{tr}\left(\vartheta_{w}^{*} \operatorname{tr} \vartheta_{L}^{*} \nabla^{\prime 2} \bar{h}\right)-\operatorname{tr}\left(\vartheta_{w} \operatorname{tr} \vartheta_{L} \nabla_{\Sigma_{\omega}}^{2} h\right) \circ \Theta_{\omega},  \tag{3.52e}\\
& \operatorname{tr}\left(\vartheta_{D v}^{*} \nabla^{2} \bar{h}\right)-\operatorname{tr}\left(\vartheta_{D v} \nabla_{\Sigma_{\omega}}^{2} h\right) \circ \Theta_{\omega},  \tag{3.52f}\\
& \operatorname{tr}\left(\vartheta_{w}^{*} \vartheta_{L}^{*} \nabla^{\prime 2} \bar{h}\right)-\operatorname{tr}\left(\vartheta_{w} \vartheta_{L} \nabla_{\Sigma_{\omega}}^{2} h\right) \circ \Theta_{\omega} .
\end{align*}
$$

Again we control lower order terms by using (3.34), (3.35), and (3.36). The differences (3.52a) to (3.52c) can be controlled by means of the estimates (3.41) and (3.42). For the terms (3.52d) to $(3.52 \mathrm{~g})$ we employ the estimates (3.43) and (3.44). We conclude that for $\varepsilon>0$ there is $\eta>0$ such that for $R, T_{1} \in(0, \infty)$ there is $T_{0} \in\left(0, T_{1}\right]$ such that

$$
\left\|\bar{z} \mapsto G_{w}\left(\bar{u}, \bar{h} ; \vartheta^{*}, \omega, \vartheta\right)\right\|_{0} \overline{\mathbb{E}}_{u}(T) \rightarrow 0 \overline{\mathbb{G}}_{w}(T) \leq \varepsilon \quad \text { for all }\left(\vartheta^{*}, \omega, \vartheta\right) \in \mathcal{P}_{M, T_{1}, \eta, R}, T \in\left(0, T_{0}\right] .
$$

(v) The relevant estimates of $G_{h}$ in ${ }_{0} \overline{\mathbb{G}}_{h}(T)$ follow from estimates (3.45) and (3.46) and we conclude that for $T_{1}>0$ and $\varepsilon>0$ there is $\eta>0$ such that for $R>0$ there is $T_{0} \in\left(0, T_{1}\right]$ such that

$$
\left\|\bar{z} \mapsto G_{h}(\bar{w} ; \omega)\right\|_{0} \overline{\mathbb{E}}(T) \rightarrow 0 \overline{\mathbb{G}}_{h}(T) \leq \varepsilon \quad \text { for all }\left(\vartheta^{*}, \omega, \vartheta\right) \in \mathcal{P}_{M, T_{1}, \eta, R}, T \in\left(0, T_{0}\right] .
$$

(vi) The preceding steps show that for given $M, T_{1}, \varepsilon, R>0$ there are $\eta=\eta\left(M, T_{1}, \varepsilon\right)>0$ and $T_{0}=T_{0}\left(M, T_{1}, \varepsilon, R\right) \in\left(0, T_{1}\right]$ such that

$$
\begin{equation*}
\left\|F\left(\vartheta^{*}, \omega, \vartheta\right)\right\|_{0 \stackrel{\mathbb{E}}{ }(T) \rightarrow 0} \overline{\mathbb{F}}(T) \leq \varepsilon \quad \text { for all }\left(\vartheta^{*}, \omega, \vartheta\right) \in \mathcal{P}_{M, T_{1}, \eta, R} . \tag{3.53}
\end{equation*}
$$

For fixed $M$ and $T_{1}$ we apply Theorem 3.14 and obtain a finite number

$$
C:=\sup \left\{\left\|\left[S\left(\vartheta^{*}\right)\right]^{-1}\right\|_{0 \overline{\mathbb{F}}(T) \rightarrow 0} \mathbb{\mathbb { E }}(T): \vartheta^{*} \in \mathcal{P}_{M}, T \in\left(0, T_{1}\right]\right\} .
$$

Next we fix $\varepsilon \in\left(0, C^{-1}\right)$ and for given $R>0$ we choose $\eta\left(M, T_{1}, \varepsilon\right)>0$ and $T_{0}\left(M, T_{1}, \varepsilon, R\right)>$ 0 such that (3.53) is valid. Then the operator (3.49) is uniformly invertible with respect to $\left(\vartheta^{*}, \omega, \vartheta\right) \in \mathcal{P}_{M, T_{1}, \eta, R}$ and $T \in\left(0, T_{0}\right]$ and the proof of Theorem 3.16 is complete.

### 3.3. Bounded domains

We consider problem (3.1) $=(\mathrm{PL})$ in a bounded domain $\Omega \subset \mathbb{R}^{n}(n \geq 2)$ with smooth boundary $\partial \Omega$ and with a compact smooth hypersurface $\Sigma \subset \Omega$ such that $\Omega \backslash \Sigma$ consists of disjoint open sets $\Omega_{+}$and $\Omega_{-}$with $\partial \Omega_{+} \cap \partial \Omega_{-}=\Sigma$. We will establish optimal regularity for this problem on some short time interval $(0, T)$ for the following class of reference velocities $u_{*}$. The involved function spaces are collected in Figure 3.7 on page 72.
3.20. Definition. Let $p \in(\max \{3,(n+2) / 2\}, \infty), M>0$, and $T>0$. The parameter set $\mathcal{P}_{M, T}$ consists of all vector fields $u_{*}=v_{*}+w_{*} \nu_{\Sigma} \in \mathbb{E}_{v}(T)+\mathbb{E}_{w}(T) \cdot \nu_{\Sigma}$ such that

$$
\left\|w_{*}\right\|_{\mathbb{G}_{w}(T)} \leq M, \quad\left\|D_{\Sigma}\left(v_{*}\right)\right\|_{\mathbb{G}_{w}(T)} \leq M,
$$

and

$$
\begin{equation*}
\inf _{(0, T) \times \Sigma} d_{0}\left(D_{\Sigma}\left(u_{*}\right)\right)=\inf _{(0, T) \times \Sigma}\left(\sigma+\left(\lambda_{s}-\mu_{s}\right) \operatorname{div}_{\Sigma} u_{*}+2 \mu_{s} \min _{\zeta \in \mathbb{R}^{n},|\zeta|=1} \zeta^{\top} D_{\Sigma}\left(u_{*}\right) \zeta\right) \geq M^{-1} . \tag{3.54}
\end{equation*}
$$

Note that condition (3.54) makes sense due to the embeddings $\mathbb{E}_{v} \hookrightarrow C\left([0, T] ; C^{1}(\Sigma ; T \Sigma)\right)$ and $\mathbb{E}_{w} \hookrightarrow C\left([0, T] ; C^{1}(\Sigma)\right)$, which are valid for $p>\max \{3,(n+2) / 2\}$.
3.21. Theorem. Let $\rho_{ \pm}, \mu_{ \pm}, \sigma, \mu_{s}, \lambda_{s}+\mu_{s}>0$, and let $p \in(\max \{5, n+2\}, \infty)$ and $M, T_{1}>0$. Then there exists $T_{0} \in\left(0, T_{1}\right]$ such that the solution-to-data map

$$
\begin{aligned}
&(u, \pi, h) \mapsto\left(f_{u}, f_{d}, g_{v}, g_{w}, g_{h}\right), \\
&{ }_{0} \mathbb{E}={ }_{0} \mathbb{E}_{u, v, w, \partial_{\nu} w} \times{ }_{0} \mathbb{E}_{\pi, \llbracket \pi]} \times{ }_{0} \mathbb{E}_{h} \rightarrow{ }_{0} \mathbb{F}={ }_{0} \mathbb{F}_{u} \times{ }_{0} \mathbb{F}_{d, \Sigma} \times{ }_{0} \mathbb{G}_{v} \times{ }_{0} \mathbb{G}_{w} \times{ }_{0} \mathbb{G}_{h}
\end{aligned}
$$

of problem (3.1) is uniformly invertible with respect to $T \in\left(0, T_{0}\right]$ and $u_{*} \in \mathcal{P}_{M, T_{1}}$.
For the proof we apply a modified version of the elliptic localization technique from Section 2.1.1, which will be presented in Section 3.3.1. As in [KPW13; Wil13], we localize problem (3.1) in both time and space and we construct both a left- and a right-inverse for the solution-to-data map. A different approach was used in [Gei+12] for a stationary Stokes problem, where the authors localize in space, establish $\mathcal{R}$-bounds for the data-to-solution map, and apply Weis' characterization of maximal $L_{p}$-regularity [Wei01].

As in Section 3.2, we employ $T$-dependent estimates for controlling lower-order perturbations; however, in order to control the commutator $\left[\nabla, \varphi_{j}\right] \pi=\nabla \varphi_{j} \pi$, the usual elliptic localization does not suffice. In addition, we employ certain projections on subspaces with vanishing momentum and divergence data. These projections are constructed similarly as in [Gei+12; KPW13; Wil13], by resolving non-trivial momentum and divergence data by means of weak

Neumann transmission problems and one-phase Stokes problems. Then, similar to [Köh13; KPW13; Wil13], we prove in Section 3.3.2 that the pressure has additional temporal regularity, if it belongs to the range of the aforementioned projection. Section 3.3.3 contains the local spaces, approximation systems, and local operators and in Section 3.3.4 we prove the relevant commutator estimates.
3.3.1. Localization technique. For $T \in\left(0, T_{1}\right]$ with some fixed number $T_{1} \in(0, \infty)$ we will consider Banach spaces $E=E(T)$ and $F=F(T)$ of functions on $(0, T)$ and linear operators $A_{T}: E(T) \rightarrow F(T)$. For the sake of brevity, we wish to omit the $T$-dependence occasionally. To justify this we always assume that the spaces and operators are compatible in the sense that for $0<T \leq T^{\prime} \leq T_{1}$ their realizations over $(0, T)$ coincide with the restrictions to $(0, T)$ of their realizations over $\left(0, T^{\prime}\right)$.

We fix a number $q \in[1, \infty)$ and an index set $I \subset \mathbb{N}_{0}$ and consider $l_{q}$-approximation systems $\left(\mathbf{E},\left(\Phi_{E, j}\right)_{j \in I},\left(\Psi_{E, j}\right)_{j \in I}\right)$ and $\left(\mathbf{F},\left(\Phi_{F, j}\right)_{j \in I},\left(\Psi_{F, j}\right)_{j \in I}\right)$ for $E$ and $F$ in the sense of Definition 2.4 on page 27. Our goal is to show that a given linear operator $A \in \mathcal{B}(E ; F)$ is uniformly invertible with respect to $T \in\left(0, T_{0}\right]$ for some $T_{0} \in\left(0, T_{1}\right]$. Let us therefore assume that
(i) there are invertible linear operators $A_{j} \in \mathcal{B}_{\text {isom }}\left(E_{j} ; F_{j}\right)$, the local operators, such that

$$
\sup _{T \in\left(0, T_{1}\right]}\left\|\left(f_{j}\right)_{j \in I} \mapsto\left(A_{j}^{-1} f_{j}\right)_{j \in I}\right\|_{l_{q}(\mathbf{F}(T)) \rightarrow l_{q}(\mathbf{E}(T))}<\infty .
$$

Indeed, these operators $A_{j}$ will correspond to certain model problems and the uniform invertibility of $A_{j}$ will follow from the boundedness of the relevant coefficients related to $\Sigma$ and $u_{*}$. We further assume that
(ii) we can find a projection $P_{F} \in \mathcal{B}(F)$ and an operator $R_{0} \in \mathcal{B}(F ; E)$ such that

$$
\left(I_{F}-P_{F}\right) A R_{0}\left(I_{F}-P_{F}\right)=I_{F}-P_{F} .
$$

We wish to choose the projection

$$
P_{F}:\left(f_{u}, f_{d}, g_{v}, g_{w}, g_{h}\right) \mapsto\left(0,0, g_{v}, g_{w}, g_{h}\right)
$$

and therefore the operator $R_{0}:\left(f_{u}, f_{d}, 0,0,0\right) \mapsto(u, \pi, 0)$ should produce functions $(u, \pi, 0) \in E$ with $\left(\rho \partial_{t}-\mu \Delta\right) u+\nabla \pi=f_{u}$ and $\operatorname{div} u=f_{d}$. Moreover, the operator

$$
P_{E}:=I_{E}-R_{0}\left(I_{F}-P_{F}\right) A
$$

is a projection in $\mathcal{B}(E)$ and we obtain

$$
\begin{aligned}
P_{F} A P_{E} & =P_{F} A\left(I_{E}-R_{0}\left(I_{F}-P_{F}\right) A\right) \\
& =A-\left(I_{F}-P_{F}\right) A+\left(I_{F}-P_{F}\right) A R_{0}\left(I_{F}-P_{F}\right) A-A R_{0}\left(I_{F}-P_{F}\right) A \\
& =A-A R_{0}\left(I_{F}-P_{F}\right) A=A P_{E} .
\end{aligned}
$$

In particular, for given $z=(u, \pi, h) \in P_{E} E$ and $A z=\left(f_{u}, f_{d}, g_{v}, g_{w}, g_{h}\right)$ we have $\left(f_{u}, f_{d}\right)=0$.
We also consider the local projections

$$
P_{F, j}:\left(f_{u j}, f_{d j}, g_{v j}, g_{w j}, g_{h j}\right) \mapsto\left(0,0, g_{v j}, g_{w j}, g_{h j}\right), \quad P_{E, j}:=A_{j}^{-1} P_{F, j} A_{j},
$$

and we assume that
(iii) the projections $P_{F, j}$ satisfy

$$
\Phi_{F, j} P_{F}=P_{F, j} \Phi_{F, j} .
$$

This property will be trivial in our situation, since it means that for given $\left(0,0, g_{v}, g_{w}, g_{h}\right) \in F$, the tuple $\Phi_{F, j}\left(0,0, g_{v}, g_{w}, g_{h}\right)$ has the form $\left(0,0, g_{v j}, g_{w j}, g_{h j}\right)$.

Now we define an approximate inverse for $A$ by

$$
R: F \rightarrow E, \quad R:=\sum_{j} \Psi_{E, j} A_{j}^{-1} \Phi_{F, j} P_{F}\left(I_{F}-A R_{0}\left(I_{F}-P_{F}\right)\right)+R_{0}\left(I_{F}-P_{F}\right) .
$$

Note that $\sum_{j} \Psi_{E, j} A_{j}^{-1} \Phi_{F, j}$ is the usual approximate inverse in the elliptic and parabolic theory and that the operator $R_{0}\left(I_{F}-P_{F}\right)$ takes care of the momentum and divergence data $\left(f_{u}, f_{d}\right)$. The latter is constructed in Lemma 3.27 on page 91 . From our assumptions (i) to (iii) we infer that

$$
\begin{align*}
& A R-I_{F}=\sum_{j}\left(A \Psi_{E, j}-\Psi_{F, j} A_{j}\right) A_{j}^{-1} \Phi_{F, j} P_{F}\left(I_{F}-A R_{0}\left(I_{F}-P_{F}\right)\right)  \tag{3.55a}\\
& R A-I_{E}=\sum_{j} \Psi_{E, j} A_{j}^{-1}\left(\Phi_{F, j} A-A_{j} \Phi_{E, j}\right) P_{E} \tag{3.55b}
\end{align*}
$$

In order to apply a Neumann series argument we wish to guarantee that

$$
\left\|A R-I_{F}\right\|_{F \rightarrow F} \leq 2^{-1}, \quad\left\|R A-I_{E}\right\|_{E \rightarrow E} \leq 2^{-1} \quad \text { for } T \in\left(0, T_{0}\right]
$$

If this is true then the operators $A R=I_{F}-\left(I_{F}-A R\right) \in \mathcal{B}(F(T))$ and $R A=I_{E}-\left(I_{E}-R A\right) \in$ $\mathcal{B}(E(T))$ are invertible for all $T \in\left(0, T_{0}\right]$ and $A$ has the inverse $R(A R)^{-1}=(R A)^{-1} R$. If further $M>0$ is a bound for $R$, then $2 M$ is a bound for $A^{-1}$.

Hence, in view of (3.55), it remains to guarantee the commutator estimates

$$
\begin{array}{r}
\left\|A \Psi_{E, j}-\Psi_{F, j} A_{j}\right\|_{P_{E, j} E_{j} \rightarrow F} \leq \varepsilon \\
\left\|\Phi_{F, j} A-A_{j} \Phi_{E, j}\right\|_{P_{E} E \rightarrow F_{j}} \leq \varepsilon \tag{3.56b}
\end{array}
$$

for every given $\varepsilon$ by choosing $A_{j}, \Phi_{j}$, and $\Psi_{j}$ suitably and $T_{0}$ sufficiently small, whereas the other operators should remain uniformly bounded.
3.3.2. Time regularity of the pressure. We consider the equations

$$
\left\{\begin{align*}
\rho \partial_{t} u-\mu \Delta u+\nabla \pi=f_{u} & \text { in } J \times \Omega \backslash \Sigma  \tag{3.57}\\
\operatorname{div} u=0 & \text { in } J \times \Omega \backslash \Sigma \\
\left.u\right|_{\partial \Omega} \cdot \nu=0 & \text { on } J \times \partial \Omega \\
\llbracket u \rrbracket \cdot \nu=0 & \text { on } J \times \Sigma
\end{align*}\right.
$$

where $\Omega \subset \mathbb{R}^{n}(n \geq 2)$ is a domain with (possibly empty) $C^{2-}$-boundary and $\Sigma \subset \Omega$ is a (possibly empty) closed $C^{2-}$-hypersurface. Let further $p \in(1, \infty)$ and assume that $\Omega$ and $\Sigma$ satisfy Assumption 2.1 where the bound $\eta=\eta\left(n, p, \rho^{-1}\right)>0$ for $\|\nabla \omega\|_{\infty}$ is chosen such that Theorem 2.2 is applicable. In this case we obtain a bounded solution operator

$$
g_{0} \mapsto \psi, \quad\left(L_{p} \cap \hat{H}_{p}^{-1}\right)(\Omega) \rightarrow\left(\dot{\mathcal{H}}_{p}^{2} \cap \dot{\mathcal{H}}_{p}^{1}\right)(\Omega \backslash \Sigma) /\left(\rho^{-1} \mathbb{K}\right)
$$

for the elliptic transmission problem

$$
-\Delta \psi=g_{0} \text { in } \Omega, \quad \partial_{\nu} \psi=0 \text { on } \partial \Omega, \quad \llbracket \partial_{\nu} \psi \rrbracket=0 \text { on } \Sigma, \quad \llbracket \rho \psi \rrbracket=0 \text { on } \Sigma .
$$

With methods from [Köh13, Proposition 7.14], [KPW13, Corollary 1], and [Wil13, Lemma 2.1.1], we will prove the following temporal regularity result for the pressure $\pi$, where we let $\langle\phi\rangle_{K}:=|K|^{-1} \int_{K} \phi d x$ denote the mean value of $\phi \in L_{1}(K)$ for a bounded domain $K$.
3.22. Lemma. Let $\rho_{1}, \rho_{2}, \mu_{1}, \mu_{2} \in(0, \infty), J=(0, T)$ with $T \in(0, \infty], p \in(1, \infty)$, and $\alpha \in$ $(0,1 / 2-1 / 2 p]$. Let $K$ be a bounded $C^{1}$-subdomain of $\Omega$ and suppose that $\left(u, \pi, f_{u}\right)$ satisfies (3.57) and

$$
\left\{\begin{align*}
u & \in \mathbb{E}_{u}=H_{p}^{1}\left(J ; L_{p}(\Omega)^{n}\right) \cap L_{p}\left(J ; H_{p}^{2}(\Omega \backslash \Sigma)^{n}\right),  \tag{3.58}\\
\pi & \in \mathbb{E}_{\pi}=L_{p}\left(J ; \dot{H}_{p}^{1}(\Omega \backslash \Sigma)\right), \\
\llbracket \pi \rrbracket & \in W_{p}^{\alpha}\left(J ; L_{p}(\Sigma)\right), \\
f_{u}=f_{u, \alpha}+\rho f_{u, \sigma} & \in W_{p}^{\alpha}\left(J ; L_{p}(\Omega)^{n}\right)+\rho L_{p}\left(J ; L_{p, \sigma}(\Omega)\right)
\end{align*}\right.
$$

Then the following estimate is valid with some $C=C(n, p, K, T)>0$.

$$
\begin{equation*}
\left\|\pi-\langle\pi\rangle_{K}\right\|_{W_{p}^{\alpha}\left(J ; L_{p}(K)\right)} \leq C\left(\|u\|_{\mathbb{E}_{u}}+\left\|f_{u, \alpha}\right\|_{W_{p}^{\alpha}\left(J ; L_{p}(\Omega)\right)}+\|\llbracket \pi \rrbracket\|_{W_{p}^{\alpha}\left(J ; L_{p}(\Sigma)\right)}\right) . \tag{3.59}
\end{equation*}
$$

Moreover, the number $C$ is uniform with respect to $T \in(0, \infty]$ under the restrictions

$$
u \in_{0} \mathbb{E}_{u}, \quad \llbracket \pi \rrbracket \in{ }_{0} W_{p}^{\alpha}\left(J ; L_{p}(\Sigma)\right), \quad f_{u, \alpha} \in{ }_{0} W_{p}^{\alpha}\left(J ; L_{p}(\Omega)^{n}\right), \quad \alpha \neq 1 / p .
$$

Proof. For $g \in L_{p^{\prime}}(K)$ we define a function $g_{0} \in L_{p^{\prime}}(\Omega)$ by $g_{0}(x):=g(x)-\langle g\rangle_{K}$ for $x \in K$ and $g_{0}(x)=0$ for $x \in \Omega \backslash K$. The Poincaré-Wirtinger inequality for $H_{p}^{1}(K)$ implies

$$
\left\|g_{0}\right\|_{\hat{H}_{p^{\prime}}^{-1}(\Omega)}=\sup _{\phi \in \mathcal{D}\left(\mathbb{R}^{n}\right) \backslash\{0\}} \frac{\left|\int_{\Omega} g_{0} \phi d x\right|}{\|\nabla \phi\|_{L_{p}(\Omega)}}=\sup _{\phi \in \mathcal{D}\left(\mathbb{R}^{n}\right) \backslash\{0\}} \frac{\left|\int_{K} g_{0}\left(\phi-\langle\phi\rangle_{K}\right) d x\right|}{\|\nabla \phi\|_{L_{p}(\Omega)}} \leq C(K)\|g\|_{L_{p^{\prime}}(K)} .
$$

By Theorem 2.2, we can find some $\rho \psi \in \dot{\mathcal{H}}_{p^{\prime}}^{2}(\Omega \backslash \Sigma) \cap \dot{\mathcal{H}}_{p^{\prime}}^{1}(\Omega)$ such that

$$
-\Delta \psi=g_{0} \text { in } \Omega, \quad \partial_{\nu} \psi=0 \text { on } \partial \Omega, \quad \llbracket \partial_{\nu} \psi \rrbracket=0 \text { on } \Sigma, \quad \llbracket \rho \psi \rrbracket=0 \text { on } \Sigma,
$$

which satisfies the estimate $\|\nabla \psi\|_{H_{p^{\prime}}^{1}(\Omega \backslash \Sigma)} \leq C\|g\|_{L_{p^{\prime}}(K)}$.
For $\pi_{0}:=\pi-\langle\pi\rangle_{K}$ and $g, g_{0}$, and $\psi$ as above, an integration by parts yields

$$
-\int_{K} \pi_{0} g d x=-\int_{\Omega} \pi_{0} g_{0} d x=\int_{\Omega} \pi_{0} \Delta \psi d x=-\int_{\Omega} \nabla \pi \cdot \nabla \psi d x-\int_{\Sigma} \llbracket \pi \rrbracket \partial_{\nu} \psi d \sigma
$$

By using the equations in (3.57) and integrating by parts, we obtain

$$
\begin{aligned}
-\int_{K} \pi_{0} g d x= & \int_{\Omega} \mu \nabla u: \nabla^{2} \psi d x-\int_{\partial \Omega} \mu \partial_{\nu} u \cdot \nabla \psi d \sigma+\int_{\Sigma} \llbracket \mu \partial_{\nu} u \cdot \nabla \psi \rrbracket d \sigma \\
& -\int_{\Omega} f_{u, \alpha} \cdot \nabla \psi d x-\int_{\Sigma} \llbracket \pi \rrbracket \partial_{\nu} \psi d \sigma=:\left\langle F_{u, f_{u, \alpha}, \llbracket \pi \rrbracket}, g\right\rangle .
\end{aligned}
$$

The duality $L_{p}(K)^{*} \cong L_{p^{\prime}}(K)$ yields the estimate

$$
\begin{aligned}
\left\|\pi_{0}(t)\right\|_{L_{p}(K)}=\left\|F_{u, f_{u, \alpha}, \llbracket \pi \rrbracket}(t)\right\|_{L_{p^{\prime}}(K)^{*}} \lesssim & \|u(t)\|_{H_{p}^{1}(\Omega)}+\left\|\partial_{\nu} u(t)\right\|_{L_{p}(\partial \Omega)}+\left\|\mu_{ \pm} \partial_{\nu} u_{ \pm}(t)\right\|_{L_{p}(\Sigma)} \\
& +\left\|f_{u, \alpha}(t)\right\|_{L_{p}(\Omega)}+\|\llbracket \pi(t) \rrbracket\|_{L_{p}(\Sigma)} .
\end{aligned}
$$

In order to apply the $W_{p}^{\alpha}(0, T)$-seminorm, we observe that

$$
\left\|\pi_{0}(t)-\pi_{0}(s)\right\|_{L_{p}(K)}=\left\|F_{u(t)-u(s), f_{u, \alpha}(t)-f_{u, \alpha}(s), \llbracket \pi(t) \rrbracket-\llbracket \pi(s) \rrbracket}\right\|_{L_{p^{\prime}}(K)^{*}} .
$$

Hence, for some number $C=C(n, p, K)$, which does not depend on $T \in(0, \infty]$, we have

$$
\begin{aligned}
\left\|\pi_{0}\right\|_{W_{p}^{\alpha}\left(J ; L_{p}(K)\right)} \leq & C\left(\|u\|_{W_{p}^{\alpha}\left(J ; H_{p}^{1}(\Omega)\right)}+\left\|\partial_{\nu} u\right\|_{W_{p}^{\alpha}\left(J ; L_{p}(\partial \Omega)\right.}+\left\|\partial_{\nu} u_{ \pm}\right\|_{W_{p}^{\alpha}\left(J ; L_{p}(\Sigma)\right)}\right) \\
& +C\left(\left\|f_{u, \alpha}\right\|_{W_{p}^{\alpha}\left(J ; L_{p}(\Omega)\right)}+\|\llbracket \pi\|_{W_{p}^{\alpha}\left(J ; L_{p}(\Sigma)\right)}\right) .
\end{aligned}
$$

Since $\alpha \leq 1 / 2-1 / 2 p$, the trace theorem (Theorem B.32) and the mixed derivative embeddings (Proposition B.44) yield a constant $C=C(n, p, K, T)$ such that

$$
\left\|\pi_{0}\right\|_{W_{p}^{\alpha}\left(J ; L_{p}(K)\right)} \leq C\left(\|u\|_{\mathbb{E}_{u}(T)}+\left\|f_{u, \alpha}\right\|_{W_{p}^{\alpha}\left(J ; L_{p}(\Omega)\right)}+\|\llbracket \pi \rrbracket\|_{W_{p}^{\alpha}\left(J ; L_{p}(\Sigma)\right)}\right) .
$$

Therefore the asserted estimate (3.59) is valid. Uniform estimates with respect to $T$ follow by extension and restriction (Lemma B.9).
3.3.3. Local operators. With the spaces from page 72, we define the space of solutions

$$
E(T):={ }_{0} \mathbb{E}(J, \Omega, \Sigma):={ }_{0} \mathbb{E}_{u, v, w, \partial_{\nu} w}(J, \Omega, \Sigma) \times{ }_{0} \mathbb{E}_{\pi, \llbracket \pi \rrbracket}(J, \Omega, \Sigma) \times{ }_{0} \mathbb{E}_{h}(J, \Sigma),
$$

and the space of data

$$
F(T):={ }_{0} \mathbb{F}(J, \Omega, \Sigma):=\mathbb{F}_{u}(J, \Omega) \times{ }_{0} \mathbb{F}_{d, \Sigma}(J, \Omega, \Sigma) \times{ }_{0} \mathbb{G}_{v}(J, \Sigma) \times{ }_{0} \mathbb{G}_{w}(J, \Sigma) \times{ }_{0} \mathbb{G}_{h}(J, \Sigma) .
$$

In order to define the local spaces $E_{j}$ and $F_{j}$ we employ Lemma 2.9 on page 29, which implies that for every $\eta>0$ there is $r_{0}(\eta)>0$ such that for every $r \in\left(0, r_{0}(\eta)\right]$ we can find an $(\eta, r)$-localization set-up for $(\Omega, \Sigma)$ in the sense of Definition 2.8. Hence for some finite set
$I=I(\eta, r)$ there exist an open covering for $\bar{\Omega}$ of balls $U_{j}=B_{r}\left(p_{j}\right)(j \in I)$ and there are rigid transformations

$$
\Theta_{j}: x \mapsto p_{j}+Q_{j} x, \quad B_{r}(0) \rightarrow U_{j},
$$

and height functions $\omega_{j} \in C_{c}^{\infty}\left(\mathbb{R}^{n-1}\right)$ with $\left\|\omega_{j}\right\|_{B C^{1} \cap H_{p}^{2}} \leq \eta$. Furthermore, the index set can be decomposed into $I=I_{1} \cup I_{2} \cup I_{3}$, where $j \in I_{1}$ corresponds to the whole space case $\Omega \cap U_{j}=$ $\Theta_{j}\left(\mathbb{R}^{n} \cap B_{r}\right), j \in I_{2}$ corresponds to the bent half-space case $\Omega \cap U_{j}=\Theta_{j}\left(\mathbb{R}_{\omega_{j}}^{n} \cap B_{r}\right)$, and $j \in I_{3}$ corresponds to the bent hyperplane case $\Sigma \cap U_{j}=\Theta_{j}\left(\Sigma_{\omega_{j}} \cap B_{r}\right)$. We define

$$
\begin{array}{lll}
\Omega_{j}:=\mathbb{R}^{n}, & \Sigma_{j}:=\emptyset & \text { for } j \in I_{1}, \\
\Omega_{j}:=\mathbb{R}_{\omega_{\omega}}^{n}, & \Sigma_{j}:=\emptyset & \text { for } j \in I_{2}, \\
\Omega_{j}:=\mathbb{R}^{n}, & \Sigma_{j}:=\Sigma_{\omega_{j}} & \text { for } j \in I_{3} .
\end{array}
$$

Then we define the local spaces

$$
E_{j}(T)={ }_{0} \mathbb{E}\left(J, \Omega_{j}, \Sigma_{j}\right), \quad F_{j}(T)={ }_{0} \mathbb{F}\left(J, \Omega_{j}, \Sigma_{j}\right) \quad \text { for } j \in I_{1} \cup I_{2} \cup I_{3},
$$

where in the case $j \in I_{1} \cup I_{2}$ we identify

$$
{ }_{0} \mathbb{E}\left(J, \Omega_{j}, \emptyset\right) \cong\left\{u \in{ }_{0} \mathbb{E}_{u}\left(J, \Omega_{j}, \emptyset\right):\left.u\right|_{\partial \Omega_{j}}=0\right\} \times \mathbb{E}_{\pi}, \quad{ }_{0} \mathbb{F}\left(J, \Omega_{j}, \emptyset\right) \cong \mathbb{F}_{u}\left(J, \Omega_{j}\right) \times{ }_{0} \mathbb{F}_{d}\left(J, \Omega_{j}\right) .
$$

We choose a partition of unity $\left(\varphi_{j}\right)_{j \in I}$ for $\bar{\Omega}$ in $\mathbb{R}^{n}$ subordinate to $\left(U_{j}\right)_{j \in I}$ and choose another family of cut-off functions $\left(\psi_{j}\right)_{j \in I}$ with $\operatorname{supp} \psi_{j} \subset U_{j}$ and $\psi_{j}=1$ on $\operatorname{supp} \varphi_{j}$. Then we have $\sum_{j} \psi_{j} \varphi_{j}=1$ in $\bar{\Omega}$ and we define approximation systems for $E$ and $F$ by

$$
\begin{aligned}
\Phi_{E, j}(u, \pi, h) & :=\left(Q_{j}^{\top}\left(\varphi_{j} u\right),\left(\varphi_{j} \pi\right),\left(\varphi_{j} h\right)\right) \circ \Theta_{j}, \\
\Psi_{E, j}\left(u_{j}, \pi_{j}, h_{j}\right): & =\left(Q_{j}\left(\psi_{j} u_{j}\right),\left(\psi_{j} \pi_{j}\right),\left(\psi_{j} h_{j}\right)\right) \circ \Theta_{j}^{-1}, \\
\Phi_{F, j}\left(f_{u}, f_{d}, g_{v}, g_{w}, g_{h}\right) & :=\left(Q_{j}^{\top}\left(\varphi_{j} f_{u}\right),\left(\varphi_{j} f_{d}\right), Q_{j}^{\top}\left(\varphi_{j} g_{v}\right),\left(\varphi_{j} g_{w}\right),\left(\varphi_{j} g_{h}\right)\right) \circ \Theta_{j}, \\
\Psi_{F, j}\left(f_{u j}, f_{d j}, g_{v j}, g_{w j}, g_{h j}\right): & :\left(Q_{j}\left(\psi_{j} f_{u j}\right),\left(\psi_{j} f_{d j}\right), Q_{j}\left(\psi_{j} g_{v j}\right),\left(\psi_{j} g_{w j}\right),\left(\psi_{j} g_{h j}\right)\right) \circ \Theta_{j}^{-1} .
\end{aligned}
$$

The relevant mapping properties of these maps follow as in Lemma 3.17.
Problem (3.1) induces a bounded linear operator $A: E \rightarrow F$ by

$$
A(u, \pi, h):=\left[\begin{array}{c}
\rho \partial_{t} u-\mu \Delta u+\nabla \pi  \tag{3.60}\\
\operatorname{div} u \\
L_{v}\left(u, h ; u_{*}\right) \\
L_{w}\left(u, \pi, h ; u_{*}\right) \\
\partial_{t} h-u \cdot \nu_{\Sigma}
\end{array}\right] \quad \text { for }(u, \pi, h) \in E
$$

For $j \in I_{1} \cup I_{2}$ we define the local operators $A_{j}: E_{j} \rightarrow F_{j}$ by

$$
A_{j}(u, \pi):=\left[\begin{array}{c}
\rho \partial_{t} u-\mu \Delta u+\nabla \pi  \tag{3.61}\\
\operatorname{div} u
\end{array}\right] \quad \text { for }(u, \pi) \in E_{j}, j \in I_{1} \cup I_{2} .
$$

The results of Bothe and Prüss ([BP07, Theorem 5.1, Theorem 6.1]) imply that $A_{j}: E_{j} \rightarrow F_{j}$ is invertible for $j \in I_{1} \cup I_{2}$ and $\omega_{j}=0$. For $j \in I_{2}$ and $\omega_{j} \neq 0$ we employ the following result.
3.23. Lemma. Let $n \geq 2, \rho, \mu>0$, and $p \in(n+2, \infty)$.

Then there exists $\eta>0$ such that for given $R>0$ we can find a number $T_{0}(R)>0$ such that the solution-to-data map $(u, \pi) \mapsto\left(f_{u}, f_{d}\right),{ }_{0} \mathbb{E}_{u} \times \mathbb{E}_{\pi} \rightarrow \mathbb{F}_{u} \times{ }_{0} \mathbb{F}_{d}$ of problem

$$
\left\{\begin{align*}
\rho \partial_{t} u-\mu \Delta u+\nabla \pi=f_{u} & \text { in } J \times \mathbb{R}_{\omega}^{n}  \tag{3.62}\\
\operatorname{div} u=f_{d} & \text { in } J \times \mathbb{R}_{\omega}^{n}
\end{align*}\right.
$$

is uniformly invertible with respect to $T \in\left(0, T_{0}\right]$ and

$$
\begin{equation*}
\omega \in B C^{3}\left(\mathbb{R}^{n-1}\right),\|\omega\|_{B C^{1}} \leq \eta,\|\omega\|_{B C^{3}} \leq R . \tag{3.63}
\end{equation*}
$$

Proof. We employ the transformation $\Theta_{\omega_{j}}: \mathbb{R}_{+}^{n} \rightarrow \mathbb{R}_{\omega_{j^{\prime}}}^{n}\left(x^{\prime}, x_{n}\right) \mapsto\left(x^{\prime}, x_{n}+\omega_{j}\left(x^{\prime}\right)\right)$ from equation (3.27) on page 72. As in Lemma 3.17, we define the transformed functions

$$
\bar{u}=\left[\partial \Theta_{\omega_{j}}\right]^{-1} u \circ \Theta_{\omega_{j}}, \quad \bar{\pi}=\pi \circ \Theta_{\omega_{j}}, \quad \bar{f}_{u}:=\left[\partial \Theta_{\omega_{j}}\right]^{-1} f_{u} \circ \Theta_{\omega_{j}} \quad \bar{f}_{d}:=f_{d} \circ \Theta_{\omega_{j}} .
$$

Then (3.62) is equivalent to

$$
\left\{\begin{align*}
\left(\rho \partial_{t}-\mu \Delta\right) \bar{u}+\nabla \bar{\pi} & =\bar{f}_{u}+F_{u}\left(\bar{u}, \bar{\pi} ; \omega_{j}\right) & & \text { in } J \times \mathbb{R}_{+}^{n},  \tag{3.64}\\
\operatorname{div} \bar{u} & =\bar{f}_{d} & & \text { in } J \times \mathbb{R}_{+}^{n},
\end{align*}\right.
$$

where the perturbation $F_{u}$ is defined by

$$
F_{u}\left(\bar{u}, \bar{\pi} ; \omega_{j}\right):=\bar{\mu}\left(\left[\partial \Theta_{\omega_{j}}\right]^{-1}(\Delta u) \circ \Theta_{\omega_{j}}-\Delta \bar{u}\right)+\left(I-\left[\partial \Theta_{\omega_{j}}\right]^{-1}\left[\partial \Theta_{\omega_{j}}\right]^{-\top}\right) \nabla \bar{\pi},
$$

and the difference $(\Delta u) \circ \Theta_{\omega_{j}}-\Delta \bar{u}$ can be expressed by (3.50). As for Theorem 3.16 it follows that the map $(\bar{u}, \bar{\pi}) \rightarrow\left(\bar{f}_{u}, \bar{f}_{d}\right)$ induced by (3.64) is uniformly invertible with respect to $T \in\left(0, T_{0}\right.$ ] (3.63) for some $\eta>0$ and $T_{0}(\eta)>0$. The proof of Lemma 3.17 shows that the transformation $\left(\bar{u}, \bar{\pi}, \bar{f}_{u}, \bar{f}_{d}\right) \rightarrow\left(u, \pi, f_{u}, f_{d}\right)$ is uniformly invertible and this yields the assertion.

For the case $j \in I_{3}$ we first define the local coefficients of $A_{j}$. These depend on the functions

$$
L_{\Sigma}=\tau_{\Sigma}^{k} \otimes \partial_{k} \nu_{\Sigma}, \quad w_{*}=\nu_{\Sigma} \cdot u_{*}, \quad D_{\Sigma}\left(v_{*}\right)=\operatorname{sym}\left(\left[\tau_{\Sigma}^{k} \otimes \partial_{k} v_{*}\right] P_{\Sigma}\right),
$$

with $v_{*}=P_{\Sigma} u_{*}$. Their transforms under the rigid map $\Theta_{j}: x \mapsto Q_{j} x+p_{j}$ are given by

$$
L_{\Sigma_{j}}=Q_{j}^{\top}\left[L_{\Sigma} \circ \Theta_{\omega_{j}}\right] Q_{j}, \quad \bar{w}_{*}=w_{*} \circ \Theta_{\omega_{j}}, \quad D_{\Sigma_{j}}\left(\bar{v}_{*}\right)=Q_{j}^{\top}\left[\left(D_{\Sigma}\left(v_{*}\right)\right) \circ \Theta_{\omega_{j}}\right] Q_{j},
$$

where $\bar{v}_{*}=Q_{j}^{\top}\left(v_{*} \circ \Theta_{\omega_{j}}\right)$.
As for the construction of $\omega_{j}$ in Lemma 2.9, we fix a cut-off function $\chi \in \mathcal{B}\left(\mathbb{R}^{n}\right)$ with $0 \leq$ $\chi \leq 1, \chi(x)=1$ for $|x| \leq 1$, and $\chi(x)=0$ for $|x| \geq 2$. For a given function $\psi$ on $J \times\left(\Sigma_{j} \cap B_{r_{0}}\right)$ and $r \in\left(0, r_{0} / 2\right]$ we define another function $\tilde{\psi}_{r}$ on $J \times \Sigma_{j}$ by

$$
\tilde{\psi}_{r}(t, x):=\left(S_{r} \psi\right)(t, x):=\psi(0,0)+ \begin{cases}\chi(x / r)(\psi(t, x)-\psi(0,0)) & \text { for }|x|<2 r \leq r_{0} \\ 0 & \text { for }|x| \geq 2 r .\end{cases}
$$

Then $\tilde{\psi}_{r}(t, x)=\psi(t, x)$ for all $(t, x) \in J \times\left(\Sigma_{j} \cap B_{r}\right)$.
3.24. Proposition (Properties of $S_{r}: \psi \mapsto \tilde{\psi}_{r}$ ). Let $\Sigma=\Sigma_{\omega}$ be a bent $C^{2}$-hyperplane in $\mathbb{R}^{n}$.
(i) For all $r_{0}>0$ there exists $C>0$ such that

$$
\begin{equation*}
\left\|\tilde{\psi}_{r}-\psi(0)\right\|_{B C(\Sigma) \cap H_{p}^{1}(\Sigma)} \leq C r^{\max \{1,(n-1) / p\}}\|\psi\|_{B C^{1}\left(\Sigma \cap B_{2 r}\right)} \tag{3.65}
\end{equation*}
$$

for all $r \in\left(0, r_{0} / 2\right]$ and $\psi \in B C^{1}\left(\Sigma \cap B_{r_{0}}\right)$.
(ii) For all $r_{0}>0$ there exists $C>0$ such that

$$
\begin{equation*}
\left\|\tilde{\psi}_{r}-\psi(0)\right\|_{B C^{1}(\Sigma) \cap H_{p}^{2}(\Sigma)} \leq C r^{\max \{1,(n-1) / p\}}\|\psi\|_{B C^{2}\left(\Sigma \cap B_{2 r}\right)} . \tag{3.66}
\end{equation*}
$$

for all $r \in\left(0, r_{0} / 2\right], \psi \in B C^{2}\left(\Sigma \cap B_{r_{0}}\right)$ with $\nabla \psi(0)=0$.
(iii) For all $T_{1}>0, r_{0}>0$, and $\gamma \in(0,1)$ there exists $C>0$ such that

$$
\begin{equation*}
\left\|\tilde{\psi}_{r}-\psi(0,0)\right\|_{C\left([0, T] ; B C(\Sigma) \cap H_{p}^{1}(\Sigma)\right)} \leq C r^{\max \{1,(n-1) / p\}}\left(1+T^{\gamma} / r\right)\|\psi\|_{C^{\gamma}\left([0, T] ; B C^{1}\left(\Sigma \cap B_{2 r}\right)\right)} \tag{3.67}
\end{equation*}
$$

for all $r \in\left(0, r_{0} / 2\right]$ and $\psi \in C^{\gamma}\left(\left[0, T_{1}\right] ; B C^{1}\left(\Sigma \cap B_{r_{0}}\right)\right)$.
Given $\eta>0$, the number $C$ in (i) to (iii) is uniform with respect to $\|\nabla \omega\|_{\infty} \leq \eta$.

Proof. With the substitution $x=r y$ we obtain the identities $\|\chi(\cdot / r)\|_{p}=r^{(n-1) / p}\|\chi\|_{p}$ and $\|(\nabla \chi)(\cdot / r)\|_{p}=r^{(n-1) / p}\|\nabla \chi\|_{p}$. We will also use the inequalities

$$
\begin{aligned}
& |\psi(x)-\psi(0)| \leq\left(1+\eta^{2}\right)^{1 / 2}|x|\left\|\nabla_{\Sigma} \psi\right\|_{\infty}, \\
& |\psi(x)-\psi(0)| \leq\left(1+\eta^{2}\right)^{1 / 2}|x|^{2}\left\|\nabla_{\Sigma}^{2} \psi\right\|_{\infty} \quad \text { if } \nabla_{\Sigma} \psi(0)=0,
\end{aligned}
$$

which follow from Proposition A. 12 on page 133. Then (i) and (ii) are readily checked.
Next, for $t \in[0, T]$ we have

$$
\begin{aligned}
\left\|\nabla \tilde{\psi}_{r}(t, \cdot)\right\|_{p} & \leq C r^{(n-1) / p}\|\chi\|_{H_{p}^{1}}\|\nabla \psi(t, \cdot)\|_{B C\left(\Sigma \cap B_{2 r}\right)}+\left\|r^{-1} \nabla \chi(\cdot / r)\right\|_{p}|\psi(t, 0)-\psi(0,0)| \\
& \leq C r^{(n-1) / p}\|\chi\|_{H_{p}^{1}}\left(1+T^{\gamma} / r\right)\|\psi\|_{C \gamma\left([0, T] ; B C\left(\Sigma \cap B_{2 r}\right)\right)} .
\end{aligned}
$$

The estimate of $\left\|\tilde{\psi}_{r}(t, \cdot)\right\|_{p}$ is similar and hence (iii) is valid.
For given $u_{*} \in \mathcal{P}_{M, T_{1}}$ and $r>0$ we define

$$
\left\{\begin{align*}
\vartheta_{j} & :=\left(\vartheta_{L, j}, \vartheta_{w, j}, \vartheta_{D v, j}\right), & \vartheta_{j}^{*} & :=\left(\vartheta_{L, j}^{*}, \vartheta_{w, j}^{*}, \vartheta_{D v, j}^{*}\right),  \tag{3.68}\\
\vartheta_{L, j} & :=S_{r}\left(Q_{j}^{\top}\left[L_{\Sigma} \circ \Theta_{\omega_{j}}\right] Q_{j}\right), & \vartheta_{L, j}^{*} & :=Q_{j}^{\top}\left[L_{\Sigma}\left(p_{j}\right)\right] Q_{j}, \\
\vartheta_{w, j} & :=S_{r}\left(w_{*} \circ \Theta_{\omega_{j}}\right), & \vartheta_{w, j}^{*} & :=w_{*}\left(0, p_{j}\right), \\
\vartheta_{D v, j} & :=S_{r}\left(Q_{j}^{\top}\left[\left(D_{\Sigma} v_{*}\right) \circ \Theta_{\omega_{j}}\right] Q_{j}\right), & \vartheta_{D v, j}^{*} & :=Q_{j}^{\top}\left[\left(D_{\Sigma} v_{*}\right)\left(0, p_{j}\right)\right] Q_{j} .
\end{align*}\right.
$$

Then the local operators $A_{j}: E_{j} \rightarrow F_{j}$ for $j \in I_{3}$ are defined by

$$
A_{j}(u, \pi, h):=\left[\begin{array}{c}
\rho \partial_{t} u-\mu \Delta u+\nabla \pi  \tag{3.69}\\
\operatorname{div} u \\
L_{v}\left(u, h ; \omega_{j}, \vartheta_{j}\right) \\
L_{w}\left(u, \pi, h ; \omega_{j}, \vartheta_{j}\right) \\
\partial_{t} h-u \cdot \nu_{\Sigma_{j}}
\end{array}\right] \quad \text { for }(u, \pi, h) \in E_{j}, j \in I_{3},
$$

where $L_{v}$ and $L_{w}$ are defined on page 71 .
3.25. Corollary. Let $p \in(n+2, \infty)$ and $M, T_{1}>0$. Then there are positive functions $r(\cdot)$ and $T_{0}(\cdot)$ such that for some $\eta_{0}>0$ and every $\eta \in\left(0, \eta_{0}\right]$, the pair $(\Omega, \Sigma)$ has an $(\eta, r(\eta))$-localization set-up and the local operators $A_{j}: E_{j} \rightarrow F_{j}\left(j \in I_{1} \cup I_{2} \cup I_{3}\right)$ defined by equations (3.61) and (3.69) are uniformly invertible with respect to $T \in\left(0, T_{0}(\eta)\right], j \in I$, and $u_{*} \in \mathcal{P}_{M, T_{1}}$.
Proof. For given $M>0$ there exists $M_{1}=M_{1}(M, \Sigma) \geq M$ such that

$$
\sup _{x \in \Sigma}\left|\left(L_{\Sigma}(x), w_{*}(0, x), D_{\Sigma} v_{*}(0, x)\right)\right| \leq M_{1}, \quad \inf _{\Sigma} d_{0}\left(D_{\Sigma}\left(\left.u_{*}\right|_{t=0}\right)\right) \geq M_{1}^{-1}
$$

for all $u_{*} \in \mathcal{P}_{M, T_{1}}$ and $T_{1} \in(0, \infty)$. For given $\eta>0$, Lemma 2.9 yields a positive number $r_{0}(\eta)$ such that for every $r \in\left(0, r_{0}(\eta)\right]$ the pair $(\Omega, \Sigma)$ has an $(\eta, r)$-localization set-up such that $\left\|\omega_{j}\right\|_{B C^{1} \cap H_{p}^{2}} \leq \eta$ for all $j \in I_{2} \cup I_{3}$ and there exists $R(r)>0$ such that $\left\|\omega_{j}\right\|_{B C^{4}} \leq R(r)$ for $j \in I_{2} \cup I_{3}$. Sobolev's embedding and the mixed derivative embeddings yield an estimate

$$
\left\|\left(w_{*}, D_{\Sigma} v_{*}\right)\right\|_{C^{\gamma}\left([0, T] ; B C^{1}(\Sigma)\right)} \leq C M \quad \text { for all } u_{*} \in \mathcal{P}_{M, T_{1}}
$$

for some $\gamma>0$ and $C \geq 1$. By Proposition 3.24 we can find a positive number $r_{1}(\eta) \leq r_{0}(\eta)$ and a positive function $r \mapsto R(r)$ such that the parameters $\left(\vartheta_{j}^{*}, \vartheta_{j}\right)$ from (3.68) satisfy

$$
\begin{aligned}
\left\|\vartheta_{L, j}-\vartheta_{L, j}^{*}\right\|_{B C\left(\Sigma_{\omega}\right) \cap H_{p}^{1}\left(\Sigma_{\omega}\right)} \leq \eta, & \left\|\vartheta_{L, j}-\vartheta_{L, j}^{*}\right\|_{B C^{2}\left(\Sigma_{\omega}\right)} \leq R(r), \\
\left\|\vartheta_{w, j}-\vartheta_{w, j}^{*}\right\|_{C\left([0, T] ; B C\left(\Sigma_{\omega}\right) \cap H_{p}^{1}\left(\Sigma_{\omega}\right)\right)} \leq \eta, & \left\|\vartheta_{w, j}-\vartheta_{w, j}^{*}\right\|_{\mathbb{G}_{w}(T)} \leq R(r), \\
\left\|\vartheta_{D v, j}-\vartheta_{D v, j}^{*}\right\|_{C\left([0, T] ; B C\left(\Sigma_{\omega}\right) \cap H_{p}^{1}\left(\Sigma_{\omega}\right)\right)} \leq \eta, & \left\|\vartheta_{D v, j}-\vartheta_{D v, j}^{*}\right\|_{\mathbb{G}_{w}(T)} \leq R(r)
\end{aligned}
$$

for all $r \in\left(0, r_{1}(\eta)\right], j \in I_{3}$, and $u_{*} \in \mathcal{P}_{M, T_{1}}$. Hence $\left(\vartheta_{j}^{*}, \omega_{j}, \vartheta_{j}\right)$ belongs to the set $\mathcal{P}_{C M_{1}, T_{1}, \eta, R(r)}$ from page 71 for all $j \in I_{3}$. By Theorem 3.16 and Lemma 3.23, there exist a positive number $\eta_{0}$
and a function $R \mapsto T_{0}(R)$ such that if $\eta \leq \eta_{0}$ and $r \in\left(0, r_{1}(\eta)\right]$, then the operators $A_{j}: E_{j}(T) \rightarrow$ $F_{j}(T)$ are uniformly invertible with respect to $T \in\left(0, T_{0}(R(r))\right], j \in I$, and $u_{*} \in \mathcal{P}_{M, T_{1}}$.
3.3.4. Commutator estimates. For proving Theorem 3.21 it remains to verify the commutator estimates (3.56) and to construct the operator $R_{0}$.
3.26. Lemma. Let $p \in(\max \{5, n+2\}, \infty)$ and let $M, T_{1},\left(U_{j}, \Theta_{j}, \omega_{j}\right), A_{j}$, and $T_{0}$ be as in Corollary 3.25. Then for all $\varepsilon>0$ there exists $T_{0}^{\prime} \in\left(0, T_{0}\right]$ such that

$$
\begin{align*}
\left\|A \Psi_{E, j}-\Psi_{F, j} A_{j}\right\|_{P_{E, j} E_{j}(T) \rightarrow F(T)} & \leq \varepsilon  \tag{3.70a}\\
\left\|\Phi_{F, j} A-A_{j} \Phi_{E, j}\right\|_{P_{E} E(T) \rightarrow F_{j}(T)} & \leq \varepsilon \tag{3.70b}
\end{align*}
$$

for all $T \in\left(0, T_{0}^{\prime}\right], j \in I$, and $u_{*} \in \mathcal{P}_{M, T_{1}}$.
Proof. It is sufficient to prove estimate (3.70a), since (3.70b) can be proved analogously.
For given $z_{j}=\left(u_{j}, \pi_{j}, h_{j}\right) \in P_{E, j} E_{j}$, the pair $\left(u_{j}, \pi_{j}\right)$ satisfies the assumptions of Lemma 3.22 in $\Omega_{j} \backslash \Sigma_{j}$ and we conclude that $\pi_{j 0}:=\pi_{j}-\left\langle\pi_{j}\right\rangle_{K_{j}}$ belongs to ${ }_{0} W_{p}^{1 / 2-1 / 2 p}\left(J ; L_{p}\left(K_{j}\right)\right)$ for every bounded smooth domain $K_{j} \subset \Omega_{j}$ which contains the support of $\nabla \psi_{j}$ and we have

$$
\begin{equation*}
\left\|\pi_{j}-\left\langle\pi_{j}\right\rangle_{K_{j}}\right\|_{0 W_{p}^{1 / 2-1 / 2 p}\left(J ; L_{p}\left(K_{j}\right)\right)} \leq C\left(K_{j}\right)\left(\left\|u_{j}\right\|_{0 \mathbb{E}_{u}\left(J, \Omega_{j}, \Sigma_{j}\right)}+\left\|\llbracket \pi_{j} \rrbracket\right\|_{0 \mathbb{G}_{w}\left(J, \Sigma_{j}\right)}\right) \tag{3.71}
\end{equation*}
$$

for all $z_{j} \in P_{E, j} E_{j}(T)$ and $T \in\left(0, T_{0}\right]$. We next deal with the cases $j \in I_{1}, I_{2}$, and $I_{3}$ separately.
(1) Perturbed whole-space problem. Let $j \in I_{1}$ be fixed and let $z=\left(u, \pi_{0}\right) \in P_{E, j} E_{j}$. Then

$$
\left(A \Psi_{E, j}-\Psi_{F, j} A_{j}\right) z=\left[\begin{array}{c}
\pi_{0} Q_{j} \nabla \psi_{j}-\mu Q_{j}\left[\Delta, \psi_{j}\right] u  \tag{3.72}\\
\nabla \psi_{j} \cdot u
\end{array}\right] \circ \Theta_{j}^{-1}=:\left[\begin{array}{c}
Q_{j} F_{u j}\left(u, \pi_{0}\right) \\
F_{d j}(u)
\end{array}\right] \circ \Theta_{j}^{-1}
$$

Here we let $[S, T]=S T-T S$ denote the commutator of linear operators $S$ and $T$.
We show that the perturbations $F_{u j}$ and $F_{d j}$ satisfy the estimate

$$
\begin{equation*}
\left\|F_{u j}\left(u, \pi_{0}\right)\right\|_{\mathbb{F}_{u}(T)}+\left\|F_{d j}(u)\right\|_{0 \mathbb{F}_{d}(T)} \leq C T^{1 / 2-1 / 2 p}\|u\|_{0 \mathbb{E}_{u}(T)} \tag{3.73}
\end{equation*}
$$

where $\mathbb{F}_{u}=\mathbb{F}_{u}\left(J, \mathbb{R}^{n}, \emptyset\right)$ and ${ }_{0} \mathbb{F}_{d}:={ }_{0} \mathbb{F}_{d}\left(J, \mathbb{R}^{n}, \emptyset\right)={ }_{0} \mathbb{F}_{d, \Sigma}\left(J, \mathbb{R}^{n}, \emptyset\right)$. From estimate (3.71), the mixed derivative embeddings and the interval dependent estimates in Lemma 3.18 we obtain the following estimates. For all $\delta \in(1 / p, 1 / 2)$ and $T_{0}>0$ we have

$$
\begin{aligned}
\left\|\left[\Delta, \psi_{j}\right] u\right\|_{\mathbb{F}_{u}(T)} & \leq C\left(T_{0}\right)\|u\|_{L_{p}\left(0, T ; H_{p}^{1}\right)} \leq C\left(\delta, T_{0}\right) T^{\delta}\|u\|_{0 W_{p}^{\delta}\left(0, T ; H_{p}^{1}\right)} \leq C\left(\delta, T_{0}\right) T^{\delta}\|u\|_{0 \mathbb{E}_{u}(T)} \\
\left\|\pi_{0} \nabla \psi_{j}\right\|_{\mathbb{F}_{u}(T)} & \leq C\left(T_{0}\right) T^{1 / 2-1 / 2 p}\left\|\pi_{0}\right\|_{0 W_{p}^{1 / 2-1 / 2 p}\left(0, T ; L_{p}\right)} \leq C\left(T_{0}\right) T^{1 / 2-1 / 2 p}\|u\|_{0} \mathbb{E}_{u}(T)
\end{aligned}
$$

The estimates in the divergence space ${ }_{0} \mathbb{F}_{d}(T)=H_{p}^{1}\left(0, T ; \dot{H}_{p}^{-1}\left(\mathbb{R}^{n}\right)\right) \cap L_{p}\left(0, T ; H_{p}^{1}\left(\mathbb{R}^{n}\right)\right)$ are obtained in two steps. First, for $\delta \in(1 / p, 1 / 2)$ and $T_{0}>0$ we have

$$
\left\|\nabla \psi_{j} \cdot u\right\|_{L_{p}\left(0, T ; H_{p}^{1}\right)} \leq C\left(T_{0}\right)\|u\|_{L_{p}\left(0, T ; H_{p}^{1}\right)} \leq T^{\delta} C\left(\delta, T_{0}\right)\|u\|_{0 \mathbb{E}_{u}(T)}
$$

Second, the term $\nabla \psi_{j} \cdot u$ acts as a functional on $\phi \in \dot{H}_{p^{\prime}}^{1}\left(\mathbb{R}^{n}\right)$ in virtue of $\phi \mapsto \int_{\Omega} \nabla \psi_{j} \cdot u \phi d x$. The condition $\operatorname{div} u=0$ yields $\int_{\mathbb{R}^{n}} \nabla \psi_{j} \cdot u d x=0$. Hence $\int_{\mathbb{R}^{n}} \nabla \psi_{j} \cdot u \phi d x=\int_{\mathbb{R}^{n}} \nabla \psi_{j} \cdot u \phi_{0} d x$, where $\phi_{0}=\phi-\langle\phi\rangle_{K_{j}}$ and thus

$$
\begin{align*}
\partial_{t} \int_{\mathbb{R}^{n}} \nabla \psi_{j} \cdot u \phi d x & =\int_{\mathbb{R}^{n}} \nabla \psi_{j} \cdot\left(\frac{\mu}{\rho} \Delta u-\frac{1}{\rho} \nabla \pi\right) \phi_{0} d x  \tag{3.74}\\
& =-\int_{\mathbb{R}^{n}} \frac{\mu}{\rho}(\nabla u)^{\top}: \nabla\left(\nabla \psi_{j} \phi_{0}\right) d x+\int_{\mathbb{R}^{n}} \frac{1}{\rho} \pi_{0} \operatorname{div}\left(\nabla \psi_{j} \phi_{0}\right) d x
\end{align*}
$$

Applying the Poincaré-Wirtinger inequality to $\phi_{0} \in H_{p^{\prime}}^{1}\left(K_{j}\right)$ and using (3.71), we obtain

$$
\left\|\nabla \psi_{j} \cdot u\right\|_{0 H_{p}^{1}\left(0, T ; \dot{H}_{p}^{-1}\right)} \leq C\left(T_{0}\right)\|u\|_{L_{p}\left(0, T ; H_{p}^{1}\right)}+C\left(T_{0}\right)\left\|\pi_{0}\right\|_{L_{p}\left(0, T ; L_{p}\right)} \leq C\left(T_{0}\right) T^{1 / 2-1 / 2 p}\|u\|_{0 \mathbb{E}_{u}(T)}
$$

Therefore estimate (3.73) is valid.
(2) Perturbed half-space problem. In the case $j \in I_{2}, \Omega_{j}=\mathbb{R}_{\omega_{j}}^{n}$, and $\Sigma_{j}=\emptyset$, the commutator is also given by (3.72) and $F_{u j}$ in $\mathbb{F}_{u}\left(J, \Omega_{j}, \Sigma_{j}\right)$ and $F_{d j}$ in $L_{p}\left(0, T ; H_{p}^{1}\left(\Omega_{j}\right)\right)$ can be estimated in the same way as above. In view of $\operatorname{div} u=0$ and $\left.u\right|_{\partial \Omega_{j}}=0$ it remains to estimate the functional

$$
F_{d j}: \phi \mapsto\left\langle\nabla \psi_{j} \cdot u, \phi\right\rangle=-\int_{\Omega_{j}} \psi_{j} u \cdot \nabla \phi d x
$$

in $H_{p}^{1}\left(0, T ; \hat{H}_{p}^{-1}\left(\Omega_{j}\right)\right)$. As for (3.74) we obtain

$$
\begin{aligned}
\partial_{t} \int_{\Omega_{j}} \nabla \psi_{j} \cdot u \phi d x= & -\int_{\Omega_{j}} \frac{\mu}{\rho}(\nabla u)^{\top}: \nabla\left(\nabla \psi_{j} \phi_{0}\right) d x+\int_{\Omega_{j}} \frac{1}{\rho} \pi_{0} \operatorname{div}\left(\nabla \psi_{j} \phi_{0}\right) d x \\
& +\int_{\partial \Omega_{j}}\left(\frac{\mu}{\rho} \nabla \psi \cdot \partial_{\nu} u \phi-\frac{\mu}{\rho} \nabla^{2} \psi: \nabla u \phi-\frac{\partial_{\nu} \psi}{\rho} \pi \phi\right) d\left(\partial \Omega_{j}\right) .
\end{aligned}
$$

For every $\delta \in(1 / p, 1 / 2-3 / 2 p)$ the trace operator

$$
\left.\pi_{0} \mapsto \pi_{0}\right|_{\partial \Omega_{j}}: W_{p}^{1 / 2-1 / 2 p}\left(0, T ; L_{p}\left(\Omega_{j}\right)\right) \cap L_{p}\left(0, T ; H_{p}^{1}\left(\Omega_{j}\right)\right) \hookrightarrow W_{p}^{\delta}\left(0, T ; L_{p}\left(\partial \Omega_{j}\right)\right)
$$

is bounded since $p>5$. Therefore the Poincaré-Wirtinger inequality and Lemma 3.22 yield

$$
\begin{aligned}
\left\|\partial_{t}\left\langle\nabla \psi_{j} \cdot u, \cdot\right\rangle\right\|_{L_{p}\left(0, T ; \hat{H}_{p}^{-1}\left(\Omega_{j}\right)\right)} & \leq C\left\|\left(\nabla u, \pi_{0}\right)\right\|_{L_{p}\left(0, T ; L_{p}\left(\Omega_{j}\right)\right)}+C\left\|\left(\partial_{\nu} u, \pi\right)\right\|_{L_{p}\left(0, T ; L_{p}\left(\partial \Omega_{j}\right)\right)} \\
& \leq C\left(\delta, T_{0}\right) T^{\delta}\|(u, \pi)\|_{0 \mathbb{E}_{u}\left(T, \Omega_{j}, \emptyset\right) \times \mathbb{E}_{\pi}\left(J, \Omega_{j}, \emptyset\right)}
\end{aligned}
$$

(3) Perturbed interface problem. Let $j \in I_{3}$ be fixed. For $z=\left(u, \pi_{0}, h\right) \in P_{E, j} E_{j}$ we have

$$
\left(A \Psi_{E, j}-\Psi_{F, j} A_{j}\right) z=\left[\begin{array}{c}
Q_{j} F_{u j}\left(u, \pi_{0}\right)  \tag{3.75}\\
F_{d j}(u) \\
Q_{j} G_{v j}(u, h) \\
G_{w j}(u) \\
0
\end{array}\right] \circ \Theta_{j}^{-1}
$$

where $F_{u j}, F_{d j}, G_{v j}$, and $G_{w j}$ are the commutators

$$
\begin{aligned}
F_{u j}\left(u, \pi_{0}\right)= & \pi_{0} \nabla \psi_{j}-\Delta \psi_{j} u-2[\nabla u]^{\top} \nabla \psi_{j}, \\
F_{d j}(u)= & \nabla \psi_{j} \cdot u, \\
G_{v j}(u, h)= & -\left[\mu_{s} \widetilde{\Delta}_{\Sigma_{j}}, \psi_{j}\right] v-\lambda_{s}\left[\nabla_{\Sigma_{j}} \operatorname{div}_{\Sigma_{j}}, \psi_{j}\right] v-\llbracket \mu \rrbracket \partial_{\nu} \psi_{j} v \\
& -\llbracket \mu \rrbracket \nabla_{\Sigma_{j}} \psi_{j} w-\left(\lambda_{s}+\mu_{s}\right) \vartheta_{w, j}\left[\nabla_{\Sigma_{j}} \Delta_{\Sigma_{j}}, \psi_{j}\right] h, \\
G_{w j}(u, h)= & -\operatorname{tr}\left(\left[\left(\lambda_{s}-\mu_{s}\right) \operatorname{tr} \vartheta_{L, j}+2 \mu_{s} \vartheta_{L, j}\right] \nabla_{\Sigma_{j}} \psi_{j}\right) v-2 \llbracket \mu \rrbracket \partial_{\nu} \psi_{j} w \\
& -\operatorname{tr}\left(\left[\sigma+\left(\lambda_{s}-\mu_{s}\right)\left(\operatorname{tr} \vartheta_{D v, j}-2 \operatorname{tr} \vartheta_{L, j} \vartheta_{w, j}\right)+2 \mu_{s}\left(\vartheta_{D v, j}-2 \vartheta_{w, j} \vartheta_{L, j}\right)\right]\left[\nabla_{\Sigma_{j}}^{2}, \psi_{j}\right] h\right) .
\end{aligned}
$$

Clearly, $F_{u j}$ in $L_{p}\left(J \times \Omega_{j}\right)$ and $F_{d j}$ in $L_{p}\left(0, T ; H_{p}^{1}\left(\Omega_{j} \backslash \Sigma_{j}\right)\right)$ can be estimated as in the case $j \in I_{1}$. Due to div $u=0$ and $\llbracket \nu_{\Sigma} \cdot u \rrbracket=0$, the functional $F_{d j} \in H_{p}^{1}\left(0, T ; \hat{H}_{p}^{-1}\left(\Omega_{j}\right)\right)$ is given by

$$
\phi \mapsto\left\langle\nabla \psi_{j} \cdot u, \phi\right\rangle=\int_{\Omega_{j}} \nabla \psi_{j} \cdot u \phi d x=-\int_{\Omega_{j}} \psi_{j} u \cdot \nabla \phi d x .
$$

Let $\delta \in(1 / p, 1 / 2-3 / 2 p)$. With

$$
\begin{aligned}
\partial_{t} \int_{\Omega_{j}} \nabla \psi_{j} \cdot u \phi d x= & -\int_{\Omega_{j}} \rho^{-1} \mu(\nabla u)^{\top}: \nabla\left(\nabla \psi_{j} \phi_{0}\right) d x+\int_{\Omega_{j}} \rho^{-1} \pi_{0} \operatorname{div}\left(\nabla \psi_{j} \phi_{0}\right) d x \\
& -\int_{\Sigma_{j}}\left[\left[\rho^{-1} \mu \nabla \psi \cdot \partial_{\nu} u \phi-\rho^{-1} \mu \nabla^{2} \psi: \nabla u \phi-\rho^{-1} \partial_{\nu} \psi \pi \phi\right]\right] d \Sigma_{j}
\end{aligned}
$$

and the pressure estimate (3.71), we obtain

$$
\begin{aligned}
\left\|\partial_{t}\left\langle\nabla \psi_{j} \cdot u, \cdot\right\rangle\right\|_{L_{p}\left(0, T ; \hat{H}_{p}^{-1}\left(\Omega_{j}\right)\right)} & \leq C\left\|\left(\nabla u, \pi_{0}\right)\right\|_{L_{p}\left(0, T ; L_{p}\left(\Omega_{j}\right)\right)}+C\left\|\left(\partial_{\nu} u, \pi_{ \pm}\right)\right\|_{L_{p}\left(0, T ; L_{p}\left(\Sigma_{j}\right)\right)} \\
& \leq C\left(\delta, T_{0}\right) T^{\delta}\|(u, \pi)\|_{0 \mathbb{E}_{u, v, w, \partial_{\nu} w}\left(J, \Omega_{j}, \Sigma_{j}\right) \times \times_{0} \mathbb{E}_{\pi,[\pi]}\left(J, \Omega_{j}, \Sigma_{j}\right)} .
\end{aligned}
$$

The remaining terms $G_{v j}$ in ${ }_{0} \mathbb{G}_{v}$ and $F_{d j \pm} \mid \Sigma$ and $G_{w j}$ in ${ }_{0} \mathbb{G}_{w}$ are lower order differential operators in ( $u, h$ ) and therefore the assertions (ii) to (iv) in Lemma 3.19 yield the estimate

$$
\left\|G_{v j}(u, h)\right\|_{o \mathbb{G}_{v}(T)}+\left\|\left(\left.F_{d j \pm}(u)\right|_{\Sigma}, G_{w j}(u, h)\right)\right\|_{o \mathbb{G}_{w}(T)} \leq T^{1 / 4} C\left(\delta, T_{0}\right)\|z\|_{E_{j}(T)} .
$$

Hence, given $\varepsilon>0$, there exists $T_{0}^{\prime}(\varepsilon) \in\left(0, T_{0}\right]$ such that (3.70a) is valid. Estimate (3.70b) follows analogously.
3.27. Lemma (Construction of $R_{0}$ ). Let $\Omega$ be a bounded domain in $\mathbb{R}^{n}(n \geq 2)$ with smooth boundary $\partial \Omega$, let $\Sigma \subset \Omega$ be a compact smooth hypersurface, and let $T_{1}, \rho_{ \pm}$, and $\mu_{ \pm}>0$ be fixed. Then the operator

$$
\begin{equation*}
(u, \pi) \mapsto\left(\left(\rho \partial_{t}-\mu \Delta\right) u+\nabla \pi, \operatorname{div} u\right), \quad{ }_{0} \mathbb{E}_{u, v, w, \partial_{\nu} w}(T) \times{ }_{0} \mathbb{E}_{\pi,[\pi \mathbb{}}(T) \rightarrow \mathbb{F}_{u}(T) \times{ }_{0} \mathbb{F}_{d, \Sigma}(T) \tag{3.76}
\end{equation*}
$$

is a retraction and it has a uniformly bounded co-retraction with respect to $T \in\left(0, T_{1}\right]$.
Proof. The spatial trace theorem, the divergence theorem and the identity $\operatorname{div} u=\operatorname{div}_{\Sigma} v-$ $H_{\Sigma} w-\partial_{\nu} w$ near $\Sigma$ imply that div: ${ }_{0} \mathbb{E}_{u, v, w, \partial_{\nu} w} \rightarrow{ }_{0} \mathbb{F}_{d, \Sigma}$ is bounded. Hence (3.76) is bounded.

From Theorem 2.3 we obtain the Helmholtz decomposition $f_{u}=\nabla F_{u}+f_{u \sigma}$ where $f_{u \sigma}:=$ $f_{u}-\nabla F_{u}$ belongs to $L_{p}\left(0, T ; L_{p, \sigma}(\Omega)\right)$ and $F_{u} \in L_{p}\left(0, T ; \dot{H}_{p}^{1}(\Omega)\right)$ is defined as the solution to the weak Neumann problem $\left\langle\nabla F_{u}, \nabla \phi\right\rangle_{\Omega}=\left\langle f_{u}, \nabla \phi\right\rangle_{\Omega}$ for all $\phi \in \dot{H}_{p^{\prime}}^{1}(\Omega)$.

Next, we define $u^{1}:=\nabla U$, where the functions $U$ solves the transmission problem

$$
\Delta U=f_{d} \text { in } J \times \Omega \backslash \Sigma,\left.\quad \partial_{\nu} U_{+}\right|_{\partial \Omega}=0, \quad \llbracket \rho U \rrbracket=0, \quad \llbracket \partial_{\nu} U \rrbracket=0 .
$$

By Theorem 2.2 and Theorem 2.3, the operator

$$
f_{d} \mapsto u^{1}=\nabla U, \quad{ }_{0} \mathbb{F}_{d, \Sigma} \rightarrow{ }_{0} H_{p}^{(1,2)}(J \times(\Omega \backslash \Sigma))={ }_{0} H_{p}^{1}\left(J ; L_{p}(\Omega)\right) \cap L_{p}\left(J ; H_{p}^{2}(\Omega \backslash \Sigma)\right)
$$

is bounded. Since the traces $\left.u^{1}\right|_{\partial \Omega}$ and $\left.u_{ \pm}^{1}\right|_{\Sigma}$ do not necessarily vanish, we construct another function $u^{2} \in{ }_{0} H_{p}^{(1,2)}(J \times(\Omega \backslash \Sigma))$ by solving the problem

$$
\begin{aligned}
\left(\rho \partial_{t}-\mu \Delta\right) u^{2} & =f_{u \sigma} & & \text { in } J \times \Omega \backslash \Sigma, \\
\operatorname{div} u^{2} & =0 & & \text { in } J \times \Omega \backslash \Sigma, \\
\left.u_{+}^{2}\right|_{\partial \Omega} & =-\left.u^{1}\right|_{\partial \Omega} & & \text { on } J \times \partial \Omega, \\
\left.u_{ \pm}^{2}\right|_{\Sigma} & =-\left.P_{\Sigma} u_{ \pm}^{1}\right|_{\Sigma} & & \text { on } J \times \Sigma .
\end{aligned}
$$

This problem can be decoupled into one-phase Stokes problems in the components of $\Omega \backslash \Sigma$ which can be solved by means of [BP07, Theorem 4.1] and the Helmholtz projection. Hence there is a bounded solution operator $\left(f_{u \sigma}, u^{1}\right) \mapsto u^{2}$, the function $u:=u^{1}+u^{2}$ satisfies $\left.u\right|_{\partial \Omega}=0$, $\left.P_{\Sigma} u\right|_{\Sigma}=0$, and $\operatorname{div} u=f_{d}$, and therefore $u$ belongs to ${ }_{0} \mathbb{E}_{u, v, w, \partial_{\nu} w}$.

The pressure $\pi$ is defined as the solution to the weak transmission problem $\langle\nabla \pi, \nabla \phi\rangle=\left\langle\nabla\left(F_{u}-\left(\rho \partial_{t}-\mu \Delta\right) U\right), \nabla \phi\right\rangle_{\Omega}$ for all $\phi \in \dot{H}_{p^{\prime}}^{1}(\Omega), \quad \llbracket \pi \rrbracket=-\llbracket\left(\rho \partial_{t}-\mu \Delta\right) U \rrbracket=\llbracket \mu \rrbracket f_{d}$, and hence belongs to ${ }_{0} \mathbb{E}_{\pi,[\pi]}$. It is now straightforward to check that the operator $R_{0}:\left(f_{u}, f_{d}\right) \mapsto$ $\left(u^{1}+u^{2}, \pi\right), \mathbb{F}_{u} \times_{0} \mathbb{F}_{d, \Sigma} \rightarrow{ }_{0} \mathbb{E}_{u, v, w, \partial_{\nu} w} \times{ }_{0} \mathbb{E}_{\pi, \llbracket \pi]}$ is a uniformly bounded co-retraction for (3.76).
Proof of Theorem 3.21. The assertions of the theorem follow from the strategy in Section 3.3.1, by applying Lemmas 3.26 and 3.27.

## CHAPTER 4

## The nonlinear problem

Let $\Omega \subset \mathbb{R}^{n}(n \geq 2)$ be a bounded smooth domain. In this chapter we transform problem (N) with compact moving interface $\Gamma(t) \subset \Omega$ to problem (T) over a fixed interface $\Sigma \subset \Omega$ and prove that problem (T) is well-posed on a sufficiently short interval $J=(0, T)$. The notion of well-posedness is based on the function spaces in Figure 4.1 on the next page, and our basically follows the strategy of Köhne, Prüß, and Wilke [KPW13]. However, we restrict our considerations to the case where the initial interface $\Gamma_{0}=\theta_{h_{0}}(\Sigma)$ is already parametrized over $\Sigma$.

In order to transform problem ( N ), we need a time-dependent diffeomorphism $\Theta(t, \cdot)$ of the underlying domain $\Omega$, which maps a fixed hypersurface $\Sigma \subset \Omega$ onto $\Gamma(t)=\Theta(\{t\} \times \Sigma)$. Such maps are studied in Section 4.1, where we construct a normal-preserving admissible map $\Theta_{h}: J \times \bar{\Omega} \rightarrow \bar{\Omega}$ induced by a height function $h(t, \cdot): \Sigma \rightarrow \mathbb{R}$ and extending the parametrization

$$
\begin{equation*}
\theta_{h}(t, x)=x+h(t, x) \nu_{\Sigma}(x) \in \Gamma(t) \quad \text { for } t \in J, x \in \Sigma \tag{4.1}
\end{equation*}
$$

to $J \times \bar{\Omega}$. The map $\Theta_{h}$ yields useful identities for the velocity transformation

$$
\begin{equation*}
u\left(t, \Theta_{h}(t, x)\right)=\left[\partial_{x} \Theta_{h}(t, x)\right] \bar{u}(t, x) . \tag{4.2}
\end{equation*}
$$

These identities are used to derive problem ( T ) in Section 4.2 and Section 4.3.
For proving well-posedness of (T), we will apply the following fixed point theorem.
4.1. Theorem (Banach's fixed point theorem, [DM07]). Let $(M, d)$ be a complete metric space, $A$ be a topological space, and $F: M \times A \rightarrow M$ be a map with the following properties:
(i) There exists $q \in(0,1)$ such that

$$
d(F(x, a), F(y, a)) \leq q d(x, y) \quad \text { for all } x, y \in M \text { and all } a \in A .
$$

(ii) For every $x \in M$, the mapping $a \mapsto F(x, a)$ is continuous on $A$.

Then for every $a \in A$ there is a unique $\varphi(a) \in M$ such that $F(\varphi(a), a)=\varphi(a)$. Moreover, the map $\varphi: A \rightarrow M$ is continuous.

This tool is applied in Section 4.4, where we prove our main result Theorem 4.33 with the following technique. First, in order to eliminate the initial condition $\left.(u, h)\right|_{t=0}=\left(u_{0}, h_{0}\right)=: z_{0}$, we will construct a triple $z_{*}=\left(u_{*}, \pi_{*}, h_{*}\right)$ with $\left.\left(u_{*}, h_{*}\right)\right|_{t=0}=z_{0}$ by means of semigroup theory and Chapter 2. Then the desired solution is given by $z=z_{\bullet}+z_{*}$, where $z_{\bullet}=\left(u_{\bullet}, \pi_{\bullet}, h_{\bullet}\right)$ should satisfy the identity $L\left(z_{\bullet} ; u_{*}\right)=N\left(z_{\bullet} ; z_{*}\right)$, the operator $L\left(\cdot ; u_{*}\right)$ is the solution-to-data map of problem (PL), and $N$ contains the nonlinear perturbations that arise during the transformation. Hence, with Theorem 3.21, we can define the map $F\left(z_{\bullet} ; z_{0}\right):=\left[L\left(\cdot ; u_{*}\right)\right]^{-1} N\left(z_{\bullet} ; z_{*}\right)$. Thus, in view of the desired identity $z_{\bullet}=F\left(z_{0} ; z_{0}\right)$, it remains to show that $F$ satisfies the assumptions of Theorem 4.1. To this end, we will show that $F\left(z_{\boldsymbol{\bullet}} ; z_{0}\right)$ and $\partial_{z_{\mathbf{0}}} F\left(z_{\boldsymbol{\bullet}} ; z_{0}\right)$ become as small as we wish, when we choose $z_{\bullet}, T$, and $h_{0}$ sufficiently small. Since $L\left(\cdot ; u_{*}\right)$ is uniformly invertible, it remains to control the perturbation $N\left(z_{\boldsymbol{\bullet}} ; z_{0}\right)$ and its derivative $\partial_{z_{\boldsymbol{\bullet}}} N\left(z_{\boldsymbol{\bullet}} ; z_{0}\right)$.

We control these perturbations in the context of their derivations. In Section 4.2, we deal with the transformed momentum balance and the transformed divergence equation, where we do not yet employ an explicit representation of $\Theta$. In Section 4.3, we control the perturbations for the transformed interface momentum balance when the moving interface is represented as $\Gamma(t)=\Theta_{h}(\{t\} \times \Sigma)$. Here we also specialize the results from Section 4.2 to the case of a normal-preserving admissible map.

For $n \geq 2, p \in(3, \infty)$, and $J=(0, T)$, we let

$$
\begin{aligned}
\mathbb{E}_{u} & =\left\{u \in H_{p}^{1}\left(J ; L_{p}(\Omega)^{n}\right) \cap L_{p}\left(J ; H_{p}^{2}(\Omega \backslash \Sigma)^{n}\right):\left.u\right|_{\partial \Omega}=0, \llbracket u \rrbracket=0 \text { on } \Sigma\right\}, \\
\mathbb{E}_{v} & =W_{p}^{1-1 / 2 p}\left(J ; L_{p}(\Sigma ; T \Sigma)\right) \cap W_{p}^{1 / 2-1 / 2 p}\left(J ; H_{p}^{2}(\Sigma ; T \Sigma)\right) \cap L_{p}\left(J ; W_{p}^{3-1 / p}(\Sigma ; T \Sigma)\right), \\
\mathbb{E}_{w} & =W_{p}^{1-1 / 2 p}\left(J ; H_{p}^{1}(\Sigma)\right) \cap L_{p}\left(J ; W_{p}^{3-1 / p}(\Sigma)\right), \\
\mathbb{E}_{u, v, w} & =\left\{u \in \mathbb{E}_{u}:\left.v\right|_{\Sigma} \in \mathbb{E}_{v},\left.w\right|_{\Sigma} \in \mathbb{E}_{w}\right\}, \\
\mathbb{E}_{u, v, w, \partial_{\nu} w} & =\left\{u \in \mathbb{E}_{u, v, w}: \partial_{\nu} w_{ \pm} \mid \Sigma \in \mathbb{G}_{w}\right\}, \\
\mathbb{E}_{\pi} & =L_{p}\left(J ; \dot{H}_{p}^{1}(\Omega \backslash \Sigma)\right), \\
\mathbb{E}_{\pi, \llbracket \pi \rrbracket} & =\left\{\pi \in \mathbb{E}_{\pi}: \llbracket \pi \rrbracket \in \mathbb{G}_{w}\right\}, \\
\mathbb{E}_{h} & =W_{p}^{2-1 / 2 p}\left(J ; H_{p}^{1}(\Sigma)\right) \cap \tilde{\mathbb{E}}_{h}, \\
\tilde{\mathbb{E}}_{h} & =H_{p}^{1}\left(J ; W_{p}^{3-1 / p}(\Sigma)\right) \cap L_{p}\left(J ; W_{p}^{4-1 / p}(\Sigma)\right), \\
\mathbb{E}_{\Theta} & =H_{p}^{3 / 2}\left(J ; H_{p}^{2}\left(\mathbb{R}^{n}\right)\right) \cap H_{p}^{1}\left(J ; H_{p}^{3}\left(\mathbb{R}^{n}\right)\right) \cap L_{p}\left(J ; H_{p}^{4}\left(\mathbb{R}^{n}\right)\right), \\
\mathbb{E}_{\partial \Theta} & =W_{p}^{2-1 / 2 p}\left(J ; L_{p}(\Sigma)\right) \cap \tilde{\mathbb{E}}_{\partial \Theta}, \\
\tilde{\mathbb{E}}_{\partial \Theta} & =H_{p}^{1}\left(J ; W_{p}^{2-1 / p}(\Sigma)\right) \cap L_{p}\left(J ; W_{p}^{3-1 / p}(\Sigma)\right), \\
\mathbb{E} & =\mathbb{E}_{u, v, w, \partial_{\nu} w} \times \mathbb{E}_{\pi, \llbracket \pi \rrbracket} \times \mathbb{E}_{h}, \\
\tilde{\mathbb{E}} & =\mathbb{E}_{u, v, w, \partial_{\nu} w} \times \mathbb{E}_{\pi, \llbracket \pi]} \times \tilde{\mathbb{E}}_{h}, \\
\mathbb{F}_{u} & =L_{p}\left(J ; L_{p}(\Omega)^{n}\right), \\
\mathbb{F}_{d} & =H_{p}^{1}\left(J ; \hat{H}_{p}^{-1}(\Omega)\right) \cap L_{p}\left(J ; H_{p}^{1}(\Omega \backslash \Sigma)\right), \\
\mathbb{F}_{d, \Sigma} & =\left\{f_{d} \in \mathbb{F}_{d}: f_{d \pm} \mid \Sigma \in \mathbb{G}_{w}\right\}, \\
\mathbb{G}_{v} & =W_{p}^{1 / 2-1 / 2 p}\left(J ; L_{p}(\Sigma ; T \Sigma)\right) \cap L_{p}\left(J ; W_{p}^{1-1 / p}(\Sigma ; T \Sigma)\right), \\
\mathbb{G}_{w} & =W_{p}^{1 / 2-1 / 2 p}\left(J ; H_{p}^{1}(\Sigma)\right) \cap L_{p}\left(J ; W_{p}^{2-1 / p}(\Sigma)\right), \\
\mathbb{G}_{h} & =W_{p}^{1-1 / 2 p}\left(J ; H_{p}^{1}(\Sigma)\right) \cap L_{p}\left(J ; W_{p}^{3-1 / p}(\Sigma)\right), \\
\tilde{\mathbb{F}} & =\mathbb{F}_{u} \times \mathbb{F}_{d, \Sigma} \times \mathbb{G}_{v} \times \mathbb{G}_{w} \times \mathbb{G}_{h} \times \mathbb{E}_{u, v, w, \partial_{\nu} w \mid t=0} \times\left.\mathbb{E}_{h}\right|_{t=0}, \\
\mathbb{F} & =\left\{\left(f_{u}, f_{d}, g_{v}, g_{w}, u_{0}, h_{0}\right) \in \tilde{\mathbb{F}}: f_{d}\left|t=0=\operatorname{div} u_{0}, L_{v}\left(u_{0}, h_{0} ;\left.u_{*}\right|_{t=0}\right)=g_{v}\right|_{t=0}\right\} .
\end{aligned}
$$

Here we decompose $u=v+w \nu_{\Sigma}$ near $\Sigma$ with $v=P_{\Sigma} u$ and $w=\nu_{\Sigma} \cdot u$. We will also write $\mathbb{E}(T)$ or $\mathbb{E}(J, \Omega, \Sigma)$ instead of $\mathbb{E}$ for indicating the dependence on $T$ or $(J, \Omega, \Sigma)$ (analogously for the other spaces).

Figure 4.1. Function spaces $\mathbb{E} . ., \mathbb{F} . .$. , and $\mathbb{G} \ldots$ on $(J, \Omega, \Sigma)$.

### 4.1. Diffeomorphism and transformation

We study time-dependent diffeomorphisms $\Theta(t, \cdot)$ in a domain $\Omega$ that map a fixed hypersurface $\Sigma \subset \Omega$ onto a moving hypersurface $\Gamma(t)$. In Section 4.1.1 we define and study admissible maps, admissible moving hypersurfaces, and normal-preserving admissible maps and derive useful identities for the velocity transformation $u(t, \Theta(t, x))=\left[\partial_{x} \Theta(t, x)\right] \bar{u}(t, x)$. In Section 4.1.2 we revisit the Hanzawa map $\Theta_{h}$ and prove that it is admissible but not normal-preserving. In Section 4.1.3 we construct a normal-preserving admissible map $\Theta_{h}: J \times \bar{\Omega} \rightarrow \bar{\Omega}$, which depends analytically on its inducing height function $h$.
4.1.1. General admissible maps. First, we consider general admissible maps, admissible moving hypersurfaces, and normal-preserving admissible maps.
4.2. Definition. Let $J \subset \mathbb{R}$ be an interval and $\Omega \subset \mathbb{R}^{n}$ be a domain.
(i) A map

$$
\Theta:(t, x) \mapsto x=\Theta(t, x), \quad J \times \bar{\Omega} \rightarrow \bar{\Omega}
$$

of class $C^{1}(J \times \bar{\Omega})^{n}$ is called an admissible map, if (a) the Jacobian $\partial_{x} \Theta(t, x)$ is invertible for all $t \in J$ and all $x \in \bar{\Omega}$, (b) the map

$$
\tilde{\Theta}:(t, x) \mapsto(t, \Theta(t, x)), \quad J \times \Omega \rightarrow J \times \Omega
$$

is a diffeomorphism, and (c) we have $\Theta(t, x)=x$ for all $t \in J$ and all $x \in \partial \Omega$.
(ii) A map $\Theta: \bar{\Omega} \rightarrow \bar{\Omega}$ is called admissible, if $(t, x) \mapsto \Theta(x), \mathbb{R} \times \bar{\Omega} \rightarrow \bar{\Omega}$ is admissible.
(iii) A moving hypersurface $\Gamma: J \rightarrow 2^{\Omega}, t \mapsto \Gamma(t)$ is called admissible, if there exist a $C^{1}$ hypersurface $\Sigma \subset \Omega$ and an admissible map $\Theta: J \times \bar{\Omega} \rightarrow \bar{\Omega}$ such that

$$
\Gamma(t)=\Theta(\{t\} \times \Sigma) \quad \text { for all } t \in J
$$

We easily obtain the following properties, which are useful for transforming problem ( N ).
4.3. Proposition. Let $J \subset \mathbb{R}$ be a compact interval, $\bar{\Omega} \subset \mathbb{R}^{n}$ be a bounded domain, and $\Theta: J \times \bar{\Omega} \rightarrow \bar{\Omega}$ be admissible. Then also $(t, x) \mapsto \Theta(t, \cdot)^{-1}(x), J \times \bar{\Omega} \rightarrow \bar{\Omega}$ is admissible.
4.4. Proposition. Let $t \mapsto \Gamma(t)=\Theta(\{t\} \times \Sigma)$ be an admissible moving hypersurface.
(i) The tangent vectors of $\Gamma(t)$ are given by

$$
\begin{equation*}
\tau_{j}^{\Gamma}(t, x)=\left[\partial_{\bar{x}} \Theta(t, \bar{x})\right] \tau_{j}^{\Sigma}(\bar{x}) \quad \text { for all } x=\Theta(t, \bar{x}), \bar{x} \in \Sigma \tag{4.3}
\end{equation*}
$$

and a continuous unit normal field on $\Gamma(t)$ is given by

$$
\begin{equation*}
\nu_{\Gamma}(t, x)=\frac{\left[\partial_{\bar{x}} \Theta(t, \bar{x})\right]^{-\top} \nu_{\Sigma}(\bar{x})}{\left|\left[\partial_{\bar{x}} \Theta(t, \bar{x})\right]^{-\top} \nu_{\Sigma}(\bar{x})\right|} \quad \text { for all } x=\Theta(t, \bar{x}), \bar{x} \in \Sigma . \tag{4.4}
\end{equation*}
$$

(ii) The normal velocity of $\Gamma(t)$ is given by

$$
V_{\Gamma}(t, x)=\nu_{\Gamma}(t, x) \cdot \partial_{t} \Theta(t, \bar{x}) \quad \text { for all } x=\Theta(t, \bar{x}), \bar{x} \in \Sigma
$$

Proof. (i) Since $\Gamma(t)$ is oriented and $\operatorname{det} \partial_{x} \Theta$ is either positive or negative in all of $\bar{\Omega}$, the hypersurface $\Sigma$ must be orientable. Let $\nu_{\Sigma}$ denote a unit normal field on $\Sigma$ and let $\varphi: \mathbb{R}^{n-1} \supset$ $U \rightarrow \Sigma$ be a parametrization for $\Sigma$. Since the restriction $\left.\Theta(t, \cdot)\right|_{\Sigma}: \Sigma \rightarrow \Gamma(t)$ is a diffeomorphism, the map $y \mapsto \Theta(t, \varphi(y)): \mathbb{R}^{n-1} \supset U \rightarrow \Gamma(t)$ is a parametrization for $\Gamma(t)$ and the vectors

$$
\tau_{j}^{\Gamma}(t, x):=\partial_{j}(\Theta(t, \cdot) \circ \varphi)(u)=\partial_{\bar{x}} \Theta(t, \bar{x}) \tau_{j}^{\Sigma}(\bar{x}) \quad \text { for } x=\Theta(t, \bar{x}), \bar{x}=\varphi(u) \in \Sigma
$$

form a basis of the tangent space $T_{x} \Gamma(t)$. Hence (4.3) is valid. Since we have $\tau_{j}^{\Gamma} \cdot \nu_{\Gamma}=0$ for all $j$, the normal $\nu_{\Gamma}(t, x)$ must be parallel to $\left[\partial_{\bar{x}} \Theta(t, \bar{x})\right]^{-\top} \nu_{\Sigma}(\bar{x})$. From the identity $\left|\nu_{\Gamma}\right|=1$, it follows that either $\nu_{\Gamma}(t, x)$ or $-\nu_{\Gamma}(t, x)$ satisfy (4.4).
(ii) The second assertion follows from Proposition 1.7, by using the trajectories $\gamma=\Theta(\cdot, \bar{x})$.

Next, we introduce normal-preserving admissible maps and study their geometric properties as well as their kinematic properties associated to the velocity transformation (4.2).
4.5. Definition. Given an admissible map $\Theta: J \times \bar{\Omega} \rightarrow \bar{\Omega}$ and a $C^{1}$-hypersurface $\Sigma \subset \Omega$, we put $\Gamma(t)=\Theta(\{t\} \times \Sigma)$ and we say that $\Theta$ is normal-preserving for $\Sigma$, if the vectors $\partial_{\nu_{\Sigma}} \Theta(t, x)$ and $\nu_{\Gamma(t)}(\Theta(t, x))$ are parallel for every $t \in J$ and every $x \in \Sigma$; that is, there exists $\beta: J \times \Sigma \rightarrow \mathbb{R} \backslash\{0\}$ such that $\partial_{\nu_{\Sigma}} \Theta(t, x)=\beta(t, x) \nu_{\Gamma(t)}(\Theta(t, x))$.
4.6. Proposition. Let $\Theta: J \times \bar{\Omega} \rightarrow \Omega$ be a normal-preserving admissible map for a $C^{1}$-hypersurface $\Sigma \subset \Omega$ and put $\Gamma(\cdot)=\Theta(\{\cdot\} \times \Sigma)$. Then the following identities are valid on $J \times \Sigma$ :

$$
\begin{align*}
\nu_{\Gamma} \circ \tilde{\Theta} & =\frac{\left[\partial_{x} \Theta\right]^{-\top} \nu_{\Sigma}}{\left|\left[\partial_{x} \Theta\right]^{-\top} \nu_{\Sigma}\right|}=\frac{\partial_{\nu_{\Sigma}} \Theta}{\left|\partial_{\nu_{\Sigma}} \Theta\right|}, \quad\left|\partial_{\nu_{\Sigma}} \Theta\right|=\left|\left[\partial_{x} \Theta^{-\top}\right] \nu_{\Sigma}\right|^{-1},  \tag{4.5a}\\
\tau_{j}^{\Gamma} \circ \tilde{\Theta} & =\left[\partial_{x} \Theta\right] \tau_{j}^{\Sigma},  \tag{4.5b}\\
\tau_{\Gamma}^{j} \circ \tilde{\Theta} & =\left[\partial_{x} \Theta\right]^{-\top} \tau_{\Sigma}^{j},  \tag{4.5c}\\
P_{\Gamma} \circ \tilde{\Theta} & =\left[\partial_{x} \Theta\right] P_{\Sigma}\left[\partial_{x} \Theta\right]^{-1}=\left[\partial_{x} \Theta\right]^{-\top} P_{\Sigma}\left[\partial_{x} \Theta\right]^{\top} . \tag{4.5d}
\end{align*}
$$

Let two vector fields $u: J \times \Omega \rightarrow \mathbb{R}^{n}$ and $\bar{u}: J \times \Omega \rightarrow \mathbb{R}^{n}$ be related by $u \circ \tilde{\Theta}=\left[\partial_{x} \Theta\right] \bar{u}$, and decompose $\left.u\right|_{\Gamma}=v+w \nu_{\Gamma}$ with $v=\left.P_{\Gamma} u\right|_{\Gamma}$ and $\left.\bar{u}\right|_{\Sigma}=\bar{v}+\bar{w} \nu_{\Sigma}$ with $\bar{v}=\left.P_{\Sigma} \bar{u}\right|_{\Sigma}$. Then we have

$$
\begin{align*}
v \circ \tilde{\Theta} & =\left[\partial_{x} \Theta\right] \bar{v}  \tag{4.6a}\\
w \circ \tilde{\Theta} & =\left|\partial_{\nu_{\Sigma}} \Theta\right| \bar{w} . \tag{4.6b}
\end{align*}
$$

Proof. Since $\Theta$ is normal-preserving, the general identity (4.4) in Proposition 4.4 yields

$$
\nu_{\Gamma} \circ \tilde{\Theta}=\frac{\partial_{\nu_{\Sigma}} \Theta}{\left(\nu_{\Gamma} \circ \tilde{\Theta} \mid \partial_{\nu_{\Sigma}} \Theta\right)}=\left|\left[\partial_{x} \Theta\right]^{-\top} \nu_{\Sigma}\right| \partial_{\nu_{\Sigma}} \Theta
$$

and therefore (4.5a) is valid. Identity (4.5b) is a repetition of (4.3). From the relations $\tau_{j}^{\Gamma} \cdot \tau_{\Gamma}^{k}=\delta_{j}^{k}$ and $\tau_{j}^{\Gamma} \cdot \nu_{\Gamma}=\tau_{\Gamma}^{j} \cdot \nu_{\Gamma}=0$ we obtain (4.5c), and then (4.5d) is readily checked.

The remaining identities can be verified as follows.

$$
\begin{aligned}
v \circ \tilde{\Theta} & =\left[P_{\Gamma} \circ \tilde{\Theta}\right]\left[\partial_{x} \Theta\right] \bar{u}=\left[\partial_{x} \Theta\right] P_{\Sigma} \bar{u}=\left[\partial_{x} \Theta\right] \bar{v} \\
w \circ \tilde{\Theta} & =\left(u \circ \tilde{\Theta} \mid \nu_{\Gamma} \circ \tilde{\Theta}\right)=\left(\bar{u} \mid\left[\partial_{x} \Theta\right]^{\top}\left(\left|\partial_{\nu_{\Sigma}} \Theta\right|\left[\partial_{x} \Theta\right]^{-\top} \nu_{\Sigma}\right)\right)=\left|\partial_{\nu_{\Sigma}} \Theta\right| \bar{w}
\end{aligned}
$$

4.1.2. The Hanzawa map. Next, we revisit the Hanzawa map $\Theta_{h}$ and prove that it is admissible but not normal-preserving. In order to construct it, we recall that

$$
x \mapsto \theta_{h}(x)=x+h(x) \nu_{\Sigma}(x), \quad \Sigma \rightarrow \Gamma_{h}
$$

is a parametrization for $\Gamma_{h}=\theta_{h}(\Sigma)$ over $\Sigma$. If $\Sigma$ is of class $C^{2}$, then for $|h(x)|<\left|L_{\Sigma}(x)\right|$, the matrix

$$
M_{h}(x):=\left[I_{x}-h(x) L_{\Sigma}(x)\right]^{-1}
$$

from page 138 is invertible, maps $T_{x} \Sigma$ onto itself, and satisfies $M_{h} \nu_{\Sigma}=\nu_{\Sigma}$. Moreover,

$$
\begin{equation*}
\nu_{\Gamma_{h}} \circ \theta_{h}=\beta_{h}\left(\nu_{\Sigma}-M_{h} \nabla_{\Sigma} h\right), \quad \text { with } \beta_{h}:=\left(\nu_{\Sigma} \mid \nu_{\Gamma_{h}} \circ \theta_{h}\right)=\left(1+\left|M_{h} \nabla_{\Sigma} h\right|^{2}\right)^{-1 / 2} \tag{4.7}
\end{equation*}
$$

A hypersurface $\Sigma \subset \mathbb{R}^{n}$ is said to have a tubular neighborhood of radius $r>0$, if the map

$$
X:(p, t) \mapsto p+t \nu_{\Sigma}(p), \quad \Sigma \times(-r, r) \rightarrow B_{r}(\Sigma):=\left\{x \in \mathbb{R}^{n}: \operatorname{dist}(x, \Sigma)<r\right\}
$$

is a homeomorphism; that is, $X$ is bijective and continuous and has a continuous inverse (see Definition A.16). The inverse of $X$ is denoted by

$$
X^{-1}(x)=(\Pi(x), d(x))=(p, t) \quad \text { for } x=p+t \nu_{\Sigma}(p) \in B_{r}(\Sigma)
$$

Proposition A. 17 implies that every compact $C^{2}$-hypersurface has a tubular neighborhood.
4.7. Definition (cf. [Han81, p. 309]). Let $\Omega \subset \mathbb{R}^{n}(n \geq 2)$ be a domain and let $\Sigma \subset \Omega$ be a closed $C^{2}$-hypersurface with tubular neighborhood $B_{r}(\Sigma) \subset \Omega$ of radius $r>0$. Choose a function $\chi \in C^{\infty}(\mathbb{R} ;[0,1])$ such that $\chi(s)=1$ if $|s| \leq r / 3$ and $\chi(s)=0$ if $|s| \geq 2 r / 3$ and $\left\|\chi^{\prime}\right\|_{\infty}<6 / r$.

Then, for a given height function $h: \Sigma \rightarrow \mathbb{R}$, we define the stationary Hanzawa map

$$
\Theta_{h}(x):= \begin{cases}x+\chi(d(x)) h(\Pi(x)) \nu_{\Sigma}(\Pi(x)) & \text { for } x \in B_{r}(\Sigma) \\ x & \text { for } x \in \bar{\Omega} \backslash B_{r}(\Sigma)\end{cases}
$$

For $J \subset \mathbb{R}$ and a height function $h: J \times \Sigma \rightarrow \mathbb{R}$, we define the time-dependent Hanzawa map

$$
\Theta_{h}(t, x):=\Theta_{h(t, \cdot)}(x) \quad \text { for }(t, x) \in J \times \bar{\Omega}
$$

4.8. Theorem. Let $\Omega \subset \mathbb{R}^{n}(n \geq 2)$ be a domain, $k \geq 1$, and let $\Sigma \subset \Omega$ be a compact $C^{k+1}$-hypersurface with tubular neighborhood $B_{r}(\Sigma) \subset \Omega$ of radius $r<\left\|L_{\Sigma}\right\|_{\infty}^{-1}$. Then, for a given height function

$$
h \in C^{k}(\Sigma) \quad \text { with }\|h\|_{\infty}<1 /\left\|\chi^{\prime}\right\|_{\infty}
$$

the stationary Hanzawa map $\Theta_{h}: \Omega \rightarrow \Omega$ is a $C^{k}$-diffeomorphism, and in $B_{r}(\Sigma)$ we have

$$
\begin{align*}
\partial_{x} \Theta_{h}= & P_{\Sigma} \circ \Pi-\chi \circ d\left[h L_{\Sigma}\right] \circ \Pi\left[P_{\Sigma} \circ \Pi-d L_{\Sigma} \circ \Pi\right]^{-1} & & \text { (purely tangential part) } \\
& +\left(1+\chi^{\prime} \circ d h \circ \Pi\right)\left[\nu_{\Sigma} \otimes \nu_{\Sigma}\right] \circ \Pi & & \text { (purely normal part) }  \tag{4.8}\\
& +\chi \circ d\left[\nu_{\Sigma} \otimes \nabla_{\Sigma} h\right] \circ \Pi\left[P_{\Sigma} \circ \Pi-d L_{\Sigma} \circ \Pi\right]^{-1} & & \text { (tangential-to-normal part) }
\end{align*}
$$

and

$$
\begin{align*}
{\left[\partial_{x} \Theta_{h}\right]^{-1}=} & {\left[P_{\Sigma} \circ \Pi-(d+\chi \circ d h \circ \Pi) L_{\Sigma} \circ \Pi\right]^{-1}\left[P_{\Sigma} \circ \Pi-d L_{\Sigma} \circ \Pi\right] } \\
& +\left(1+\chi^{\prime} \circ d h \circ \Pi\right)^{-1}\left[\nu_{\Sigma} \otimes \nu_{\Sigma}\right] \circ \Pi  \tag{4.9}\\
& -\chi \circ d\left(1+\chi^{\prime} \circ d h \circ \Pi\right)^{-1}\left[\nu_{\Sigma} \otimes \nabla_{\Sigma} h\right] \circ \Pi\left[P_{\Sigma} \circ \Pi-(d+\chi \circ d h \circ \Pi) L_{\Sigma} \circ \Pi\right]^{-1}
\end{align*}
$$

In particular, the following identities are valid on $\Sigma$ :

$$
\begin{align*}
\left.\partial_{x} \Theta_{h}\right|_{\Sigma} & =P_{\Sigma}-h L_{\Sigma}+\nu_{\Sigma} \otimes \nu_{\Sigma}+\nu_{\Sigma} \otimes \nabla_{\Sigma} h  \tag{4.10a}\\
{\left[\partial_{x} \Theta_{h} \mid \Sigma\right]^{-1} } & =\left[P_{\Sigma}-h L_{\Sigma}\right]^{-1} P_{\Sigma}+\nu_{\Sigma} \otimes \nu_{\Sigma}-\nu_{\Sigma} \otimes \nabla_{\Sigma} h\left[P_{\Sigma}-h L_{\Sigma}\right]^{-1}
\end{align*}
$$

Proof. Local invertibility. We check that the inverse $\left[\partial_{x} \Theta_{h}\right]^{-1}$ exists everywhere in $\Omega$. Clearly, it suffices to consider the case $x \in B_{r}(\Sigma)$. Proposition A. 20 and a straightforward calculation show that (4.8) is valid in $B_{r}(\Sigma)$. The purely tangential part of $\partial_{x} \Theta_{h}$ can be written as

$$
\left[P_{\Sigma} \circ \Pi-(d+\chi \circ d h \circ \Pi) L_{\Sigma} \circ \Pi\right]\left[I-d L_{\Sigma} \circ \Pi\right]^{-1}
$$

The conditions $\|h\|_{\infty}<r / 3$ and $\chi(s)=0$ for $|s| \geq 2 r / 3$ yield $|d+\chi \circ d h \circ \Pi|_{\infty} \leq r<\left\|L_{\Sigma}\right\|_{\infty}^{-1}$ in $B_{r}(\Sigma)$. Hence the purely tangential part is a linear isomorphism of $T_{\Pi(x)} \Sigma$. The purely normal part is a linear isomorphism of $\mathbb{R} \nu_{\Sigma}(x)$, since $\|h\|_{\infty}<\left\|\chi^{\prime}\right\|_{\infty}^{-1}$. Therefore $\partial_{x} \Theta_{h}(x)$ is an isomorphism of $\mathbb{R}^{n}$. For every invertible $A \in \mathbb{C}^{n \times n}$ and $a, b \in \mathbb{C}^{n}$ we have

$$
\begin{equation*}
[A+b \otimes a]^{-1}=A^{-1}-\frac{A^{-1} b \otimes A^{-\top} a}{1+A^{-1} b \cdot a} \tag{4.11}
\end{equation*}
$$

Then also (4.9) follows by straightforward calculations. Hence $\left[\partial_{x} \Theta_{h}(\cdot)\right]^{-1}$ is bounded in $\Omega$ and, by the implicit function theorem, $\Theta_{h}: \Omega \rightarrow \Omega$ is a local $C^{k}$-diffeomorphism.

Surjectivity. Since the map $\Theta_{h}: \Omega \rightarrow \Omega$ is a local homeomorphism, the set $\Theta_{h}(\Omega)$ is an open subset of $\Omega$. We now show that it is closed as a subset of $\Omega$. Let $\left(y_{n}\right)_{n} \subset \Theta_{h}(\Omega)$ converge to $y \in \Omega$. Since $\bar{B}_{r}(\Sigma)$ is compact and $\Theta_{h}(x)=x$ in $\Omega \backslash B_{r}(\Sigma)$, the preimages $x_{n}=\Theta_{h}^{-1}\left(y_{n}\right)$ have a convergent subsequence $x_{n_{k}} \rightarrow x \in \Omega$. Therefore $y=\lim _{k} y_{n_{k}}=\lim _{k} \Theta_{h}\left(x_{n_{k}}\right)=\Theta_{h}(x)$ also belongs to $\Theta_{h}(\Omega)$. Consequently, $\Theta_{h}(\Omega)$ is open, closed, and nonempty in $\Omega$, thus $\Theta_{h}(\Omega)=\Omega$, which implies that $\Theta_{h}: \Omega \rightarrow \Omega$ is surjective.

Injectivity. It suffices to show that the restriction of $\Theta_{h}$ to $B_{r}(\Sigma)$ is injective. If $\Theta_{h}(x)=$ $\Theta_{h}(y)$, then the tubular neighborhood property of $\Sigma$ implies

$$
\Pi(x)+(d(x)+\chi(d(x)) h(\Pi(x))) \nu_{\Sigma}(\Pi(x))=\Pi(y)+(d(y)+\chi(d(y)) h(\Pi(y))) \nu_{\Sigma}(\Pi(y))
$$

and hence $\Pi(x)=\Pi(y)$ and $d(x)+\chi(d(x)) h(\Pi(x))=d(y)+\chi(d(y)) h(\Pi(y))$. Since $s \mapsto s+$ $\chi(s) h(\Pi(x))$ is injective by $\left|h \chi^{\prime}\right|<1$, we obtain $d(x)=d(y)$, and thus $x=y$.

We conclude that the stationary Hanzawa map $\Theta_{h}$ is a $C^{k}$-diffeomorphism of $\Omega$.
4.9. Corollary. Let $\Sigma \subset \Omega$ be a compact $C^{2}$-hypersurface with tubular neighborhood $B_{r}(\Sigma) \subset \Omega$ of radius $r<\left\|L_{\Sigma}\right\|_{\infty}^{-1}$, and let $\Theta_{h}$ denote the time-dependent Hanzawa map induced by

$$
\begin{equation*}
h \in C^{1}([0, T] \times \Sigma) \quad \text { with }\|h\|_{\infty}<\left\|\chi^{\prime}\right\|_{\infty}^{-1} . \tag{4.12}
\end{equation*}
$$

Then the map $\Theta_{h}:[0, T] \times \bar{\Omega} \rightarrow \bar{\Omega}$ and the moving hypersurface $t \mapsto \Gamma_{h}(t)=\Theta_{h}(\{t\} \times \Sigma)$ are admissible, but the map $\Theta_{h}$ is not normal-preserving unless $\nabla_{\Sigma} h=0$.

Proof. Admissibility follows from Theorem 4.8 and identity (4.10) yields

$$
\partial_{\nu_{\Sigma}} \Theta_{h}=\nu_{\Sigma}, \quad \nu_{\Gamma} \circ \tilde{\Theta}=\frac{\nu_{\Sigma}-\left[I_{x}-h L_{\Sigma}\right]^{-1} \nabla_{\Sigma} h}{\left(1+\left|\left[I_{x}-h L_{\Sigma}\right]^{-1} \nabla_{\Sigma} h\right|^{2}\right)^{1 / 2}}
$$

Thus, $\partial_{\nu_{\Sigma}} \Theta$ and $\nu_{\Gamma} \circ \tilde{\Theta}_{h}$ are not parallel for $\nabla_{\Sigma} h \neq 0$.
4.1.3. A normal-preserving admissible map. We will construct a normal-preserving admissible map $\Theta_{h}: J \times \bar{\Omega} \rightarrow \bar{\Omega}$, which, considered as an element of some Banach space $\mathbb{E}_{\Theta}$, depends analytically on its inducing height function $h \in \mathbb{E}_{h}$. We first construct a diffeomorphism $\Theta_{h}: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ which maps a compact smooth hypersurface $\Sigma$ onto $\Gamma:=\Theta_{h}(\Sigma)$ such that $\partial_{\nu_{\Sigma}} \Theta_{h}(x)$ is parallel to $\nu_{\Gamma}\left(\Theta_{h}(x)\right)$. For this construction, we employ a co-retraction $\mathfrak{S}$ for the trace operator

$$
u \mapsto\left(\left.u\right|_{\Sigma},\left.\partial_{\nu} u\right|_{\Sigma}, \ldots,\left.\partial_{\nu}^{k} u\right|_{\Sigma}\right), \quad W_{p}^{s}\left(\mathbb{R}^{n}\right) \rightarrow \prod_{j=0}^{k} W_{p}^{s-j-1 / p}(\Sigma)
$$

4.10. Lemma. Let $\Omega \subset \mathbb{R}^{n}(n \geq 2)$ be a domain, $\Sigma \subset \Omega$ be a compact smooth hypersurface, and let $p \in(1, \infty), k \in \mathbb{N}_{0}$, and $s \in(k+1 / p, \infty)$. Then there exists a bounded linear operator

$$
\mathfrak{S}: \prod_{j=0}^{k} W_{p}^{s-j-1 / p}(\Sigma) \rightarrow W_{p}^{s}\left(\mathbb{R}^{n}\right)
$$

with the properties

$$
\left.\mathfrak{S}\left(f_{0}, \ldots, f_{k}\right)\right|_{\mathbb{R}^{n} \backslash \Omega}=0,\left.\quad \partial_{\nu_{\Sigma}}^{j}\left(\mathfrak{S}\left(f_{0}, \ldots, f_{k}\right)\right)\right|_{\Sigma}=f_{j} \quad \text { for } j \in\{0,1, \ldots, k\}
$$

for all

$$
\left(f_{0}, \ldots, f_{k}\right) \in \prod_{j=0}^{k} W_{p}^{s-j-1 / p}(\Sigma)
$$

The operator $\mathfrak{S}$ only depends on $\Sigma$ and $k$ but not on $s$ or $p$.
Proof. With Corollary A. 19 we decompose $\Omega \backslash \Sigma=\Omega_{+} \dot{U} \Omega_{-}$such that $\Sigma=\partial \Omega_{-}$and $\nu_{\Sigma}=$ $\nu_{\partial \Omega_{-}}=-\nu_{\partial \Omega_{+} .}$Let $\Omega^{\prime} \subset \Omega$ be a bounded smooth domain which still contains $\Sigma$ and let $\Omega_{ \pm}^{\prime}:=$ $\Omega^{\prime} \cap \Omega_{ \pm}$. Triebel [Tri10, p.3.3.3] has shown that there exist bounded linear operators

$$
S_{ \pm}: \prod_{j=0}^{k} W_{p}^{s-j-1 / p}\left(\partial \Omega_{ \pm}^{\prime}\right) \rightarrow W_{p}^{s}\left(\Omega^{\prime}\right)
$$

with the property

$$
\partial_{j} S_{ \pm}\left(g_{0}, \ldots, g_{k}\right)=g_{j} \text { on } \partial \Omega_{ \pm}^{\prime} \quad \text { for all } j \in\{0, \ldots, k\},\left(g_{0}, \ldots, g_{k}\right) \in \prod_{j=0}^{k} W_{p}^{s-j-1 / p}\left(\partial \Omega_{ \pm}^{\prime}\right)
$$

For a tuple $\left(f_{0}, \ldots, f_{k}\right) \in \prod_{j=0}^{k} W_{p}^{s-j-1 / p}(\Sigma)$ we define a linear operator $\mathfrak{S}$ by

$$
\mathfrak{S}\left(f_{0}, \ldots, f_{k}\right):= \begin{cases}0 & \text { in } \mathbb{R}^{n} \backslash \Omega^{\prime} \\ S_{+}\left(\chi_{\Sigma} \cdot\left((-1)^{j} f_{j}\right)_{j=0}^{k}+\chi_{\partial \Omega^{\prime}} \cdot(0, \ldots, 0)\right) & \text { in } \Omega_{+}^{\prime} \\ S_{-}\left(f_{0}, \ldots, f_{k}\right) & \text { in } \Omega_{-}^{\prime}\end{cases}
$$

This operator has the asserted properties.
We are ready to construct a normal-preserving map.
4.11. Definition. Let $\Omega \subset \mathbb{R}^{n}(n \geq 2)$ be a domain and $\Sigma \subset \Omega$ be a compact smooth hypersurface.
(i) Let $p \in(1, \infty)$ and $s \in(1+n / p)$, and let $h \in W_{p}^{s-1 / p}(\Sigma)$ be a height function satisfying

$$
\begin{equation*}
\|h\|_{\infty}<\left\|L_{\Sigma}\right\|_{\infty}^{-1} \tag{4.13}
\end{equation*}
$$

Then we define the stationary normal-preserving map

$$
\begin{equation*}
\Theta_{h}(x):=x+\mathfrak{S}\left(h \nu_{\Sigma}, g_{h}\right)(x) \quad \text { for } x \in \mathbb{R}^{n} \tag{4.14}
\end{equation*}
$$

where $\mathfrak{S}$ denotes the linear operator from Lemma 4.10 with $k=1$, and

$$
\begin{equation*}
g_{h}:=\left[\left(\nu_{\Gamma_{h}} \otimes \nu_{\Gamma_{h}}\right) \circ \theta_{h}-I\right] \nu_{\Sigma}=\beta_{h} \nu_{\Gamma_{h}} \circ \theta_{h}-\nu_{\Sigma} \tag{4.15}
\end{equation*}
$$

(ii) Let $J=(0, T)$ with $T \in(0, \infty)$, and let $h: J \times \Sigma \rightarrow \mathbb{R}$ be a height function such that $h(t, \cdot)$ satisfies the assumptions of (i) for almost all $t \in J$. Then we define the time-dependent normal-preserving map

$$
\begin{equation*}
\Theta_{h}(t, x):=\Theta_{h(t, \cdot)}(x)=x+\mathfrak{S}\left(h(t, \cdot) \nu_{\Sigma}, g_{h(t, \cdot)}\right)(x) \quad \text { for } t \in J, x \in \mathbb{R}^{n} \tag{4.16}
\end{equation*}
$$

4.12. Proposition. Let $\Theta_{h}$ denote the stationary normal-preserving map.
(i) $\Theta_{h}$ maps $\Sigma$ onto $\Gamma_{h}:=\theta_{h}(\Sigma)$.
(ii) $\partial_{\nu_{\Sigma}} \Theta_{h}=\left[\partial_{x} \Theta_{h}\right] \nu_{\Sigma}=\left[\left(\nu_{\Gamma_{h}} \otimes \nu_{\Gamma_{h}}\right) \circ \theta_{h}\right] \nu_{\Sigma}=\beta_{h} \nu_{\Gamma_{h}} \circ \theta_{h}$ on $\Sigma$.
(iii) $\Theta_{h}=\operatorname{Id}_{x}$ in $\mathbb{R}^{n} \backslash \Omega$.
(iv) For $M_{h}:=\left(I-h L_{\Sigma}\right)^{-1}$ as on page 138 , we have

$$
\begin{align*}
\left.\partial_{x} \Theta_{h}\right|_{\Sigma} & =P_{\Sigma}-h L_{\Sigma}-\beta_{h}^{2} M_{h} \nabla_{\Sigma} h \otimes \nu_{\Sigma}+\nu_{\Sigma} \otimes \nabla_{\Sigma} h+\beta_{h}^{2} \nu_{\Sigma} \otimes \nu_{\Sigma}  \tag{4.17}\\
{\left[\left.\partial_{x} \Theta_{h}\right|_{\Sigma}\right]^{-1} } & =M_{h}-\beta_{h}^{2} M_{h}^{2} \nabla_{\Sigma} h \otimes M_{h} \nabla_{\Sigma} h+\beta_{h}^{2} M_{h}^{2} \nabla_{\Sigma} h \otimes \nu_{\Sigma}-\nu_{\Sigma} \otimes M_{h} \nabla_{\Sigma} h \tag{4.18}
\end{align*}
$$

Proof. All assertions are obvious, except for (4.18), which can be derived from (4.11).
It remains to prove that the stationary normal-preserving map is a diffeomorphism of $\mathbb{R}^{n}$ and that its time-dependent version is an admissible map. Compared to the Hanzawa map whose Jacobian has an explicit inverse (4.9) in all of $\Omega$, the normal-preserving map $\Theta_{h}$ as in (4.14) and (4.16) lacks such a representation. We therefore want to show that

$$
\sup _{t, x}\left|I_{x}-\left[\partial_{x} \Theta_{h}(t, x)\right]^{-1}\right|<1
$$

and this can be shown for height functions which are sufficiently small in an appropriate norm. As in Chapter 3 we consider height functions in the class

$$
\mathbb{E}_{h}=W_{p}^{2-1 / 2 p}\left(J ; H_{p}^{1}(\Sigma)\right) \cap H_{p}^{1}\left(J ; W_{p}^{3-1 / p}(\Sigma)\right) \cap L_{p}\left(J ; W_{p}^{4-1 / p}(\Sigma)\right)
$$

Then the Jacobian $\left.\partial_{x} \Theta_{h}\right|_{\Sigma}$ belongs to the space

$$
\mathbb{E}_{\partial \Theta}:=W_{p}^{2-1 / 2 p}\left(J ; L_{p}(\Sigma)\right) \cap H_{p}^{1}\left(J ; W_{p}^{2-1 / p}(\Sigma)\right) \cap L_{p}\left(J ; W_{p}^{3-1 / p}(\Sigma)\right)
$$

The space $\mathbb{E}_{\partial \Theta}$ is considered as the target space of the nonlinear map $\left.h \mapsto \partial_{x} \Theta_{h}\right|_{\Sigma}$, and for proving analyticity of the latter, we employ the following properties of $\mathbb{E}_{\partial \Theta}$.
4.13. Lemma. Let $\Sigma \subset \mathbb{R}^{n}(n \geq 2)$ be a compact smooth hypersurface, $J=(0, T)$ be bounded, $p \in((n+2) / 2, \infty)$, and $m \in \mathbb{N}$. Then the following assertions are valid:
(i) The space $\mathbb{E}_{\partial \Theta}$ is continuously embedded into $C^{1}(\bar{J} \times \Sigma)$.
(ii) $\mathbb{E}_{\partial \Theta}$ is a multiplication algebra and there exists $C(T) \geq 1$ such that

$$
\begin{equation*}
\|f g\|_{\mathbb{E}_{\partial \Theta}} \leq C(T)\left(\|f\|_{\mathbb{E}_{\partial \Theta}}\|g\|_{C^{1}(\bar{J} \times \Sigma)}+\|f\|_{C^{1}(\bar{J} \times \Sigma)}\|g\|_{\mathbb{E}_{\partial \Theta}}\right) \quad \text { for all } f, g \in \mathbb{E}_{\partial \Theta} \tag{4.19}
\end{equation*}
$$

(iii) The operator $A \mapsto A^{-1},\left\{A \in \mathbb{E}_{\partial \Theta}^{m \times m}: \sup _{J \times \Sigma}\left|A^{-1}(\cdot)\right|<\infty\right\} \rightarrow \mathbb{E}_{\partial \Theta}$ is analytic.
(iv) The operator $f \mapsto f^{1 / 2},\left\{f \in \mathbb{E}_{\partial \Theta}: \inf _{J \times \Sigma} \operatorname{dist}\left(f(\cdot), \mathbb{R}_{-}\right)>0\right\} \rightarrow \mathbb{E}_{\partial \Theta}$ is analytic.

Proof. (i) We abbreviate $W_{p}^{t}\left(W_{p}^{s}\right):=W_{p}^{t}\left(J ; W_{p}^{s}(\Sigma)\right), C^{k}\left(C^{l}\right):=C^{k}\left(\bar{J} ; C^{l}(\Sigma)\right)$, and similarly for the other spaces. The mixed derivative embeddings and Sobolev's embedding (B.1) imply

$$
W_{p}^{2-1 / 2 p}\left(L_{p}\right) \cap H_{p}^{1}\left(W_{p}^{2-1 / p}\right) \hookrightarrow W_{p}^{1+\theta}\left(H_{p}^{2-1 / p-2 \theta)}\right) \hookrightarrow W_{p}^{1+1 / p+\varepsilon_{t}}\left(W_{p}^{(n-1) / p+\varepsilon_{s}}\right) \hookrightarrow C^{1}(C),
$$

for sufficiently small $\varepsilon_{t}, \varepsilon_{s}>0$, provided that $\theta \in(0,1-1 / 2 p)$ satisfies $1+\theta>1+1 / p$ and $2-1 / p-2 \theta>(n-1) / p$. Such a number $\theta$ exists if $1 / p<1-n / 2 p$, and this is true for $p>(n+2) / 2$. Analogously, we obtain

$$
H_{p}^{1}\left(W_{p}^{2-1 / p}\right) \cap L_{p}\left(W_{p}^{3-1 / p}\right) \hookrightarrow W_{p}^{\theta}\left(H_{p}^{3-1 / p-\theta}\right) \hookrightarrow C\left(C^{1}\right),
$$

provided that $\theta \in(0,1)$ satisfies $\theta>1 / p$ and $3-1 / p-\theta>1+(n-1) / p$. Such a number $\theta$ exists if $1 / p<2-n / p$, and this is true if $p>(n+1) / 2$. Hence we have $\mathbb{E}_{\partial \Theta} \hookrightarrow C^{1}(\bar{J} \times \Sigma)$.
(ii) The norm of $\mathbb{E}_{\partial \Theta}$ consists of the semi-norms

$$
\llbracket \partial_{t} \cdot \rrbracket_{1-1 / 2 p, p ; p}, \quad \llbracket\left(\partial_{t} \partial_{x}, \partial_{x}^{2}, \partial_{x}\right) \cdot \rrbracket_{p ; 1-1 / p, p}, \quad\left\|\left(1, \partial_{t}, \partial_{t} \partial_{x}, \partial_{x}, \partial_{x}^{2}\right) \cdot\right\|_{p},
$$

where we recall the following abbreviations from page 78:

$$
\llbracket \cdot \rrbracket_{t, p ; p}:=\llbracket \cdot \rrbracket_{W_{p}^{t}\left(L_{p}\right)}, \quad \llbracket \cdot \rrbracket_{p ; s, p}:=\llbracket \cdot \rrbracket_{L_{p}\left(W_{p}^{s}\right)}, \quad\|\cdot\|_{p}:=\|\cdot\|_{L_{p}\left(L_{p}\right)} .
$$

With Lemma B. 81 and Lemma B. 10 we control some of the leading-order terms of $\|f g\|_{\mathbb{E}_{\partial \Theta}}$ by

$$
\begin{aligned}
\llbracket \partial_{t} f g \rrbracket_{1-1 / 2 p, p ; p} & \lesssim \llbracket \partial_{t} f \rrbracket_{1-1 / 2 p, p ; p}\|g\|_{\infty}+\left\|\partial_{t} f\right\|_{\infty} \llbracket g \rrbracket_{1-1 / 2 p, p ; p}, \\
\llbracket \partial_{t} \partial_{x} f g \rrbracket_{p ; 1-1 / p, p} & \lesssim \llbracket \partial_{t} \partial_{x} f \rrbracket_{p ; 1-1 / p, p}\|g\|_{\infty}+\left\|\partial_{t} \partial_{x} f\right\|_{p}\left\|\left(g, \partial_{x} g\right)\right\|_{\infty}, \\
\llbracket \partial_{x}^{2} f g \rrbracket_{p ; 1-1 / p, p} & \lesssim \llbracket \partial_{x}^{2} f \rrbracket_{p ; 1-1 / p, p}\|g\|_{\infty}+\left\|\partial_{x}^{2} f\right\|_{p}\left\|\left(g, \partial_{x} g\right)\right\|_{\infty} .
\end{aligned}
$$

These terms and the remaining ones can be estimated by the right-hand side of (4.19). Therefore the pointwise multiplication estimate (4.19) is valid and $\mathbb{E}_{\partial \Theta}$ is a multiplication algebra.
(iii) Let us check that $A^{-1}$ belongs to $\mathbb{E}_{\partial \Theta}^{m \times m}$ for every $A \in \mathbb{E}_{\partial \Theta}^{m \times m}$ with $A^{-1} \in L_{\infty}(J \times \Sigma)$. Abbreviating $\tau=1-1 / 2 p$ and using Lemma B.81, we obtain

$$
\llbracket \partial_{t} A^{-1} \rrbracket_{\tau, p ; p}=\llbracket A^{-1}\left[\partial_{t} A\right] A^{-1} \rrbracket_{\tau, p ; p} \lesssim \llbracket \partial_{t} A \rrbracket_{\tau, p ; p}\left\|A^{-1}\right\|_{\infty}^{2}+\llbracket A^{-1} \rrbracket_{\tau, p ; p}\left\|A^{-1}\right\|_{\infty}\left\|\partial_{t} A\right\|_{\infty}
$$

Next, from the inequality

$$
\left|A(t, x)^{-1}-A\left(t^{\prime}, x\right)^{-1}\right| \leq\left\|A^{-1}\right\|_{\infty}^{2}\left|A(t, x)-A\left(t^{\prime}, x\right)\right|,
$$

we infer that $\llbracket A^{-1} \rrbracket_{\tau, p ; p} \lesssim\left\|A^{-1}\right\|_{\infty}^{2} \llbracket A \rrbracket_{\tau, p ; p}$ is finite and therefore $\llbracket \partial_{t} A^{-1} \rrbracket_{\tau, p ; p}$ is finite. Analogously,

$$
\llbracket A^{-1} \rrbracket_{p ; \sigma, p} \lesssim\left\|A^{-1}\right\|_{\infty}^{2} \llbracket A \rrbracket_{p ; \sigma, p}<\infty,
$$

with $\sigma=1-1 / p$. Hence, for $j \in\{1, \ldots, n-1\}$, we obtain

$$
\begin{aligned}
& \llbracket \partial_{t} \partial_{j} A^{-1} \rrbracket_{p ; \sigma, p}=\llbracket A^{-1}\left[\partial_{j} A\right] A^{-1}\left[\partial_{t} A\right] A^{-1}+A^{-1}\left[\partial_{t} A\right] A^{-1}\left[\partial_{j} A\right] A^{-1}-A^{-1}\left[\partial_{t} \partial_{j} A\right] A^{-1} \rrbracket_{p ; \sigma, p} \\
& \lesssim\left\|A^{-1}\right\|_{\infty}^{4}\|A\|_{p ; \sigma, p}\left\|\partial_{j} A\right\|_{\infty}\left\|\partial_{t} A\right\|_{\infty}+\left\|A^{-1}\right\|_{\infty}^{3} \llbracket \partial_{j} A \rrbracket_{p ; \sigma, p}\left\|\partial_{t} A\right\|_{\infty}+\left\|A^{-1}\right\|_{\infty}^{3} \llbracket \partial_{t} A \rrbracket_{p ; \sigma, p}\left\|\partial_{j} A\right\|_{\infty} \\
& \quad+\left\|A^{-1}\right\|_{\infty}^{2} \llbracket \partial_{t} \partial_{j} A \rrbracket_{p ; \sigma, p}+\left\|\left(A^{-1}, \nabla_{\Sigma} A^{-1}\right)\right\|_{\infty}\left\|\partial_{t} \partial_{j} A\right\|_{p}\left\|A^{-1}\right\|_{\infty}<\infty .
\end{aligned}
$$

The semi-norm $\llbracket \partial_{j} \partial_{k} A^{-1} \rrbracket_{p ; \sigma, p}$ can be estimated analogously. We further have

$$
\left\|A^{-1}\right\|_{p} \leq T^{1 / p}|\Sigma|^{1 / p}\left\|A^{-1}\right\|_{\infty}<\infty
$$

and the remaining terms in $\left\|A^{-1}\right\|_{\mathbb{E}_{\partial \Theta}}$ are also finite. Therefore $A^{-1}$ belongs to $\mathbb{E}_{\partial \Theta}^{m \times m}$, and Proposition B. 88 yields analyticity of the inversion operator $A \mapsto A^{-1}$.
(iv) Every bounded function $f$ with $\inf \operatorname{dist}\left(f(\cdot), \mathbb{R}_{-}\right)>0$ satisfies both $\inf |f|>0$ and $\sup |\arg f|<\pi$. Hence $f^{1 / 2}$ and $f^{-1}$ are bounded. The Cauchy-Schwarz inequality yields

$$
\left\|f^{1 / 2}\right\|_{p}=\left(\int_{0}^{T} \int_{\Sigma} 1 \cdot|f|^{p / 2} d \Sigma d t\right)^{1 / p} \leq T^{1 / 2 p}|\Sigma|^{1 / 2 p}\|f\|_{p}^{1 / 2}<\infty
$$

Since sup $\left|\arg f^{1 / 2}\right|<\pi / 2$, there exists $c>0$ such that

$$
\left|f(t, x)^{1 / 2}+f\left(t^{\prime}, x^{\prime}\right)^{1 / 2}\right| \geq c\left(|f(t, x)|^{1 / 2}+\left|f\left(t^{\prime}, x^{\prime}\right)\right|^{1 / 2}\right)
$$

by Lemma B.54. Then the estimate

$$
\left|f(t, x)^{1 / 2}-f\left(t^{\prime}, x\right)^{1 / 2}\right|=\frac{\left|f(t, x)-f\left(t^{\prime}, x\right)\right|}{\left|f(t, x)^{1 / 2}+f\left(t^{\prime}, x\right)^{1 / 2}\right|} \lesssim \frac{\left|f(t, x)-f\left(t^{\prime}, x\right)\right|}{\inf |f|^{1 / 2}},
$$

yields

$$
\llbracket f^{1 / 2} \rrbracket_{\tau, p ; p} \lesssim(\inf |f|)^{-1 / 2} \llbracket f \rrbracket_{\tau, p ; p}<\infty
$$

for $\tau=1-1 / 2 p$, and therefore

$$
\begin{aligned}
\llbracket \partial_{t} f^{1 / 2} \rrbracket_{\tau, p ; p} & =\llbracket 2^{-1} f^{-1} f^{1 / 2} \partial_{t} f \rrbracket_{\tau, p ; p} \\
& \lesssim \llbracket f^{-1} \rrbracket_{\tau, p ; p}\|f\|_{\infty}^{1 / 2}\left\|\partial_{t} f\right\|_{\infty}+\left\|f^{-1}\right\|_{\infty}\left(\left\|f^{1 / 2}\right\|_{\tau, p ; p}\left\|\partial_{t} f\right\|_{\infty}+\left\|f^{1 / 2}\right\|_{\infty} \llbracket \partial_{t} f \rrbracket_{\tau, p ; p}\right)
\end{aligned}
$$

is finite. For $\sigma=1-1 / p$ we similarly obtain

$$
\llbracket f^{1 / 2} \rrbracket_{p ; \sigma, p} \lesssim(\mathrm{inf}|f|)^{-1 / 2} \llbracket f \rrbracket_{p ; \sigma, p}<\infty .
$$

Next,

$$
\begin{aligned}
\llbracket \nabla_{\Sigma} \partial_{t} f^{1 / 2} \rrbracket_{p ; \sigma, p}= & 2^{-1} \llbracket 2^{-1} f^{-3 / 2} \nabla_{\Sigma} f \partial_{t} f+f^{-1 / 2} \nabla_{\Sigma} \partial_{t} f \rrbracket_{p ; \sigma, p} \\
& \lesssim\left\|f^{-1}\right\|_{\mathbb{E}_{\partial \Theta}}^{2}\left(\left\|f^{1 / 2}\right\|_{\infty}+\left\|f^{1 / 2}\right\|_{p ; \sigma, p}\right)\|f\|_{\mathbb{E}_{\partial \Theta}}^{2} \\
& +\left\|\left(1, \nabla_{\Sigma}\right) f^{-1 / 2}\right\|_{\infty}\left(\llbracket \nabla_{\Sigma} \partial_{t} f \rrbracket_{p ; \sigma, p}+\left\|\nabla_{\Sigma} \partial_{t} f\right\|_{p}\right)<\infty .
\end{aligned}
$$

The remaining terms in $\left\|f^{1 / 2}\right\|_{\mathbb{E}_{\partial \Theta}}$ can be estimated similarly. Hence $f^{1 / 2}$ belongs to $\mathbb{E}_{\partial \Theta}$ for every $f \in \mathbb{E}_{\partial \Theta}$ with $\inf _{J \times \Sigma} \operatorname{dist}\left(f(\cdot), \mathbb{R}_{-}\right)>0$, and Proposition B. 89 yields analyticity of $f \mapsto$ $f^{1 / 2}$.

The next step towards analyticity of $h \mapsto \Theta_{h}$ is to show that $h \mapsto \nu_{\Gamma_{h}} \circ \theta_{h}$ is analytic.
4.14. Lemma. Let $\Sigma \subset \mathbb{R}^{n}(n \geq 2)$ be a compact smooth hypersurface.
(i) Let $p \in(1, \infty), s \in(1+n / p, \infty)$, and $\tau \in(1+n / p, s]$. Then there exists $\delta_{h_{0}}=\delta_{h_{0}}(\Sigma, p, \tau)>0$ such that all height functions

$$
h \in W_{p}^{s-1 / p}(\Sigma) \cap \mathcal{U}_{h_{0}}, \quad \text { with } \mathcal{U}_{h_{0}}:=\left\{h \in W_{p}^{\tau-1 / p}(\Sigma):\|h\|_{W_{p}^{\tau-1 / p}(\Sigma)}<\delta_{h_{0}}\right\}
$$

satisfy (4.13). In this case the map

$$
h \mapsto \nu_{\Gamma_{h}} \circ \theta_{h}, \quad W_{p}^{s-1 / p}(\Sigma) \cap \mathcal{U}_{h_{0}} \rightarrow W_{p}^{s-1-1 / p}(\Sigma)^{n}
$$

is analytic.
(ii) Let $p \in((n+2) / 2, \infty), \tau \in(1+n / p, 4-1 / p]$, and let $J=(0, T)$ be bounded. Then there exists $\delta_{h}=\delta_{h}(\Sigma, p, \tau)>0$ such that all height functions

$$
\begin{equation*}
h \in \mathbb{E}_{h} \cap \mathcal{U}_{h}, \quad \text { with } \mathcal{U}_{h}:=\left\{h \in L_{\infty}\left(0, T ; W_{p}^{\tau-1 / p}(\Sigma)\right):\|h\|_{L_{\infty}\left(J ; W_{p}^{\tau-1 / p}(\Sigma)\right)}<\delta_{h}\right\} \tag{4.20}
\end{equation*}
$$

satisfy (4.13). In this case the map

$$
h \mapsto \nu_{\Gamma_{h}} \circ \tilde{\Theta}_{h}, \quad \mathbb{E}_{h} \cap \mathcal{U}_{h} \rightarrow \mathbb{E}_{\partial \Theta}^{n}
$$

is analytic.
Proof. (i) The identity (4.15) shows that the values of $h \nu_{\Sigma}$ and $g_{h}$ depend analytically on those of $\left(h, \nabla_{\Sigma} h\right) \in \mathbb{R} \times \mathbb{R}^{n}$ such that $|h|<\left\|L_{\Sigma}\right\|_{\infty}^{-1}$ and $\left|\nabla_{\Sigma} h\right|<\left(1-|h|\left\|L_{\Sigma}\right\|_{\infty}^{-1}\right)^{-1}$. From Sobolev's embedding $W_{p}^{s-1-1 / p}(\Sigma) \hookrightarrow B C(\Sigma)$ we infer that there exists $\delta_{h_{0}}$ such that (4.13)
is satisfied if $\|h\|_{W_{p}^{\tau-1 / p}(\Sigma)}<\delta_{h_{0}}$. By Remark B.80, the space $W_{p}^{s-1-1 / p}(\Sigma)$ is a multiplication algebra, and since $\nu_{\Sigma}$ and $L_{\Sigma}$ are smooth, we infer from Lemma B. 10 that

$$
h \mapsto\left(h \nu_{\Sigma}, I-h L_{\Sigma}\right), \quad W_{p}^{s-1 / p}(\Sigma) \rightarrow W_{p}^{s-1 / p}(\Sigma)^{n} \times W_{p}^{s-1-1 / p}(\Sigma)^{n \times n}
$$

is affine and continuous. The inversion operator

$$
A \mapsto A^{-1}, \quad\left\{A \in W_{p}^{s-1-1 / p}(\Sigma)^{n \times n}:\left\|A^{-1}\right\|_{\infty}<\infty\right\} \rightarrow W_{p}^{s-1-1 / p}(\Sigma)^{n \times n}
$$

is analytic by Lemma B. 90 . Hence, by Corollary B.86, the map

$$
h \mapsto\left(I-h L_{\Sigma}\right)^{-1}, \quad W_{p}^{s-1 / p}(\Sigma) \cap \mathcal{U}_{h_{0}} \rightarrow W_{p}^{s-1-1 / p}(\Sigma)^{n \times n}
$$

is analytic, and therefore also

$$
h \mapsto\left|\left(I-h L_{\Sigma}\right)^{-1} \nabla_{\Sigma} h\right|^{2}, \quad W_{p}^{s-1 / p}(\Sigma) \cap \mathcal{U}_{h_{0}} \rightarrow W_{p}^{s-1-1 / p}(\Sigma)^{n \times n}
$$

is analytic. Again by Lemma B. 90 and Sobolev's embedding, the square root operator and the inversion operator are analytic operators from $\left\{u \in W_{p}^{s-1-1 / p}(\Sigma): \inf _{\Sigma} \operatorname{dist}\left(u(\cdot), \mathbb{R}_{-}\right)>0\right\}$ to $W_{p}^{s-1-1 / p}(\Sigma)$. Thus, in view of (4.15), we conclude that

$$
h \mapsto \nu_{\Gamma_{h}} \circ \theta_{h}, \quad W_{p}^{s-1 / p}(\Sigma) \cap \mathcal{U}_{h_{0}} \rightarrow W_{p}^{s-1-1 / p}(\Sigma)
$$

is analytic.
(ii) The temporal trace theorem yields the embedding

$$
\mathbb{E}_{h}(T) \hookrightarrow C\left([0, T] ; W_{p}^{4-2 / p}(\Sigma)\right) \hookrightarrow C\left([0, T] ; W_{p}^{\tau-1 / p}(\Sigma)\right) .
$$

By employing Lemma 4.13 instead of Lemma B.90, assertion (ii) follows analogously.
Now we can prove that the normal-preserving map $\Theta_{h}$ is a diffeomorphism and that it depends analytically on the height function $h$. We consider $\Theta_{h}$ as an element of $\operatorname{Id}_{x}+\mathbb{E}_{\Theta}{ }^{n}$, where

$$
\mathbb{E}_{\Theta}:=H_{p}^{3 / 2}\left(J ; H_{p}^{2}\left(\mathbb{R}^{n}\right)\right) \cap H_{p}^{1}\left(J ; H_{p}^{3}\left(\mathbb{R}^{n}\right)\right) \cap L_{p}\left(J ; H_{p}^{4}\left(\mathbb{R}^{n}\right)\right) .
$$

4.15. Theorem. Let $\Omega \subset \mathbb{R}^{n}(n \geq 2)$ be a domain, $\Sigma \subset \Omega$ be a compact smooth hypersurface, and $p \in(1, \infty)$.
(i) Let $s \in(1+n / p, \infty)$ and $\tau \in(1+n / p, s]$. Then for some $\delta_{h_{0}}>0$ and all height functions

$$
h \in W_{p}^{s-1 / p}(\Sigma) \cap \mathcal{U}_{h_{0}} \quad \text { with } \mathcal{U}_{h_{0}}:=\left\{h \in W_{p}^{\tau-1 / p}(\Sigma):\|h\|_{W_{p}^{\tau-1 / p}(\Sigma)}<\delta_{h_{0}}\right\},
$$

the inequality (4.13) is satisfied, the stationary normal-preserving map $\Theta_{h}: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ from (4.14) is an admissible map, and the map

$$
h \mapsto \Theta_{h}-\operatorname{Id}_{x}, \quad W_{p}^{s-1 / p}(\Sigma) \cap \mathcal{U}_{h_{0}} \rightarrow W_{p}^{s}\left(\mathbb{R}^{n}\right)^{n}
$$

is analytic.
(ii) Let $p \in((n+2) / 2, \infty)$ and $\tau \in(1+n / p, 4-1 / p]$. Then there exists $\delta_{h}>0$ such that for all $T \in(0, \infty)$ and all height functions
(4.21) $h \in \mathbb{E}_{h}(T) \cap \mathcal{U}_{h} \quad$ with $\mathcal{U}_{h}:=\left\{h \in L_{\infty}\left(0, T ; W_{p}^{\tau-1 / p}(\Sigma)\right):\|h\|_{L_{\infty}\left(0, T ; W_{p}^{\tau-1 / p}(\Sigma)\right)}<\delta_{h}\right\}$,
the following assertions are true:
(ii.a) The inequality (4.13) is satisfied by $h(t, \cdot)$ for all $t \in[0, T]$.
(ii.b) $\Theta_{h}:[0, T] \times \bar{\Omega} \rightarrow \bar{\Omega}$ is a normal-preserving admissible map for $\Sigma$.
(ii.c) The following maps are analytic:

$$
\begin{array}{ll}
h \mapsto \Theta_{h}-\operatorname{Id}_{x}, & \mathbb{E}_{h}(T) \cap \mathcal{U}_{h} \rightarrow \mathbb{E}_{\Theta}(T), \\
\left.h \mapsto\left[\partial_{x} \Theta_{h}\right]\right|_{\Sigma}=I_{x}+\partial_{j}\left(h \nu_{\Sigma}\right) \otimes \tau_{\Sigma}^{j}+g_{h} \otimes \nu_{\Sigma}, & \mathbb{E}_{h}(T) \cap \mathcal{U}_{h} \rightarrow \mathbb{E}_{\partial \Theta}^{n \times n}(T) .
\end{array}
$$

Proof. (i) From Lemma 4.14.(i) we infer that the map

$$
h \mapsto g_{h}, \quad W_{p}^{\tau-1 / p}(\Sigma) \cap \mathcal{U}_{h_{0}} \rightarrow W_{p}^{\tau-1-1 / p}(\Sigma)
$$

is analytic for every $\tau \in(1+n / p, s]$. Moreover,

$$
\left\|h \nu_{\Sigma}\right\|_{W_{p}^{\tau-1 / p}(\Sigma)}+\left\|g_{h}\right\|_{W_{p}^{\tau-1-1 / p}(\Sigma)} \rightarrow 0 \quad \text { as }\|h\|_{W_{p}^{\tau-1 / p}(\Sigma)} \rightarrow 0
$$

Therefore the map $\Theta_{h}-\operatorname{Id}_{x}$ belongs to $W_{p}^{s}\left(\mathbb{R}^{n}\right)^{n} \cap C^{1}\left(\mathbb{R}^{n}\right)^{n}$, satisfies $\left.\Theta_{h}\right|_{\Sigma}=\theta_{h}$ and $\left.\partial_{\nu_{\Sigma}} \Theta_{h}\right|_{\Sigma}=$ $\left[\left(\nu_{\Gamma_{h}} \otimes \nu_{\Gamma_{h}}\right) \circ \theta_{h}\right] \nu_{\Sigma}$, and depends analytically on $h \in W_{p}^{s-1 / p}(\Sigma) \cap \mathcal{U}_{h_{0}}$. From Lemma 4.10 we infer that $\mathfrak{S}\left(h \nu_{\Sigma}, g_{h}\right)$ has compact support in $\bar{\Omega}$, and hence $\Theta_{h}=\operatorname{Id}_{x}$ in $\mathbb{R}^{n} \backslash \Omega$. In order to guarantee that $\Theta_{h}$ is a diffeomorphism, we observe that

$$
\left\|\partial_{x} \Theta_{h}-I_{x}\right\|_{L_{\infty}\left(\mathbb{R}^{n}\right)} \lesssim\left\|\mathfrak{S}\left(h \nu_{\Sigma}, g_{h}\right)\right\|_{W_{p}^{\tau}\left(\mathbb{R}^{n}\right)} \lesssim\left\|h \nu_{\Sigma}\right\|_{W_{p}^{\tau-1 / p}(\Sigma)}+\left\|g_{h}\right\|_{W_{p}^{\tau-1-1 / p}(\Sigma)}
$$

Hence, if $\|h\|_{W_{p}^{\tau-1 / p}(\Sigma)}$ is sufficiently small, then $\left\|\partial_{x} \Theta_{h}-I\right\|_{\infty}<1$, and thus $\Theta_{h}$ is a global diffeomorphism of $\mathbb{R}^{n}$.
(ii) It is shown in Lemma 4.13 that $\mathbb{E}_{\partial \Theta}$ is a multiplication algebra and the subset $\{u \in$ $\left.\mathbb{E}_{\partial \Theta}: \inf _{\Sigma} \operatorname{dist}\left(u(\cdot), \mathbb{R}_{+}\right)>0\right\}$ is invariant under pointwise inversion and square root. Let $J=(0, T)$. From Lemma 4.14.(ii) we infer that $h \nu_{\Sigma} \in \mathbb{E}_{h}$ and $g_{h} \in \mathbb{E}_{\partial \Theta}$ defined by (4.15) depend analytically on $h \in \mathbb{E}_{h}(T) \cap \mathcal{U}_{h}$. The mixed derivative embeddings yield

$$
h \nu_{\Sigma} \in W_{p}^{2-1 / 2 p-\rho}\left(J ; H_{p}^{1+2 \rho}(\Sigma)\right)^{n}, \quad g_{h} \in W_{p}^{2-1 / 2 p-\rho}\left(J ; H_{p}^{2 \rho}(\Sigma)\right)^{n} \quad \text { for all } \rho \in[0,1-1 / 2 p)
$$

We choose $\rho:=1 / 2-1 / 2 p$. By Lemma 4.10, the map (4.22) is well-defined and analytic. Analyticity of (4.23) follows from analyticity of $h \mapsto\left(h \nu_{\Sigma}, g_{h}\right)$ and Lemma B.10. The diffeomorphism property follows from assertion (i).
(iii) From Sobolev's embedding (B.1) we deduce

$$
W_{p}^{\theta}\left(J ; W_{p}^{4-\theta}\left(\mathbb{R}^{n}\right)\right) \hookrightarrow C\left(\bar{J} ; B C^{2}\left(\mathbb{R}^{n}\right)\right)
$$

provided that $\theta>1 / p$ and $4-\theta-n / p>2$. Since $p>(n+2) / 2$, we have

$$
H_{p}^{3 / 2}\left(J ; H_{p}^{2}\left(\mathbb{R}^{n}\right)\right) \cap H_{p}^{1}\left(J ; H_{p}^{3}\left(\mathbb{R}^{n}\right)\right) \hookrightarrow C^{1}\left(\bar{J} ; B C^{1}\left(\mathbb{R}^{n}\right)\right)
$$

Hence the $\operatorname{map} \Theta_{h}$ is admissible.
We complete this section with a collection of useful transformation identities for the normalpreserving admissible map $\Theta_{h}$ and the velocity transformation $u \circ \tilde{\Theta}_{h}=\left[\partial_{x} \Theta_{h}\right] \bar{u}$.
4.16. Lemma. The relations $(4.24)$ and (4.25) on the next page are valid.

Proof. Most identities follow from Propositions 4.6 and 4.12 and equation (4.15). The remaining identity $(4.25 \mathrm{c}$ ) can be verified as follows.

$$
\left(\partial_{\nu_{\Gamma}} w\right) \circ \tilde{\Theta}_{h}=\left(\nabla\left(\left(\beta_{h} \bar{w}\right) \circ \tilde{\Theta}_{h}^{-1}\right) \mid \nu_{\Gamma}\right) \circ \Theta_{h}=\nabla\left(\beta_{h} \bar{w}\right) \cdot\left[\partial_{x} \Theta_{h}\right]^{-1} \beta_{h}^{-1}\left[\partial_{x} \Theta_{h}\right] \nu_{\Sigma}=\partial_{\nu_{\Sigma}} \bar{w}
$$

### 4.2. The transformed bulk equations

In this section we transform the momentum balance and the divergence equation

$$
\begin{align*}
\rho \partial_{t} u+\rho(u \cdot \nabla) u-\mu \Delta u+\nabla \pi=0 & \text { for } t \in J, x \in \Omega \backslash \Gamma(t),  \tag{4.26}\\
\operatorname{div} u=0 & \text { for } t \in J, x \in \Omega \backslash \Gamma(t) . \tag{4.27}
\end{align*}
$$

Here $J=(0, T)$ is a bounded interval, $\Omega \subset \mathbb{R}^{n}(n \geq 2)$ is a domain, and $\Gamma$ is an admissible moving hypersurface in $\Omega$, which is induced by an admissible map $\Theta: J \times \bar{\Omega} \rightarrow \bar{\Omega}$ and a compact smooth hypersurface $\Sigma \subset \Omega$. We do not yet employ an explicit representation of $\Theta$.

Let $\Theta_{h}: J \times \bar{\Omega} \rightarrow \bar{\Omega}$ denote the normal-preserving map for $\Sigma \subset \Omega$ as defined in Theorem 4.15, $\Gamma(t)=\Theta_{h(t)}(\Sigma)$, and $M_{h}=\left(I_{x}-h L_{\Sigma}\right)^{-1}$. Then the following identities are valid on $J \times \Sigma$ :

$$
\begin{array}{rlrl}
(4.24 \mathrm{a}) & & \beta_{h} & =\left(\nu_{\Gamma} \circ \tilde{\Theta}_{h} \mid \nu_{\Sigma}\right)=\left|\left[\partial_{x} \Theta_{h}\right] \nu_{\Sigma}\right|=\left|\left[\partial_{x} \Theta_{h}\right]^{-\top} \nu_{\Sigma}\right|^{-1}, \\
\text { (4.24b) } & \nu_{\Gamma} \circ \tilde{\Theta}_{h} & =\beta_{h}^{-1}\left[\partial_{x} \Theta_{h}\right] \nu_{\Sigma}=\beta_{h}\left[\partial_{x} \Theta_{h}\right]^{-\top} \nu_{\Sigma}, \\
\text { (4.24c) } & \tau_{j}^{\Gamma_{h}} \circ \tilde{\Theta}_{h} & =\left[\partial_{x} \Theta_{h}\right] \tau_{j}^{\Sigma}, \\
\text { (4.24d) } & \tau_{\Gamma_{h}}^{j} \circ \tilde{\Theta}_{h} & =\left[\partial_{x} \Theta_{h}\right]^{-\top} \tau_{\Sigma}^{j}, \\
\text { (4.24e) } & P_{\Gamma} \circ \tilde{\Theta}_{h} & =\left[\partial_{x} \Theta_{h}\right] P_{\Sigma}\left[\partial_{x} \Theta_{h}\right]^{-1}, \\
\text { (4.24f) } & \partial_{x} \Theta_{h} & =P_{\Sigma}-h L_{\Sigma}-\beta_{h}^{2} M_{h} \nabla_{\Sigma} h \otimes \nu_{\Sigma}+\nu_{\Sigma} \otimes \nabla_{\Sigma} h+\beta_{h}^{2} \nu_{\Sigma} \otimes \nu_{\Sigma}, \tag{4.24f}
\end{array}
$$

Let $u=v+w \nu_{\Gamma}$ and $\bar{u}=\bar{v}+\bar{w} \nu_{\Sigma}$ be related by $u \circ \tilde{\Theta}_{h}=\left[\partial_{x} \Theta_{h}\right]$. Then we also have

$$
\begin{align*}
v \circ \tilde{\Theta}_{h} & =\left[\partial_{x} \Theta_{h}\right] \bar{v},  \tag{4.25a}\\
w \circ \tilde{\Theta}_{h} & =\beta_{h} \bar{w},  \tag{4.25b}\\
\partial_{\nu_{\Gamma}} w \circ \tilde{\Theta}_{h} & =\partial_{\nu_{\Sigma}} \bar{w},  \tag{4.25c}\\
V_{\Gamma_{h}} \circ \tilde{\Theta}_{h} & =\beta_{h} \partial_{t} h . \tag{4.25d}
\end{align*}
$$

Figure 4.2. Transformation identities for the normal-preserving map $\Theta_{h}$.

Our first task is to derive the transformed equations

$$
\begin{align*}
\bar{\rho} \partial_{t} \bar{u}-\bar{\mu} \Delta \bar{u}+\nabla \bar{\pi} & =F_{u}(\bar{u}, \bar{\pi}, \Theta) & & \text { in } J \times(\Omega \backslash \Sigma),  \tag{4.28}\\
\operatorname{div} \bar{u} & =F_{d}(\bar{u}, \Theta) & & \text { in } J \times(\Omega \backslash \Sigma), \tag{4.29}
\end{align*}
$$

for the transformed velocity

$$
\begin{equation*}
\bar{u}(t, \bar{x})=\left[\partial_{\bar{x}} \Theta(t, \bar{x})\right]^{-1} u(t, \Theta(t, \bar{x})), \tag{4.30}
\end{equation*}
$$

and the transformed pressure

$$
\begin{equation*}
\bar{\pi}(t, \bar{x}):=\pi(t, \Theta(t, \bar{x})) . \tag{4.31}
\end{equation*}
$$

The nonlinear perturbations $F_{u}$ and $F_{d}$ are derived in Lemmas 4.17 and 4.19.
Second, for proving well-posedness of the transformed problem (T) with Banach's fixed point theorem, we have to control the perturbations $F_{u}$ and $F_{d}$. To be precise, we will show that their values and their first order Fréchet derivatives can be deemed as small as we wish, by choosing $T$ sufficiently small and by requiring that $\left.\Theta\right|_{t=0}$ is sufficiently close to the identity (Lemmas 4.21 and 4.23). These perturbations are polynomial Nemytskiĭ operators with respect to the functions ( $\bar{u}, \bar{\pi}, \Theta,\left[\partial_{\bar{x}} \Theta\right]^{-1}$ ) and some of their derivatives. In order to prove their analyticity, we employ their polynomial structure and certain $T$-dependent embedding estimates.
4.17. Lemma. Assume that $\Theta$ is of class $C^{1}\left(J ; C^{1}(\bar{\Omega})\right) \cap C\left(J ; C^{3}(\bar{\Omega})\right)$ and put

$$
X(t, \bar{x}):=\Theta(t, \bar{x}), \quad \bar{X}(t, x):=\Theta(t, \cdot)^{-1}(x) .
$$

For given $\pi \in L_{1, l o c}\left(J ; H_{1, l o c}^{1}(\Omega \backslash \Gamma)\right)$ and $u \in H_{1, l o c}^{1}\left(J ; L_{1, l o c}\left(\Omega ; \mathbb{R}^{n}\right)\right) \cap L_{1, \text { loc }}\left(J ; H_{1, l o c}^{2}\left(\Omega \backslash \Gamma ; \mathbb{R}^{n}\right)\right)$ we define $\bar{u}$ and $\bar{\pi}$ as in (4.30) and (4.31). Then the identities in Figure 4.3 on the facing page are valid.

For an admissible map $\Theta: J \times \bar{\Omega} \rightarrow \bar{\Omega}$ and for $u(t, x)=\left[\partial_{\bar{x}} \Theta(t, \bar{x})\right] \bar{u}(t, \bar{x})$ and $\pi(t, x)=$ $\bar{\pi}(t, \bar{x})$ with $x=X(t, \bar{x})$ and $\bar{x}=\bar{X}(t, x)$ as in Lemma 4.17, we have

$$
\begin{aligned}
\partial_{t} \bar{X}_{m}= & -\partial_{i} \bar{X}_{m} \partial_{t} X_{i}, \\
\partial_{j} u_{k}= & \partial_{j} \bar{u}_{k}+\left(\partial_{l} X_{k} \partial_{j} \bar{X}_{m}-\delta_{k l} \delta_{j m}\right) \partial_{m} \bar{u}_{l}+\partial_{l} \partial_{m} X_{k} \partial_{j} \bar{X}_{m} \bar{u}_{l}, \\
\partial_{t} u_{k}= & \partial_{t} \bar{u}_{k}+\left(\partial_{l} X_{k}-\delta_{k l}\right) \partial_{t} \bar{u}_{l}-\partial_{t} X_{i} \partial_{l} X_{k} \partial_{i} \bar{X}_{m} \partial_{m} \bar{u}_{l} \\
& +\left(\partial_{t} \partial_{l} X_{k}-\partial_{t} X_{i} \partial_{l} \partial_{m} X_{k} \partial_{i} \bar{X}_{m}\right) \bar{u}_{l}, \\
\operatorname{div} u= & \operatorname{div} \bar{u}+\partial_{l} \partial_{m} X_{j} \partial_{j} \bar{X}_{m} \bar{u}_{l}, \\
\Delta u_{k}= & \Delta \bar{u}_{k}+\left(\partial_{l} X_{k} \partial_{j} \bar{X}_{m} \partial_{j} \bar{X}_{i}-\delta_{k l} \delta_{j m} \delta_{j i}\right) \partial_{m} \partial_{i} \bar{u}_{l} \\
& +\left(\partial_{l} \partial_{i} X_{k} \partial_{j} \bar{X}_{m} \partial_{j} \bar{X}_{i}-\partial_{i} \bar{X}_{m} \partial_{j} \partial_{r} X_{i} \partial_{j} \bar{X}_{r} \partial_{l} X_{k}\right) \partial_{m} \bar{u}_{l} \\
& +\left(\partial_{l} \partial_{m} \partial_{i} X_{k} \partial_{j} \bar{X}_{m} \partial_{j} \bar{X}_{i}-\partial_{i} \bar{X}_{m} \partial_{j} \partial_{r} X_{i} \partial_{j} \bar{X}_{r} \partial_{l} \partial_{m} X_{k}\right) \bar{u}_{l}, \\
u_{j} \partial_{j} u_{k}= & \partial_{l} X_{k} \partial_{i} X_{j} \partial_{j} \bar{X}_{m} \bar{u}_{i} \partial_{m} \bar{u}_{l}+\partial_{l} \partial_{m} X_{k} \partial_{i} X_{j} \partial_{j} \bar{X}_{m} \bar{u}_{l} \bar{u}_{i}, \\
\partial_{j} \pi= & \partial_{j} \bar{\pi}+\left(\partial_{j} \bar{X}_{m}-\delta_{j m}\right) \partial_{m} \bar{\pi} .
\end{aligned}
$$

Here the values of $u, \pi$, and $\bar{X}$ are taken at $(t, x)$, and those of $\bar{u}, \bar{\pi}$, and $X$ at $(t, \bar{x})$.
Figure 4.3. Transformed differential operators.

Proof. By the inverse function theorem we have

$$
\begin{aligned}
\partial_{x} \bar{X}(t, x) & =\partial_{x}(\Theta(t, \cdot))^{-1}(x)=\left[\partial_{\bar{x}} \Theta(t, \bar{x})\right]^{-1}, \\
\partial_{t} \bar{X}(t, x) & =-\left[\partial_{\bar{x}} \Theta(t, x)\right]^{-1} \partial_{t} \Theta(t, \bar{x})=-\partial_{x} \bar{X}(t, x) \partial_{t} X(t, \bar{x}), \quad \partial_{t} \bar{X}_{m}=-\partial_{n} \bar{X}_{m} \partial_{t} X_{n} .
\end{aligned}
$$

In order to transform $\partial_{j} u_{k}$, we apply the chain rule for weak derivatives ([Hun13, Proposition 3.21]). Neglecting the dependence on $t$, we obtain

$$
\begin{aligned}
\partial_{x_{j}} u_{k}(x)= & \partial_{x_{j}}\left(\partial_{\bar{x}_{l}} X_{k}(\bar{X}(x)) \bar{u}_{l}(\bar{X}(x))\right) \\
= & \partial_{\bar{x}_{j}} \bar{u}_{k}(\bar{x})+\left(\partial_{\bar{x}_{l}} X_{k}(\bar{x}) \partial_{x_{j}} \bar{X}_{m}(x)-\delta_{k l} \delta_{j m}\right) \partial_{\bar{x}_{m}} \bar{u}_{l}(\bar{x}) \\
& +\partial_{\bar{x}_{l}} \partial_{\bar{x}_{m}} X_{k}(x) \partial_{x_{j}} \bar{X}_{m}(x) \bar{u}_{l}(\bar{x}) .
\end{aligned}
$$

The remaining equations follow by straightforward computations.
In every transformation formula in Figure 4.3, the first summand on the right-hand side is the principal part and the remaining summands are treated as perturbations. In order to abstract their polynomial structure, we employ the following convention.
4.18. Convention. For a map $f: E_{1} \times \cdots \times E_{k} \rightarrow Y$ between Banach spaces $E_{1}, \ldots, E_{k}$, and $Y$ over $\mathbb{K} \in\{\mathbb{R}, \mathbb{C}\}$, and the induced Nemytskií operator

$$
F: u \mapsto f \circ u, \quad F\left(u_{1}, \ldots, u_{k}\right)(x)=f\left(u_{1}(x), \ldots, u_{k}(x)\right),
$$

acting on $E_{1} \times \cdots \times E_{k}$-valued maps $u=\left(u_{1}, \ldots, u_{k}\right)$, we write

\[

\]

The symbols $L, M, P, P_{\alpha}$, and $P_{\alpha, 0}$ may denote a different mapping at every occurrence.

For $x=X(t, \bar{x})=\Theta(t, \bar{x}), \partial_{x} \bar{X}(t, x)=[\partial \Theta(t, \bar{x})]^{-1}$, and $\tilde{\Theta}(t, \bar{x})=(t, \Theta(t, \bar{x}))$ we have

$$
\begin{aligned}
F_{u}(\bar{u}, \bar{\pi}, \Theta): & -\left(\left(\rho \partial_{t} u\right) \circ \tilde{\Theta}-\bar{\rho} \partial_{t} \bar{u}\right) \\
& -((\rho u \cdot \nabla) u) \circ \tilde{\Theta} \\
& -((\nabla \pi) \circ \tilde{\Theta}-\nabla \bar{\pi}) \\
& +((\mu \Delta u) \circ \tilde{\Theta}-\bar{\mu} \Delta \bar{u}),
\end{aligned}
$$

where

$$
\begin{aligned}
\left(\partial_{t} u_{k}\right) \circ \tilde{\Theta}-\partial_{t} \bar{u}_{k}= & \left(\partial_{l} X_{k}-\delta_{l k}\right) \partial_{t} \bar{u}_{l}-\partial_{l} X_{k} \partial_{t} X_{n} \partial_{n} \bar{X}_{m} \partial_{m} \bar{u}_{l} \\
& +\left(\partial_{t} \partial_{l} X_{k}-\partial_{l} \partial_{m} X_{k} \partial_{t} X_{n} \partial_{n} \bar{X}_{m}\right) \bar{u}_{l} \\
= & M\left(\partial_{\bar{x}} \Theta-I, \partial_{t} \bar{u}\right)+M\left(\partial_{t} X, \partial_{\bar{x}} X, \partial_{x} \bar{X}, \partial_{\bar{x}} \bar{u}\right) \\
& +M\left(\left(\partial_{t} \partial_{\bar{x}} X, \partial_{t} X\right), P\left(\partial_{\bar{x}}^{2} X, \partial_{x} \bar{X}\right), \bar{u}\right), \\
\left(u_{j} \partial_{j} u_{k}\right) \circ \tilde{\Theta}= & \partial_{l} X_{k} \partial_{n} X_{j} \partial_{j} \bar{X}_{m} \bar{u}_{n} \partial_{m} \bar{u}_{l}+\partial_{l} \partial_{m} X_{k} \partial_{n} X_{j} \partial_{j} \bar{X}_{m} \bar{u}_{l} \bar{u}_{n} \\
= & M\left(P\left(\partial_{\bar{x}}^{2} X, \partial_{\bar{x}} X, \partial_{x} \bar{X}\right), \bar{u},\left(\bar{u}, \partial_{\bar{x}} \bar{u}\right)\right), \\
\Delta u \circ \tilde{\Theta}-\Delta \bar{u}= & \left(\partial_{l} X_{k} \partial_{j} \bar{X}_{m} \partial_{j} \bar{X}_{n}-\delta_{l k} \delta_{j m} \delta_{j n}\right) \partial_{m} \partial_{n} \bar{u}_{l} \\
& +\left(2 \partial_{l} \partial_{n} X_{k} \partial_{j} \bar{X}_{m} \partial_{j} \bar{X}_{n}+\partial_{l} X_{k} \Delta \bar{X}_{m}\right) \partial_{m} \bar{u}_{l} \\
& +\left(\partial_{l} \partial_{m} \partial_{n} X_{k} \partial_{j} \bar{X}_{m} \partial_{j} \bar{X}_{n}-\partial_{j} \partial_{l} X_{k} \partial_{l} \partial_{m} X_{k} \partial_{k} \bar{X}_{m} \partial_{j} \bar{X}_{l}\right) \bar{u}_{l} \\
= & M\left(P_{(1,2), 0}\left(\partial_{\bar{x}} X-I, \partial_{x} \bar{X}-I\right), \partial_{\bar{x}}^{2} \bar{u}\right) \\
& +M\left(P\left(\partial_{\bar{x}}^{2} X, \partial_{\bar{x}} X, \partial_{x} \bar{X}\right),\left(\bar{u}, \partial_{\bar{x}} \bar{u}\right)\right)+M\left(\partial_{\bar{x}}^{3} X,\left(\partial_{x} \bar{X}\right)^{2}, \bar{u}\right), \\
\left(\partial_{j} \pi\right) \circ \tilde{\Theta}-\partial_{j} \bar{\pi}= & -\left(\partial_{j} \bar{X}_{m}-\delta_{j m}\right) \partial_{m} \bar{\pi}=M\left(\partial_{\bar{x}} X-I, \partial_{\bar{x}} \bar{\pi}\right) .
\end{aligned}
$$

Here the values of $u$, $\pi$, and $\bar{X}$ are taken at $(t, x) \in J \times(\Omega \backslash \Gamma(t))$, and those of $\bar{u}, \bar{\pi}$, and $X$ at $(t, \bar{x}) \in J \times(\Omega \backslash \Sigma)$ with $x=X(t, \bar{x})$.

FIgURE 4.4. The perturbations in the transformed momentum equation.
4.2.1. The transformed momentum equation. In the next Lemma 4.19, we derive the transformed momentum equation for admissible diffeomorphisms $\Theta(t, \cdot)$. The map $F_{u}(u, \pi, \Theta)$ is a polynomial operator in $\left(u, \pi, \Theta,\left[\partial_{x} \Theta\right]^{-1}\right)$. For suitable $\Theta \in \mathcal{U}_{\Theta}$ we obtain analyticity of $F_{u}$ and smallness of a certain Fréchet derivative of $F_{u}$ in Lemma 4.21. Sufficient $T$-dependent embeddings are given in Lemma 4.20. Later on we will specialize this result to normal-preserving diffeomorphisms $\Theta_{h}(t, \cdot)$ with $h \in \mathcal{U}_{h}$ (Corollary 4.27).
4.19. Lemma. The momentum equation (4.26) corresponds to the transformed momentum equation (4.28), where the vector-field $F_{u}(\bar{u}, \bar{\pi}, \Theta): J \times(\Omega \backslash \Sigma) \rightarrow \mathbb{R}^{n}$ is given in Figure 4.4 on the current page. Therefore $F_{u}$ is a polynomial Nemytskiŭ operator with respect to ( $\bar{u}, \bar{\pi}, \Theta,\left[\partial_{\bar{x}} \Theta\right]^{-1}$ ) of the form

$$
\begin{aligned}
F_{u}(\bar{u}, \bar{\pi}, \Theta)= & M\left(P_{(1,2), 0}\left(\partial_{\bar{x}} \Theta-I,\left[\partial_{\bar{x}} \Theta\right]^{-1}-I\right),\left(\partial_{t} \bar{u}, \partial_{\bar{x}}^{2} \bar{u}\right)\right)+M\left(\partial_{\bar{x}} \Theta-I, \partial_{\bar{x}} \bar{\pi}\right) \\
& +M\left(P\left(\partial_{\bar{x}}^{2} \Theta, \partial_{\bar{x}} \Theta,\left[\partial_{\bar{x}} \Theta\right]^{-1}\right), \bar{u},\left(\bar{u}, \partial_{\bar{x}} \bar{u}\right)\right)+M\left(P\left(\partial_{\bar{x}}^{2} \Theta, \partial_{\bar{x}} \Theta,\left[\partial_{\bar{x}} \Theta\right]^{-1}\right), \partial_{\bar{x}} \bar{u}\right) \\
& +M\left(\left(\partial_{t} \partial_{\bar{x}} \Theta, \partial_{t} \Theta, \partial_{\bar{x}}^{3} \Theta\right), P\left(\partial_{\bar{x}}^{2} \Theta,\left[\partial_{\bar{x}} \Theta\right]^{-1}\right), \bar{u}\right) .
\end{aligned}
$$

Proof. This follows from Lemma 4.17 by straightforward calculations.

In the remainder of this section we omit the bars on $\bar{u}$ and $\bar{\pi}$; that is, $u$ and $\pi$ denote the transformed velocity and pressure. We recall from Figure 4.1 that

$$
\begin{aligned}
\mathbb{E}_{u} & =\left\{u \in H_{p}^{1}\left(J ; L_{p}(\Omega)^{n}\right) \cap L_{p}\left(J ; H_{p}^{2}(\Omega \backslash \Sigma)^{n}\right):\left.u\right|_{\partial \Omega}=0, \llbracket u \rrbracket=0 \text { on } \Sigma\right\}, \\
\mathbb{E}_{\Theta} & =H_{p}^{3 / 2}\left(J ; H_{p}^{2}\left(\mathbb{R}^{n}\right)\right) \cap H_{p}^{1}\left(J ; H_{p}^{3}\left(\mathbb{R}^{n}\right)\right) \cap L_{p}\left(J ; H_{p}^{4}\left(\mathbb{R}^{n}\right)\right) .
\end{aligned}
$$

We consider the map $(u, \pi, \Theta) \mapsto F_{u}(u, \pi, \Theta)$ with target space $\mathbb{F}_{u}$, defined for $u \in \mathbb{E}_{u}, \pi \in \mathbb{E}_{\pi}$, and $\Theta \in \mathbb{E}_{\Theta}$, provided that $\left[\partial_{x} \Theta\right]^{-1}$ is bounded on $J \times \Omega$. Therefore we let

$$
\begin{equation*}
\mathcal{U}_{\Theta}:=\left\{\Theta \in \mathbb{E}_{\Theta}^{n}:\left.\Theta\right|_{J \times \bar{\Omega}}: J \times \bar{\Omega} \rightarrow \bar{\Omega} \text { is an admissible map }\right\} . \tag{4.32}
\end{equation*}
$$

From Proposition 4.3 we infer that for $\Theta \in \mathbb{E}_{\Theta} \cap \mathcal{U}_{\Theta}$, the map $\left[\partial_{x} \Theta\right]^{-1}$ is bounded on $J \times \Omega$. In order to control the nonlinearities on small time intervals, we consider the closed subspaces

$$
{ }_{0} \mathbb{E}_{u}:=\left\{u_{\bullet} \in \mathbb{E}_{u}:\left.u_{\bullet}\right|_{t=0}=0\right\}, \quad{ }_{0} \mathbb{E}_{\Theta}:=\left\{\eta_{\bullet} \in \mathbb{E}_{\Theta}:\left.\eta_{\bullet}\right|_{t=0}=0,\left.\partial_{t} \eta_{\bullet}\right|_{t=0}=0\right\} .
$$

4.20. Lemma. Let $p \in(1, \infty) \backslash\{3 / 2,3\}$ and $T \in(0, \infty)$. Then the continuous embeddings

$$
\begin{array}{rlrl}
\mathbb{E}_{u}(T) & \hookrightarrow C([0, T] \times \bar{\Omega})^{n} & & \text { if } p>(n+2) / 2, \\
\mathbb{E}_{\Theta}(T) \hookrightarrow H_{p}^{1}\left(0, T ; H_{p}^{1}(\Omega)^{n}\right) \cap C\left([0, T] ; C^{2}(\bar{\Omega})^{n}\right) \cap L_{p}\left(0, T ; H_{p}^{3}(\Omega)^{n}\right) & & \text { if } p>(n+1) / 2, \\
\mathbb{E}_{\Theta} & \hookrightarrow C^{1}\left([0, T] ; C^{1}(\bar{\Omega})^{n}\right) \cap C\left([0, T] ; C^{3}(\bar{\Omega})^{n}\right) & & \text { if } p>n+2, \tag{4.35}
\end{array}
$$

are valid, and for some $\delta_{0}>0$ and all $\delta \in\left(0, \delta_{0}\right], T_{0}>0$, and $T \in\left(0, T_{0}\right]$ we have

$$
\begin{array}{rlrl}
\|u \cdot\|_{L_{p}\left(0, T ; H_{p}^{1}(\Omega \backslash \Sigma)\right)} \leq T^{\delta} C\left(\delta, T_{0}\right)\left\|u_{\bullet}\right\|_{0 \mathbb{E}_{u}(T)} & & \text { if } p>2, \\
\left\|u_{\bullet}\right\|_{0} C([0, T] \times \bar{\Omega}) & \leq T^{\delta} C\left(\delta, T_{0}\right)\left\|u_{\bullet}\right\|_{0 \mathbb{E}_{u}(T)} & & \text { if } p>(n+2) / 2, \tag{4.37}
\end{array}
$$

Proof. We proceed as in the proof of Lemma 3.19.
Assertions (4.33) and (4.37) follow from Proposition B.44, (B.2), (3.32c), and (3.32d), since

$$
\begin{aligned}
\mathbb{E}_{u} & \hookrightarrow W_{p}^{\rho-\tau}\left(0, T ; W_{p}^{2(1-\rho)}(\Omega \backslash \Sigma)\right)^{n} \\
& \hookrightarrow W_{p}^{1 / p+\varepsilon_{t}}\left(0, T ; W_{p}^{n / p+\varepsilon_{s}}(\Omega \backslash \Sigma)\right)^{n} \hookrightarrow C([0, T] ; B U C(\Omega \backslash \Sigma))^{n}
\end{aligned}
$$

for some $\tau, \varepsilon_{t}, \varepsilon_{s}>0$, and $\rho \in(0,1)$, if $\rho>1 / p$ and $2(1-\rho)>n / p$, and this is possible if $p>(n+2) / 2$. Since $u \in \mathbb{E}_{u}$ satisfies $\llbracket u \rrbracket=0$, we obtain $u \in C([0, T] \times \bar{\Omega})^{n}$. Estimate (4.36) follows from Lemma 3.19.(i). Again by Proposition B. 44 and Sobolev's embedding (B.1) we have

$$
\mathbb{E}_{\Theta} \hookrightarrow W_{p}^{\theta}\left(J ; W_{p}^{4-\theta}\left(\mathbb{R}^{n}\right)\right) \hookrightarrow C\left(\bar{J} ; B C^{2}\left(\mathbb{R}^{n}\right)\right),
$$

provided that $\theta>1 / p$ and $4-\theta-n / p>2$, and this is possible if $p>(n+1) / 2$. Therefore (4.34) is valid, and estimate (4.38) also follows from (B.2), (3.32c), and (3.32d). The embedding (4.35) can be verified similarly.

We are ready to control the perturbation $F_{u}$.
4.21. Lemma. Let $p \in(n+2, \infty)$ and $T \in(0, \infty)$. Then the map

$$
F_{u}:\left\{(u, \pi, \Theta) \in \mathbb{E}_{u}(T) \times \mathbb{E}_{\pi}(T) \times \mathbb{E}_{\Theta}(T): \Theta \in \mathcal{U}_{\Theta}\right\} \rightarrow \mathbb{F}_{u}(T)
$$

is analytic and has the following properties:
(i) For given $T_{0} \in(0, \infty), R \in[1, \infty)$, $u \in \mathbb{E}_{u}\left(T_{0}\right), \pi \in \mathbb{E}_{\pi}\left(T_{0}\right)$, and $\Theta \in \mathbb{E}_{\Theta}\left(T_{0}\right) \cap \mathcal{U}_{\Theta}$, we have

$$
\left\|F_{u}(u, \pi, \Theta)\right\|_{\mathbb{F}_{u}(T)} \rightarrow 0 \quad \text { as } T \rightarrow 0,\left\|\partial_{x} \Theta-I_{x}\right\|_{C([0, T] \times \bar{\Omega})} \rightarrow 0,
$$

and this convergence is uniform with respect to

$$
\begin{equation*}
\|u\|_{\mathbb{E}_{u}\left(T_{0}\right)}+\|\pi\|_{\mathbb{E}_{\pi}\left(T_{0}\right)}+\|\Theta\|_{\mathbb{E}_{\Theta}\left(T_{0}\right)}+\left\|\left[\partial_{x} \Theta\right]^{-1}\right\|_{C\left(\left[0, T_{0}\right] \times \bar{\Omega}\right)} \leq R . \tag{4.39}
\end{equation*}
$$

(ii) For given $T_{0} \in(0, \infty)$, $R \in[1, \infty)$, $u_{*} \in \mathbb{E}_{u}\left(T_{0}\right)$, and $\Theta_{*} \in \mathbb{E}_{\Theta}\left(T_{0}\right)$, the map

$$
\begin{aligned}
& \left(u_{\bullet}, \pi, \eta_{\bullet}\right) \mapsto F_{u}\left(u_{*}+u_{\bullet}, \pi, \Theta_{*}+\eta_{\bullet}\right), \\
& \left\{\left(u_{\bullet}, \pi, \eta_{\bullet}\right) \in{ }_{0} \mathbb{E}_{u}(T) \times \mathbb{E}_{\pi}(T) \times{ }_{0} \mathbb{E}_{\Theta}(T): \Theta_{*}+\eta_{\bullet} \in \mathcal{U}_{\Theta}\right\} \rightarrow \mathbb{F}_{u}(T),
\end{aligned}
$$

satisfies

$$
\begin{aligned}
\left\|\partial_{\left(u_{\bullet}, \pi, \eta_{\bullet}\right)} F_{u}\left(u_{*}+u_{\bullet}, \pi, \Theta_{*}+\eta_{\bullet}\right)\right\|_{\mathcal{B}\left(\mathbb{E}_{u}(T) \times \mathbb{E}_{\pi}(T) \times_{0} \mathbb{E}_{\Theta}(T) ; \mathbb{F}_{u}(T)\right)} & \rightarrow 0 \\
\text { as } T & \rightarrow 0,\left\|\partial_{x} \Theta-I_{x}\right\|_{C([0, T] \times \bar{\Omega})}
\end{aligned}>0,
$$

and this convergence is uniform with respect to

$$
\begin{equation*}
\left\|\left(u_{*}, u_{\bullet}\right)\right\|_{\mathbb{E}_{u}\left(T_{0}\right)}+\|\pi\|_{\mathbb{E}_{\pi}\left(T_{0}\right)}+\left\|\left(\Theta_{*}, \eta_{\bullet}\right)\right\|_{\mathbb{E}_{\ominus}\left(T_{0}\right)}+\left\|\left[\partial_{x}\left(\Theta_{*}+\eta_{\bullet}\right)\right]^{-1}\right\|_{C\left(\left[0, T_{0}\right] \times \bar{\Omega}\right)} \leq R . \tag{4.40}
\end{equation*}
$$

Proof. Lemma 4.19 shows that $F_{u}(u, \pi, \Theta)$ depends polynomially on $\left(u, \pi, \Theta,\left[\partial_{x} \Theta\right]^{-1}\right)$. With Theorem 4.15 and the embeddings (4.33) and (4.34), it is straightforward to check that $F_{u}$ is analytic. It is also bounded and uniformly continuous with respect to (4.39).
(i) Assume that $T \leq 1$. First, Hardy's inequality (B.4) yields the estimate

$$
\begin{align*}
\|u\|_{\mathbb{F}_{u}(T)} & \leq\left(\int_{0}^{T}\left\|u_{0}\right\|_{L_{p}(\Omega)}^{p} d t\right)^{1 / p}+\left(\int_{0}^{T}\left\|\frac{T}{t} \int_{0}^{t} \partial_{s} u(s) d s\right\|_{L_{p}(\Omega)}^{p} d t\right)^{1 / p}  \tag{4.41}\\
& \leq C T^{1 / p}\left(\left\|u_{0}\right\|_{L_{p}(\Omega)}+\left\|\partial_{t} u\right\|_{\mathbb{F}_{u}\left(T_{0}\right)}\right) \leq C T^{1 / p} R .
\end{align*}
$$

Here the latter inequality follows from $\mathbb{E}_{u}\left(T_{0}\right) \hookrightarrow C\left(\left[0, T_{0}\right] \times \bar{\Omega}\right) \hookrightarrow C\left(\left[0, T_{0}\right] ; L_{p}(\Omega)\right)$ (see also (4.33)). Second, with inequality (3.32c) and the mixed derivative embeddings we obtain

$$
\begin{equation*}
\left\|\partial_{x} u\right\|_{\mathbb{F}_{u}(T)} \leq C T^{1 / p}\left\|u_{0}\right\|_{H_{p}^{1}(\Omega)}+C T^{\alpha} \llbracket u-u_{0} \rrbracket_{0 W_{p}^{\alpha}\left(0, T ; H_{p}^{1}(\Omega)\right)} \leq C T^{1 / p} R \tag{4.42}
\end{equation*}
$$

for some $\alpha \in(1 / p, 1 / 2)$. Hence, by using embedding (4.35) and choosing $T>0$ sufficiently small, we can control those terms in $F_{u}(u, \pi, \Theta)$ which contain a lower-order factor $u$ or $\partial_{x} u$. The leading-order terms $\partial_{t} u, \partial_{x}^{2} u$, and $\partial_{x} \pi$ only appear in products with a factor $\partial_{x} \Theta-I_{x}$ or $\left[\partial_{x} \Theta\right]^{-1}-I_{x}=\left[\partial_{x} \Theta\right]^{-1}\left[I_{x}-\partial_{x} \Theta\right]$, and can therefore be controlled with the smallness of $\left\|\partial_{x} \Theta-I_{x}\right\|_{C([0, T] \times \bar{\Omega})}$.
(ii) For given $u_{*} \in \mathbb{E}_{u}\left(T_{0}\right)$, $u_{\bullet} \in{ }_{0} \mathbb{E}_{u}\left(T_{0}\right), \Theta_{*} \in \mathbb{E}_{\Theta}\left(T_{0}\right)$, and $\eta_{\bullet} \in{ }_{0} \mathbb{E}_{\Theta}\left(T_{0}\right)$, we let

$$
u=u_{*}+u_{\bullet} \in \mathbb{E}_{u}\left(T_{0}\right), \quad \Theta=\Theta_{*}+\eta_{\bullet} \in \mathbb{E}_{\Theta}\left(T_{0}\right)
$$

With Convention 4.18 we can express the derivative of $u_{\bullet} \mapsto F_{u}\left(u_{*}+u_{\bullet}, \pi, \Theta\right)$ applied to $\tilde{u}_{\bullet} \in$ ${ }_{0} \mathbb{E}_{u}(T)$ as

$$
\begin{aligned}
{\left[\partial_{u_{\bullet}} F_{u}(u, \pi, \Theta)\right] \tilde{u}_{\bullet}=} & M\left(P_{(1,2), 0}\left(\partial_{x} \Theta-I,\left[\partial_{x} \Theta\right]^{-1}-I\right),\left(\partial_{t} \tilde{u}_{\bullet}, \partial_{x}^{2} \tilde{u}_{\bullet}\right)\right) \\
& +M\left(P\left(\partial_{x}^{2} \Theta, \partial_{x} \Theta,\left[\partial_{x} \Theta\right]^{-1}\right), \tilde{u}_{\bullet},\left(u, \partial_{x} u\right)\right) \\
& +M\left(P\left(\partial_{x}^{2} \Theta, \partial_{x} \Theta,\left[\partial_{x} \Theta\right]^{-1}\right), u,\left(\tilde{u}_{\bullet}, \partial_{x} \tilde{\bullet}_{\bullet}\right)\right) \\
& +M\left(P\left(\partial_{x}^{2} \Theta, \partial_{x} \Theta,\left[\partial_{x} \Theta\right]^{-1}\right), \partial_{x} \tilde{u}_{\bullet}\right) \\
& +M\left(P\left(\partial_{t} \partial_{x} \Theta, \partial_{t} \Theta, \partial_{x}^{3} \Theta, \partial_{x}^{2} \Theta,\left[\partial_{x} \Theta\right]^{-1}\right), \tilde{u}_{\bullet}\right) .
\end{aligned}
$$

Together with (4.36) and (4.37), a straightforward estimation yields

$$
\left\|\partial_{u} . F_{u}(u, \pi, \Theta)\right\|_{\mathcal{B}\left(o \mathbb{E}_{u}(T) ; \mathbb{F}_{u}(T)\right)} \rightarrow 0 \quad \text { as } T \rightarrow 0,\left\|\partial_{x} \Theta-I_{x}\right\|_{C([0, T] \times \bar{\Omega})} \rightarrow 0,
$$

uniformly with respect to (4.40).
Next, the derivative of $\mathbb{E}_{\pi}(T) \ni \pi \mapsto F_{u}(u, \pi, \Theta)$ applied to $\tilde{\pi} \in \mathbb{E}_{\pi}(T)$ is given by

$$
\left[\partial_{\pi} F_{u}(u, \pi, \Theta)\right] \tilde{\pi}=M\left(\left[\partial_{x} \Theta\right]^{-1}-I, \partial_{x} \tilde{\pi}\right),
$$

and therefore

$$
\left\|\partial_{\pi} F_{u}(u, \pi, \Theta)\right\|_{\mathcal{B}\left(\mathbb{E}_{\pi}(T) ; \mathbb{F}_{u}(T)\right)} \rightarrow 0 \quad \text { as }\left\|\partial_{x} \Theta-I_{x}\right\|_{C([0, T] \times \bar{\Omega})} \rightarrow 0 .
$$

Finally, we study the derivative of ${ }_{0} \mathbb{E}_{\Theta}(T) \ni \eta_{\bullet} \mapsto F_{u}\left(u, \pi, \Theta_{*}+\eta_{\bullet}\right)$. With

$$
\begin{equation*}
\left[\partial_{\eta_{\bullet}}\left(\left[\partial_{x} \Theta\right]^{-1}\right)\right] \tilde{\eta}_{\bullet}=-\left[\partial_{x} \Theta\right]^{-1}\left[\partial_{x} \tilde{\eta}_{\bullet}\right]\left[\partial_{x} \Theta\right]^{-1}=M\left(\partial_{x} \tilde{\eta}_{\bullet},\left(\left[\partial_{x} \Theta\right]^{-1}\right)^{2}\right) \tag{4.43}
\end{equation*}
$$

we obtain

$$
\left[\partial_{\eta_{\bullet}} P_{(1,2), 0}\left(\partial_{x} \Theta-I,\left[\partial_{x} \Theta\right]^{-1}-I\right)\right] \tilde{\eta}_{\bullet}=M\left(\partial_{x} \tilde{\eta}, P\left(\partial_{x} \Theta,\left[\partial_{x} \Theta\right]^{-1}\right)\right)
$$

Hence

$$
\begin{aligned}
{\left[\partial_{\eta_{\bullet}} F_{u}(u, \pi, \Theta)\right] \tilde{\eta}_{\bullet}=} & M\left(\partial_{x} \tilde{\eta}_{\bullet}, P\left(\partial_{x} \Theta,\left[\partial_{x} \Theta\right]^{-1}\right),\left(\partial_{t} u, \partial_{x}^{2} u, \partial_{x} \pi\right)\right) \\
& +M\left(\left(\partial_{x}^{2} \tilde{\eta}_{\bullet}, \partial_{x} \tilde{\eta}_{\bullet}\right), P\left(\partial_{x} \Theta, \partial_{x} \Theta,\left[\partial_{x} \Theta\right]^{-1}\right), P_{1}(u),\left(u, \partial_{x} u\right)\right) \\
& +M\left(\left(\partial_{t} \partial_{x} \tilde{\eta}_{\bullet}, \partial_{t} \tilde{\eta}_{\bullet}, \partial_{x}^{3} \tilde{\eta}_{\bullet}\right), P\left(\partial_{x}^{2} \Theta,\left[\partial_{x} \Theta\right]^{-1}\right), u\right) \\
& +M\left(\left(\partial_{t} \partial_{x} \Theta, \partial_{t} \Theta, \partial_{x}^{3} \Theta\right),\left(\partial_{x}^{2} \tilde{\eta}_{\bullet}, \partial_{x} \tilde{\eta}_{\bullet}\right), P\left(\partial_{x}^{2} \Theta,\left[\partial_{x} \Theta\right]^{-1}\right), u\right)
\end{aligned}
$$

By a straightforward estimation and by using (4.33), (4.34), and (4.38) we conclude that

$$
\left\|\partial_{\eta_{\bullet}} F_{u}(u, \pi, \Theta)\right\|_{\mathcal{B}\left(0 \mathbb{E}_{\Theta}(T) ; \mathbb{F}_{u}(T)\right)} \rightarrow 0 \quad \text { as } T \rightarrow 0
$$

4.2.2. The transformed divergence equation. We have transformed the equation

$$
\operatorname{div} u(t, x)=0 \quad \text { for } t \in J, x \in \Omega \backslash \Gamma(t)
$$

to the following equation for $(\bar{u}, \Theta)$ with $u \circ \tilde{\Theta}=\left[\partial_{x} \Theta\right] \bar{u}$ :

$$
\operatorname{div} \bar{u}=F_{d}(\bar{u}, \Theta) \quad \text { in } J \times(\Omega \backslash \Sigma)
$$

Here the perturbation $F_{d}(\bar{u}, \Theta): J \times \Omega \rightarrow \mathbb{R}$ is given by

$$
\begin{equation*}
F_{d}(\bar{u}, \Theta)(t, \bar{x})=-\partial_{l} \partial_{m} \Theta_{j}(t, \bar{x})\left[\partial_{x} \Theta(t, \bar{x})\right]_{m j}^{-1} \bar{u}_{l}(t, \bar{x})=M\left(\partial_{x}^{2} \Theta,\left[\partial_{x} \Theta\right]^{-1}, \bar{u}\right) \tag{4.44a}
\end{equation*}
$$

Again we replace $\bar{u}$ by $u$. Abels and Wilke [AW13] noticed that the identities $\operatorname{div} u=F_{d}(u, \Theta)$ and $\int_{\Omega} \operatorname{div} u d x=0$ imply that the integral $\int_{\Omega} F_{d}(u, \Theta) d x$ vanishes, but this might be false for arbitrary $u \in \mathbb{E}_{u}$. Therefore we replace $F_{d}(u, \Theta)$ by its part

$$
\begin{equation*}
\tilde{F}_{d}(u, \Theta):=F_{d}(u, \Theta)-\frac{1}{|\Omega|} \int_{\Omega} F_{d}(u, \Theta) d x \tag{4.44b}
\end{equation*}
$$

with vanishing mean value $|\Omega|^{-1} \int_{\Omega} \tilde{F}_{d}(u, \Theta) d x=0$. We will exploit the fact that $F_{d}(u, \Theta)$ is trilinear in $\left(u, \partial_{x}^{2} \Theta,\left[\partial_{x} \Theta\right]^{-1}\right)$, and with the embeddings in Lemma 4.22 we will show that $\tilde{F}_{d}$ is analytic and can be controlled in a similar way as $F_{u}$ (Lemma 4.23).

### 4.22. Lemma. The embedding

$$
\begin{equation*}
\mathbb{E}_{\Theta} \hookrightarrow H_{p}^{1}\left(0, T ; C^{2}(\bar{\Omega})\right)^{n} \cap C\left([0, T] ; H_{p}^{3}(\Omega)\right)^{n} \quad \text { if } p>n \tag{4.45}
\end{equation*}
$$

is continuous, and for some $\delta_{0}>0$ and all $\delta \in\left(0, \delta_{0}\right], T_{0}>0$, and $T \in\left(0, T_{0}\right]$, we have

$$
\begin{equation*}
\left\|\eta_{\bullet}\right\|_{0 H_{p}^{1}\left(0, T ; C^{2}(\bar{\Omega})\right) \cap_{0} C\left([0, T] ; H_{p}^{3}(\bar{\Omega})\right)} \leq T^{\delta} C\left(\delta, T_{0}\right)\left\|\eta_{\bullet}\right\|_{0 \mathbb{E}_{\Theta}(T)} \quad \text { if } p>n+2 \tag{4.46}
\end{equation*}
$$

Proof. Sobolev's embedding (B.1) yields (4.45). Next, Hölder's inequality yields

$$
\left\|\eta_{\bullet}\right\|_{0 C\left([0, T] ; H_{p}^{3}(\Omega)\right)} \leq T^{1-1 / p}\left\|\partial_{t} \eta_{\bullet}\right\|_{L_{p}\left(0, T ; H_{p}^{3}(\Omega)\right)} \leq T^{1-1 / p}\left\|\eta_{\bullet}\right\|_{0 \mathbb{E}_{\Theta}(T)}
$$

With the embeddings (B.3) and (B.1), the estimates (3.32b), (3.32c), and (3.32d), and the mixed derivative embeddings (Proposition B.44) we obtain

$$
\begin{aligned}
\left\|\eta_{\bullet}\right\|_{0 H_{p}^{1}\left(0, T ; C^{2}(\bar{\Omega})\right)} & \lesssim T^{1 / p+\varepsilon}\left\|\eta_{\bullet}\right\|_{0 W_{p}^{1+1 / p+\varepsilon}\left(0, T ; C^{2}(\bar{\Omega})\right)} \\
& \lesssim T^{1 / p+\varepsilon}\left\|\eta_{\bullet}\right\|_{0 H_{p}^{3 / 2-\rho}\left(0, T ; H_{p}^{2+2 \rho}(\Omega)\right)} \lesssim T^{1 / p+\varepsilon}\left\|\eta_{\bullet}\right\|_{0 \mathbb{E}_{\Theta}(T)},
\end{aligned}
$$

with suitable numbers $\varepsilon>0$ and $\rho \in(0,1 / 2)$ which exist if $p>n+2$. Thus (4.46) is valid.
4.23. Lemma. For $p \in(n, \infty)$ and $T \in(0, \infty)$, the map

$$
\tilde{F}_{d}:\left\{(u, \Theta) \in \mathbb{E}_{u}(T) \times \mathbb{E}_{\Theta}(T): \Theta \in \mathcal{U}_{\Theta}\right\} \rightarrow \mathbb{F}_{d}(T)
$$

is analytic. Assume in addition that $p>n+2$. Then the following assertions are valid:
(i) For given $T_{0} \in(0, \infty), R \in[1, \infty), u \in \mathbb{E}_{u}\left(T_{0}\right)$, and $\Theta \in \mathbb{E}_{\Theta}\left(T_{0}\right) \cap \mathcal{U}_{\Theta}$, we have

$$
\left\|\tilde{F}_{d}(u, \Theta)\right\|_{\mathbb{F}_{d}(T)} \rightarrow 0 \quad \text { as } T \rightarrow 0,\left\|\partial_{x} \Theta-I_{x}\right\|_{C\left([0, T] ; C^{1}(\bar{\Omega})\right)} \rightarrow 0,
$$

and this convergence is uniform with respect to

$$
\|u\|_{\mathbb{E}_{u}\left(T_{0}\right)}+\|\Theta\|_{\mathbb{E}_{\Theta}\left(T_{0}\right)}+\left\|\left[\partial_{x} \Theta\right]^{-1}\right\|_{C\left(\left[0, T_{0}\right] \times \bar{\Omega}\right)} \leq R .
$$

(ii) For given $T_{0} \in(0, \infty), R \in[1, \infty)$, $u_{*} \in \mathbb{E}_{u}\left(T_{0}\right), u_{\bullet} \in{ }_{0} \mathbb{E}_{u}\left(T_{0}\right), \Theta=\Theta_{*}+\eta_{\bullet} \in \mathbb{E}_{\Theta}\left(T_{0}\right) \cap \mathcal{U}_{\Theta}$ with $\Theta_{*} \in \mathbb{E}_{\Theta}\left(T_{0}\right)$, and $\eta_{\bullet} \in{ }_{0} \mathbb{E}_{\Theta}\left(T_{0}\right)$, we have

$$
\left\|\partial_{\left(u_{\bullet}, \eta_{\bullet}\right)} \tilde{F}_{d}\left(u_{*}+u_{\bullet}, \Theta_{*}+\eta_{\bullet}\right)\right\|_{\mathcal{B}\left(0 \mathbb{E}_{u}(T) \times_{0} \mathbb{E}_{\Theta}(T) ; \mathbb{F}_{d}(T)\right)} \rightarrow 0 \quad \text { as } T \rightarrow 0,\left\|\partial_{x}^{2} \Theta\right\|_{C\left([0, T] ; C^{1}(\bar{\Omega})\right)} \rightarrow 0,
$$

and this convergence is uniform with respect to

$$
\left\|\left(u_{*}, u_{\bullet}\right)\right\|_{\mathbb{E}_{u}\left(T_{0}\right)}+\left\|\left(\Theta_{*}, \eta_{\bullet}\right)\right\|_{\mathbb{E}_{\ominus}\left(T_{0}\right)}+\left\|\left[\partial_{x}\left(\Theta_{*}+\eta_{\bullet}\right)\right]^{-1}\right\|_{C\left(\left[0, T_{0}\right] \times \bar{\Omega}\right)} \leq R .
$$

Proof. The divergence theorem implies

$$
\int_{\Omega} \operatorname{div} f \varphi d x=-\int_{\Omega} f \cdot \nabla \varphi d x, \quad \text { for } \varphi \in \dot{H}_{p^{\prime}}^{1}(\Omega), f \in H_{p, 0}^{1}(\Omega) .
$$

Therefore we can extend the divergence operator to a bounded operator on $L_{p}(\Omega)$ such that

$$
\begin{equation*}
\|\operatorname{div} f\|_{\dot{H}_{p^{\prime}}^{1}(\Omega)^{*}} \leq\|f\|_{L_{p}(\Omega)} \quad \text { for } f \in L_{p}(\Omega) . \tag{4.47}
\end{equation*}
$$

For $f \in L_{p}(\Omega)$ and $\varphi \in C_{c}^{\infty}\left(\mathbb{R}^{n}\right)$, we have

$$
\int_{\Omega}\left(f-\langle f\rangle_{\Omega}\right) \varphi d x=\int_{\Omega}\left(f-\langle f\rangle_{\Omega}\right)\left(\varphi-\langle\varphi\rangle_{\Omega}\right) d x
$$

Hence the Poincaré-Wirtinger inequality $\left\|\varphi-\langle\varphi\rangle_{\Omega}\right\|_{p^{\prime}} \leq C_{P W}\|\nabla \varphi\|_{p^{\prime}}$ for $\varphi \in \dot{H}_{p^{\prime}}^{1}(\Omega)$ implies

$$
\begin{equation*}
\left\|f-\langle f\rangle_{\Omega}\right\|_{\dot{H}_{p^{\prime}}^{1}(\Omega)^{*}} \leq C_{P W}\left\|f-\langle f\rangle_{\Omega}\right\|_{p} \leq C_{P W}\left(1+|\Omega|^{-1 / p}\right)\|f\|_{p} \quad \text { for } f \in L_{p}(\Omega) \tag{4.48}
\end{equation*}
$$

The inequality (4.48) implies

$$
\begin{equation*}
\left\|f-\langle f\rangle_{\Omega}\right\|_{H_{p}^{1}\left(0, T ; \dot{H}_{p^{\prime}}^{1}(\Omega)^{*}\right)} \leq C_{P W}\left(1+|\Omega|^{-1 / p}\right)\|f\|_{H_{p}^{1}\left(0, T ; L_{p}(\Omega)\right)} . \tag{4.49}
\end{equation*}
$$

From the embedding $H_{p}^{1}(0, T ; X) \hookrightarrow C([0, T] ; X)$ we infer that pointwise multiplication

$$
\bullet: H_{p}^{1}\left(0, T ; L_{\infty}(\Omega)\right) \times H_{p}^{1}\left(0, T ; L_{p}(\Omega)\right) \rightarrow H_{p}^{1}\left(0, T ; L_{p}(\Omega)\right)
$$

is continuous and that $H_{p}^{1}\left(0, T ; L_{\infty}(\Omega)\right)$ is a multiplication algebra. Moreover,

$$
\bullet: L_{\infty}\left(0, T ; H_{p}^{1}(\Omega)\right) \times L_{p}\left(0, T ; H_{p}^{1}(\Omega)\right) \rightarrow L_{p}\left(0, T ; H_{p}^{1}(\Omega)\right)
$$

is continuous and $L_{\infty}\left(0, T ; H_{p}^{1}(\Omega)\right)$ is a multiplication algebra for $p>n$. Thus

$$
Y:=H_{p}^{1}\left(0, T ; L_{\infty}(\Omega)\right) \cap L_{\infty}\left(0, T ; H_{p}^{1}(\Omega)\right)
$$

also is a multiplication algebra, and from (4.49) we infer that

$$
\begin{equation*}
\left\|g f-\langle g f\rangle_{\Omega}\right\|_{\mathbb{F}_{d}} \lesssim\|g f\|_{H_{p}^{1}(J \times \Omega)} \lesssim\|g\|_{Y}\|f\|_{H_{p}^{1}(J \times \Omega)} . \tag{4.50}
\end{equation*}
$$

Estimate (4.50) and multiplication in $Y$ imply that the trilinear map

$$
(A, B, u) \mapsto a_{l m j} b_{m j} u_{l}-\left\langle a_{l m j} b_{m j} u_{l}\right\rangle_{\Omega}, \quad Y^{n \times n \times n} \times Y^{n \times n} \times H_{p}^{1}((0, T) \times \Omega)^{n} \rightarrow \mathbb{F}_{d}(T)
$$

is continuous. The map $\Theta \mapsto A:=\partial_{x}^{2} \Theta, \mathbb{E}_{\Theta} \rightarrow Y^{n \times n \times n}$ is linear and bounded by (4.45), and the $\operatorname{map} \Theta \mapsto B:=\left[\partial_{x} \Theta\right]^{-1}, \mathcal{U}_{\Theta} \rightarrow Y^{n \times n}$ is analytic by Proposition B.88. Therefore $\tilde{F}_{d}$ is analytic.
(i) With estimate (4.42) we obtain

$$
\begin{align*}
\left\|a_{l m j} b_{m j} u_{l}\right\|_{L_{p}\left(J ; H_{p}^{1}(\Omega)\right)} & \leq\|A\|_{Y}\|B\|_{Y} \cdot T^{1 / p} C\left(T_{0}\right)\|u\|_{\mathbb{E}_{u}\left(T_{0}\right)},  \tag{4.51a}\\
\left\|\partial_{t}\left(a_{l m j} b_{m j}\right) u_{l}\right\|_{L_{p}(J \times \Omega)} & \leq\left(\left\|A_{t}\right\|_{p}\|B\|_{\infty}+\|A\|_{\infty}\left\|B_{t}\right\|_{p}\right) \cdot T^{1 / p} C\left(T_{0}\right)\|u\|_{\mathbb{E}_{u}\left(T_{0}\right)},  \tag{4.51b}\\
\left\|a_{l m j} b_{m j} \partial_{t} u_{l}\right\|_{L_{p}(J \times \Omega)} & \leq\|A\|_{\infty}\|B\|_{\infty}\|u\|_{\mathbb{E}_{u}(T)} . \tag{4.51c}
\end{align*}
$$

Hence the first estimate in (4.50) yields the assertion.
(ii) With (4.44a) and (4.43) we obtain the partial Fréchet derivatives

$$
\begin{aligned}
& {\left[\partial_{u_{\bullet}} F_{d}\left(u_{*}+u_{\bullet}, \Theta_{*}+\eta_{\bullet}\right)\right] \tilde{u}_{\bullet}=F_{d}\left(\tilde{u}_{\bullet}, \Theta\right)=M\left(\partial_{x}^{2} \Theta,\left[\partial_{x} \Theta\right]^{-1}, \tilde{u}_{\bullet}\right),} \\
& {\left[\partial_{\eta_{\bullet}} F_{d}\left(u_{*}+u_{\bullet}, \Theta_{*}+\eta_{\bullet}\right)\right] \tilde{\eta}_{\bullet}=M\left(\partial_{x}^{2} \tilde{\eta}_{\bullet},\left[\partial_{x} \Theta\right]^{-1}, u\right)+M\left(\partial_{x}^{2} \Theta,\left(\left[\partial_{x} \Theta\right]^{-1}\right)^{2}, \partial_{x} \tilde{\eta}_{\bullet}, u\right) .}
\end{aligned}
$$

From (4.46) with $p>n+2$ we infer that

$$
\left\|\partial_{x}^{j} \eta_{\bullet}\right\|_{Y} \leq T^{\delta} C\left(\delta, T_{0}\right)\left\|\eta_{\bullet}\right\|_{0 \mathbb{E}_{\Theta}(T)} \text { for } j \in\{1,2\} .
$$

This estimate and (4.51) yield the assertion.

### 4.3. The transformed interface equations

In this section we transform the interface momentum balance

$$
\begin{equation*}
-\llbracket T(u, \pi) \rrbracket \nu_{\Gamma}-\operatorname{div}_{\Gamma} T_{\Gamma}(u)=0 \quad \text { on } \Gamma(t), t \in J, \tag{4.52}
\end{equation*}
$$

which was derived on page 19 . We assume that the unknown moving interface is represented as $\Gamma(t)=\Gamma_{h}(t)=\Theta_{h}(\{t\} \times \Sigma)$ in terms of the unknown height function $h$ and the normalpreserving map $\Theta_{h}$ from page 99 and Theorem 4.15.(ii). Our goal is to decompose (4.52) into a principal linear part and a remaining nonlinear part, and to handle the latter as a perturbation with respect to the function spaces on page 94 . An explicit description of these perturbations is given on the following page. As the main result of this section we prove that the nonlinear perturbations can be deemed as small as we wish provided that the time interval $J=(0, T)$ and the initial height function $h_{0}=\left.h\right|_{t=0}$ are sufficiently small (Lemma 4.26). In Corollaries 4.27 and 4.28 we prove the corresponding results for the transformed bulk equations by specializing Lemmas 4.21 and 4.23.

Let us take a closer look at (4.52). From the identities (1.19) and (1.22) we recall that

$$
\left\{\begin{align*}
T(u, \pi) & =2 \mu D(u)-\pi I,  \tag{4.53}\\
D(u) & =2^{-1}\left(\nabla u+[\nabla u]^{\top}\right), \\
T_{\Gamma}(u) & =\sigma P_{\Gamma}+\left(\lambda_{s}-\mu_{s}\right)\left(\operatorname{div}_{\Gamma} u\right) P_{\Gamma}+2 \mu_{s} D_{\Gamma}(u), \\
D_{\Gamma}(u) & =2^{-1} P_{\Gamma}\left(\nabla_{\Gamma} u+\left[\nabla_{\Gamma} u\right]^{\top}\right) P_{\Gamma} .
\end{align*}\right.
$$

Define a tangential vector field $N_{v}(u, \Gamma)$ and a scalar field $N_{w}(u, \pi, \Gamma)$ by

$$
\begin{aligned}
& N_{v}(u, \Gamma):=-P_{\Gamma} \llbracket T(u, 0) \rrbracket \nu_{\Gamma}-P_{\Gamma} \operatorname{div}_{\Gamma} T_{\Gamma}(u) \\
&=-\left(\llbracket \mu \underline{\left.\partial_{\nu} v \rrbracket+\llbracket \mu \rrbracket\left[\nabla_{\Gamma} v\right] \nu_{\Gamma}+\llbracket \mu \rrbracket \nabla_{\Gamma} w\right)}\right. \\
&-\left(\mu_{s} \widetilde{\Delta}_{\Gamma} v\right. \\
&\left.+\lambda_{s} \underline{\nabla_{\Gamma} \operatorname{div}_{\Gamma} v}+\left(\mu_{s}+\lambda_{s}\right) w \underline{\nabla_{\Gamma} H_{\Gamma}}+\left[\left(\mu_{s}-\lambda_{s}\right) H_{\Gamma}-2 \mu_{s} L_{\Gamma}\right] \nabla_{\Gamma} w\right), \\
& N_{w}(u, \pi, \Gamma):-\nu_{\Gamma} \cdot \llbracket T(u, \pi) \rrbracket \nu_{\Gamma}-\nu_{\Gamma} \cdot \operatorname{div}_{\Gamma} T_{\Gamma}(u) \\
&=-\left(2 \llbracket \mu \underline{\left.\partial_{\nu} w \rrbracket-\llbracket \pi \rrbracket\right)}\right. \\
&-\left(\sigma \underline{H_{\Gamma}}+\left(\lambda_{s}-\mu_{s}\right) \underline{\operatorname{div}_{\Gamma} u H_{\Gamma}}+2 \mu_{s} \underline{\left.D_{\Gamma}(u): L_{\Gamma}\right) .}\right.
\end{aligned}
$$

Here we have again used the decomposition $u=v+w \nu_{\Gamma}$ near $\Gamma$, and the underlined terms are considered as the principal part with respect to the chosen function spaces. Thus, the interface momentum balance (4.52) can be written as

$$
N_{v}(u, \Gamma)+N_{w}(u, \pi, \Gamma) \nu_{\Gamma}=0 \quad \text { on } \Gamma(t), t \in J
$$

For given

$$
z_{*}=\left(\bar{u}_{*}, \bar{\pi}_{*}, h_{*}\right) \in \tilde{\mathbb{E}}, \quad z_{\bullet}=\left(\bar{u}_{\bullet}, \bar{\pi}_{\bullet}, h_{\bullet}\right) \in_{0} \tilde{\mathbb{E}}, \quad \text { with } h=h_{*}+h_{\bullet} \in \mathcal{U}_{h},
$$

we define $\Theta_{h}$ as in Theorem 4.15, $\tilde{\Theta}_{h}(t, x):=\left(t, \Theta_{h}(t, x)\right), \Gamma_{h}(t):=\Theta_{h}(t, \Sigma)$, and

$$
\bar{u}:=\bar{u}_{*}+\bar{u}_{\bullet}, \quad u_{h}:=\left(\left[\partial_{x} \Theta_{h}\right] \bar{u}\right) \circ \tilde{\Theta}_{h}^{-1},
$$

and decompose $u_{h}=v_{h}+w_{h} \nu_{\Gamma_{h}}$ near $\Gamma_{h}$, and $\bar{u}_{*}=\bar{v}_{*}+\bar{w}_{*} \nu_{\Sigma}$ and $\bar{u}_{\bullet}=\bar{v}_{\bullet}+\bar{w}_{\bullet} \nu_{\Sigma}$ near $\Sigma$. Then the maps $G_{v}$ and $G_{w}$ from (4.57) have the representations

$$
\begin{equation*}
=\llbracket \mu\left\{\left[\partial_{x} \Theta_{h}\right]^{-1}\left(\partial_{\nu_{\Gamma_{h}}} v_{h}\right) \circ \tilde{\Theta}_{h}-\partial_{\nu} \bar{v}_{\bullet}\right\} \rrbracket \tag{4.55a}
\end{equation*}
$$

$$
\begin{equation*}
+\mu_{s}\left\{\left[\partial_{x} \Theta_{h}\right]^{-1}\left(\widetilde{\Delta}_{\Gamma_{h}} v_{h}\right) \circ \tilde{\Theta}_{h}-\widetilde{\Delta}_{\Sigma} \bar{v}_{\bullet}\right\} \tag{4.55b}
\end{equation*}
$$

$$
G_{v}\left(z_{\bullet} ; z_{*}\right)=G_{v}\left(\bar{u}_{\bullet}, h_{\bullet} ; \bar{u}_{*}, h_{*}\right)
$$

$$
\begin{equation*}
+\lambda_{s}\left\{\left[\partial_{x} \Theta_{h}\right]^{-1}\left(\nabla_{\Gamma_{h}} \operatorname{div}_{\Gamma_{h}} v_{h}\right) \circ \tilde{\Theta}_{h}-\nabla_{\Sigma} \operatorname{div}_{\Sigma} \bar{v}_{\bullet}\right\} \tag{4.55c}
\end{equation*}
$$

$$
\begin{equation*}
+\left(\lambda_{s}+\mu_{s}\right)\left\{\left[\partial_{x} \Theta_{h}\right]^{-1}\left(w_{h} \nabla_{\Gamma_{h}} H_{\Gamma_{h}}\right) \circ \tilde{\Theta}_{h}-\bar{w}_{*} \nabla_{\Sigma} \Delta_{\Sigma} h_{\bullet}\right\} \tag{4.55d}
\end{equation*}
$$

$$
\begin{equation*}
+\left[\partial_{x} \Theta_{h}\right]^{-1}\left(\left[\left(\mu_{s}-\lambda_{s}\right) H_{\Gamma_{h}}-2 \mu_{s} L_{\Gamma_{h}}\right] \nabla_{\Gamma_{h}}\left(w_{h}\right)\right) \circ \tilde{\Theta}_{h} \tag{4.55e}
\end{equation*}
$$

$$
\begin{equation*}
+\llbracket \mu \rrbracket\left[\partial_{x} \Theta_{h}\right]^{-1}\left(\left[\nabla_{\Gamma_{h}} v_{h}\right] \nu_{\Gamma_{h}}\right) \circ \tilde{\Theta}_{h} \tag{4.55f}
\end{equation*}
$$

$$
\begin{equation*}
+\llbracket \mu \rrbracket\left\{\left[\partial_{x} \Theta_{h}\right]^{-1}\left(\nabla_{\Gamma_{h}} w_{h}\right) \circ \tilde{\Theta}_{h}-\nabla_{\Sigma} \bar{w}_{\bullet}\right\} \tag{4.55~g}
\end{equation*}
$$

$$
\begin{aligned}
G_{w} & \left(z_{\bullet} ; z_{*}\right)=G_{w}\left(\bar{u}_{\bullet}, h_{\bullet} ; \bar{u}_{*}, \bar{\pi}_{*}, h_{*}\right) \\
= & 2 \llbracket \mu \partial_{\nu_{\Sigma}} \bar{w}_{*} \rrbracket+\llbracket \bar{\pi}_{*} \rrbracket \\
& +\sigma\left\{H_{\Gamma_{h}} \circ \tilde{\Theta}_{h}-\Delta_{\Sigma} h_{\bullet}\right\} \\
& +\left(\lambda_{s}-\mu_{s}\right)\left\{\left(\operatorname{div}_{\Gamma_{h}} u_{h} H_{\Gamma_{h}}\right) \circ \tilde{\Theta}_{h}-H_{\Sigma} \operatorname{div}_{\Sigma} \bar{v}_{\bullet}-\left(\operatorname{div}_{\Sigma} \bar{v}_{*}-2 H_{\Sigma} \bar{w}_{*}\right) \Delta_{\Sigma} h_{\bullet}\right\} \\
& +2 \mu_{s}\left\{\left(D_{\Gamma_{h}}\left(u_{h}\right): L_{\Gamma_{h}}\right) \circ \tilde{\Theta}_{h}-D_{\Sigma}(\bar{v}): L_{\Sigma}-\left[D_{\Sigma}\left(\bar{v}_{*}\right)-2 \bar{w}_{*} L_{\Sigma}\right]: \nabla_{\Sigma}^{2} h_{\bullet}\right\} .
\end{aligned}
$$

Figure 4.5. The perturbations $G_{v}$ and $G_{w}$.

Next, we derive the transformed version of (4.52). For given transformed functions $\bar{u} \in \mathbb{E}_{u}$, $\bar{\pi} \in \mathbb{E}_{\pi}$, and $h \in \mathbb{E}_{h} \cap \mathcal{U}_{h}$, and with $\tilde{\Theta}_{h}(t, x)=\left(t, \Theta_{h}(t, x)\right)$, we define

$$
\begin{aligned}
u_{h}: & :=\left(\left[\partial_{x} \Theta_{h}\right] \bar{u}\right) \circ \tilde{\Theta}_{h}^{-1}, & \bar{N}_{v}(\bar{u}, h) & :=\left[\partial_{x} \Theta_{h}\right]^{-1} N_{v}\left(u_{h}, \Gamma_{h}\right) \circ \tilde{\Theta}_{h}, \\
\pi_{h}: & =\bar{\pi} \circ \tilde{\Theta}_{h}^{-1}, & \bar{N}_{w}(\bar{u}, \bar{\pi}, h) & :=N_{w}\left(u_{h}, \pi_{h}, \Gamma_{h}\right) \circ \tilde{\Theta}_{h} .
\end{aligned}
$$

Then the transformed interface momentum balance is given by

$$
\begin{equation*}
\bar{N}_{v}(\bar{u}, h)+\bar{N}_{w}(\bar{u}, \bar{\pi}, h) \nu_{\Sigma}=0 \quad \text { on } J \times \Sigma . \tag{4.54}
\end{equation*}
$$

In order to resolve this interface condition, we decompose both $\bar{N}_{j}(j \in\{v, w\})$ into a principal linear part $L_{j}$ and a nonlinear perturbation $G_{j}$. By means of Lemma 4.16, it is straightforward to compute more explicit representations of $G_{v}$ and $G_{w}$, and we will employ the identities (4.55) and (4.56) in Figure 4.5 on the current page.

For controlling the perturbations $G_{v}$ and $G_{w}$ and for proving their analyticity, we first provide some estimates for the lower-order terms in Lemma 4.24. Then we study pointwise multiplication, inversion, and square root in the function spaces $\mathbb{G}_{v}$ and $\mathbb{G}_{w}$ in Lemma 4.25. It is sufficient to consider the larger class of height functions

$$
\tilde{\mathbb{E}}_{h}:=H_{p}^{1}\left(J ; W_{p}^{3-1 / p}(\Sigma)\right) \cap L_{p}\left(J ; W_{p}^{4-1 / p}(\Sigma)\right),
$$

which contains $\mathbb{E}_{h}$. Then $\left.\partial_{x} \Theta_{h}\right|_{\Sigma}$ belongs to the space

$$
\tilde{\mathbb{E}}_{\partial \Theta}:=H_{p}^{1}\left(J ; W_{p}^{2-1 / p}(\Sigma)\right) \cap L_{p}\left(J ; W_{p}^{3-1 / p}(\Sigma)\right),
$$

which satisfies $\mathbb{E}_{\partial \Theta} \hookrightarrow \tilde{\mathbb{E}}_{\partial \Theta} \hookrightarrow \mathbb{G}_{v} \cap \mathbb{G}_{w}$.
We will consider triples $z=z_{*}+z_{\bullet}$ of the form

$$
\begin{aligned}
& z_{*}=\left(\bar{u}_{*}, \bar{\pi}_{*}, h_{*}\right) \in \tilde{\mathbb{E}}\left(T_{0}\right):=\mathbb{E}_{u, v, w, \partial_{\nu} w}\left(T_{0}\right) \times \mathbb{E}_{\pi, \llbracket \pi \rrbracket}\left(T_{0}\right) \times \tilde{\mathbb{E}}_{h}\left(T_{0}\right), \\
& z_{\bullet}=\left(\bar{u}_{\bullet}, \bar{\pi}_{\bullet}, h_{\bullet}\right) \in_{0} \tilde{\mathbb{E}}(T):={ }_{0} \mathbb{E}_{u, v, w, \partial_{\nu} w}(T) \times{ }_{0} \mathbb{E}_{\pi, \llbracket \pi]}(T) \times{ }_{0} \tilde{\mathbb{E}}_{h}(T) .
\end{aligned}
$$

The operators $L_{j}$ are chosen as follows:

$$
\begin{aligned}
L_{v}\left(\bar{u}_{\bullet}, h_{\bullet} ; \bar{u}_{*}\right)= & -\mu_{s} \widetilde{\Delta}_{\Sigma} \bar{v}_{\bullet}-\lambda_{s} \nabla_{\Sigma} \operatorname{div}_{\Sigma} \bar{v}_{\bullet}-\llbracket \mu \partial_{\nu} \bar{v}_{\bullet} \rrbracket-\llbracket \mu \rrbracket \nabla_{\Sigma} \bar{w}_{\bullet}-\left(\lambda_{s}+\mu_{s}\right) \bar{w}_{*} \nabla_{\Sigma} \Delta_{\Sigma} h_{\bullet}, \\
L_{w}\left(\bar{u}_{\bullet}, \bar{\pi}_{\bullet}, h_{\bullet} ; \bar{u}_{*}\right)= & -\operatorname{tr}\left(\left[\left(\lambda_{s}-\mu_{s}\right) H_{\Sigma}+2 \mu_{s} L_{\Sigma}\right] \nabla_{\Sigma} \bar{v}_{\bullet}\right)-2 \llbracket \mu \partial_{\nu} \bar{w}_{\bullet} \rrbracket+\llbracket \bar{\pi}_{\bullet} \rrbracket \\
& -\operatorname{tr}\left(\left[\sigma+\left(\lambda_{s}-\mu_{s}\right)\left(\operatorname{div}_{\Sigma} \bar{v}_{*}-2 \bar{w}_{*} H_{\Sigma}\right)+2 \mu_{s}\left(D_{\Sigma}\left(\bar{v}_{*}\right)-2 \bar{w}_{*} L_{\Sigma}\right)\right] \nabla_{\Sigma}^{2} h_{\bullet}\right) .
\end{aligned}
$$

These operators are linear with respect to $z_{0}$. The nonlinear perturbations $G_{j}$ are given by

$$
\begin{align*}
G_{v}\left(\bar{u}_{\bullet}, h_{\bullet} ; \bar{u}_{*}, h_{*}\right) & :=L_{v}\left(\bar{u}_{\bullet}, h_{\bullet} ; \bar{u}_{*}\right)-\bar{N}_{v}\left(\bar{u}_{*}+\bar{u}_{\bullet}, h_{*}+h_{\bullet}\right),  \tag{4.57a}\\
G_{w}\left(\bar{u}_{\bullet}, h_{\bullet} ; \bar{u}_{*}, \bar{\pi}_{*}, h_{*}\right) & :=L_{w}\left(\bar{u}_{\bullet}, 0, h_{\bullet} ; \bar{u}_{*}\right)-\bar{N}_{w}\left(\bar{u}_{*}+\bar{u}_{\bullet}, \bar{\pi}_{*}, h_{*}+h_{\bullet}\right) .
\end{align*}
$$

Note that the right-hand side of (4.57b) satisfies

$$
\begin{aligned}
& L_{w}\left(\bar{u}_{\bullet}, \bar{\pi}_{\bullet}, h_{\bullet} ; \bar{u}_{*}\right)-\bar{N}_{w}\left(\bar{u}_{*}+\bar{u}_{\bullet}, \bar{\pi}_{*}+\bar{\pi}_{\bullet}, h_{*}+h_{\bullet}\right) \\
& =L_{w}\left(\bar{u}_{\bullet}, 0, h_{\bullet} ; \bar{u}_{*}\right)-\bar{N}_{w}\left(\bar{u}_{*}+\bar{u}_{\bullet}, \bar{\pi}_{*}, h_{*}+h_{\bullet}\right)
\end{aligned}
$$

and hence does not depend on $\bar{\pi}_{\bullet}$.
The lower-order terms of $G_{v}$ and $G_{w}$ will be controlled with the following estimates.
4.24. Lemma. Let $p \in(n+2, \infty)$.
(i) There exists $\delta_{0}>0$ such that for all $\delta \in\left(0, \delta_{0}\right], T_{0}>0$, and $T \in\left(0, T_{0}\right]$, we have

$$
\begin{align*}
& \left\|\left(h_{\bullet}, \nabla_{\Sigma} h_{\bullet}, \nabla_{\Sigma}^{2} h_{\bullet}\right)\right\|_{0 \mathbb{G}_{v}(T)} \leq C\left(\delta, T_{0}\right) T^{\delta}\left\|h_{\bullet}\right\|_{0 \tilde{\mathbb{E}}_{h}(T)},  \tag{4.58a}\\
& \left\|\left(h_{\bullet}, \nabla_{\Sigma} h_{\bullet}\right)\right\|_{0 \mathbb{G}_{w}(T)} \leq C\left(\delta, T_{0}\right) T^{\delta}\left\|h_{\bullet}\right\|_{0 \tilde{\mathbb{E}}_{h}(T)},  \tag{4.58b}\\
& \left\|\left(u_{\bullet}, \nabla_{\Sigma} u_{\bullet}\right)\right\|_{0_{\mathbb{G}_{v}(T)}} \leq C\left(\delta, T_{0}\right) T^{\delta}\left\|u_{\bullet}\right\|_{0 \mathbb{E}_{v}(T)},  \tag{4.58c}\\
& \left\|u_{\bullet}\right\|_{0 \mathbb{G}_{w}(T)} \leq C\left(\delta, T_{0}\right) T^{\delta}\left\|u_{\bullet}\right\|_{0 \mathbb{E}_{v}(T)}, \tag{4.58d}
\end{align*}
$$

for all $h_{\bullet} \in{ }_{0} \tilde{\mathbb{E}}_{h}(T)$ and all not necessarily tangential vector fields $u_{\bullet} \in{ }_{0} \mathbb{E}_{v}(T)$.
(ii) Given $\varepsilon>0$, there exists $\delta_{0}>0$ such that for all $\delta \in\left(0, \delta_{0}\right], T_{0}>0$, and $T \in\left(0, T_{0}\right]$, we have

$$
\begin{align*}
&\|h \bullet\|_{0 C\left([0, T] ; C^{3}(\Sigma)\right)} \leq C\left(\delta, T_{0}\right) T^{\delta}\left\|h_{\bullet}\right\|_{0 \tilde{\mathbb{E}}_{h}(T)},  \tag{4.59a}\\
&\left\|u_{\bullet}\right\|_{C\left([0, T] ; C^{1}(\Sigma)\right)} \leq C\left(\delta, T_{0}\right) T^{\delta}\left\|u_{\bullet}\right\|_{0 \mathbb{E}_{v}(T)}, \tag{4.59b}
\end{align*}
$$

for all $h_{\bullet} \in{ }_{0} \tilde{\mathbb{E}}_{h}(T)$ and $u_{\bullet} \in{ }_{0} \mathbb{E}_{v}(T)$, and

$$
\begin{align*}
\left\|\left(h, \nabla_{\Sigma} h\right)\right\|_{\mathbb{G}_{v}(T)} & \leq C\left(\delta, T_{0}\right)\left(T^{\delta}\|h\|_{\tilde{\mathbb{E}}_{h}(T)}+\left\|h_{0}\right\|_{W_{p}^{2-2 / p}(\Sigma)}\right),  \tag{4.60a}\\
\|h\|_{\left.C(0, T] ; C^{3}(\Sigma)\right)} & \leq C\left(\delta, T_{0}\right)\left(T^{\delta}\|h\|_{\tilde{\mathbb{E}}_{h}(T)}+\left\|h_{0}\right\|_{W_{p}^{3+(n-1) / p+\varepsilon}(\Sigma)}\right),  \tag{4.60b}\\
\left\|\left(h, \nabla_{\Sigma} h\right)\right\|_{\mathbb{G}_{w}(T)} & \leq C\left(\delta, T_{0}\right)\left(T^{\delta}\|h\|_{\tilde{\mathbb{E}}_{h}(T)}+\left\|h_{0}\right\|_{W_{p}^{3-2 / p}(\Sigma)}\right), \tag{4.60c}
\end{align*}
$$

for all $T \in\left(0, T_{0}\right]$ and $h \in \tilde{\mathbb{E}}_{h}(T)$ with $\left.h\right|_{t=0}=h_{0}$.
Proof. We proceed as in the proofs of Lemmas 3.19 and 4.20 and we also employ temporal extension operators of initial values from Corollaries B.26, B. 58 and B. 59 on pages 155, 163 and 164.
(i) For proving (4.58), we first observe that (with all spaces considered over $(0, T) \times \Sigma)$ )

$$
\left(h_{\bullet}, \nabla_{\Sigma} h_{\bullet}, \nabla_{\Sigma}^{2} h_{\bullet}\right) \in_{0} H_{p}^{1}\left(W_{p}^{1-1 / p}\right) \hookrightarrow_{0} W_{p}^{1 / 2-1 / 2 p}\left(L_{p}\right) \cap L_{p}\left(W_{p}^{1-1 / p}\right)={ }_{0} \mathbb{G}_{v},
$$

and estimate (4.58a) follows by using the inequalities (3.32b) and (3.32e). Similarly,

$$
\left(h_{\bullet}, \nabla_{\Sigma} h_{\bullet}\right) \in_{0} H_{p}^{1}\left(W_{p}^{2-1 / p}\right) \hookrightarrow_{0} W_{p}^{3 / 4-1 / 2 p}\left(L_{p}\right) \cap_{0} W_{p}^{1 / 2-1 / 2 p}\left(H_{p}^{1}\right) \cap L_{p}\left(W_{p}^{2-1 / p}\right)={ }_{0} \mathbb{G}_{w},
$$

and hence (4.58b) is valid. Next,

$$
\left(u_{\bullet}, \nabla_{\Sigma} u_{\bullet}\right) \in{ }_{0} W_{p}^{3 / 4-1 / 2 p}\left(L_{p}\right) \cap_{0} W_{p}^{1 / 2-1 / p}\left(H_{p}^{1}\right) \hookrightarrow{ }_{0} W_{p}^{1 / 2-1 / 2 p}\left(L_{p}\right) \cap L_{p}\left(W_{p}^{1-1 / p}\right)={ }_{0} \mathbb{G}_{v},
$$

and therefore estimate (4.58c) follows from Lemma 3.18. Similarly, (4.58d) follows from

$$
u_{\bullet} \in{ }_{0} \mathbb{E}_{v} \hookrightarrow{ }_{0} W_{p}^{1-1 / 2 p}\left(L_{p}\right) \cap_{0} W_{p}^{3 / 4-1 / 2 p}\left(H_{p}^{1}\right) \cap_{0} H_{p}^{1 / 2-1 / 4 p}\left(W_{p}^{2-1 / p}\right) \hookrightarrow{ }_{0} \mathbb{G}_{w} .
$$

(ii) Estimates (4.59) follow similarly, by using Sobolev's embedding (B.1).

In order to prove the estimates (4.60) for $h \in \tilde{\mathbb{E}}_{h}(T)$ we employ the decomposition

$$
h=\mathcal{E}_{T} h_{0}+\left(h-\mathcal{E}_{T} h_{0}\right), \quad\left(\mathcal{E}_{T} h_{0}\right)(t):=e^{-t \sqrt{\mu-\Delta_{\Sigma}}} h_{0}, \quad h_{0}:=\left.h\right|_{t=0} .
$$

From Corollaries B. 26 and B. 58 we infer that the realizations

$$
\begin{aligned}
& \mathcal{E}_{T}: W_{p}^{2-2 / p}(\Sigma) \rightarrow W_{p}^{1-1 / p}\left(0, T ; H_{p}^{1}(\Sigma)\right) \cap L_{p}\left(0, T ; W_{p}^{2-1 / p}(\Sigma)\right), \\
& \mathcal{E}_{T}: W_{p}^{4-2 / p}(\Sigma) \rightarrow \tilde{\mathbb{E}}_{h}(T)
\end{aligned}
$$

are bounded, uniformly with respect to $T \in\left(0, T_{0}\right]$. With estimate (4.58a) we obtain

$$
\begin{aligned}
\left\|\nabla_{\Sigma} h\right\|_{\mathbb{G}_{v}(T)} & \leq\left\|\nabla_{\Sigma}\left(h-\mathcal{E}_{T} h_{0}\right)\right\|_{0^{G_{v}}(T)}+\left\|\nabla_{\Sigma} \mathcal{E}_{T} h_{0}\right\|_{\mathbb{G}_{v}(T)} \\
& \leq C\left(\delta, T_{0}\right)\left(T^{\delta}\|h\|_{\tilde{\mathbb{E}}^{2}(T)}+\left\|h_{0}\right\|_{W_{p}^{2-2 / p}(\Sigma)}\right) .
\end{aligned}
$$

Therefore (4.60a) is valid. Next, the realization

$$
\mathcal{E}_{T}: W_{p}^{3+(n-1) / p+\varepsilon}(\Sigma) \rightarrow H_{p}^{1}\left(0, T ; W_{p}^{2+n / p+\varepsilon}(\Sigma)\right) \cap L_{p}\left(0, T ; W_{p}^{3+n / p+\varepsilon}(\Sigma)\right)
$$

is also bounded and its target space is embedded into $C\left([0, T] ; C^{3}(\Sigma)\right)$. This yields an estimate

$$
\begin{aligned}
\|h\|_{C\left([0, T] ; C^{3}(\Sigma)\right)} & \leq\left\|h-\mathcal{E}_{T} h_{0}\right\|_{0 C\left([0, T] ; C^{3}(\Sigma)\right)}+\left\|\mathcal{E}_{T} h_{0}\right\|_{C\left([0, T] ; C^{3}(\Sigma)\right)} \\
& \leq C\left(\delta, T_{0}\right)\left(T^{\delta}\|h\|_{\tilde{\mathbb{E}}_{h}(T)}+\left\|h_{0}\right\|_{W_{p}^{3+(n-1) / p+\varepsilon}(\Sigma)}\right),
\end{aligned}
$$

which proves (4.60b). With the boundedness of

$$
\mathcal{E}_{T}: W_{p}^{3-2 / p}(\Sigma) \rightarrow H_{p}^{1}\left(0, T ; W_{p}^{2-1 / p}(\Sigma)\right) \cap L_{p}\left(0, T ; W_{p}^{3-1 / p}(\Sigma)\right),
$$

and with estimate (4.58b) we obtain

$$
\begin{aligned}
\left\|\left(h, \nabla_{\Sigma} h\right)\right\|_{\mathbb{G}_{w}(T)} & \leq\left\|\left(1, \nabla_{\Sigma}\right)\left(h-\mathcal{E}_{T} h_{0}\right)\right\|_{0 \mathbb{G}_{w}(T)}+\left\|\left(1, \nabla_{\Sigma}\right) \mathcal{E}_{T} h_{0}\right\|_{\mathbb{G}_{w}(T)} \\
& \leq C\left(\delta, T_{0}\right)\left(T^{\delta}\|h\|_{\tilde{\mathbb{E}}_{h}(T)}+\left\|h_{0}\right\|_{W_{p}^{3-2 / p}(\Sigma)}\right)
\end{aligned}
$$

and thus (4.60c) is valid.
Next, we provide estimates for controlling products with leading-order terms. Let $X(T)$ denote the scalar version of one of the spaces $\tilde{\mathbb{E}}_{\partial \Theta}(T), \mathbb{G}_{v}(T)$, or $\mathbb{G}_{w}(T)$ from Figure 4.1 on page 94. Analogously as for $\mathbb{E}_{\partial \Theta}(T)$ in Lemma 4.13 , we will show that $X(T)$ is multiplication algebra, and that pointwise inversion and square root are analytic operators in suitable subsets of $X(T)$. We also consider certain larger spaces $Y(T)=C\left([0, T] ; C^{k}(\Sigma)\right) \supset X(T)$ with the property

$$
\|f\|_{0 Y(T)} \leq T^{\delta} C\left(\delta, T_{0}\right)\|f\|_{0 X(T)} \quad \text { for } f \in{ }_{0} X(T), T \in\left(0, T_{0}\right],
$$

where ${ }_{0} X(T):=\left\{f \in X(T):\left.f\right|_{t=0}=0\right\}$ and ${ }_{0} Y(T):=\left\{f \in Y(T):\left.f\right|_{t=0}=0\right\}$. Moreover, the temporal trace space $\gamma_{0} X$ of $X(T)$ is embedded into a larger space $Z$ for which we obtain a $T$-dependent estimate

$$
\|f\|_{X(T)} \leq T^{\delta} C\left(\delta, T_{0}\right)\left(\|f\|_{X(T)}+\left\|\left.f\right|_{t=0}\right\|_{\gamma_{0} X}\right)+C\left(T_{0}\right)\left\|\left.f\right|_{t=0}\right\|_{Z} \quad \text { for } f \in X(T), T \in\left(0, T_{0}\right] .
$$

Hence, together with a bilinear estimate

$$
\|f g\|_{X(T)} \leq C(T)\left(\|f\|_{X(T)}\|g\|_{Y(T)}+\|f\|_{Y(T)}\|g\|_{X(T)}\right),
$$

we can control $\|f g\|_{X(T)}$ by choosing $T,\left\|\left.f\right|_{t=0}\right\|_{Z}$, and $\left\|\left.g\right|_{t=0}\right\|_{Z}$ sufficiently small.
4.25. Lemma. Let $\Sigma \subset \mathbb{R}^{n}(n \geq 2)$ be a compact smooth hypersurface and let

$$
\begin{array}{rlll} 
& X(T)=\tilde{\mathbb{E}}_{\partial \Theta}(T), & Y(T)=C\left([0, T] ; C^{2}(\Sigma)\right), & Z=W_{p}^{2+(n-1) / p+\varepsilon}(\Sigma), \\
\text { or } & X(T)=\mathbb{G}_{v}(T), & Y(T)=C([0, T] ; C(\Sigma)), & Z=W_{p}^{(n-1) / p+\varepsilon}(\Sigma), \\
\text { or } & X(T)=\mathbb{G}_{w}(T), & Y(T)=C\left([0, T] ; C^{1}(\Sigma)\right), & Z=W_{p}^{1+(n-1) / p+\varepsilon}(\Sigma), \tag{4.61c}
\end{array}
$$

where $p \in(n+2, \infty), T \in(0, \infty)$, and $\varepsilon \in(0,1-(n+2) / p]$. Then the following assertions are valid:
(i) We have $X(T) \hookrightarrow Y(T)$, and for some $\delta_{0}>0$ and all $\delta \in\left(0, \delta_{0}\right], T_{0}>0$, and $T \in\left(0, T_{0}\right]$ we have

$$
\begin{equation*}
\|f\|_{Y(T)} \leq T^{\delta} C\left(\delta, T_{0}\right)\|f\|_{0 X(T)} \quad \text { for } f \in_{0} X(T)=\left\{f \in X(T):\left.f\right|_{t=0}=0\right\} \tag{4.62}
\end{equation*}
$$

(ii) For $\varepsilon \in(0,1-(n+2) / p]$ there is $\delta_{0}>0$ such that for all $\delta \in\left(0, \delta_{0}\right], T_{0}>0$, and $T \in\left(0, T_{0}\right]$ we have

$$
\begin{equation*}
\|f\|_{Y(T)} \leq T^{\delta} C\left(\delta, T_{0}\right)\left(\|f\|_{X(T)}+\left\|\left.f\right|_{t=0}\right\|_{\gamma_{0} X}\right)+C\left(T_{0}\right)\left\|\left.f\right|_{t=0}\right\|_{Z} \quad \text { for } f \in X(T) . \tag{4.63}
\end{equation*}
$$

(iii) $X(T)$ is a multiplication algebra, and there exists $C(T) \geq 1$ such that

$$
\begin{equation*}
\|f g\|_{X(T)} \leq C(T)\left(\|f\|_{X(T)}\|g\|_{Y(T)}+\|f\|_{Y(T)}\|g\|_{X(T)}\right) \quad \text { for } f, g \in X(T), \tag{4.64}
\end{equation*}
$$

and for given $T_{0} \in(0, \infty)$ there exists $C\left(T_{0}\right)$ such that for all $T \in\left(0, T_{0}\right]$ we have

$$
\begin{equation*}
\|f g\|_{0 X(T)} \leq C\left(T_{0}\right)\left(\|f\|_{0 X(T)}\|g\|_{Y\left(T_{0}\right)}+\|f\|_{0 Y(T)}\|g\|_{X\left(T_{0}\right)}\right) \quad \text { for } f \in{ }_{0} X(T), g \in X\left(T_{0}\right) . \tag{4.65}
\end{equation*}
$$

(iv) The inversion operator $A \mapsto A^{-1},\left\{A \in X^{m \times m}: \sup _{J \times \Sigma}\left|A^{-1}\right|<\infty\right\} \rightarrow X$ is analytic.
(v) The square root operator $f \mapsto f(\cdot)^{1 / 2},\left\{f \in X: \inf _{J \times \Sigma} \operatorname{dist}\left(f(\cdot), \mathbb{R}_{-}\right)>0\right\} \rightarrow X$ is analytic.

Proof. We only deal with (4.61a) since the remaining assertions can be proved analogously.
(i) We abbreviate $W_{p}^{t}\left(W_{p}^{s}\right):=W_{p}^{t}\left(J ; W_{p}^{s}(\Sigma)\right), C^{k}\left(C^{l}\right):=C^{k}\left(\bar{J} ; C^{l}(\Sigma)\right)$, and similarly for the other spaces. The mixed derivative embeddings and Sobolev's embedding (B.1) imply

$$
H_{p}^{1}\left(W_{p}^{2-1 / p}\right) \cap L_{p}\left(W_{p}^{3-1 / p}\right) \hookrightarrow W_{p}^{\theta}\left(H_{p}^{3-1 / p-\theta}\right) \hookrightarrow C\left(C^{2}\right),
$$

provided that $\theta \in(0,1)$ satisfies $\theta>1 / p$ and $3-1 / p-\theta>2+(n-1) / p$. Such a number $\theta$ exists if $1 / p<1-n / p$, and this is true if $p>n+1$. Moreover, Lemma 3.18 yields the estimate (4.62).
(ii) By Corollaries B. 26 and B. 58 on pages 155 and 163, the extension operator

$$
R_{A}: x \mapsto\left[t \mapsto e^{-A t} x\right], \quad W_{p}^{s+2-2 / p}(\Sigma) \rightarrow H_{p}^{1}\left(0, T ; W_{p}^{s}(\Sigma)\right) \cap L_{p}\left(0, T ; W_{p}^{s+2}(\Sigma)\right)
$$

for $A=1-\Delta_{\Sigma}$ and $s \in[0, \infty)$ is uniformly bounded with respect to $T \in\left(0, T_{0}\right]$. By decomposing $f(t)=\left(f(t)-e^{-t A} f(0)\right)+e^{-t A} f(0)$ and applying (4.62), we obtain (4.63).
(iii) The norm of $X(T)$ consists of the semi-norms

$$
\llbracket\left(\nabla_{\Sigma}, \partial_{t} \nabla_{\Sigma}, \nabla_{\Sigma}^{2}\right) \cdot \rrbracket_{p ; 1-1 / p, p}, \quad\left\|\left(1, \partial_{t}, \partial_{t} \nabla_{\Sigma}, \nabla_{\Sigma}, \nabla_{\Sigma}^{2}\right) \cdot\right\|_{p},
$$

where $\llbracket \cdot \rrbracket_{p ; \sigma, p}:=\llbracket \cdot \rrbracket_{L_{p}\left(W_{p}^{\sigma}\right)}$ and $\|\cdot\|_{p}:=\|\cdot\|_{L_{p}\left(L_{p}\right)}$. With Lemma B. 10 we control the leading-order terms of $\|f g\|_{X}$ by

$$
\begin{aligned}
\llbracket \partial_{t} \nabla_{\Sigma} f g \rrbracket_{p ; 1-1 / p, p} & \lesssim \llbracket \partial_{t} \nabla_{\Sigma} f \rrbracket_{p ; 1-1 / p, p}\|g\|_{\infty}+\left\|\partial_{t} \nabla_{\Sigma} f\right\|_{p}\left\|\left(g, \nabla_{\Sigma} g\right)\right\|_{\infty}, \\
\llbracket \partial_{t} f \nabla_{\Sigma} g \rrbracket_{p ; 1-1 / p, p} & \lesssim \llbracket \partial_{t} f \rrbracket_{p ; 1-1 / p, p}\left\|\nabla_{\Sigma} g\right\|_{\infty}+\left\|\partial_{t} f\right\|_{p}\left\|\left(\nabla_{\Sigma} g, \nabla_{\Sigma}^{2} g\right)\right\|_{\infty}, \\
\llbracket \nabla_{\Sigma}^{2} f g \rrbracket_{p ; 1-1 / p, p} & \lesssim \llbracket \nabla_{\Sigma}^{2} f \rrbracket_{p ; 1-1 / p, p}\|g\|_{\infty}+\left\|\nabla_{\Sigma}^{2} f\right\|_{p}\left\|\left(g, \nabla_{\Sigma} g\right)\right\|_{\infty} .
\end{aligned}
$$

These terms and the remaining ones can be estimated by the right-hand side of (4.64). Therefore (4.64) is valid and the uniform estimate (4.65) follows by means of extension ${ }_{0} X(T) \rightarrow{ }_{0} X(\infty)$ and restriction ${ }_{0} X(\infty) \rightarrow{ }_{0} X\left(T_{0}\right)$, where the temporal extension operator ${ }_{0} X(T) \rightarrow{ }_{0} X(\infty)$ from Lemma B. 9 on page 148 is uniformly bounded with respect to $T \in\left(0, T_{0}\right]$.
(iv) Let us check that $A^{-1}$ belongs to $X^{m \times m}$ for every $A \in X^{m \times m}$ with $A^{-1} \in C(\bar{J} \times$ $\Sigma$ ). From the inequality $\left|A(t, x)^{-1}-A\left(t, x^{\prime}\right)^{-1}\right| \leq\left\|A^{-1}\right\|_{\infty}^{2}\left|A(t, x)-A\left(t, x^{\prime}\right)\right|$ we infer that $\llbracket A^{-1} \rrbracket_{p ; \sigma, p} \lesssim\left\|A^{-1}\right\|_{\infty}^{2} \llbracket A \rrbracket_{p ; \sigma, p}<\infty$ for $\sigma=1-1 / p$. For given $j \in\{1, \ldots, n-1\}$, Lemma B. 10 yields

$$
\begin{aligned}
& \llbracket \partial_{t} \partial_{j} A^{-1} \rrbracket_{p ; \sigma, p}=\left[\left[A^{-1}\left[\partial_{j} A\right] A^{-1}\left[\partial_{t} A\right] A^{-1}+A^{-1}\left[\partial_{t} A\right] A^{-1}\left[\partial_{j} A\right] A^{-1}-A^{-1}\left[\partial_{t} \partial_{j} A\right] A^{-1}\right]\right]_{p ; \sigma, p} \\
& \lesssim\left\|A^{-1}\right\|_{\infty}^{3}\left\|\partial_{j} A\right\|_{\infty} \llbracket \partial_{t} A \rrbracket_{p ; \sigma, p}+\left(\left\|A^{-1}\right\|_{\infty}^{4}\left\|\left(A, \nabla_{\Sigma} A\right)\right\|_{\infty}^{2}+\left\|A^{-1}\right\|_{\infty}^{3} \llbracket\left(\nabla_{\Sigma} A, \nabla_{\Sigma}^{2} A\right) \rrbracket_{\infty}\right)\left\|\partial_{t} A\right\|_{p} \\
& \quad+\left\|A^{-1}\right\|_{\infty}^{2} \llbracket \partial_{t} \partial_{j} A \rrbracket_{p ; \sigma, p}+\left\|\left(A^{-1}, \nabla_{\Sigma} A^{-1}\right)\right\|_{\infty}\left\|A^{-1}\right\|_{\infty}\left\|\partial_{t} \partial_{j} A\right\|_{p}<\infty .
\end{aligned}
$$

The semi-norm $\llbracket \partial_{j} \partial_{k} A^{-1} \rrbracket_{p ; \sigma, p}$ can be estimated analogously. We further have

$$
\left\|A^{-1}\right\|_{p} \leq T^{1 / p}|\Sigma|^{1 / p}\left\|A^{-1}\right\|_{\infty}<\infty
$$

and the remaining terms in $\left\|A^{-1}\right\|_{X}$ are also finite. Therefore $A^{-1}$ belongs to $X^{m \times m}$, and then Proposition B. 88 on page 172 yields analyticity of the inversion operator $A \mapsto A^{-1}$.
(v) Assertion (v) follows by a similar proof as on page 100.

We are ready to control the perturbations $G_{v}$ and $G_{w}$. The triple $z_{\bullet}=\left(\bar{u}_{\bullet}, \bar{\pi}_{\bullet}, h_{\bullet}\right) \in_{0} \tilde{\mathbb{E}}(T)$ has vanishing initial values, and $z_{*}=\left(\bar{u}_{*}, \bar{\pi}_{*}, h_{*}\right) \in \tilde{\mathbb{E}}\left(T_{0}\right)$ should satisfy the compatibility conditions

$$
\begin{equation*}
\left.G_{j}\left(0 ; z_{*}\right)\right|_{t=0}=0 \quad \text { for } j \in\{v, w\} . \tag{4.66}
\end{equation*}
$$

Then we can control $G_{j}\left(z_{\bullet} ; z_{*}\right)$ in ${ }_{0} \mathbb{G}_{j}(T)$ by choosing $T \in\left(0, T_{0}\right]$ and $h_{*} \mid t=0$ sufficiently small. Even without requiring (4.66) we can control the partial Fréchet derivative

$$
\partial_{z_{\bullet}} G_{j}\left(z_{\bullet} ; z_{*}\right) \in \mathcal{B}\left(0 \tilde{\mathbb{E}}(T) ;{ }_{0} \mathbb{G}_{j}(T)\right) .
$$

4.26. Lemma. Let $p \in(n+2, \infty), \tau \in(1+n / p, 4-1 / p], T_{0} \in(0, \infty), T \in\left(0, T_{0}\right]$, and

$$
\mathcal{U}_{h}=\left\{h \in L_{\infty}\left(0, T ; W_{p}^{\tau-1 / p}(\Sigma)\right):\|h\|_{L_{\infty}\left(0, T ; W_{p}^{\tau-1 / p}(\Sigma)\right)}<\delta_{h}\right\}
$$

with $\delta_{h}(\Omega, \Sigma, p, \tau)>0$ as in Theorem 4.15. Then the maps

$$
\begin{array}{ll}
\left(z_{\bullet}, z_{*}\right) \mapsto L_{v}\left(\bar{u}_{\bullet}, h_{\bullet} ; u_{*}\right), & { }_{0} \tilde{\mathbb{E}}(T) \times \tilde{\mathbb{E}}\left(T_{0}\right) \rightarrow{ }_{0} \mathbb{G}_{v}(T), \\
\left(z_{\bullet}, z_{*}\right) \mapsto L_{w}\left(\bar{u}_{\bullet}, \bar{\pi}_{\bullet}, h_{\bullet} ; u_{*}\right), & \\
\left(z_{\bullet}, z_{*}\right) \mapsto G_{v}(T) \times \tilde{\mathbb{E}}\left(\bar{u}_{\bullet}, h_{\bullet} ; \bar{u}_{*}, h_{*}\right), & \left\{\left(z_{\bullet} ; z_{*}\right) \in{ }_{0} \tilde{0_{0}}(T) \times \tilde{\mathbb{E}}\left(\mathbb{G}_{w}(T),\right.\right. \\
\left.\left.\left(z_{\bullet}, z_{*}\right) \mapsto h_{\bullet}\right) \mapsto h_{w}\left(h_{\bullet}, h_{\bullet} ; \bar{U}_{h}\right\} \rightarrow \bar{u}_{*}, \bar{\pi}_{*}, h_{*}\right), & \left\{\left(z_{\bullet} ; z_{*}\right) \in{ }_{0} \tilde{\mathbb{E}}(T) \times \tilde{\mathbb{E}}\left(T_{0}\right): h_{\bullet}+h_{*} \in \mathcal{U}_{h}\right\} \rightarrow \mathbb{G}_{w}(T)
\end{array}
$$

are analytic and depend polynomially on $z_{\bullet}, z_{*}, \partial_{x} \Theta_{h_{*}+h_{\bullet},}\left[\partial_{x} \Theta_{h_{*}+h_{\bullet}}\right]^{-1}, \beta_{h_{*}+h_{\bullet}}$, and $\beta_{h_{*}+h_{\bullet}}^{-1}$.
In addition, let $\tau \in(3+n / p, 4-1 / p)$. Then $G_{v}$ and $G_{w}$ have the following properties:
(i) For $z_{\bullet} \in{ }_{0} \tilde{\mathbb{E}}\left(T_{0}\right), z_{*} \in \tilde{\mathbb{E}}\left(T_{0}\right)$ with $h=h_{*}+h_{\bullet} \in \mathcal{U}_{h}\left(T_{0}\right)$ and (4.66), we have

$$
\left\|G_{v}\left(z_{\bullet} ; z_{*}\right)\right\|_{\mathbb{G}_{v}(T)}+\left\|G_{w}\left(z_{\bullet} ; z_{*}\right)\right\|_{\mathbb{G}_{w}(T)} \rightarrow 0 \quad \text { as } T \rightarrow 0, h_{0}=\left.h_{*}\right|_{t=0} \rightarrow 0 \text { in } W_{p}^{\tau-1 / p}(\Sigma)
$$

Given $R \in[1, \infty)$, this convergence is uniform with respect to

$$
\begin{equation*}
\left\|\left(z_{\bullet}, z_{*}\right)\right\|_{\tilde{\mathbb{E}}\left(T_{0}\right)}+\left\|\left(\partial_{x} \Theta_{h_{*}+h_{\bullet}},\left[\partial_{x} \Theta_{h_{*}+h_{\bullet}}\right]^{-1}\right)\right\|_{\tilde{\mathbb{E}}_{\partial \Theta}\left(T_{0}\right)}+\left\|\left(\beta_{h_{*}+h_{\bullet},}, \beta_{h_{*}+h_{\bullet}}^{-1}\right)\right\|_{\tilde{\mathbb{E}}_{\partial \Theta}\left(T_{0}\right)} \leq R . \tag{4.67}
\end{equation*}
$$

(ii) For given $T_{0} \in(0, \infty), z_{\bullet} \in{ }_{0} \tilde{\mathbb{E}}\left(T_{0}\right)$, and $z_{*} \in \tilde{\mathbb{E}}\left(T_{0}\right)$ with $h=h_{*}+h_{\bullet} \in \mathcal{U}_{h}\left(T_{0}\right)$, we have

$$
\begin{equation*}
\left\|\partial_{z_{\bullet}} G_{v}\left(z_{\bullet} ; z_{*}\right)\right\|_{0 \tilde{\mathbb{E}}(T) \rightarrow 0} \mathbb{G}_{v}(T)+\left\|\partial_{z_{\bullet}} G_{w}\left(z_{\bullet} ; z_{*}\right)\right\|_{0 \tilde{\mathbb{E}}(T) \rightarrow 0} \mathbb{G}_{w}(T) \rightarrow 0, \tag{4.68}
\end{equation*}
$$

as $T \rightarrow 0, h_{0} \rightarrow 0$ in $W_{p}^{\tau-1 / p}(\Sigma)$. Given $R \geq 1$, this convergence is uniform with respect to (4.67).
Proof. Analyticity. We first note that the scalar-valued versions of the spaces $\mathbb{G}_{v}$ and $\mathbb{G}_{w}$ are multiplication algebras by Lemma 4.25. The maps $L_{v}$ and $L_{w}$ consist of linear and bilinear differential operators, and hence their analyticity follows from the mixed derivative embeddings and the spatial trace theorem. In order to prove the analyticity of $G_{v}$ and $G_{w}$, it is sufficient to prove that the map

$$
(\bar{u}, \bar{\pi}, h) \mapsto\left(\bar{N}_{v}(\bar{u}, h), \bar{N}_{w}(\bar{u}, \bar{\pi}, h)\right): \mathbb{E}_{u, v, w, \partial_{\nu} w} \times \mathbb{E}_{\pi, \llbracket \pi \rrbracket} \times \tilde{\mathbb{E}}_{h} \cap \mathcal{U}_{h} \rightarrow \mathbb{G}_{v} \times \mathbb{G}_{w}
$$

is analytic. Theorem 4.15, the identities (4.24) and Lemma 4.25 imply that the quantities

$$
\beta_{h}, \beta_{h}^{-1}, \nu_{\Gamma_{h}} \circ \tilde{\Theta}_{h},\left.\left[\partial_{x} \Theta_{h}\right]\right|_{\Sigma},\left.\left[\partial_{x} \Theta_{h}\right]^{-1}\right|_{\Sigma}, \tau_{j}^{\Gamma_{h}} \circ \tilde{\Theta}_{h}, \tau_{\Gamma_{h}}^{j} \circ \tilde{\Theta}_{h}
$$

considered in $\tilde{\mathbb{E}}_{\partial \Theta}$ depend analytically on $h \in \tilde{\mathbb{E}}_{h} \cap \mathcal{U}_{h}$. Next, the Weingarten tensor $L_{\Gamma}$ and the mean curvature $H_{\Gamma}$ are given by

$$
\begin{equation*}
L_{\Gamma}=-\nabla_{\Gamma} \nu_{\Gamma}=-\tau_{\Gamma}^{j} \otimes \partial_{j} \nu_{\Gamma}=l_{j k}^{\Gamma} \tau_{\Gamma}^{j} \otimes \tau_{\Gamma}^{k}=l_{\Gamma}^{j k} \tau_{j}^{\Gamma} \otimes \tau_{k}^{\Gamma}, \quad H_{\Gamma}=\operatorname{tr} L_{\Gamma}=-\operatorname{div}_{\Gamma} \nu_{\Gamma} . \tag{4.69}
\end{equation*}
$$

Therefore the maps $h \mapsto L_{\Gamma_{h}} \circ \tilde{\Theta}_{h}, \tilde{\mathbb{E}}_{h} \cap \mathcal{U}_{h} \rightarrow\left(\mathbb{G}_{v} \cap \mathbb{G}_{w}\right)^{n \times n}$ and $h \mapsto \nabla_{\Gamma_{h}} H_{\Gamma_{h}} \circ \tilde{\Theta}_{h}, \tilde{\mathbb{E}}_{h} \cap \mathcal{U}_{h} \mapsto \mathbb{G}_{v}^{n}$ are analytic, and their values depend polynomially on $\left(\partial_{x} \Theta_{h},\left[\partial_{x} \Theta_{h}\right]^{-1}, \beta_{h}, \beta_{h}^{-1}\right)$. Lemma 4.16 yields

$$
\begin{equation*}
v_{h} \circ \tilde{\Theta}_{h}=\left[\partial_{x} \Theta_{h}\right] \bar{v}, \quad w_{h} \circ \tilde{\Theta}_{h}=\beta_{h} \bar{w}, \quad\left(\partial_{\nu_{\Gamma_{h}}} w_{h}\right) \circ \tilde{\Theta}_{h}=\partial_{\nu_{\Sigma}} \bar{w}, \tag{4.70}
\end{equation*}
$$

and we conclude that, given $h \in \tilde{\mathbb{E}}_{h} \cap \mathcal{U}_{h}$, the map $\bar{u} \mapsto \bar{N}_{v}(\bar{u}, h), \mathbb{E}_{u, v, w} \rightarrow \mathbb{G}_{v}$ is linear and continuous, and $(\bar{u}, \bar{\pi}) \mapsto \bar{N}_{w}(\bar{u}, \bar{\pi}, h), \mathbb{E}_{u, v, w, \partial_{\nu} w} \times \mathbb{E}_{\pi, \llbracket \pi \rrbracket} \rightarrow \mathbb{G}_{w}$ is affine and continuous. Therefore $\bar{N}_{v}$ and $\bar{N}_{w}$ are analytic and depend polynomially on $\left(\bar{u}, \bar{\pi}, h, \partial_{x} \Theta_{h},\left[\partial_{x} \Theta_{h}\right]^{-1}, \beta_{h}, \beta_{h}^{-1}\right)$.
(i) Smallness of $G_{j}\left(z_{\bullet} ; z_{*}\right)$. With $\left.G_{j}\left(0 ; z_{*}\right)\right|_{t=0}=0$ we rewrite

$$
\begin{equation*}
G_{j}\left(z_{\bullet} ; z_{*}\right)=G_{j}\left(z_{\bullet} ; z_{*}\right)-R_{j}\left(\left.G_{j}\left(0 ; z_{*}\right)\right|_{t=0}\right), \tag{4.71}
\end{equation*}
$$

where we employ the temporal extension operators

$$
R_{v}: g_{v 0} \mapsto\left[t \mapsto e^{-t\left(1-\widetilde{\Delta}_{\Sigma}\right)} g_{v 0}\right], \quad R_{w}: g_{w 0} \mapsto\left[t \mapsto e^{-\left(1-\Delta_{\Sigma}\right)} g_{w 0}\right]
$$

from Corollaries B.26, B. 58 and B.59. Then we can rewrite the representations (4.55) and (4.56) for $G_{j}$ in such a way that every difference has a vanishing initial value. For instance, with $h_{0}=\left.h_{*}\right|_{t=0}$, the first difference (4.55a) in $G_{v}\left(z_{\bullet} ; z_{*}\right)$ becomes

$$
\left[\left[\mu\left\{\left[\partial_{x} \Theta_{h}\right]^{-1}\left(\partial_{\nu_{\Gamma_{h}}} v_{h}\right) \circ \tilde{\Theta}_{h}-\partial_{\nu_{\Sigma}} \bar{v}_{\bullet}-R_{v}\left(\left[\partial_{x} \Theta_{h_{0}}\right]^{-1}\left(\partial_{\nu_{\Gamma_{h_{0}}}} v_{h_{0}}\right) \circ \Theta_{h_{0}}\right)\right\}\right]\right] .
$$

With (4.24) and (4.25) we rewrite and decompose the difference in the curled brackets as

$$
\begin{aligned}
& {\left[\partial_{x} \Theta_{h}\right]^{-1} \partial_{x}\left(\left[\partial_{x} \Theta_{h}\right]\left(\bar{v}_{*}+\bar{v}_{\bullet}\right)\right)\left[\partial_{x} \Theta_{h}\right]^{-1}\left(\beta_{h}^{-1}\left[\partial_{x} \Theta_{h}\right] \nu_{\Sigma}\right)} \\
& \quad-\partial_{\nu_{\Sigma}} \bar{v}_{\bullet}-R_{v}\left(\left[\partial_{x} \Theta_{h_{0}}\right]^{-1}\left(\partial_{\nu_{\Gamma_{h_{0}}}} v_{h_{0}}\right) \circ \Theta_{h_{0}}\right) \\
& =\partial_{\nu_{\Sigma}} \bar{v}_{\bullet}\left(\beta_{h}^{-1}-1\right) \\
& \quad+\beta_{h}^{-1} \partial_{\nu_{\Sigma}} \bar{v}_{*}-R_{v}\left(\beta_{h_{0}}^{-1} \partial_{\nu_{\Sigma}} \bar{v}_{0}\right) \\
& \left.\quad+\beta_{h}^{-1}\left[\partial_{x} \Theta_{h}\right]^{-1}\left[\partial_{x}\left[\partial_{x} \Theta_{h}\right]\left(\bar{v}_{*}+\bar{v}_{\bullet}\right)\right]\left[\partial_{x} \Theta_{h}\right] \nu_{\Sigma}\right]-R_{v}\left(\beta_{h_{0}}\left[\partial_{x} \Theta_{h_{0}}\right]^{-1}\left[\partial_{x}\left[\partial_{x} \Theta_{h_{0}}\right] \bar{v}_{0}\right]\left[\partial_{x} \Theta_{h_{0}}\right] \nu_{\Sigma}\right) .
\end{aligned}
$$

These differences belong to ${ }_{0} \mathbb{G}_{v}(T)$, and from the estimates (4.63) and (4.65) we infer that they tend to zero in ${ }_{0} \mathbb{G}_{v}(T)$ as $T \rightarrow 0$ and $\left[\partial_{x} \Theta_{h_{0}}\right]^{-1} \rightarrow I_{x}$ in $W_{p}^{2+(n+1) / p+\varepsilon}(\Sigma)$ for some $\varepsilon \in(0,1-$ $(n+1) / p]$. The latter follows from $h_{0} \rightarrow 0$ in $W_{p}^{\tau-1 / p}(\Sigma)$ since $\tau-1 / p \geq 3+(n+1) / p+\varepsilon$ for some $\varepsilon>0$. The remaining differences in (4.71) can be estimated similarly, and therefore assertion (i) is valid.
(ii) Smallness of $\partial_{\bar{u}_{\bullet}} G_{v}$. For proving estimate (4.68) we first investigate the directional derivative $\partial_{\bar{u}} G_{v}\left(\bar{u}_{\bullet}, h_{\bullet} ; \bar{u}_{*}, h_{*}\right)$ applied to $\overline{\bar{u}}_{\bullet} \in{ }_{0} \mathbb{E}_{u, v, w, \partial_{\nu} w}(T)$. The map $\tilde{\bar{u}}_{\bullet} \mapsto G_{v}\left(\bar{u}_{\bullet}+\tilde{\bar{u}}_{\bullet}, h_{\bullet} ; \bar{u}_{*}, h_{*}\right)$ is affine, and therefore $G_{v}$ satisfies

$$
\left[\partial_{\bar{u}_{\bullet}} G_{v}\left(\bar{u}_{\bullet}, h_{\bullet} ; \bar{u}_{*}, h_{*}\right)\right] \tilde{\bar{u}}_{\bullet}=G_{v}\left(\tilde{u}_{\bullet}, h_{\bullet} ; 0, h_{*}\right),
$$

and has the lower-order terms (4.55d) to $(4.55 \mathrm{~g})$ with respect to $\bar{u}=\bar{u}_{*}+\bar{u}_{\bullet}$. Their directional derivatives with respect to $\bar{u}_{\bullet}$ applied to $\tilde{\bar{u}}_{\bullet}$ only depend on the values of $\tilde{\bar{u}}_{\bullet}$ and $\left.\nabla_{\Sigma} \tilde{\bar{u}}_{\bullet}\right|_{\Sigma}$, and with estimate (4.58c) we can control these terms by choosing $T$ small. Applying the identities in Figure 4.2, the leading-order terms in the $\bar{u}_{\bullet}$-derivatives of (4.55a) to (4.55c) are given by

$$
\left[\partial_{x} \tilde{\bar{v}}_{\bullet}\right]\left\{\beta_{h}^{-1}\left[\partial_{x} \Theta_{h}\right] \nu_{\Sigma}-\nu_{\Sigma}\right\},\left\{\left(g_{\Gamma_{h}}^{i j} \circ \tilde{\Theta}_{h}\right)-g_{\Sigma}^{i j}\right\} \partial_{i} \partial_{j} \tilde{\bar{v}}_{\bullet},\left\{\left[\tau_{\Gamma_{h}}^{i} \otimes \tau_{\Gamma_{h}}^{j}\right] \circ \tilde{\Theta}_{h}-\left[\tau_{\Sigma}^{i} \otimes \tau_{\Sigma}^{j}\right]\right\} \partial_{i} \partial_{j} \tilde{\bar{v}}_{\bullet} .
$$

By means of estimate (4.60a), we can further control $\left(h, \nabla_{\Sigma} h\right)$ in the $\mathbb{G}_{v}(T)$-norm and obtain

$$
\left\|\left(\beta_{h}-1, \beta_{h}^{-1}-1\right)\right\|_{\mathbb{G}_{v}(T)}+\left\|\left(\left[\partial_{x} \Theta_{h}\right]-I_{x},\left[\partial_{x} \Theta_{h}\right]^{-1}-I_{x}\right)\right\|_{\mathbb{G}_{v}(T)} \rightarrow 0
$$

as $T \rightarrow 0, h_{0} \rightarrow 0$ in $W_{p}^{2-2 / p}(\Sigma)$, and therefore

$$
\left\|\partial_{\bar{u}_{\bullet}} G_{v}\left(\bar{u}_{\bullet}, h_{\bullet} ; \bar{u}_{*}, h_{*}\right)\right\|_{0 \mathbb{E}_{u, v, w, \partial_{\nu w}}(T) \rightarrow 0} \mathbb{G}_{v}(T) \rightarrow 0 \quad \text { as } T \rightarrow 0, h_{0} \rightarrow 0 \text { in } W_{p}^{2-2 / p}(\Sigma),
$$

uniformly with respect to (4.67).
Smallness of $\partial_{\bar{u}_{\bullet}} G_{w}$. For the computation of $\left.\partial_{\bar{u}_{\bullet}} G_{w}\left(\bar{u}_{\bullet}, h_{\bullet} ; \bar{u}_{*}, \bar{\pi}_{*}, h_{*}\right)\right] \overline{u_{\bullet}}$ we only have to consider the differences (4.56c) and (4.56d) where $\left.\nabla_{\Sigma} \tilde{\bar{u}}_{\bullet}\right|_{\Sigma}$ and $\nabla_{\Sigma}^{2} h$ are of leading order. The lower-order terms can be controlled with estimate (4.58d). Concerning the leading-order terms we note that by using the identities (4.7), (4.53), and (4.69), it remains to control the products $\left[\nabla_{\Sigma} \tilde{\tilde{v}}_{\bullet}\right]\left[\nabla_{\Sigma}^{2} h\right]$ and $\left[\nabla_{\Sigma} \tilde{\tilde{v}}_{\bullet}\right] \nabla_{\Sigma} h$ in the ${ }_{0} \mathbb{G}_{w}(T)$-norm. From Lemma 4.25 we infer that

$$
\left\|\left[\nabla_{\Sigma} \bar{v}\right]\left(\left[\nabla_{\Sigma}^{2} h\right], \nabla_{\Sigma} h\right)\right\|_{0 \mathbb{G}_{w}(T)} \lesssim\left\|\nabla_{\Sigma} \bar{v}\right\|_{0 C\left(\bar{J} ; C^{1}(\Sigma)\right)}\|h\|_{\tilde{\mathbb{E}}_{h}\left(T_{0}\right)}+\|\bar{v}\|_{0 \mathbb{E}_{v}(T)}\|h\|_{C\left(\bar{J} ; C^{3}(\Sigma)\right)},
$$

and therefore, by using the estimates (4.59a) and (4.60b), we obtain

$$
\left\|\partial_{\bar{u}} G_{w}\left(\bar{u}_{\bullet}, h_{\bullet} ; \bar{u}_{*}, \bar{\pi}_{*}, h_{*}\right)\right\|_{0 \mathbb{E}_{u, v, w}(T) \rightarrow 0} \mathbb{G}_{v}(T) \rightarrow 0 \quad \text { as } T \rightarrow 0, h_{0} \rightarrow 0 \text { in } W_{p}^{\tau-1 / p}(\Sigma),
$$

for some $\tau>3+n / p$, and this convergence is uniform with respect to (4.67).
Smallness of $\partial_{h_{\bullet}} G_{v}$. We control $\left[\partial_{h_{\bullet}} G_{v}\left(\bar{u}_{\bullet}, h_{\bullet} ; \bar{u}_{*}, h_{*}\right)\right] \tilde{h}_{\bullet}$ in the ${ }_{0} \mathbb{G}_{v}(T)$-norm for $\tilde{h}_{\bullet} \in{ }_{0} \tilde{\mathbb{E}}_{h}(T)$. Estimate (4.58a) allows to control all the terms in $\left[\partial_{h_{\bullet}} G_{v}\left(\bar{u}_{\bullet}, h_{\bullet} ; \bar{u}_{*}, h_{*}\right)\right] \tilde{h}_{\bullet}$ by $T$, except for

$$
\left[\partial_{h_{\bullet}}\left(\left[\partial_{x} \Theta_{h}\right]^{-1}\left(w_{h} \nabla_{\Gamma_{h}} H_{\Gamma_{h}}\right) \circ \tilde{\Theta}_{h}\right)\right] \tilde{h}_{\bullet}-\bar{w} \nabla_{\Sigma} \Delta_{\Sigma} \tilde{h}_{\bullet}, \quad h:=h_{*}+h_{\bullet},
$$

which contains the leading-order term

$$
\left[\partial_{x} \Theta_{h}\right]^{-1} \beta_{h} \bar{w}\left[\partial_{x} \Theta_{h}\right]^{-\top} \nabla_{\Sigma}\left(\left[\partial_{x} \Theta_{h}\right]^{-\top} \tau_{\Sigma}^{j} \mid \partial_{j}\left(\beta_{h}\left[I-h L_{\Sigma}\right]^{-1} \nabla_{\Sigma} \tilde{h}_{\bullet}\right)\right)-\bar{w} \nabla_{\Sigma} \operatorname{div}_{\Sigma} \nabla_{\Sigma} \tilde{h}_{\bullet}
$$

In order to control this term in ${ }_{0} \mathbb{G}_{v}(T)$, it suffices to control $\partial_{x} \Theta_{h}-I_{x}$ and thus $\left(h, \nabla_{\Sigma} h\right)$ in $\mathbb{G}_{v}(T)$, and that was already done in (4.60a). We conclude that

$$
\left\|\partial_{h_{\bullet}} G_{v}\left(\bar{u}_{\bullet}, h_{\bullet} ; \bar{u}_{*}, h_{*}\right)\right\|_{0 \tilde{\mathbb{E}}_{h}(T) \rightarrow_{0} \mathbb{G}_{v}(T)} \rightarrow 0 \quad \text { as } T \rightarrow 0, h_{0} \rightarrow 0 \text { in } W_{p}^{1+(n-1) / p+\varepsilon}(\Sigma),
$$

for some $\varepsilon>0$, uniformly with respect to (4.67).
Smallness of $\partial_{h_{\bullet}} G_{w}$. It remains to control $\left[\partial_{h_{\bullet}} G_{w}\left(\bar{u}_{\bullet}, h_{\bullet} ; \bar{u}_{*}, \bar{\pi}_{*}, h_{*}\right)\right] \tilde{h}_{\bullet}$ in ${ }_{0} \mathbb{G}_{w}(T)$. All its summands which only contain ( $\tilde{h}_{\bullet}, \nabla_{\Sigma} \tilde{h}_{\bullet}$ ) but not $\nabla_{\Sigma}^{2} \tilde{h}_{\mathbf{\bullet}}$ can be controlled by $T$ with estimate (4.58b). With estimate (4.60c) we can also control all terms which contain $\left(h, \nabla_{\Sigma} h\right)$ but not $\nabla_{\Sigma}^{2} h$. Furthermore, with the estimates (4.60b) and (4.64), we can also control the bilinear leadingorder term

$$
\begin{aligned}
\left\|\left[\nabla_{\Sigma}^{2} h\right]\left[\nabla_{\Sigma}^{2} \tilde{h}_{\bullet}\right]\right\|_{0 \mathbb{G}_{w}(T)} & \lesssim\left\|\nabla_{\Sigma}^{2} h\right\|_{\mathbb{G}_{w}\left(T_{0}\right)}\left\|\nabla_{\Sigma}^{2} \tilde{h}_{\bullet}\right\|_{0 C\left([0, T] ; C^{1}\right)}+\left\|\nabla_{\Sigma}^{2} h\right\|_{C\left(\left[0, T_{0}\right] ; C^{1}\right)}\left\|\nabla_{\Sigma}^{2} \tilde{h}_{\bullet}\right\|_{0 \mathbb{G}_{w}(T)} \\
& \lesssim\|h\|_{\tilde{\mathbb{E}}_{h}\left(T_{0}\right)} \cdot T^{\delta}\left\|\tilde{h}_{\bullet}\right\|_{0 \tilde{\mathbb{E}}_{h}(T)}+\left\|h_{0}\right\|_{W_{p}^{3+(n-1) / p+\varepsilon}}\left\|\tilde{h}_{\bullet}\right\|_{0} \tilde{\mathbb{E}}_{h}(T)
\end{aligned}
$$

Among the leading-order terms, we consider the directional derivative

$$
\left[\partial_{h_{\bullet}}\left(H_{\Gamma_{h}} \circ \tilde{\Theta}_{h}\right)\right] \tilde{h}_{\bullet}=-\left[\partial_{h_{\bullet}}\left(\left(\tau_{\Gamma_{h}}^{j} \mid \partial_{j} \nu_{\Gamma_{h}}\right) \circ \tilde{\Theta}_{h}\right)\right] \tilde{h}_{\bullet}, \quad h:=h_{*}+h_{\bullet}
$$

Its leading-order part containing $\nabla_{\Sigma}^{2} \tilde{h}_{\bullet}$ is given by

$$
\left[\partial_{x} \Theta_{h}\right]^{-\top} \tau_{\Sigma}^{j} \cdot\left(\beta_{h}\left[I-h L_{\Sigma}\right]^{-1} \partial_{j} \nabla_{\Sigma} \tilde{h}_{\bullet}+\partial_{j}\left(\left[\partial_{h_{\bullet}} \beta_{h}\right] \tilde{h}_{\bullet}\right)\left[I-h L_{\Sigma}\right]^{-1} \nabla_{\Sigma} h\right) .
$$

With estimate (4.60c) we can estimate the first summand by

$$
\left\|\tilde{h}_{\bullet} \mapsto\left(\left[\partial_{x} \Theta_{h}\right]^{-\top} \tau_{\Sigma}^{j} \cdot\left(\left[I-h L_{\Sigma}\right]^{-1} \partial_{j} \nabla_{\Sigma} \tilde{h}_{\bullet}\right)-\Delta_{\Sigma} \tilde{h}_{\bullet}\right)\right\|_{0 \tilde{\mathbb{E}}_{h}(T) \rightarrow 0 \mathbb{G}_{w}(T)} \rightarrow 0,
$$

as $T \rightarrow 0$ and $h_{0} \rightarrow 0$ in $W_{p}^{3-2 / p}(\Sigma)$. For the second summand we use

$$
\left[\partial_{h_{\bullet}} \beta_{h}\right] \tilde{h}_{\bullet}=-\beta_{h}^{2}\left(\nabla_{\Sigma} h \mid\left(I-h L_{\Sigma}\right)^{-2}\left(\nabla_{\Sigma} \tilde{h}_{\bullet}-\tilde{h}_{\bullet} L_{\Sigma}\left(I-h L_{\Sigma}\right)^{-1} \nabla_{\Sigma} h\right)\right)
$$

and therefore the estimates (4.60b) and (4.60c) yield $\partial_{j}\left(\left[\partial_{h_{\bullet}} \beta_{h}\right] \tilde{h}_{\bullet}\right) \rightarrow 0$ in ${ }_{0} \mathbb{G}_{w}(T)$ as $T \rightarrow 0$ and $h_{0} \rightarrow 0$ in $W_{p}^{3+(n-1) / p+\varepsilon}(\Sigma)$. Therefore

$$
\begin{equation*}
\left\|\left[\partial_{h_{\bullet}}\left(L_{\Gamma_{h}} \circ \tilde{\Theta}_{h}\right)\right] \tilde{h}_{\bullet}-\nabla_{\Sigma}^{2} \tilde{h}_{\bullet}\right\|_{0 \mathbb{G}_{w}(T)} \rightarrow 0, \tag{4.72}
\end{equation*}
$$

as $T \rightarrow 0$ and $h_{0} \rightarrow 0$ in $W_{p}^{3+(n-1) / p+\varepsilon}(\Sigma)$. This allows to control the directional derivative of (4.56b). Concerning the remaining terms (4.56c) and (4.56d), we note that (see (A.17) on page 140)

$$
\operatorname{div}_{\Gamma} u=\left(\tau_{\Gamma}^{j} \mid \partial_{j} v\right)-w H_{\Gamma}, \quad D_{\Gamma}(u)=D_{\Gamma}(v)-w L_{\Gamma}=\operatorname{sym}\left(\tau_{\Gamma}^{j} \otimes P_{\Gamma} \partial_{j} v\right)-w L_{\Gamma} .
$$

Therefore it is sufficient to consider the differences

$$
\begin{align*}
& {\left[\partial_{h \bullet}\left(\left(\operatorname{div}_{\Gamma_{h}}\left(u_{h}\right)\right) \circ \tilde{\Theta}_{h}\right)\right] \tilde{h}_{\bullet} H_{\Gamma_{h}} \circ \tilde{\Theta}_{h}+\bar{w} \Delta_{\Sigma} \tilde{h}_{\bullet} H_{\Sigma},}  \tag{4.73a}\\
& \operatorname{div}_{\Gamma_{h}}\left(u_{h}\right) \circ \tilde{\Theta}_{h}\left[\partial_{h \bullet}\left(H_{\Gamma_{h}} \circ \tilde{\Theta}_{h}\right)\right] \tilde{h}_{\bullet}-\operatorname{div}_{\Sigma}(\bar{u}) \Delta_{\Sigma} \tilde{h}_{\bullet},  \tag{4.73b}\\
& {\left[\partial_{h \bullet}\left(\left(D_{\Gamma_{h}}\left(u_{h}\right) \circ \tilde{\Theta}_{h}\right)\right] \tilde{h}_{\bullet}:\left[L_{\Gamma_{h}} \circ \tilde{\Theta}_{h}\right]+\bar{w} \nabla_{\Sigma}^{2} \tilde{h}_{\bullet}: L_{\Sigma},\right.}  \tag{4.73c}\\
& {\left[D_{\Gamma_{h}}\left(u_{h}\right) \circ \tilde{\Theta}_{h}\right]:\left[\partial_{h \bullet}\left(L_{\Gamma_{h}} \circ \tilde{\Theta}_{h}\right)\right] \tilde{h}_{\bullet}-D_{\Sigma}(\bar{u}): \nabla_{\Sigma}^{2} \tilde{h}_{\bullet} .} \tag{4.73d}
\end{align*}
$$

With the estimates (4.60b), (4.60c), (4.64), and (4.72), we can control the directional derivatives of (4.73) in ${ }_{0} \mathbb{G}_{w}(T)$ with $T \rightarrow 0$ and $h_{0} \rightarrow 0$ in $W_{p}^{3+(n-1) / p+\varepsilon}(\Sigma)$. The proof of the lemma is complete.

We also have to specialize the corresponding results for $F_{u}$ and $F_{d}$ to the case of a normalpreserving admissible map. Theorem 4.15 implies that the map

$$
h \mapsto \Theta_{h}-\operatorname{Id}_{x}=\mathfrak{S}\left(h \nu_{\Sigma}, g_{h}\right), \quad \mathbb{E}_{h} \cap \mathcal{U}_{h} \rightarrow \mathbb{E}_{\Theta} \cap \mathcal{U}_{\Theta}
$$

is analytic for $p \in((n+2) / 2, \infty)$ and $\tau \in(1+n / p, 4-1 / p]$, where the subsets $\mathcal{U}_{h}$ and $\mathcal{U}_{\Theta}$ were defined on pages 102 and 107. For $h=h_{*}+h_{\bullet}$ with $h_{*} \in \mathbb{E}\left(T_{0}\right)$ and $h_{\bullet} \in{ }_{0} \mathbb{E}\left(T_{0}\right)$, the Fréchet derivative of $\Theta_{h}-\mathrm{Id}_{x}$ is given by

$$
\left[\partial_{h}\left(\Theta_{h}-\operatorname{Id}_{x}\right)\right] \tilde{h}_{\bullet}=\mathfrak{S}\left(\tilde{h}_{\bullet} \nu_{\Sigma},\left[\partial_{h_{\bullet}}\left(g_{h}\right)\right] \tilde{h}_{\bullet}\right),
$$

and becomes $\mathfrak{S}\left(\tilde{h}_{\bullet} \nu_{\Sigma},-\nabla_{\Sigma} \tilde{h}_{\bullet}\right)$ at $h=0$.
4.27. Corollary. Let $p \in(n+2, \infty) \backslash\{3\}, \tau \in(1+n / p, 4-1 / p), T \in(0, \infty)$, and

$$
F_{u}(u, \pi, h):=F_{u}\left(u, \pi, \Theta_{h}\right) \quad \text { for } u \in \mathbb{E}_{u}, \pi \in \mathbb{E}_{\pi}, h \in \mathbb{E}_{h} \cap \mathcal{U}_{h} .
$$

Then $F_{u}:\left\{(u, \pi, h) \in \mathbb{E}_{u} \times \mathbb{E}_{\pi} \times \mathbb{E}_{h}: h \in \mathcal{U}_{h}\right\} \rightarrow \mathbb{F}_{u}$ is analytic and has the following properties:
(i) Given $T_{0} \in(0, \infty)$, $R \in\left(\delta_{h}^{-1}, \infty\right)$, $u \in \mathbb{E}_{u}\left(T_{0}\right)$, $\pi \in \mathbb{E}_{\pi}\left(T_{0}\right)$, and $h \in \mathbb{E}_{h}\left(T_{0}\right) \cap \mathcal{U}_{h}$, we have

$$
\left\|F_{u}(u, \pi, h)\right\|_{\mathbb{F}_{u}(T)} \rightarrow 0 \quad \text { as } T \rightarrow 0, h_{0}:=\left.h\right|_{t=0} \rightarrow 0 \text { in } W_{p}^{\tau-1 / p}(\Sigma),
$$

uniformly with respect to

$$
\|u\|_{\mathbb{E}_{u}\left(T_{0}\right)}+\|\pi\|_{\mathbb{E}_{\pi}\left(T_{0}\right)}+\|h\|_{\mathbb{E}_{h}\left(T_{0}\right)} \leq R, \quad\|h\|_{L_{\infty}\left(0, T_{0} ; W_{p}^{\tau-1 / p}(\Sigma)\right)} \leq \delta_{h}-R^{-1} .
$$

(ii) Given $T_{0} \in(0, \infty), R \in\left(\delta_{h}^{-1}, \infty\right), u_{*} \in \mathbb{E}_{u}\left(T_{0}\right), u_{\bullet} \in{ }_{0} \mathbb{E}_{u}\left(T_{0}\right), \pi \in \mathbb{E}_{\pi}\left(T_{0}\right), h_{*} \in \mathbb{E}_{h}\left(T_{0}\right)$, and $h_{\bullet} \in{ }_{0} \mathbb{E}_{h}\left(T_{0}\right)$ with $h=h_{*}+h_{\bullet} \in \mathcal{U}_{h}$, we have

$$
\left\|\partial_{\left(u_{\bullet}, \pi, h_{\bullet}\right)} F_{u}\left(u_{*}+u_{\bullet}, \pi, h_{*}+h_{\bullet}\right)\right\|_{0 \mathbb{E}_{u}(T) \times \mathbb{E}_{\pi}(T) \times_{0} \mathbb{E}_{h}(T) \rightarrow \mathbb{F}_{u}(T)} \rightarrow 0,
$$

as $T \rightarrow 0$ and $h_{0} \rightarrow 0$ in $W_{p}^{\tau-1 / p}(\Sigma)$. This convergence is uniform with respect to

$$
\begin{aligned}
\left\|\left(u_{*}, u_{\bullet}\right)\right\|_{\mathbb{E}_{u}\left(T_{0}\right)}+\|\pi\|_{\mathbb{E}_{\pi}\left(T_{0}\right)}+\left\|\left(h_{*}, h_{\bullet}\right)\right\|_{\mathbb{E}_{h}\left(T_{0}\right)} & \leq R, \\
\left\|h_{*}+h_{\bullet}\right\|_{L_{\infty}\left(0, T_{0} ; W_{p}^{\tau-1 / p}(\Sigma)\right)} & \leq \delta_{h}-R^{-1} .
\end{aligned}
$$

Proof. In order to apply Theorem 4.15 and Lemma 4.21, it remains to show that $\left\|\partial_{x} \Theta_{h}-I_{x}\right\|_{\infty}=$ $\left\|\mathfrak{S}\left(h \nu \Sigma, g_{h}\right)\right\|_{\infty} \rightarrow 0$ as $T \rightarrow 0$ and $h_{0} \rightarrow 0$ in $W_{p}^{\tau-1 / p}(\Sigma)$. Given $\tau>1+n / p$ we have

$$
\begin{aligned}
\left\|\partial_{x} \Theta_{h}-I_{x}\right\|_{L_{\infty}((0, T) \times \Omega)} & =\left\|\mathfrak{S}\left(h \nu_{\Sigma}, g_{h}\right)\right\|_{L_{\infty}((0, T) \times \Omega)} \\
& \lesssim\left\|h \nu_{\Sigma}\right\|_{L_{\infty}\left(0, T ; W_{p}^{\tau-1 / p}(\Sigma)\right)}+\left\|g_{h}\right\|_{L_{\infty}\left(0, T ; W_{p}^{\tau-1-1 / p}(\Sigma)\right)} \rightarrow 0,
\end{aligned}
$$

as $\|h\|_{L_{\infty}\left(0, T ; W_{p}^{\tau-1 / p}(\Sigma)\right)} \rightarrow 0$. As on page 114, we decompose $h=\left(h-\mathcal{E}_{T} h_{0}\right)+\mathcal{E}_{T} h_{0}$. The embedding

$$
C\left(W_{p}^{\tau-1 / p}\right) \supset H_{p}^{1 / p+\varepsilon}\left(W_{p}^{\tau-1 / p}\right) \supset H_{p}^{1}\left(W_{p}^{\tau-1+\varepsilon}\right) \cap L_{p}\left(W_{p}^{\tau+\varepsilon}\right) \supset \mathbb{E}_{h},
$$

and Lemma 3.18 yield an estimate

$$
\begin{equation*}
\|h\|_{C\left([0, T], W_{p}^{\tau-1 / p}(\Sigma)\right)} \leq C\left(\delta, T_{0}\right)\left(T^{\delta}\|h\|_{\mathbb{E}_{h}\left(T_{0}\right)}+\left\|h_{0}\right\|_{W_{p}^{\tau-1 / p}(\Sigma)}\right), \tag{4.74}
\end{equation*}
$$

for some $\delta_{0}>0$ and all $\delta \in\left(0, \delta_{0}\right]$ and $T \in\left(0, T_{0}\right]$, provided that $\tau<4-1 / p$. This yields the required convergence $\|h\|_{L_{\infty}\left(0, T ; W_{p}^{\tau-1 / p}(\Sigma)\right)} \rightarrow 0$ as $T \rightarrow 0$ and $h_{0} \rightarrow 0$ in $W_{p}^{\tau-1 / p}(\Sigma)$.

For an estimation of the divergence perturbation $\tilde{F}_{d}$ we also use the compatibility condition

$$
\begin{equation*}
\operatorname{div} u_{0}=\tilde{F}_{d}\left(u_{0}, h_{0}\right) \quad \text { in } \Omega . \tag{4.75}
\end{equation*}
$$

4.28. Corollary. Let $p \in(n+2, \infty), \tau \in(3+n / p, 4-1 / p), T \in(0, \infty)$, and

$$
\tilde{F}_{d}(u, h):=\tilde{F}_{d}\left(u, \Theta_{h}\right) \quad \text { for } u \in \mathbb{E}_{u, v, w, \partial_{\nu} w}, h \in \mathbb{E}_{h} \cap \mathcal{U}_{h} .
$$

Then the map $\tilde{F}_{d}:\left\{(u, h) \in \mathbb{E}_{u, v, w, \partial_{\nu} w}(T) \times \mathbb{E}_{h}(T): h \in \mathcal{U}_{h}\right\} \rightarrow \mathbb{F}_{d, \Sigma}(T)$ is analytic and has the following properties:
(i) For given $T_{0} \in(0, \infty), R \in\left(\delta_{h}^{-1}, \infty\right), u \in \mathbb{E}_{u, v, w, \partial_{\nu} w}\left(T_{0}\right), h \in \mathbb{E}\left(T_{0}\right) \cap \mathcal{U}_{h}$, and $h_{0}=\left.h\right|_{t=0}$, we have

$$
\left\|\tilde{F}_{d}(u, h)-\operatorname{div} u\right\|_{\mathbb{F}_{d, \Sigma}(T)} \rightarrow 0 \quad \text { as } T \rightarrow 0,\left\|h_{0}\right\|_{W_{p}^{\tau-1 / p}(\Sigma)} \rightarrow 0
$$

and this convergence is uniform with respect to

$$
\|u\|_{\mathbb{E}_{u, v, w, \partial_{\nu} w}\left(T_{0}\right)}+\|h\|_{\mathbb{E}_{h}\left(T_{0}\right)} \leq R, \quad\|h\|_{L_{\infty}\left(0, T_{0} ; W_{p}^{\tau-1 / p}(\Sigma)\right)} \leq \delta_{h}-R^{-1} .
$$

(ii) For given $T_{0} \in(0, \infty), R \in\left(\delta_{h}^{-1}, \infty\right), u_{*} \in \mathbb{E}_{u, v, w, \partial_{\nu} w}\left(T_{0}\right), u_{\bullet} \in{ }_{0} \mathbb{E}_{u, v, w, \partial_{\nu} w}\left(T_{0}\right), h_{*} \in$ $\mathbb{E}\left(T_{0}\right), h_{\bullet} \in{ }_{0} \mathbb{E}_{h}\left(T_{0}\right)$ with $h=h_{*}+h_{\bullet} \in \mathcal{U}_{h}$, and $h_{0}=h_{*} \mid t=0$, we have

$$
\left\|\partial_{\left(u_{\bullet}, h_{\bullet}\right)} \tilde{F}_{d}\left(u_{*}+u_{\bullet}, h_{*}+h_{\bullet}\right)\right\|_{0 \mathbb{E}_{u, v, w, \partial_{\nu} w}(T) \times_{0} \mathbb{E}_{h}(T) \rightarrow \mathbb{F}_{d, \Sigma}(T)} \rightarrow 0,
$$

as $T \rightarrow 0$ and $\left\|h_{0}\right\|_{W_{p}^{\tau-1 / p}(\Sigma)} \rightarrow 0$, and this convergence is uniform with respect to

$$
\begin{align*}
\left\|\left(u_{*}, u_{\bullet}\right)\right\|_{\mathbb{E}_{u, v, w, \partial_{\nu} w}\left(T_{0}\right)}+\left\|\left(h_{*}, h_{\bullet}\right)\right\|_{\mathbb{E}_{h}\left(T_{0}\right)} & \leq R,  \tag{4.76}\\
\left\|h_{*}+h_{\bullet}\right\|_{L_{\infty}\left(0, T_{0} ; W_{p}^{\tau-1 / p}(\Sigma)\right)} & \leq \delta_{h}-R^{-1} . \tag{4.77}
\end{align*}
$$

Proof. We recall from pages 94 and 109 that $\tilde{F}_{d}(u, h)$ is trilinear in $\left(\partial_{x}^{2} \Theta_{h},\left[\partial_{x} \Theta_{h}\right]^{-1}, u\right)$, and that

$$
\left\|f_{d}\right\|_{\mathbb{F}_{d, \Sigma}}=\left\|f_{d}\right\|_{\mathbb{F}_{d}}+\left\|\left(f_{d+}\left|\Sigma, f_{d-}\right| \Sigma\right)\right\|_{\mathbb{G}_{w}},
$$

where $J=(0, T)$. The assertions for the $\mathbb{F}_{d}$-norm follow from Theorem 4.15 and Lemma 4.23, as soon as we have ensured that the difference $\partial_{x} \Theta_{h}-I_{x}=\partial_{x} \mathfrak{S}\left(h \nu_{\Sigma}, g_{h}\right)$ tends to zero in the $C\left([0, T] ; C^{1}(\bar{\Omega})\right)$-norm. Since $\tau>2+n / p$, we have

$$
\left\|\partial_{x} \Theta_{h}-I_{x}\right\|_{L_{\infty}((0, T) \times \Omega)} \lesssim\left\|h \nu_{\Sigma}\right\|_{L_{\infty}\left(0, T ; W_{p}^{\tau-1 / p}(\Sigma)\right)}+\left\|g_{h}\right\|_{L_{\infty}\left(0, T ; W_{p}^{\tau-1-1 / p}(\Sigma)\right)} \rightarrow 0,
$$

as $\|h\|_{L_{\infty}\left(0, T ; W_{p}^{\tau-1 / p}(\Sigma)\right)} \rightarrow 0$. By using (4.74) and $\tau<4-1 / p$, the assertions for the $\mathbb{F}_{d}$-norm follow.

The space $\mathbb{G}_{w}$ is a multiplication algebra by Lemma 4.25 , and from the mixed derivative embeddings and the $T$-dependent estimates in Lemma 3.18, we obtain the estimate

$$
\begin{equation*}
\left\|u_{\bullet} \mid \Sigma\right\|_{{ }_{0} G_{w}(T)} \leq T^{\delta} C\left(\delta, T_{0}\right)\left\|u_{\bullet}\right\|_{\mathbb{E}_{u, v, w, \partial_{\nu} w}(T)} . \tag{4.78}
\end{equation*}
$$

With $\left.\partial_{\nu_{\Sigma}}^{2} \mathfrak{S}\left(h \nu_{\Sigma}, g_{h}\right)\right|_{\Sigma}=0$, it follows that the values of $\left.\partial_{x}^{2} \mathfrak{S}\left(h \nu_{\Sigma}, g_{h}\right)\right|_{\Sigma}$ depend linearly on those of $\left(h, \nabla_{\Sigma} h, \nabla_{\Sigma}^{2} h, g_{h}, \nabla_{\Sigma} g_{h}\right)$. Therefore $h \mapsto \partial_{x}^{2} \mathfrak{S}\left(h \nu_{\Sigma}, g_{h}\right), \tilde{\mathbb{E}}_{h}(T) \cap \mathcal{U}_{h} \rightarrow \mathbb{G}_{w}(T)$ is analytic. Moreover, with the estimates (4.60b), (4.65), and (4.78), we can control $\tilde{F}_{d}(u, h)-\operatorname{div} u$ in ${ }_{0} \mathbb{G}_{w}(T)$ and $\partial_{\left(u_{\bullet}, h_{\bullet}\right)} \tilde{F}_{d}(u, h)$ in $\mathcal{B}\left({ }_{0} \mathbb{E}_{u, v, w, \partial_{\nu} w}(T) \times{ }_{0} \tilde{\mathbb{E}}_{h}(T) ;{ }_{0} \mathbb{G}_{w}(T)\right)$ by choosing $T$ and $h_{0}$ in $W_{p}^{3+(n-1) / p+\varepsilon}(\Sigma)$ sufficiently small. Therefore both assertions of Corollary 4.28 are true.

### 4.4. Local well-posedness of the transformed problem

Finally, we prove well-posedness for the transformed problem (T), which we restate as

$$
\begin{aligned}
\rho \partial_{t} u-\mu \Delta u+\nabla \pi & =F_{u}(u, \pi, h) & & \text { in } J \times \Omega \backslash \Sigma, \\
\operatorname{div} u & =F_{d}(u, h) & & \text { in } J \times \Omega \backslash \Sigma, \\
\llbracket u \rrbracket & =0 & & \text { on } J \times \Sigma, \\
\bar{N}_{v}(u, h)+\bar{N}_{w}(u, \pi, h) \nu_{\Sigma} & =0 & & \text { on } J \times \Sigma, \\
\partial_{t} h-u \cdot \nu_{\Sigma} & =0 & & \text { on } J \times \Sigma, \\
\left.u\right|_{\partial \Omega} & =0 & & \text { on } J \times \partial \Omega, \\
\left.h\right|_{t=0} & =h_{0} & & \text { on } \Sigma, \\
\left.u\right|_{t=0} & =u_{0} & & \text { in } \Omega \backslash \Sigma .
\end{aligned}
$$

Here $J=(0, T)$ is a bounded interval and $\Omega$ is a bounded smooth domain in $\mathbb{R}^{n}(n \geq 2)$ that contains a compact smooth hypersurface $\Sigma$. We employ the operators $F_{u}$ from page 106, $F_{d}$ from page 109, and $N_{v}$ and $N_{w}$ from page 112. We decompose $u=v+w \nu_{\Sigma} \circ \Pi_{\Sigma}$ near $\Sigma$. Both $u$ and $\pi$ denote transformed quantities; that is, we omit the bars over $u$ and $\pi$.

An $\mathbb{E}$-solution of problem (4.79) $=(\mathrm{T})$ on $J=(0, T)$ is a triple

$$
(u, \pi, h) \in \mathbb{E}(T):=\mathbb{E}_{u, v, w, \partial_{\nu} w}(T) \times \mathbb{E}_{\pi, \llbracket \pi \rrbracket}(T) \times\left(\mathbb{E}_{h}(T) \cap \mathcal{U}_{h}\right),
$$

which satisfies (4.79) pointwise almost everywhere. The relevant function spaces are collected in Figure 4.1. The nonlinearities are well-defined if the height function satisfies the smallness condition $h \in \mathcal{U}_{h}$ from Theorem 4.15.(ii) on page 102.

We will consider $\mathbb{E}$-solutions of the form

$$
(u, \pi, h)=\left(u_{*}+u_{\bullet}, \pi_{*}+\pi_{\bullet}, h_{*}+h_{\bullet}\right) \quad \text { with }\left(u_{\bullet}, \pi_{\bullet}, h_{\bullet}\right) \in{ }_{0} \mathbb{E},\left(u_{*}, \pi_{*}, h_{*}\right) \in \mathbb{E},
$$

where $\left(u_{\bullet}, \pi_{\bullet}, h_{\bullet}\right)$ has vanishing initial values and $\left(u_{*}, \pi_{*}, h_{*}\right)$ satisfies the initial conditions. In Definition 4.29 we define the state space $X_{p}$ of initial data ( $u_{0}, h_{0}$ ), which is a subset of some Banach space $\mathbb{X}_{p}$. It is shown in Lemma 4.31 that for every $\left(u_{0}, h_{0}\right) \in X_{p}$, there exists a tuple $\left(u_{*}, \pi_{*}, h_{*}\right) \in \mathbb{E}(T)$ which satisfies $\left.\left(u_{*}, h_{*}\right)\right|_{t=0}=\left(u_{0}, h_{0}\right)$ and depends linearly and continuously on ( $u_{0}, h_{0}$ ). Then it remains to solve a variant of problem (4.79) with vanishing initial values. In Theorem 4.33 we finally prove that (4.79) is locally well-posed in $X_{p}$ with respect to $\mathbb{E}$ in the sense of Definition 4.32 on page 125, and that the trajectories $t \mapsto(u(t), h(t))$ remain in $X_{p}$.

First, we deal with the non-homogeneous initial conditions in problem (4.79). For $p>3$ and a given tuple $(u, \pi, h) \in \mathbb{E}(T)$, the temporal trace theorem yields

$$
\begin{aligned}
u & \in C\left(\bar{J} ; W_{p}^{2-2 / p}(\Omega \backslash \Sigma)^{n}\right),\left.u\right|_{\partial \Omega}=0, \llbracket u \rrbracket=0, \\
\left.u\right|_{\Sigma} & \in C\left(\bar{J} ; W_{p}^{3-3 / p}(\Sigma)^{n}\right), \\
\left.\partial_{\nu} w\right|_{\Sigma}, \llbracket \pi \rrbracket & \in C\left(\bar{J} ; W_{p}^{2-3 / p}(\Sigma)\right), \\
h & \in C\left(\bar{J} ; W_{p}^{4-2 / p}(\Sigma)\right) .
\end{aligned}
$$

This observation motivates the following definition of initial states.
4.29. Definition (State space $X_{p}(\tau, M) \subset \mathbb{X}_{p}$ ). (i) Given $p \in(3, \infty)$, we let $\mathbb{X}_{p}$ denote the Banach space of all pairs $\left(u_{0}, h_{0}\right)$, which satisfy the conditions

$$
\begin{aligned}
u_{0} & \in W_{p}^{2-2 / p}\left(\Omega \backslash \Sigma ; \mathbb{R}^{n}\right),\left.u_{0}\right|_{\partial \Omega}=0, \llbracket u_{0} \rrbracket=0, \\
\left.u_{0}\right|_{\Sigma} & \in W_{p}^{3-3 / p}\left(\Sigma ; \mathbb{R}^{n}\right), \\
\left.\partial_{\nu} w_{0}\right|_{\Sigma} & \in W_{p}^{2-3 / p}(\Sigma ; \mathbb{R}), \\
h_{0} & \in W_{p}^{4-2 / p}(\Sigma ; \mathbb{R}) ;
\end{aligned}
$$

and $\mathbb{X}_{p}$ is equipped with the norm

$$
\left\|\left(u_{0}, h_{0}\right)\right\|_{\mathbb{X}_{p}}:=\left\|u_{0}\right\|_{W_{p}^{2-2 / p}(\Omega \backslash \Sigma)}+\left\|\left.u_{0}\right|_{\Sigma}\right\|_{W_{p}^{3-3 / p}(\Sigma)}+\left\|\left.\partial_{\nu} w_{0}\right|_{\Sigma}\right\|_{W_{p}^{2-3 / p}(\Sigma)}+\left\|h_{0}\right\|_{W_{p}^{4-2 / p}(\Sigma)} .
$$

(ii) Given $p \in(\max \{3,(n+2) / 2\}, \infty)$ and $\tau \in(1+n / p, 4-1 / p]$, we choose the number $\delta_{h}=$ $\delta_{h}(\Omega, \Sigma, p, \tau)>0$ such that both assertions of Theorem 4.15 are valid. For given $M \in\left(\delta_{h}^{-1}, \infty\right]$, the (nonlinear) state space $X_{p}(\tau, M)$ consists of all pairs $\left(u_{0}, h_{0}\right) \in \mathbb{X}_{p}$ with

$$
\left\|\left(u_{0}, h_{0}\right)\right\|_{\mathbb{X}_{p}}<M
$$

which satisfy the compatibility conditions

$$
\begin{align*}
\operatorname{div} u_{0}=F_{d}\left(u_{0}, h_{0}\right) & =-\sum_{j, l, m}\left(u_{0}\right)_{l} \partial_{l} \partial_{m}\left(\Theta_{h_{0}}\right)_{j} \partial_{j}\left(\Theta_{h_{0}}^{-1}\right)_{m},  \tag{4.80a}\\
G_{v}\left(0,0 ; u_{0}, h_{0}\right) & =0 \tag{4.80b}
\end{align*}
$$

the smallness condition

$$
\begin{equation*}
\left\|h_{0}\right\|_{W_{p}^{\tau-1 / p}(\Sigma)}<\delta_{h}-M^{-1} \tag{4.81}
\end{equation*}
$$

and the well-posedness condition

$$
\begin{equation*}
\inf _{x \in \Sigma} d_{0}\left(D_{\Sigma}\left(u_{0}\right)(x)\right)=\inf _{\Sigma}\left(\sigma+\left(\lambda_{s}-\mu_{s}\right) \operatorname{tr} D_{\Sigma}\left(u_{0}\right)+2 \mu_{s} \min _{\zeta \in \mathbb{R}^{n},|\zeta|=1} \zeta^{\top}\left[D_{\Sigma}\left(u_{0}\right)\right] \zeta\right)>M^{-1} . \tag{4.82}
\end{equation*}
$$

For given $\eta \in(0, \infty)$, we further let

$$
X_{p}(\tau, M, \eta):=\left\{\left(u_{0}, h_{0}\right) \in X_{p}(\tau, M):\left\|h_{0}\right\|_{W_{p}^{\tau-1 / p}(\Sigma)}<\eta\right\} .
$$

4.30. Remark. The compatibility conditions (4.80) arise since both spaces $\mathbb{F}_{d}$ and $\mathbb{G}_{v}$ have welldefined initial traces. There is no compatibility condition for $L_{w}$ and $G_{w}$ since the initial value $\left.\llbracket \pi \rrbracket\right|_{t=0}$ is not prescribed. Condition (4.81) and $p>(n+2) / 3$ allow to define the diffeomorphism $\Theta_{h_{0}}: \Omega \rightarrow \Omega$ with $\Theta_{h_{0} \mid \Sigma}: \Sigma \rightarrow \Gamma_{h_{0}}$ by means of Theorem 4.15. With $p>\max \{3,(n+2) / 2\}$ we obtain $u_{0} \mid \Sigma \in C^{1}(\Sigma)$, and condition (4.82) will be used to employ the linear solution operator from Theorem 3.21. Equipped with the induced metric of $\mathbb{X}_{p}$, the space $X_{p}(\tau, M)$ is a metric space. If $M_{0} \leq M$ and $\eta_{0} \leq \eta$, then $X_{p}\left(\tau, M_{0}, \eta_{0}\right) \subset X_{p}(\tau, M, \eta)$.

Next, we construct functions $(u, \pi, h) \in \mathbb{E}\left(T_{0}\right)$ satisfying the initial condition $\left.(u, h)\right|_{t=0}=$ $\left(u_{0}, h_{0}\right)$ for given $\left(u_{0}, h_{0}\right) \in X_{p} \in\left\{X_{p}(\tau, M), X_{p}(\tau, M, \eta)\right\}$, together with corresponding interior data $\left(f_{u}, f_{d}\right) \in \mathbb{F}_{u} \times \mathbb{F}_{d, \Sigma}$. We also show that the trajectories $t \mapsto(u(t, \cdot), h(t, \cdot))$ remain in $X_{p}$.
4.31. Lemma. Let $p \in(\max \{3,(n+2) / 2\}, \infty)$ and $\tau \in(1+n / p, 4-1 / p]$.
(i) For every $T_{0} \in(0, \infty)$, there exists a bounded linear operator

$$
\begin{aligned}
\left(u_{0}, h_{0}\right) & \mapsto\left(u, \pi, h, f_{u}, f_{d}\right), \\
X_{p}(\tau, \infty) & \rightarrow \mathbb{E}_{u, v, w, \partial_{\nu} w}\left(T_{0}\right) \times \mathbb{E}_{\pi, \llbracket \pi \rrbracket}\left(T_{0}\right) \times \mathbb{E}_{h}\left(T_{0}\right) \times \mathbb{F}_{u}\left(T_{0}\right) \times \mathbb{F}_{d, \Sigma}\left(T_{0}\right)
\end{aligned}
$$

whose values satisfy

$$
\begin{align*}
\left(\rho \partial_{t}-\mu \Delta\right) u+\nabla \pi & =f_{u} & & \text { in } J \times \Omega,  \tag{4.83a}\\
\operatorname{div} u & =f_{d} & & \text { in } J \times \Omega,  \tag{4.83b}\\
\partial_{t} h-\left.\nu_{\Sigma} \cdot u\right|_{\Sigma} & =0 & & \text { on } J \times \Sigma,  \tag{4.83c}\\
\left.u\right|_{t=0} & =u_{0} & & \text { in } \Omega,  \tag{4.83d}\\
\left.h\right|_{t=0} & =h_{0} & & \text { on } \Sigma . \tag{4.83e}
\end{align*}
$$

(ii) Moreover, if $\tau<4-1 / p$ and $M_{0}<M$, then there exists $T \in\left(0, T_{0}\right]$ such that

$$
\begin{equation*}
(u, h) \in C\left([0, T] ; X_{p}(\tau, M)\right) \quad \text { for all }\left(u_{0}, h_{0}\right) \in X_{p}\left(\tau, M_{0}\right), \tag{4.84}
\end{equation*}
$$

and for $M_{0}<M$ and $\eta_{0}<\eta$, there exists $T \in\left(0, T_{0}\right]$ such that

$$
\begin{equation*}
(u, h) \in C\left([0, T] ; X_{p}(\tau, M, \eta)\right) \quad \text { for all }\left(u_{0}, h_{0}\right) \in X_{p}\left(\tau, M_{0}, \eta_{0}\right) . \tag{4.85}
\end{equation*}
$$

Proof. (i) Let $\left(u_{0}, h_{0}\right) \in X_{p}(\tau, \infty)$ be given. With $w_{0}:=\left.\nu_{\Sigma} \cdot u_{0}\right|_{\Sigma}$, we define

$$
h(t):=h_{A}(t)+h_{B}(t):=\left(2 e^{-t A}-e^{-2 t A}\right) h_{0}+\left(e^{-t B}-e^{-2 t B}\right) B^{-1} w_{0}, \quad w(t):=\partial_{t} h(t),
$$

where the operators $A=\sqrt{1-\Delta_{\Sigma}}$ and $B=1-\Delta_{\Sigma}$ are realized in $L_{p}(\Sigma)$. With $h_{0} \in D_{A}(4-$ $2 / p, p)$ and Corollaries B. 26 and B. 58 on pages 155 and 163, we see that $h_{A}$ belongs to $\mathbb{E}_{h}$. Similarly, with $B^{-1} w_{0} \in W_{p}^{4-3 / p}(\Sigma)=D_{B}(2-3 / 2 p, p)$ and $w_{0} \in W_{p}^{3-3 / p}(\Sigma)=D_{B}(3 / 2-$ $3 / 2 p, p)$, we obtain $h_{B} \in W_{p}^{4-1 / p}(J \times \Sigma)$ and $\partial_{t} h_{B} \in W_{p}^{3 / 2-1 / 2 p}\left(J ; L_{p}(\Sigma)\right) \cap L_{p}\left(J ; W_{p}^{3-1 / p}(\Sigma)\right)$, and thus $h_{B}$ belongs to $\mathbb{E}_{h}$. We conclude that $h$ is well-defined in $\mathbb{E}_{h}$, the function $w=\partial_{t} h$ belongs to $\mathbb{E}_{w}$, both functions depend linearly and continuously on $w_{0} \in W_{p}^{3-3 / p}(\Sigma)$ and $h_{0} \in$ $W_{p}^{4-2 / p}(\Sigma)$, and (4.83c) is satisfied.

Next, with the operator $\widetilde{\Delta}_{\Sigma}=g^{\alpha \beta} \widetilde{\nabla}_{\alpha} \widetilde{\nabla}_{\beta}: H_{p}^{2}(\Sigma ; T \Sigma) \rightarrow L_{p}(\Sigma ; T \Sigma)$, we define

$$
v(t, \cdot):=e^{-t\left(1-\widetilde{\Delta}_{\Sigma}\right)}\left(\left.P_{\Sigma} u_{0}\right|_{\Sigma}\right)
$$

The complexification of $1-\widetilde{\Delta}_{\Sigma}$ belongs to the class $\mathcal{R} \mathcal{S}\left(L_{p}\left(\Sigma ;(T \Sigma)_{\mathbb{C}}\right)\right)$ with $\mathcal{R}$-angle zero by Corollary B.59, and to $\mathcal{H}^{\infty}\left(W_{p}^{1-1 / p}\left(\Sigma ;(T \Sigma)_{\mathbb{C}}\right)\right)$ by Theorem B.27. Hence the semigroup $e^{t \widetilde{\Delta}_{\Sigma}}$ is analytic in $W_{p}^{1-1 / p}\left(\Sigma ;(T \Sigma)_{\mathbb{C}}\right)$, and from Theorem B. 25 we infer that $v$ belongs to

$$
W_{p}^{3 / 2-1 / 2 p}\left(J ; L_{p}(\Sigma ; T \Sigma)\right) \cap L_{p}\left(J ; W_{p}^{3-1 / p}(\Sigma ; T \Sigma)\right) \hookrightarrow \mathbb{E}_{v} .
$$

Let us construct the divergence data $f_{d}$ on $J \times(\Omega \backslash \Sigma)$. From $p>(n+2) / 2$, Sobolev's embedding (B.1), and Lemma B.81, we deduce that $W_{p}^{2-2 / p}(\Omega \backslash \Sigma)$ and $W_{p}^{2-2 / p}(\Sigma)$ are multiplication algebras. The compatibility condition (4.80a) implies $\operatorname{div} u_{0} \in W_{p}^{2-2 / p}(\Omega \backslash \Sigma)$ and $\left.\operatorname{div} u_{0}\right|_{\Sigma} \in$ $W_{p}^{2-2 / p}(\Sigma)$. By Corollary B. 58 and Theorem B.25, the function $f_{d \Sigma}(t):=e^{-t\left(1-\Delta_{\Sigma}\right)}\left(\operatorname{div} u_{0} \mid \Sigma\right)$ belongs to $H_{p}^{1}\left(J ; L_{p}(\Sigma)\right) \cap L_{p}\left(J ; H_{p}^{2}(\Sigma)\right) \hookrightarrow \mathbb{G}_{w}$. Let $\tilde{f}_{d}$ solve the heat problem $\left(\partial_{t}-\Delta\right) \tilde{f}_{d}=0$ in $J \times$ $\Omega_{ \pm}, \tilde{f}_{d}\left|\partial \Omega=0, \tilde{f}_{d}\right|_{\Sigma}=f_{d \Sigma}$, and $\tilde{f}_{d} \mid t=0=\operatorname{div} u_{0}$. Then we let $\left.f_{d}\right|_{\Omega_{ \pm}}:=\tilde{f}_{d}\left|\Omega_{ \pm}-\left\langle\tilde{f}_{d}\right\rangle_{\Omega_{ \pm}} \mp\right| \Sigma \mid\langle w\rangle_{\Sigma}$, and this function $f_{d}$ belongs to $H_{p}^{1}\left(J ; L_{p}(\Omega)\right) \cap L_{p}\left(J ; H_{p}^{2}(\Omega \backslash \Sigma)\right)$. In view of $\int_{\Omega} f_{d} d x=0$ and the Poincaré-Wirtinger inequality, it also belongs to $H_{p}^{1}\left(J ; \hat{H}_{p}^{-1}(\Omega)\right)$, and hence $f_{d} \in \mathbb{F}_{d, \Sigma}$.

Next, we obtain the bulk velocity field $u \in \mathbb{E}_{u, v, w, \partial_{\nu} w}$ from the solution $(u, q)$ of the onephase Stokes problems $\left(\rho_{ \pm} \partial_{t}-\mu_{ \pm} \Delta\right) u_{ \pm}+\nabla q_{ \pm}=0$ in $\Omega_{ \pm}$, $\operatorname{div} u=f_{d}$ in $\Omega,\left.u_{+}\right|_{\partial \Omega}=0$, and $\left.u_{ \pm}\right|_{\Sigma}=v+w \nu_{\Sigma}$ with [BP07, Theorem 4.1]. For the construction of $\pi$, we first define $g_{0}=\llbracket q_{0} \rrbracket$ by eliminating $\llbracket q_{0} \rrbracket$ from the equation $G_{w}\left(0,0 ; u_{0}, q_{0}, h_{0}\right)=0$. Then $t \mapsto g(t):=e^{-t\left(1-\Delta_{\Sigma}\right)} g_{0}$ belongs to $\mathbb{G}_{w}$. The function $\pi \in \mathbb{E}_{\pi,[\pi \pi]}$ is defined with Theorem 2.3 as the unique solution of the weak transmission problem $\langle\nabla \pi, \nabla \varphi\rangle_{\Omega}=-\left\langle\left(\rho \partial_{t}-\mu \Delta\right) u, \nabla \varphi\right\rangle_{\Omega}$ for all $\varphi \in \dot{H}_{p^{\prime}}^{1}(\Omega)$ and $\llbracket \pi \rrbracket=g$. Finally, the function $f_{u}:=\left(\rho \partial_{t}-\mu \Delta\right) u+\nabla \pi$ belongs to $\mathbb{F}_{u}$. Therefore assertion (i) is valid.
(ii) We employ the following estimates, which follow from Sobolev's embedding, Lemma 3.18, the mixed derivative embeddings, and Theorem B.25. For $\delta \in(0,1-1 / p]$, we have

$$
\begin{aligned}
\left\|h_{A}\right\|_{H_{p}^{1 / p+\delta}\left(0, T_{0} ; W_{p}^{s-1 / p}(\Sigma)\right)} & \leq C\left(T_{0}\right)\left\|h_{A}\right\|_{H_{p}^{1}\left(0, T_{0} ; W_{p}^{s-1+\delta}(\Sigma)\right) \cap L_{p}\left(0, T_{0} ; W_{p}^{s+\delta}(\Sigma)\right)} \\
& \leq C\left(T_{0}\right)\left\|h_{0}\right\|_{W_{p}^{s-1 / p+\delta}(\Sigma)} \leq C\left(T_{0}\right) M_{0},
\end{aligned}
$$

provided that $1-\delta \leq s<4-1 / p$ and $\delta \leq 4-1 / p-s$; and

$$
\begin{aligned}
\left\|h_{B}\right\|_{H_{p}^{1 / p+\delta}\left(0, T_{0} ; W_{p}^{s-1 / p}(\Sigma)\right)} & \leq C\left(T_{0}\right)\left\|h_{B}\right\|_{H_{p}^{1}\left(0, T_{0} ; W_{p}^{s-2+2 \delta}(\Sigma)\right) \cap L_{p}\left(0, T_{0} ; W_{p}^{s+1 / p+2 \delta}(\Sigma)\right)} \\
& \leq C\left(T_{0}\right)\left\|B^{-1} w_{0}\right\|_{W_{p}^{s-1 / p+2 \delta}(\Sigma)} \leq C\left(T_{0}\right) M_{0},
\end{aligned}
$$

provided that $2-2 \delta \leq s<5-2 / p$ and $\delta \leq 5 / 2-1 / p-s / 2$. Thus, for $s=\tau$, some $\delta_{0}>0$, and all $\delta \in\left(0, \delta_{0}\right], T \in\left(0, T_{0}\right]$, and $\left(u_{0}, h_{0}\right) \in X_{p}\left(\tau, M_{0}\right)$, we have

$$
\left\|h(\cdot)-h_{0}\right\|_{0 C\left([0, T] ; W_{p}^{\tau-1 / p}(\Sigma)\right)} \leq T^{\delta / 2} C\left(\delta, T_{0}\right)\left\|h(\cdot)-h_{0}\right\|_{0 H_{p}^{1 / p+\delta}\left(0, T ; W_{p}^{\tau-1 / p}\right)} \leq T^{\delta / 2} C\left(\delta, T_{0}\right) M_{0}
$$

Next, recall that $w=\partial_{t} h_{A}+\partial_{t} h_{B}$. For $\delta \in(0,1-1 / p]$, we have

$$
\begin{aligned}
\left\|\partial_{t} h_{A}\right\|_{H_{p}^{1 / p+\delta}\left(0, T_{0} ; W_{p}^{s-1 / p}(\Sigma)\right)} & \leq C\left(T_{0}\right)\left\|A h_{A}\right\|_{H_{p}^{1}\left(0, T_{0} ; W_{p}^{s-1+\delta}(\Sigma)\right) \cap L_{p}\left(0, T_{0} ; W_{p}^{s+\delta}(\Sigma)\right)} \\
& \leq C\left(T_{0}\right)\left\|A h_{0}\right\|_{W_{p}^{s-1 / p+\delta}(\Sigma)} \leq C\left(T_{0}\right) M_{0},
\end{aligned}
$$

provided that $1-\delta \leq s<3-1 / p$ and $\delta \leq 3-1 / p-s ;$

$$
\begin{aligned}
\left\|\partial_{t} h_{B}\right\|_{H_{p}^{1 / p+\delta}\left(0, T_{0} ; W_{p}^{s-1 / p}(\Sigma)\right)} & \leq C\left(T_{0}\right)\left\|B h_{B}\right\|_{H_{p}^{1}\left(0, T_{0} ; W_{p}^{s-2+2 \delta}(\Sigma)\right) \cap L_{p}\left(0, T_{0} ; W_{p}^{s+1 / p+2 \delta}(\Sigma)\right)} \\
& \leq C\left(T_{0}\right)\left\|w_{0}\right\|_{W_{p}^{s-1 / p+2 \delta}(\Sigma)} \leq C\left(T_{0}\right) M_{0}
\end{aligned}
$$

provided that $2-2 \delta \leq s<3-2 / p$ and $\delta \leq 3 / 2-1 / p-s / 2$; and

$$
\begin{aligned}
\|v\|_{H_{p}^{1 / p+\delta}\left(0, T_{0} ; W_{p}^{s-1 / p}(\Sigma)\right)} & \leq C\left(T_{0}\right)\|v\|_{H_{p}^{1}\left(0, T_{0} ; W_{p}^{s-2+2 \delta}(\Sigma)\right) \cap L_{p}\left(0, T_{0} ; W_{p}^{s+1 / p+2 \delta}(\Sigma)\right)} \\
& \leq C\left(T_{0}\right)\left\|v_{0}\right\|_{W_{p}^{s-1 / p+2 \delta}(\Sigma)} \leq C\left(T_{0}\right) M_{0}
\end{aligned}
$$

provided that $2-2 \delta \leq s<3-2 / p$ and $\delta \leq 3 / 2-1 / p-s / 2$. Thus, for given $s \in(1+n / p, 3-2 / p)$ there exists $\delta_{0}>0$ such that for all $\delta \in\left(0, \delta_{0}\right], T \in\left(0, T_{0}\right]$, and $\left(u_{0}, h_{0}\right) \in X_{p}\left(\tau, M_{0}\right)$, we have

$$
\begin{aligned}
\left\|u(\cdot)-u_{0}\right\|_{0} C\left([0, T] ; C^{1}(\Sigma)\right) & \leq T^{\delta / 2} C\left(\delta, T_{0}\right)\left\|v(\cdot)-v_{0}+\left(w(\cdot)-w_{0}\right) \nu_{\Sigma}\right\|_{0 W_{p}^{1 / p+\delta}\left(0, T ; W_{p}^{s-2 / p}\right)(\Sigma)} \\
& \leq T^{\delta / 2} C\left(\delta, T_{0}\right) M_{0} .
\end{aligned}
$$

We conclude that there exists $T=T\left(M_{0}, M\right)>0$ such that

$$
\sup _{t \leq T}\|h(t, \cdot)\|_{W_{p}^{\tau-1 / p}(\Sigma)}<\delta_{h}-M^{-1}, \quad \inf _{t \leq T} \inf _{\Sigma} d_{0}(u(t, \cdot))>M^{-1}
$$

for all $\left(u_{0}, h_{0}\right) \in X_{p}\left(\tau, M_{0}\right)$. The assertion for $X_{p}(\tau, M, \eta)$ follows similarly.
In order to formulate our main result, we first specify our notion of well-posedness.
4.32. Definition. Let $X_{p}=X_{p}(\tau, M, \eta)$ and $\mathbb{E}$ have the same meaning as on pages 94 and 122. Problem $(4.79)=(T)$ is called locally well-posed in $X_{p}$ with respect to $\mathbb{E}$, if
(i) for every $z_{0 *} \in X_{p}$ there exist $T>0$ and $\delta>0$ such for all $z_{0}=\left(u_{0}, h_{0}\right) \in X_{p} \cap B_{\delta}^{\mathbb{X}_{p}}\left(z_{0 *}\right)$, problem (4.79) has a unique $\mathbb{E}$-solution $(u, \pi, h)$ on $(0, T)$,
(ii) the map $z_{0} \mapsto(u, \pi, h), X_{p} \cap B_{\delta}^{\mathbb{X}_{p}}\left(z_{0 *}\right) \rightarrow \mathbb{E}(T)$ is continuous,
(iii) the trajectory $t \mapsto(u(t), h(t))$ belongs to $C\left([0, T] ; X_{p}\right)$, and the map $z_{0} \mapsto(u, h), X_{p} \cap$ $B_{\delta}^{\mathbb{X}_{p}}\left(z_{0 *}\right) \rightarrow C\left([0, T] ; X_{p}\right)$ is continuous.
Our main result for the transformed problem $(4.79)=(\mathrm{T})$ is the following.
4.33. Theorem (Main result). Let $p>n+2, \tau \in(3+n / p, 4-1 / p)$, and $M<\infty$. Then there exists $\eta>0$ such that problem (4.79) is locally well-posed in $X_{p}(\tau, M, \eta)$ with respect to $\mathbb{E}$.

Proof. For given $z_{0}=\left(u_{0}, h_{0}\right) \in X_{p}(\tau, \infty)$ we seek a solution of the form

$$
z=(u, \pi, h)=z_{*}+z_{\bullet} \in \mathbb{E}(T) \quad \text { with } z_{\bullet}=\left(u_{\bullet}, \pi_{\bullet}, h_{\bullet}\right) \in{ }_{0} \mathbb{E}(T), z_{*}=\left(u_{*}, \pi_{*}, h_{*}\right) \in \mathbb{E}(T)
$$

on some small time interval $J=(0, T)$ such that

$$
\left.z_{\bullet}\right|_{t=0}=\left.\left(u_{\bullet}, h_{\bullet}\right)\right|_{t=0}=(0,0),\left.\quad z_{*}\right|_{t=0}=\left.\left(u_{*}, h_{*}\right)\right|_{t=0}=\left(u_{0}, h_{0}\right)=z_{0} .
$$

We abbreviate the transformed problem (4.79) as

$$
\begin{equation*}
L\left(z_{\bullet} ; z_{*}\right)=N\left(z_{\bullet} ; z_{*}\right),\left.\quad z_{\bullet}\right|_{t=0}=0,\left.\quad z_{*}\right|_{t=0}=z_{0}, \tag{4.86}
\end{equation*}
$$

where $L\left(z_{\bullet} ; z_{*}\right)=L\left(u_{\bullet}, \pi_{\bullet}, h_{\bullet} ; u_{*}\right)$ and $N\left(z_{\bullet} ; z_{*}\right)=N\left(u_{\bullet}, \pi_{\bullet}, h_{\bullet} ; u_{*}, \pi_{*}, h_{*}\right)$ are given by

$$
L\left(z_{\bullet} ; z_{*}\right)=\left[\begin{array}{c}
\left(\rho \partial_{t}-\mu \Delta\right) u_{\bullet}+\nabla \pi_{\bullet} \\
\operatorname{div} u_{\bullet} \\
L_{v}\left(u_{\bullet}, h_{\bullet} ; u_{*}\right) \\
L_{w}\left(u_{\bullet}, \pi_{\bullet}, h_{\bullet} ; u_{*}\right) \\
\partial_{t} h_{\bullet}-w_{\bullet}
\end{array}\right], \quad N\left(z_{\bullet} ; z_{*}\right)=\left[\begin{array}{c}
F_{u}\left(u_{*}+u_{\bullet}, \pi_{*}+\pi_{\bullet}, h_{*}+h_{\bullet}\right) \\
F_{d}\left(u_{*}+u_{\bullet}, h_{*}+h_{\bullet}\right)-\operatorname{div} u_{*} \\
G_{v}\left(u_{\bullet}, h_{\bullet} ; u_{*}, h_{*}\right) \\
G_{w}\left(u_{\bullet}, h_{\bullet} ; u_{*}, \pi_{*}, h_{*}\right) \\
0
\end{array}\right] .
$$

(i) Construction of $z_{*}$. Let $M \in\left(\delta_{h}^{-1}, \infty\right]$, let $\left(u_{0}, h_{0}\right) \mapsto\left(u_{*}, \pi_{*}, h_{*}, f_{u}, f_{d}\right)$ denote the bounded linear operator from Lemma 4.31, put $E_{T}\left(u_{0}, h_{0}\right):=\left(u_{*}, \pi_{*}, h_{*}\right)$, and let $\mathcal{P}_{R, T_{0}}$ denote the set of admissible parameters $u_{*}$ from page 82 . Then for given $M_{0}<M$ and $\eta_{0}<\eta$, there are $T_{0}>0$ and $R \geq 1$ such that the realizations

$$
\begin{aligned}
& E_{T}: X_{p}\left(\tau, M_{0}\right) \rightarrow\left\{\left(u_{*}, \pi_{*}, h_{*}\right) \in \mathbb{E}(T): u_{*} \in \mathcal{P}_{R, T_{0}},\left(u_{*}, h_{*}\right) \in C\left([0, T] ; X_{p}(\tau, M)\right)\right\}, \\
& E_{T}: X_{p}\left(\tau, M_{0}, \eta_{0}\right) \rightarrow\left\{\left(u_{*}, \pi_{*}, h_{*}\right) \in \mathbb{E}(T): u_{*} \in \mathcal{P}_{R, T_{0}},\left(u_{*}, h_{*}\right) \in C\left([0, T] ; X_{p}(\tau, M, \eta)\right)\right\}
\end{aligned}
$$

are linear and bounded for every $T \in\left(0, T_{0}\right]$.
(ii) Strategy to determine $z_{0}$. It remains to determine the solution $z_{\bullet} \in{ }_{0} \mathbb{E}(T)$ of the equation $L\left(z_{\bullet} ; E_{T}\left(z_{0}\right)\right)=N\left(z_{\bullet} ; E_{T}\left(z_{0}\right)\right)$. In Theorem 3.21, we have shown that $L\left(\cdot ; z_{*}\right):{ }_{0} \mathbb{E}(T) \rightarrow_{0} \mathbb{F}(T)$ is uniformly invertible with respect to $T \in\left(0, T_{0}\right]$ and $u_{*} \in \mathcal{P}_{R, T_{0}}$, for given $T_{0} \in(0, \infty)$ and $R \in[1, \infty)$. Therefore we intend to apply Banach's fixed point theorem to the map

$$
\begin{aligned}
F: & \left(z_{\bullet}, z_{0}\right) \mapsto\left[L\left(\cdot ; E_{T}\left(z_{0}\right)\right)\right]^{-1} N\left(z_{\bullet} ; E_{T}\left(z_{0}\right)\right), \\
& \left\{\left(z_{\bullet}, z_{0}\right) \in_{0} \mathbb{E}(T) \times X_{p}(\tau, M, \eta)\right\} \rightarrow{ }_{0} \mathbb{E}(T),
\end{aligned}
$$

with suitable $\eta>0$ and $T>0$, depending on $M \in\left(\delta_{h}^{-1}, \infty\right)$. To this end we will show that

$$
\begin{equation*}
\left\|F\left(z_{\bullet} ; z_{0}\right)\right\|_{o \mathbb{E}(T)}+\left\|\partial_{z_{\bullet}} F\left(z_{\bullet} ; z_{0}\right)\right\|_{\mathcal{B}(0 \mathbb{E}(T))} \rightarrow 0 \quad \text { as } T \rightarrow 0,\left\|h_{0}\right\|_{W_{p}^{\tau-1 / p}(\Sigma)} \rightarrow 0, \tag{4.87}
\end{equation*}
$$

uniformly with respect to $z_{0} \in X_{p}(\tau, M)$ and $z_{\bullet} \in \overline{\mathbb{B}}_{r}(T)$, where

$$
\overline{\mathbb{B}}_{r}(T):=\left\{z_{\bullet}=\left(u_{\bullet}, \pi_{\bullet}, h_{\bullet}\right) \in{ }_{0} \mathbb{E}(T):\left\|z_{\bullet}\right\|_{0 \mathbb{E}}(T) \leq r\right\} .
$$

(iii) Properties of $F$. Let $T_{0} \in(0, \infty), T \in\left(0, T_{0}\right]$, and $R \in[1, \infty)$. The map $u_{*} \mapsto L\left(\cdot ; u_{*}\right)$, $\mathbb{E}_{u, v, w}\left(T_{0}\right) \rightarrow \mathcal{B}\left({ }_{0} \mathbb{E}(T) ;{ }_{0} \mathbb{F}(T)\right)$ is affine and therefore uniformly continuous on $B_{R}^{\mathbb{E}_{u, v, w}\left(T_{0}\right)}$. Theorem 3.21 implies that $u_{*} \mapsto\left[L\left(\cdot ; u_{*}\right)\right]^{-1}, \mathcal{P}_{R, T_{0}} \rightarrow \mathcal{B}\left({ }_{0} \mathbb{F}(T) ;{ }_{0} \mathbb{E}(T)\right)$ is uniformly bounded. Since $E_{T}$ is linear and bounded, the map $z_{0} \mapsto\left[L\left(\cdot ; E_{T}\left(z_{0}\right)\right)\right]^{-1}, X_{p}(\tau, M) \rightarrow \mathcal{B}\left({ }_{0} \mathbb{F}(T) ;{ }_{0} \mathbb{E}(T)\right)$ is uniformly continuous. The function $N\left(z_{\bullet} ; E_{T}\left(z_{0}\right)\right)$ and its derivative with respect to $z_{\bullet}$ depend polynomially on the functions $z_{\bullet}, z_{*}=E_{T}\left(z_{0}\right), \partial_{x} \Theta_{h},\left[\partial_{x} \Theta_{h}\right]^{-1}, \beta_{h}$, and $\beta_{h}^{-1}$, where $h=h_{*}+h_{\bullet}$. From estimate (4.74) we infer that there exists $\delta>0$ such that

$$
\left\|h_{\bullet}\right\|_{0 C\left([0, T] ; W_{p}^{\tau-1 / p}(\Sigma)\right)} \leq T^{\delta} C\left(\delta, T_{0}\right)\left\|h_{\bullet}\right\|_{0 \mathbb{E}_{h}(T)} \quad \text { for } h_{\bullet} \in{ }_{0} \mathbb{E}_{h}(T), T \in\left(0, T_{0}\right] .
$$

Hence for given $M \in\left(\delta_{h}^{-1}, \infty\right)$ there exist $T_{0}, r \in(0, \infty)$ such that the maps

$$
\begin{array}{ll}
\left(z_{\bullet} ; z_{0}\right) \mapsto F\left(z_{\bullet} ; z_{0}\right), & \left\{\left(z_{\bullet}, z_{0}\right) \in \overline{\mathbb{B}}_{r}(T) \times X_{p}(\tau, M)\right\} \rightarrow{ }_{0} \mathbb{E}(T), \\
\left(z_{\bullet} ; z_{0}\right) \mapsto \partial_{z_{\bullet}} F\left(z_{\bullet} ; z_{0}\right), & \left\{\left(z_{\bullet}, z_{0}\right) \in \overline{\mathbb{B}}_{r}(T) \times X_{p}(\tau, M)\right\} \rightarrow \mathcal{B}\left({ }_{0} \mathbb{E}(T)\right)
\end{array}
$$

are well-defined and uniformly continuous for every $T \in\left(0, T_{0}\right]$.
Corollaries 4.27 and 4.28 and Lemma 4.26 yield

$$
\left\|N\left(z_{\bullet} ; E_{T}\left(z_{0}\right)\right)\right\|_{0 \mathbb{F}(T)}+\left\|\partial_{\bullet} N\left(z_{\bullet} ; E_{T}\left(z_{0}\right)\right)\right\|_{0 \mathbb{E}(T) \rightarrow 0} \mathbb{F}(T) \rightarrow 0 \quad \text { as } T \rightarrow 0,\left\|h_{0}\right\|_{W_{p}^{\tau-1 / p}(\Sigma)} \rightarrow 0,
$$

uniformly with respect to $z_{0} \in X_{p}(\tau, M)$ and $z_{\bullet} \in \overline{\mathbb{B}}_{r}(T)$. Therefore (4.87) is valid.
(iv) Strict contraction. Let us prove that $F\left(\cdot ; z_{0}\right)$ is a strict contraction within the closed set $\overline{\mathbb{B}}_{r}(T)$ for some $r, T>0$. From estimate (4.87) we infer that for given $q \in(0,1)$ and $M \in$ $\left(\delta_{h}^{-1}, \infty\right)$ there exist positive numbers $\eta, T_{0}$, and $r$ such that

$$
\left\|\partial_{z_{\mathbf{\bullet}}} F\left(z_{\bullet} ; z_{0}\right)\right\|_{\mathcal{B}(0 \mathbb{E}(T))} \leq q \quad \text { for all } z_{\bullet} \in \overline{\mathbb{B}}_{r}(T), T \in\left(0, T_{0}\right], z_{0} \in X_{p}(\tau, M, \eta) .
$$

Estimate (4.87) and the differentiability of $F\left(\cdot, z_{0}\right)$ imply that there are positive numbers $\eta, T_{0}$, and $r$ such that

$$
\left\|F\left(z_{\bullet} ; z_{0}\right)\right\|_{0 \mathbb{E}(T)} \leq q\left\|z_{\bullet}\right\|_{0 \mathbb{E}(T)}+\left\|F\left(0 ; z_{0}\right)\right\|_{0 \mathbb{E}(T)} \leq r
$$

for all $z_{\bullet} \in \overline{\mathbb{B}}_{r}(T), T \in\left(0, T_{0}\right]$, and $z_{0} \in X_{p}(\tau, M, \eta)$. Therefore $F\left(\cdot ; z_{0}\right)$ maps $\overline{\mathbb{B}}_{r}(T)$ into itself as a $q$-contraction.
(v) Banach's fixed point theorem implies that $F\left(\cdot ; z_{0}\right)$ has a unique fixed point $z_{0}$ within $\overline{\mathbb{B}}_{r}(T) \subset{ }_{0} \mathbb{E}(T)$ and $z_{\bullet}$ depends continuously on $z_{0} \in X_{p}(\tau, M, \eta)$. Moreover, $z_{\bullet}+E_{T}\left(z_{0}\right)$ is an $\mathbb{E}$ solution of problem (4.79) on $(0, T)$ and $(u, h)=\left(u_{*}+u_{\bullet}, h_{*}+h_{\bullet}\right)$ belongs to $C\left([0, T] ; X_{p}(\tau, \infty)\right)$. Let us show that its trajectory remains in $X_{p}(\tau, M, \eta)$. For given $z_{0 *} \in X_{p}(\tau, M, \eta)$ there are $M_{0}<M$ and $\eta_{0}<\eta$ such that $z_{0 *}$ belongs to $X_{p}\left(\tau, M_{0}, \eta_{0}\right)$. Therefore Lemma 4.31 yields some numbers $T_{1} \in\left(0, T_{0}\right]$ and $r_{1} \leq r$ such that, given $z_{0} \in X_{p}(\tau, M, \eta) \cap B_{\delta}^{\mathbb{X}_{p}}\left(z_{0 *}\right), T \in\left(0, T_{1}\right]$, and $z_{\bullet} \in \overline{\mathbb{B}}_{r_{1}}(T)$, the solution $z=z_{\bullet}+E_{T}\left(z_{0}\right)$ satisfies $(u, h) \in C\left([0, T] ; X_{p}(\tau, M, \eta)\right)$.
(vi) In order to prove uniqueness within the larger space $\mathbb{E}(T)=E_{T}\left(z_{0}\right)+{ }_{0} \mathbb{E}(T)$ for $T<T_{1}$, we assume that there is a different $\mathbb{E}$-solution $z^{1}=z_{*}+z_{\boldsymbol{\bullet}}^{1}$ on $(0, T)$. Since $F\left(\cdot, z_{0}\right)$ has at most one fixed point within $\overline{\mathbb{B}}_{r}(T)$, the triple $z_{0}^{1} \in{ }_{0} \mathbb{E}(T)$ does not belong to $\overline{\mathbb{B}}_{r}(T)$. But since the norm of $\mathbb{E}(T)$ consists of integrals over $(0, T)$, we have $\left\|z_{\bullet}^{1}\right\|_{\mathbb{E}\left(T^{\prime}\right)} \rightarrow 0$ as $T^{\prime} \rightarrow 0$, and hence $z_{\boldsymbol{\bullet}}$ and $z_{\boldsymbol{0}}^{1}$ coincide on some interval $\left(0, T^{\prime}\right)$. We may assume that this interval is maximal in the sense that for every $\varepsilon>0$, the triples $z_{\bullet}$ and $z_{\bullet}^{1}$ do not coincide on $\left(T^{\prime}, T^{\prime}+\varepsilon\right)$. Since $\left.z^{1}\right|_{t=T^{\prime}}=\left.z\right|_{t=T^{\prime}}$ belongs to $X_{p}(\tau, M, \eta)$, we can repeat the fixed point argument and obtain a contradiction. Hence problem (4.79) has at most one $\mathbb{E}$-solution. The proof of Theorem 4.33 is complete.

## APPENDIX A

## Differential geometry of hypersurfaces in $\mathbb{R}^{n}$

We provide results on hypersurfaces in the $n$-dimensional Euclidean space $\mathbb{R}^{n}$ that are used in the main part of this thesis. Kimura [Kim08] and Prüss and Simonett [PS13] developed such a theory of hypersurfaces that is applicable for moving boundary problems.

## A.1. Classes of hypersurfaces in $\mathbb{R}^{n}$

We will define hypersurfaces in terms of parametrizations over hyperplanes, where the hypersurface is locally represented as a translated and rotated graph of a scalar height function. The regularity of that surface is defined by the regularity of its height functions. We next introduce tangent vectors, normal vectors, and differential operators and characterize the regularity of a hypersurface by the regularity of its normal. According to Einstein's summation convention, we always sum over repeated greek indices $\alpha, \beta, \ldots \in\{1, \ldots, n-1\}$, whereas latin indices $i, j, \ldots \in\{1, \ldots, n-1\}$ denote free indices.
A.1. Definition. We say that $\Sigma \subset \mathbb{R}^{n}$ is a Lipschitz hypersurface or hypersurface of class $C^{1-}$, if every point $p \in \Sigma$ has a neighborhood $U$ of $p$ in $\Sigma$, such that there are a hyperplane

$$
\nu_{0}^{\perp}:=\left\{x \in \mathbb{R}^{n}: x \cdot \nu_{0}=0\right\} \quad \text { with } \nu_{0} \in \mathbb{R}^{n},\left|\nu_{0}\right|=1,
$$

a point $p_{0} \in \mathbb{R}^{n}$, a number $r>0$, and a Lipschitz function

$$
h: \nu_{0}^{\perp} \cap \bar{B}_{r}(0):=\left\{u \in \nu_{0}^{\perp}:|u| \leq r\right\} \rightarrow \mathbb{R},
$$

such that $U$ is parametrized by

$$
\varphi: \nu_{0}^{\perp} \cap B_{r}(0) \subset \nu_{0}^{\perp} \rightarrow \Sigma, \quad u \mapsto p_{0}+u+h(u) \nu_{0} .
$$

We call $\varphi$ a parametrization of $\Sigma$ over the hyperplane $\nu_{0}^{\perp}$ with height function $h$.
(i) $\Sigma$ is called $C^{k}$-hypersurface $(k \geq 1)$ or hypersurface of class $C^{k}$, if the height function in every parametrization satisfies $h \in C^{k}\left(\bar{B}_{r}(0)\right)$.
(ii) The notions of $C^{k-}$-hypersurfaces, analytic or $C^{\omega}$-hypersurfaces, and $W_{p}^{s}$-hypersurfaces are defined accordingly, where $k \geq 1, s \geq 0$, and $p \in[1, \infty]$.
(iii) $\Sigma$ is called closed resp. compact if it is closed resp. compact as a subset of $\mathbb{R}^{n}$.
A.2. Remark. Our definition of hypersurfaces exhibits the following topological properties.
(i) Clearly, every hypersurface is a $C^{1-}$-submanifold of $\mathbb{R}^{n}$ of dimension $n-1$. Therefore it may have a boundary and even an infinite number of connected components, but no selfintersections. A closed hypersurface has no boundary and a compact hypersurface has a finite number of connected components.
(ii) If $\Sigma$ is a connected compact hypersurface, then Jordan's theorem asserts that $\Sigma$ separates $\mathbb{R}^{n}$ in a bounded and an exterior domain, both having the same boundary $\Sigma$ [Bro11]. Any connected closed hypersurface $\Sigma$ separates $\mathbb{R}^{n}$ in precisely two domains [cf. Sam69].
A.3. Definition (Tangents, normal, differential operators). Let $\Sigma$ be a $C^{1}$-hypersurface in $\mathbb{R}^{n}$ and $\varphi: B_{r}(0) \subset \nu_{0}^{\perp} \rightarrow \Sigma, u \mapsto p_{0}+u+h(u) \nu_{0}$ be a local parametrization.
(i) Let $f: \Sigma \rightarrow X$ be a map with values in a Banach space $X$. For a given parametrization $\varphi$ and a basis $\left\{e_{i}\right\}_{i}$ of the hyperplane $\nu_{0}^{\perp}$, we define the partial derivatives of $f$ by

$$
\partial_{i}^{\Sigma} f(p)=\partial_{i} f(p):=\partial u_{i}(f \circ \varphi)(u) \quad \text { for } p=\varphi(u) .
$$

(ii) The tangent space $T_{p} \Sigma$ is the vector space $\left\{\varphi^{\prime}(0) u: u \in \nu_{0}^{\perp}\right\}$, its elements are the tangent vectors. In particular, $\partial_{i} \varphi(u)=: \tau_{i}(p)$ with $p=\varphi(u)$ are tangent vectors and the set $\left\{\tau_{i}(p)\right\}_{i}$ is a basis for $T_{p} \Sigma$. The dual basis $\left\{\tau^{i}(p)\right\}_{i}$ of the cotangent vectors $\tau^{i}(p)=\tau_{\Sigma}^{i}(p)$ is defined by the relation $\tau_{i}(p) \cdot \tau^{j}(p)=\delta_{i}^{j}$. We will also use the notation $\tau_{j}^{\Sigma}(p)$ to indicate the dependence on the hypersurface $\Sigma$. In terms of the parametrization $\varphi$, we can choose the tangent vectors

$$
\tau_{j}(\varphi(u))=\tau_{j}^{\Sigma}(\varphi(u))=e_{j}+\partial_{j} h(u) \nu_{0} .
$$

(iii) For a closed connected $C^{1}$-hypersurface $\Sigma \subset \mathbb{R}^{n}$, there exists a continuous unit normal field $\nu_{\Sigma}: \Sigma \rightarrow \mathbb{R}^{n}$, also called Gauss map [cf. Sam69]. Locally, the unit normal can be chosen as

$$
\begin{equation*}
\nu_{\Sigma}(\varphi(u))=\frac{\nu_{0}-\nabla_{u} h(u)}{\sqrt{1+\left|\nabla_{u} h(u)\right|^{2}}}, \tag{A.1}
\end{equation*}
$$

where the $(n-1)$-dimensional gradient $\nabla_{u} h:=e_{\alpha} \partial_{u_{\alpha}} h$ is considered as an element of $\mathbb{R}^{n}$.
(iv) The tangential projection $P_{\Sigma}(p): \mathbb{R}^{n} \rightarrow T_{p} \Sigma$ onto $T_{p} \Sigma$ is given by

$$
P_{\Sigma}=\tau^{\alpha} \otimes \tau_{\alpha}=\tau_{\alpha} \otimes \tau^{\alpha}=I-\nu_{\Sigma} \otimes \nu_{\Sigma}
$$

where $I \in \mathbb{R}^{n \times n}$ denotes the identity matrix.
(v) For a scalar function $f: \Sigma \rightarrow \mathbb{K}$, a possibly non-tangential vector field $u: \Sigma \rightarrow \mathbb{R}^{n}$, and a matrix field $S: \Sigma \rightarrow \mathbb{K}^{n \times n}$, we define the surface gradient

$$
\nabla_{\Gamma} f:=\tau^{\alpha} \partial_{\alpha} f, \quad \nabla_{\Gamma} u:=\tau^{\alpha} \otimes \partial_{\alpha} u
$$

and the surface divergence

$$
\operatorname{div}_{\Gamma} u:=\left(\partial_{\alpha} u \mid \tau^{\alpha}\right), \quad \operatorname{div}_{\Gamma} S:=\left(\partial_{\alpha} S\right) \tau^{\alpha} .
$$

If $\Sigma$ is of class $C^{2-}$, then we define the scalar Laplace-Beltrami operator

$$
\Delta_{\Gamma} f:=\operatorname{div}_{\Gamma} \nabla_{\Gamma} f=g^{\alpha \beta}\left(\partial_{\alpha} \partial_{\beta} f-\Lambda_{\alpha \beta}^{\gamma} \partial_{\gamma} f\right),
$$

where $g_{i j}=\tau_{i} \cdot \tau_{j}$ are the compontents of the Riemannian metric tensor, the components $g^{i j}$ are defined via $g^{i \alpha} g_{\alpha j}=\delta_{j}^{i}$, and $\Lambda_{i j, k}=\partial_{i} \tau_{j} \cdot \tau_{k}$ and $\Lambda_{i j}^{k}=g^{k \alpha} \Lambda_{i j, \alpha}$ are the Christoffel symbols.
A.4. Remark. We shall use further relations between height function $h$ and normal $\nu_{\Sigma}$ of a $C^{1}$ hypersurface $\Sigma$ in order to extend a given parametrization $\varphi: B_{r}(0) \subset \nu_{0}^{\perp} \rightarrow \Sigma, u \mapsto p_{0}+$ $u+h(u) \nu_{0}$. With the projection $P_{0}:=I-\nu_{0} \otimes \nu_{0}$ of $\mathbb{R}^{n}$ onto the hyperplane $\nu_{0}^{\perp}$, we obtain $\left|P_{0} \nu_{\Sigma}(\varphi(u))\right|^{2}=1-\left(\nu_{0} \mid \nu_{\Sigma}(\varphi(u))\right)^{2}$ and therefore (A.1) implies

$$
\begin{equation*}
|\nabla h(u)|^{2}=\frac{\left|P_{0} \nu_{\Sigma}(p)\right|^{2}}{1-\left|P_{0} \nu_{\Sigma}(p)\right|^{2}}=\frac{1-\left(\nu_{0} \mid \nu_{\Sigma}(p)\right)^{2}}{\left(\nu_{0} \mid \nu_{\Sigma}(p)\right)^{2}} \quad \text { for } p=\varphi(u), u \in B_{r}(0) \subset \nu_{0}^{\perp} . \tag{A.2}
\end{equation*}
$$

This shows that, if we want to extend the domain $B_{r}(0)$, we have to ensure that $\nabla h$ remains bounded, which is equivalent to $\nu_{\Sigma}(\varphi(u)) \cdot \nu_{0} \geq \eta$ for some $\eta \in(0,1]$ and all $u$, where the optimal $\eta$ and the Lipschitz constant $\|\nabla h\|_{\infty}$ are related by

$$
\eta=\left(1+\|\nabla h\|_{\infty}^{2}\right)^{-1 / 2}, \quad\|\nabla h\|_{\infty}=\left(\eta^{-2}-1\right)^{1 / 2}
$$

For the height function $h$ we obtain

$$
\begin{equation*}
\nabla h(u)=-\frac{P_{0} \nu_{\Sigma}(p)}{\nu_{0} \cdot \nu_{\Sigma}(p)} \quad \text { for } p=\varphi(u) \tag{A.3}
\end{equation*}
$$

For a fixed basis $\left\{e_{i}\right\}_{i}$ of $\nu_{0}^{\perp}$, the parametrization $\varphi$ induces the tangent vectors

$$
\tau_{j}^{\Sigma}(p)=\partial_{u_{j}} \varphi(u)=e_{j}+\partial_{j} h(u) \nu_{0}=e_{j}-\frac{e_{j} \cdot \nu_{\Sigma}(p)}{\nu_{0} \cdot \nu_{\Sigma}(p)} \nu_{0} \quad \text { for } p=\varphi(u) .
$$

Further properties of $\varphi$ in terms of the intrinsic distance are given in Proposition A.12.
In the spirit of Prüss and Simonett [PS13], we can also characterize the regularity of a hypersurface by the regularity of its normal.
A.5. Proposition. For a compact $C^{1-}$-hypersurface $\Sigma \subset \mathbb{R}^{n}$ with normal $\nu_{\Sigma} \in L_{\infty}\left(\Sigma ; \mathbb{R}^{n}\right)$ and $k \in \mathbb{N}$, the following characterizations are valid.
(i) $\Sigma$ is a $C^{k+1}$-hypersurface if and only if $\nu_{\Sigma} \in C^{k}\left(\Sigma ; \mathbb{R}^{n}\right)$.
(ii) $\Sigma$ is a $C^{k+1--h y p e r s u r f a c e ~ i f ~ a n d ~ o n l y ~ i f ~} \nu_{\Sigma} \in C^{k-}\left(\Sigma ; \mathbb{R}^{n}\right)$.
(iii) $\Sigma$ is an analytic hypersurface if and only if $\nu_{\Sigma} \in C^{\omega}\left(\Sigma ; \mathbb{R}^{n}\right)$.

Proof. We employ local coordinates $u \in U \subset \mathbb{R}^{n-1}$ and, for simplicity, we neglect the rotation and translation; that is, we assume $Q=I$ and $p_{0}=0$. Then we can express the normal $\nu_{\Sigma}$ as

$$
\begin{equation*}
\nu(u, h(u))=\beta(u)(-\nabla h(u), 1), \quad \beta(u)=\left(1+|\nabla h(u)|^{2}\right)^{-1 / 2} \quad \text { for } u \in U \subset \mathbb{R}^{n-1} . \tag{A.4}
\end{equation*}
$$

(i) If $\Sigma \in C^{k+1}$, then we have $h \in C^{k+1}$. With identity (A.1), this implies that $u \mapsto \nu(u, h(u))$ is $C^{k}$ in local coordinates, which means that $\nu_{\Sigma} \in C^{k}\left(\Sigma ; \mathbb{R}^{n}\right)$. Conversely, let $\nu_{\Sigma} \in C^{k}\left(\Sigma ; \mathbb{R}^{n}\right)$. Then $\beta=\nu \cdot e_{n}$ is $C^{k}$ and therefore also $\nabla h=-\beta^{-1} P_{1, \ldots, n-1} \nu$ belongs to $C^{k}$ by (A.3). Together with $h \in C^{1-}$ this gives $h \in C^{k+1}$.
(ii) This equivalence follows analogously.
(iii) It is sufficient to note that $h \in C^{1-}$ and $\nabla h \in C^{\omega}$ imply $h \in C^{\omega}$, since

$$
h(u)-h\left(u_{0}\right)=\int_{0}^{1} \nabla h\left(u_{0}+s\left(u-u_{0}\right)\right) \cdot\left(u-u_{0}\right) d s=\sum_{k=0}^{\infty} \frac{\nabla h^{(k)}\left(u_{0}\right)}{(k+1)!}\left(u-u_{0}\right)^{k} \cdot\left(u-u_{0}\right) .
$$

A.6. Remark. For a $C^{2-}$-hypersurface $\Sigma$ in $\mathbb{R}^{n}$, we define the Weingarten tensor

$$
L:=L_{\Sigma}:=-\nabla_{\Sigma} \nu_{\Sigma}=-\tau^{\alpha} \otimes \partial_{\alpha} \nu_{\Sigma}: \Sigma \rightarrow \mathbb{R}^{n \times n} .
$$

For every $p \in \Sigma$, the matrix $L(p)$ is symmetric and vanishes on $\mathbb{R} \nu_{\Sigma}(p)$ so that $L(p) T_{p} \Sigma \subset T_{p} \Sigma$. The Weingarten tensor induces the second fundamental form $I I_{p}(v, w)=L_{\Sigma}(p) v \cdot w=l_{\alpha \beta}(p) v^{\alpha} w^{\beta}$ for $v, w \in T_{p} \Sigma$. The eigenvalues $\kappa_{j}(p)$ of $L(p)$ are the principal curvatures of $\Sigma$ at $p$ and the corresponding eigenvectors are the principal directions [Kim08, Theorem 2.10]. For every $C^{2-}{ }^{2}$ path $\gamma:[a, b] \rightarrow \Sigma$ with $\left|\gamma^{\prime}(t)\right|=1$ for all $t \in[a, b]$, the curvature of $\gamma$ at $\gamma(t)$ is $\gamma^{\prime \prime}(t)$ and we have $\left|\gamma^{\prime \prime}(t)\right| \leq|L(\gamma(t))|$. The $(n-1)$-fold mean curvature is given by

$$
H_{\Sigma}:=\kappa_{1}+\cdots+\kappa_{n-1}=\operatorname{tr} L_{\Sigma}=-\operatorname{div}_{\Sigma} \nu_{\Sigma} .
$$

The Christoffel symbols satisfy the following relations [PS13, (12), (14)],

$$
\partial_{i} \tau_{j}=\Lambda_{i j}^{\alpha} \tau_{\alpha}+l_{i j} \nu, \quad \partial_{i} \tau^{j}=-\Lambda_{i \alpha}^{j} \tau^{\alpha}+l_{i}^{j} \nu .
$$

We recall a well-known fact from differential geometry.
A.7. Theorem (see e. g. [Ale62]). Let $\Sigma$ be a compact connected $C^{2}$-hypersurface in $\mathbb{R}^{n}(n \geq 2)$ with constant mean curvature. Then $\Sigma$ is a sphere.

## A.2. The intrinsic distance of a hypersurface

In this thesis, a $C^{1-}$-hypersurface $\Sigma$ is equipped with the Riemannian metric induced by the scalar product of $\mathbb{R}^{n}$; that is, for $p \in \Sigma$, a scalar product in $T_{p} \Sigma$ is defined by $(\tau \mid \tilde{\tau})_{p}:=\tau \cdot \tilde{\tau}$ and therefore $T_{p} \Sigma$ has the induced norm $|\tau|=\left(\sum_{j=1}^{n}\left(e_{j} \cdot \tau\right)^{2}\right)^{1 / 2}$. The intrinsic distance dist $\Sigma(p, q)$ for $p, q \in \Sigma$ is defined as the infimum of the lengths of all $C^{1-}$-curves in $\Sigma$ joining $p$ to $q$. In

Proposition A.12, we prove that the intrinsic distance and the induced norm of $\mathbb{R}^{n}$ are equivalent for compact connected $C^{1-}$-hypersurfaces. If $\Sigma$ is of class $C^{3-}$, then we can find a curve $\gamma$ from $p$ to $q$ with minimal length $l(\gamma)=\operatorname{dist}_{\Sigma}(p, q)$, which is a minimizing geodesic [see e. g. Car92, Chapter 3].
A.8. Remark. Geodesics can be defined for every $C^{3-}$-hypersurface $\Sigma \subset \mathbb{R}^{n}(n \geq 2)$. Let $\gamma:[a, b] \rightarrow \Sigma$ be a $C^{1}$-curve and let $v:[a, b] \rightarrow T \Sigma$ be a tangential vector field along $\gamma$; that is, $v(t)$ belongs to the $T_{\gamma(t)} \Sigma$ and therefore has a representation $v(t)=v^{\alpha}(t) \tau_{\alpha}(\gamma(t))$. The covariant derivative of $v$ along $\gamma$ is defined by

$$
\frac{D v(t)}{D t}:=P_{\Sigma}(\gamma(t)) \frac{d v(t)}{d t}=\partial_{t} v^{\alpha}(t) \tau_{\alpha}(\gamma(t))+v^{\alpha}(t) \partial_{t} \gamma^{\beta}(t) P_{\Sigma}(\gamma(t)) \partial_{\beta} \tau_{\alpha}(\gamma(t))
$$

We call a $C^{2}$-curve $\gamma:[a, b] \rightarrow \Sigma$ a geodesic, if it satisfies the geodesic equation

$$
\frac{D}{D t} \frac{d \gamma}{d t}=\left(P_{\Sigma} \circ \gamma\right) \gamma^{\prime \prime}=0 \quad \text { in }(a, b) .
$$

In local coordinates $x^{i}(t)=e^{i} \cdot \varphi^{-1}(\gamma(t))$, the geodesic equation becomes a system of ordinary differential equations

$$
\begin{equation*}
\frac{d^{2} x^{i}}{d t^{2}}+\left(\Lambda_{\alpha \beta}^{i} \circ x\right) \frac{d x^{\alpha}}{d t} \frac{d x^{\beta}}{d t}=0 . \tag{A.5}
\end{equation*}
$$

Here the Christoffel symbols $\Lambda_{i j}^{k}=g^{k \alpha}\left(\partial_{i} \tau_{j} \mid \tau_{\alpha}\right)$ are locally Lipschitz continuous. From the theory of ordinary differential equations, we infer that (A.5) has a unique local $C^{3-}$-solution $t \mapsto x(t)=x\left(t ; x_{0}, x_{1}\right)$ that satisfies prescribed initial conditions $x\left(t_{0}\right)=x_{0}$ and $x^{\prime}\left(t_{0}\right)=x_{1}$. Moreover, $x\left(t ; x_{0}, x_{1}\right)$ depends continuously on ( $x_{0}, x_{1}$ ). Consequently, for given $p \in \Sigma$ and $v \in T_{p} \Sigma$, there exists a unique geodesic $t \mapsto \gamma(t ; p, v)$ such that $\gamma(0 ; p, v)=p$ and $\gamma^{\prime}(0 ; p, v)=v$.

The geodesics are homogeneous in the sense that for a geodesic $\gamma(\cdot ; p, v)$ on $(-\delta, \delta)$ and every $\alpha>0$, the map $t \mapsto \gamma(\alpha t ; p, v)$ is also a geodesic on $(-\delta / \alpha, \delta / \alpha)$ and the identity $\gamma(\alpha t ; p, v)=\gamma(t ; p, \alpha v)$ applies to all $t \in(-\delta, \delta)$ [Car92, Lemma 3.2.6]. Moreover, the map

$$
(t,(p, v)) \mapsto\left(\gamma(t ; p, v), \gamma^{\prime}(t ; p, v)\right)
$$

is a local flow on $T \Sigma$, called the geodesic flow. We say that $\Sigma$ is (geodesically) complete, if every geodesic $\gamma:[a, b] \rightarrow \Sigma$ can be extended to a geodesic $\tilde{\gamma}: \mathbb{R} \rightarrow \Sigma$. In this case, the geodesic flow is global with respect to $t \in \mathbb{R}, p \in \Sigma$, and $v \in T_{p} \Sigma$.

A geodesic locally minimizes the distance between two points in the sense that for every $t \in[a, b]$ there is $\varepsilon>0$ such that $\operatorname{dist}_{\Sigma}\left(\gamma\left(t_{1}\right), \gamma\left(t_{2}\right)\right)=\int_{t_{1}}^{t_{2}}\left|\gamma^{\prime}(s)\right| d s$ for all $t_{1}, t_{2} \in[a, b] \cap(t-\varepsilon, t+\varepsilon)$ with $t_{1}<t_{2}$ [Car92, Remark 3.3.8]. Conversely, if $p, q \in \Sigma$ are given, then every piecewise differentiable curve joining $p$ to $q$ with minimal length is a geodesic [Car92, Corollary 3.3.9].
A.9. Remark. For a $C^{3-}$-hypersurface $\Sigma$, we define the exponential map

$$
\exp _{p}(v):=\gamma(1 ; p, v)=\gamma(|v| ; p, v /|v|) \quad \text { for } p \in \Sigma, v \in B_{r}(0) \subset T_{p} \Sigma
$$

for some $r>0$ [see Car92, Chapters 3,13]. In view of

$$
\left(\frac{d \exp _{p}}{d v}(0)\right) v=\left.\frac{d}{d t} \exp _{p}(t v)\right|_{t=0}=\left.\frac{d}{d t} \gamma(t ; p, v)\right|_{t=0}=\gamma^{\prime}(0 ; p, v)=v,
$$

we see that $(d / d v) \exp _{p}(0)=P_{\Sigma}(p)$. Therefore, by the inverse function theorem, $\exp _{p}$ is a local $C^{1}$-diffeomorphism at $0 \in T_{p} \Sigma$ into $\Sigma$. The number

$$
i(\Sigma, p)=\sup \left\{r>0: \exp _{p}: B_{r}(0) \subset T_{p} \Sigma \rightarrow \exp _{p}\left(B_{r}(0)\right) \subset \Sigma \text { is a diffeomorphism }\right\}
$$

is called the injectivity radius of $\Sigma$ at $p$ and $i(\Sigma):=\inf \{i(\Sigma, p): p \in \Sigma\}$ is the injectivity radius of $\Sigma$. Clearly, if $\Sigma$ is compact, then $i(\Sigma)>0$. If $q \in \exp _{p}\left(B_{i(\Sigma)}(0)\right)$, then there exists a unique geodesic joining $p$ and $q$ that minimizes $\operatorname{dist}_{\Sigma}(p, q)$ [Car92, Corollary 13.2.8].

The Hopf-Rinow theorem [HR31] characterizes geodesic completeness of general $C^{3-}$-class Riemannian manifolds.
A.10. Theorem (Hopf-Rinow [cf. Car92, Theorem 7.2.8]). Let $M$ be a Riemannian $C^{3-}$-manifold. Then the following assertions are equivalent.
(i) The exponential map $\exp _{p}$ is defined on all of $T_{p} M$ for every $p \in M$.
(ii) Every closed and bounded subset of $M$ is compact.
(iii) The metric space $\left(M, \operatorname{dist}_{M}(\cdot, \cdot)\right)$ is complete.
(iv) $M$ is geodesically complete.
(v) There exists a sequence of compact subsets $K_{j} \subset M$ with $K_{j} \subset K_{j+1}$ and $\bigcup_{j} K_{j}=M$ such that if $q_{j} \notin K_{j}$, then $\operatorname{dist}_{M}\left(p, q_{j}\right) \rightarrow \infty$ for every $p \in M$.
If, in addition, $M$ is connected, then any of the statements above implies that
(vi) For any $p, q \in M$, there exists a geodesic $\gamma$ joining $p$ and $q$ with $l(\gamma)=\operatorname{dist}_{M}(p, q)$.
A.11. Corollary ([Car92, Corollaries 7.2.9, 7.2.10]). Every compact Riemannian $C^{3-}$-manifold is complete and every closed $C^{3-}$-submanifold of a complete Riemannian $C^{3-}$-manifold is complete in the induced metric. In particular, every closed $C^{3-}$-hypersurface in $\mathbb{R}^{n}(n \geq 2)$ is complete.

The following relations between intrinsic distance and Euclidean distance will be used later on for dealing with the intrinsic Slobodeckiĭ semi-norm.
A.12. Proposition. Let $\Sigma \subset \mathbb{R}^{n}$ be a $C^{1-}$-hypersurface.
(i) Let $p \in \Sigma$ be fixed, let $\varphi: T_{p} \Sigma \supset U \rightarrow \Sigma, u \mapsto p+u+h(u) \nu_{\Sigma}(p)$ be a parametrization over the tangent space where $U$ is convex, and let $h \in C^{1-}(\bar{U})$ so that $\|\nabla h\|_{\infty}<\infty$. Then

$$
|u-v| \leq|\varphi(u)-\varphi(v)| \leq \operatorname{dist}_{\Sigma}(\varphi(u), \varphi(v)) \leq\left(1+\|\nabla h\|_{\infty}^{2}\right)^{1 / 2}|u-v| \quad \text { for all } u, v \in U .
$$

(ii) Let $\Sigma \in C^{1-}$ be compact and connected. Then $\operatorname{dist}_{\Sigma}(\cdot, \cdot)$ is bounded and for some $C \geq 1$ we have

$$
|p-q| \leq \operatorname{dist}_{\Sigma}(p, q) \leq C|p-q| \quad \text { for all } p, q \in \Sigma .
$$

(iii) Let $\Sigma \in C^{2-}$ and $\left\|L_{\Sigma}\right\|_{\infty}<\infty$. If $p, q \in \Sigma$ satisfy $\operatorname{dist}_{\Sigma}(p, q)<\sqrt{2}\left\|L_{\Sigma}\right\|_{\infty}^{-1}$, then

$$
\left|\nu_{\Sigma}(p)-\nu_{\Sigma}(q)\right| \leq \operatorname{dist}_{\Sigma}(p, q)\left\|L_{\Sigma}\right\|_{\infty}, \quad \nu_{\Sigma}(p) \cdot \nu_{\Sigma}(q)>0 .
$$

(iv) Let $\Sigma \in C^{3-}$ and $\left\|L_{\Sigma}\right\|_{\infty}<\infty$. If $p, q \in \Sigma$ satisfy $\operatorname{dist}_{\Sigma}(p, q)<2\left\|L_{\Sigma}\right\|_{\infty}^{-1}$, then

$$
|p-q| \leq \operatorname{dist}_{\Sigma}(p, q) \leq \frac{|p-q|}{1-\frac{1}{2} \operatorname{dist}_{\Sigma}(p, q)\left\|L_{\Sigma}\right\|_{\infty}}
$$

Proof. (i) For almost all $u \in U$, we have $\varphi^{\prime}(u)=P_{\Sigma}(p)+\nu_{\Sigma}(p) \otimes \nabla h(u)$ and $\left|\varphi^{\prime}(u)\right|^{2}=$ $1+|\nabla h(u)|^{2}$. The map $[0,1] \mapsto \varphi(u+(v-u) t)$ is a curve from $\varphi(u)$ to $\varphi(v)$ in $\Sigma$ and therefore

$$
\begin{aligned}
& \operatorname{dist}_{\Sigma}(\varphi(u), \varphi(v)) \leq\left\|\varphi^{\prime}\right\|_{\infty}|u-v| \leq\left(1+\|\nabla h\|_{\infty}^{2}\right)^{1 / 2}|u-v| \\
& \operatorname{dist}_{\Sigma}(\varphi(u), \varphi(v)) \geq|\varphi(u)-\varphi(v)|=\left(|u-v|^{2}+\left|(h(u)-h(v)) \nu_{\Sigma}(p)\right|^{2}\right)^{1 / 2} \geq|u-v| .
\end{aligned}
$$

(ii) For every $p \in \Sigma$, there exists $r(p)>0$ such that $\Sigma \cap B_{r(p)}(p)$ can be parametrized over $T_{p} \Sigma$ via $\varphi_{p}(u)=p+u+h_{p}(u) \nu_{\Sigma}(p)$, where $h$ satisfies $\left\|\nabla h_{p}\right\|_{\infty} \leq 1$. From (i) we obtain that

$$
|q-\tilde{q}| \leq \operatorname{dist}_{\Sigma}(q, \tilde{q}) \leq \sqrt{2}|q-\tilde{q}| \quad \text { for } q, \tilde{q} \in \Sigma \cap B_{r(p)}(p) .
$$

In particular, we have $\operatorname{dist}_{\Sigma}(p, q)<\sqrt{2} r(p)$ for every $q \in \Sigma \cap B_{r(p)}(p)$. By compactness, there exist finitely many sets $\Sigma \cap B_{r_{j}}(p)$ with the above properties and $r_{j}=r\left(p_{j}\right)$ such that $\Sigma$ is the union of these sets. Since $\Sigma$ is connected, the numbers $\operatorname{dist}_{\Sigma}\left(p_{j}, p_{k}\right)$ have a finite maximum $R$. Therefore $\operatorname{dist}_{\Sigma}(\cdot, \cdot)$ is bounded by $R+2 \sqrt{2} \max r_{j}$.

Assume that the assertion is false. Then there exist $p_{n}, q_{n} \in \Sigma$ with $\operatorname{dist}_{\Sigma}\left(p_{n}, q_{n}\right)>n\left|p_{n}-q_{n}\right|$. Since $\Sigma$ is compact we may assume that $p_{n} \rightarrow p \in \Sigma$ and $q_{n} \rightarrow q \in \Sigma$ and since dist $\Sigma$ is bounded, the $p_{n}$ and $q_{n}$ converge to the same limit $p=q$. But then almost all $p_{n}, q_{n}$ are contained in some
$\Sigma \cap B_{r_{0}}\left(p_{0}\right)$ and hence $n\left|p_{n}-q_{n}\right|<\operatorname{dist}_{\Sigma}\left(p_{n}, q_{n}\right) \leq \sqrt{2}\left|p_{n}-q_{n}\right|$, a contradiction. Therefore (ii) is valid.
(iii) For a curve $\gamma$ joining $q$ to $p$ in $\Sigma$, we have

$$
\left|\nu_{\Sigma}(p)-\nu_{\Sigma}(q)\right|=\left|\int_{0}^{l(\gamma)} \frac{d}{d t} \nu_{\Sigma}(\gamma(t)) d t\right|=\left|\int_{0}^{l(\gamma)} L_{\Sigma}(\gamma(t)) \gamma^{\prime}(t) d t\right| \leq l(\gamma)\left\|L_{\Sigma}\right\|_{\infty}
$$

The inequality $\nu_{\Sigma}(p) \cdot \nu_{\Sigma}(q)>0$ is valid if and only if $\left|\nu_{\Sigma}(p)-\nu_{\Sigma}(q)\right|^{2}=2-2 \nu_{\Sigma}(p) \cdot \nu_{\Sigma}(q)<2$ and this is true if $\operatorname{dist}_{\Sigma}(p, q)<\sqrt{2}\left\|L_{\Sigma}\right\|_{\infty}^{-1}$.
(iv) Let $\gamma$ be a geodesic from $q$ to $p$ with minimal length $l(\gamma)=\operatorname{dist}_{\Sigma}(p, q)$ and $\left|\gamma^{\prime}\right|=1$. Since $P_{\Sigma}(\gamma(t)) \gamma^{\prime \prime}(t)=0$ and $\nu_{\Sigma}(\gamma(t)) \cdot \gamma^{\prime}(t)=0$, we obtain $\gamma^{\prime \prime}(t)=\nu_{\Sigma}(\gamma(t))\left(L_{\Sigma}(\gamma(t)) \gamma^{\prime}(t) \cdot \gamma^{\prime}(t)\right)$. Hence

$$
\begin{aligned}
p-q & =\int_{0}^{l(\gamma)} \gamma^{\prime}(t) d t=l(\gamma) \gamma^{\prime}(0)+\int_{0}^{l(\gamma)} \int_{0}^{t} \gamma^{\prime \prime}(s) d s d t \\
|p-q| & \geq l(\gamma)-\int_{0}^{l(\gamma)} \int_{0}^{t}\left|L_{\Sigma}(\gamma(s))\right| d s d t \geq l(\gamma)-\frac{l(\gamma)^{2}}{2}\left\|L_{\Sigma}\right\|_{\infty}
\end{aligned}
$$

This yields the assertion.
The next result provides parametrizations that are defined on balls with uniform radius.
A.13. Proposition. Let $\Sigma \subset \mathbb{R}^{n}(n \geq 2)$ be a $C^{2-}$-hypersurface such that $L_{\Sigma}$ is bounded and put $R_{*}:=\sqrt{2}\left\|L_{\Sigma}\right\|_{\infty}^{-1} \in(0, \infty], \delta(R):=1-R^{2}\left\|L_{\Sigma}\right\|_{\infty}^{2} / 2 \in(0,1]$, and $r(R):=R \delta(R)$ for $R \in\left(0, R_{*}\right)$. Then for every $x \in \Sigma$, there exists a parametrization

$$
\varphi_{x}: B_{r(R)}(0) \subset T_{x} \Sigma \rightarrow B_{R}^{\Sigma}(x), \quad u \mapsto x+u+h_{x}(u) \nu_{\Sigma}(x)
$$

with height function $h_{x} \in C^{2-}\left(\overline{B_{r(R)}(0)}\right)$ such that $h_{x}(0)=\left|\nabla h_{x}(u)\right|=0$.
Proof. Given $x \in \Sigma$, there exists a parametrization $\varphi_{x}: V_{x} \rightarrow \Sigma$ on some small neighborhood $V_{x}$ of the origin such that $\varphi_{x}(u)=x+u+h_{x}(u) \nu_{\Sigma}(x)$ for some $h_{x} \in C^{2-}\left(\overline{V_{x}}\right)$ with $h_{x}(0)=$ $\left|\nabla h_{x}(0)\right|=0$. Our goal is to show that $\varphi_{x}$ can be extended to map with the asserted properties. Such an extension is uniquely determined by the representation (A.3) of $\nabla h_{x}$ in terms of $\nu_{\Sigma}$. The identity (A.2) shows that

$$
\left|\nabla h_{x}(u)\right|^{2}=\left(\nu_{\Sigma}(x) \mid \nu_{\Sigma}\left(\varphi_{x}(u)\right)\right)^{-2}-1 \quad \text { for } u \in \overline{V_{x}}
$$

With Remark A. 4 we obtain

$$
\left(\nu_{\Sigma}(x) \mid \nu_{\Sigma}\left(\varphi_{x}(u)\right)\right) \geq \delta, \quad\left|\nabla h_{x}(u)\right|^{2} \leq \frac{1-\delta^{2}}{\delta^{2}} \quad \text { for all } u \in \overline{V_{x}}, x \in \Sigma
$$

Therefore we can extend $h_{x}$ and $\varphi_{x}$ uniquely onto $\overline{B_{r(R)}(0)}$.

## A.3. Neighborhoods of hypersurfaces

We show that every compact hypersurface of class $C^{2-}$ satisfies a uniform ball condition and has a tubular neighborhood with uniform thickness. Within such a neighborhood, we study further hypersurfaces that are induced by height functions. For the corresponding diffeomorphism between the original and the new hypersurface, we derive an integral transformation formula that does not use local coordinates (see (A.12)). We also provide a level function for a possibly disconnected compact hypersurface.
A.14. Definition. A hypersurface $\Sigma \subset \mathbb{R}^{n}$ satisfies the ball condition of radius $r>0$ at the point $p \in \Sigma$, if the open balls $B_{r}\left(p-r \nu_{\Sigma}(p)\right)$ and $B_{r}\left(p+r \nu_{\Sigma}(p)\right)$ do not intersect $\Sigma$. We say that $\Sigma$ satisfies the uniform ball condition of radius $r>0$, if it satisfies the ball condition of the same radius $r$ at every $p \in \Sigma$.
A.15. Remark. Let $S$ be a closed subset of $\mathbb{R}^{n}$ of the form $S=\mathbb{R}^{n} \backslash\left(\Omega_{+} \cup \Omega_{-}\right)$with disjoint open subsets $\Omega_{+}$and $\Omega_{-}$of $\mathbb{R}^{n}$. As in [Kim08, Section 3.1], we define the signed distance

$$
d(x)= \begin{cases}\operatorname{dist}(x, S) & \text { for } x \in \Omega_{+} \\ 0 & \text { for } x \in S \\ -\operatorname{dist}(x, S) & \text { for } x \in \Omega_{-}\end{cases}
$$

By [Kim08, Theorem 3.2], both maps $\operatorname{dist}(\cdot, S)$ and $d(\cdot)$ are Lipschitz continuous.
A.16. Definition ([cf. PS13]). A hypersurface $\Sigma \subset \mathbb{R}^{n}$ has a tubular neighborhood of radius $r>0$, if the map

$$
\begin{equation*}
X:(p, t) \mapsto p+t \nu_{\Sigma}(p), \quad \Sigma \times(-r, r) \rightarrow B_{r}(\Sigma):=\left\{x \in \mathbb{R}^{n}: \operatorname{dist}(x, \Sigma)<r\right\} \tag{A.6}
\end{equation*}
$$

is a homeomorphism; that is, $X$ is continuous and bijective and therefore has a continuous inverse. We say that the tubular neighborhood is of class $C^{k}(k \geq 1)$, if $X$ is a $C^{k}$-diffeomorphism; that is, $X$ is of class $C^{k}$ and $\partial X(p, t): T_{p} \Sigma \times \mathbb{R} \rightarrow \mathbb{R}^{n}$ is invertible for all $(p, t)$. We decompose

$$
X^{-1}(x)=(\Pi(x), d(x)) \quad \text { with } \Pi(x)=p \in \Sigma, d(x)=t \in(-r, r), x=\Pi(x)+d(x) \nu_{\Sigma}(\Pi(x))
$$

A.17. Proposition. The following assertions are valid.
(i) A closed hypersurface $\Sigma \subset \mathbb{R}^{n}$ has a tubular neighborhood of radius $r$ if and only if it satisfies the uniform ball condition of radius $r$. In this case, it also has a tubular neighborhood of radius

$$
r_{\Sigma}:=\sup \{r>0: \Sigma \text { has a tubular neighborhood of radius } r\}
$$

(ii) If $\Sigma$ is a compact $C^{2-}$-hypersurface in $\mathbb{R}^{n}$, then it has a tubular neighborhood of radius $r_{\Sigma}>0$, the homeomorphism $X$ in (A.6) has a locally Lipschitz continuous inverse and the principal curvatures of $\Sigma$ and the Weingarten map $L_{\Sigma}$ are bounded by $1 / r_{\Sigma}$ almost everywhere.
(iii) If $\Sigma$ is a compact $C^{k+1}$-hypersurface $(k \geq 1)$, then $X$ is a $C^{k}$-diffeomorphism with derivative

$$
\begin{equation*}
\partial_{(p, t)} X(p, t)(v, s)=v+s \nu_{\Sigma}(p)-t L_{\Sigma}(p) v, \quad(p, t) \in \Sigma \times(-r, r), \quad(v, s) \in T_{p} \Sigma \times \mathbb{R} \tag{A.7}
\end{equation*}
$$

Proof. (i) Ball condition $\Rightarrow$ tubular neighborhood. Suppose that $\Sigma$ satisfies the uniform ball condition of radius $r$. It remains to show that the continuous map $X:(p, t) \mapsto p+t \nu_{\Sigma}(p)$, $\Sigma \times(-r, r) \rightarrow B_{r}(\Sigma):=\left\{x \in \mathbb{R}^{n}: \operatorname{dist}(x, \Sigma)<r\right\}$ is bijective.

Surjectivity. Given $x \in B_{r}(\Sigma)$, we put $\delta:=\operatorname{dist}(x, \Sigma)<r$. Then $B_{\delta}(x) \cap \Sigma$ is empty and, since $\Sigma$ is closed, there exists $p \in \Sigma$ with $|x-p|=\delta$, so that $p \in \partial B_{\delta}(x)$. Since none of the balls $B_{\delta^{\prime}}(x)\left(\delta^{\prime}<\delta\right)$ intersects $\Sigma$, the vector $p-x$ is normal to $T_{p} \Sigma$ and hence $x$ belongs to the image of $X$.

Injectivity. Suppose that for some $x \in B_{r}(\Sigma)$ there are two different points $p, q \in \Sigma$ such that $X(p, s)=X(q, t)=x$ for some $s, t \in(-r, r)$. Then we must have $|s|=|p-x|=|q-x|=|t|$. Therefore the balls $B_{\sigma}\left(p+\sigma \nu_{\Sigma}(p)\right)$ with $\sigma \in(s, r)$ if $s>0$ and $\sigma \in(-r, s)$ if $s<0$ are tangent to $\Sigma$ at $p$ and contain the point $q \in \Sigma$. But this means $\Sigma \cap B_{r}\left(p+(\operatorname{sign} s) r \nu_{\Sigma}(p)\right) \neq \emptyset$, a contradiction. Hence $q=p$. The map $X$ is therefore bijective and continuous and therefore $\Sigma$ has a tubular neighborhood of radius $r$.

Tubular neighborhood $\Rightarrow$ ball condition. Suppose that $\Sigma$ has a tubular neighborhood of radius $r$. We show that none of the balls $B_{r}\left(p \pm r \nu_{\Sigma}(p)\right)(p \in \Sigma)$ intersects $\Sigma$. Assuming the contrary, there exist two different $p, q \in \Sigma$ such that $q \in B_{r}\left(x_{0}\right)$ with $x_{0}:=p+r \nu_{\Sigma}(p)$ (the case $p-r \nu_{\Sigma}(p)$ can be handled analogously). Then for some $\delta<r$, the ball $B_{\delta}\left(x_{0}\right)$ touches $\Sigma$; that is, there exists $q_{0} \in \Sigma \cap \partial B_{\delta}\left(x_{0}\right)$ with $\delta:=\left|q_{0}-x_{0}\right|=\operatorname{dist}\left(x_{0}, \Sigma\right)<r$ and thus $T_{q_{0}} \Sigma$ coincides with $T_{q_{0}} \partial B_{\delta}\left(x_{0}\right)$ and hence $x_{0}=q_{0}+\delta \nu_{\Sigma}\left(q_{0}\right)$. But since $X: \Sigma \times(-r, r) \rightarrow B_{r}(\Sigma)$ is bijective, we have $q_{0}=\Pi\left(x_{0}\right)=p$ which implies $p \in \overline{B_{\delta}}\left(p+r \nu_{\Sigma}(p)\right)$, a contradiction to $p \notin B_{r}\left(p+\nu_{\Sigma}(p)\right)$. Hence $\Sigma$ satisfies the uniform ball condition of radius $r$.

The number $r_{\Sigma}$. Suppose that $B_{r}\left(p+\operatorname{sr} \nu_{\Sigma}(p)\right) \cap \Sigma=\emptyset$ for all $p \in \Sigma, s \in\{-1,1\}$, and $r \in\left(0, r_{\Sigma}\right)$. For fixed $s \in\{-1,1\}$ and $p \in \Sigma$, the ball $B_{r_{\Sigma}}\left(p+s r_{\Sigma} \nu_{\Sigma}(p)\right)$ is the union of the
balls $B_{r}\left(p+\operatorname{sr} \nu_{\Sigma}(p)\right)\left(r \in\left(0, r_{\Sigma}\right)\right)$ and hence we obtain $B_{r_{\Sigma}}\left(p+s r_{\Sigma} \nu_{\Sigma}(p)\right) \cap \Sigma=\emptyset$. Therefore $\Sigma$ satisfies the uniform ball condition of radius $r_{\Sigma}$ and has a tubular neighborhood of radius $r_{\Sigma}$.
(ii) We show that every compact $C^{2-}$-hypersurface $\Sigma$ satisfies a uniform ball condition.
(ii.a) For given $\varepsilon>0$, we first prove that there is a number $\delta>0$ such that every surface piece $\Sigma \cap B_{\delta}\left(p_{0}\right)(p \in \Omega)$ satisfies a ball condition at $p_{0}$. It is sufficient to consider the case $p_{0}=0$ and $\nu_{\Sigma}\left(p_{0}\right)=e_{n}=(0, \ldots, 0,1)$, since the assertion for general $p_{0} \in \Sigma$ and $\nu_{\Sigma}\left(p_{0}\right)$ follows by using an appropriate rotation and translation. We choose a local coordinate system $\Sigma \cap B_{\delta}(0)=$ $\left\{(u, h(u)): u \in U_{\delta}\right\}$ for some open subset $U_{\delta} \subset \mathbb{R}^{n-1}$ and a $C^{1,1}$-function $h: U_{\delta} \rightarrow \mathbb{R}$. The lower half sphere of $\partial B_{r}\left(r e_{n}\right)$ and the upper half sphere of $\partial B_{r}\left(-r e_{n}\right)$ are parametrized by $\left(u, k_{r}(u)\right)$ and $\left(u,-k_{r}(u)\right)$, respectively, where $k_{r}$ is defined by

$$
k_{r}(u):=r-\sqrt{r^{2}-|u|^{2}} \quad \text { for } u \in B_{r}(0) \subset \mathbb{R}^{n-1} .
$$

The ball condition requires that the following inequality is satisfied.

$$
\begin{equation*}
-k_{r}(u) \leq h(u) \leq k_{r}(u) \quad \text { for } u \in B_{r}(0) \cap U_{\delta} \tag{A.8}
\end{equation*}
$$

In order to guarantee this condition, we seek a sufficient upper bound for the radius $r$. Fix $v \in \mathbb{R}^{n-1}$ with $|v|=1$ and consider the rescaled functions $\tilde{h}(s):=h(r s v)$ and $\tilde{k}(s):=k_{r}(r s v)=$ $r\left(1-\left(1-s^{2}\right)^{1 / 2}\right)$ for $s \in(-1,1)$. Then $\tilde{h}(0)=\tilde{h}^{\prime}(0)=\tilde{k}(0)=\tilde{k}^{\prime}(0)=0$ and $\tilde{h}^{\prime \prime}(s)=r^{2} v$. $\left(\nabla^{2} h(s r v)\right) v$ and $\tilde{k}^{\prime \prime}(s)=r\left(1-s^{2}\right)^{-3 / 2}$. For $t \in(0,1)$ this yields

$$
\begin{aligned}
\tilde{k}(t)-\tilde{h}(t) & =\int_{0}^{t} \int_{0}^{s}\left(\tilde{k}^{\prime \prime}\left(s^{\prime}\right)-\tilde{h}^{\prime \prime}\left(s^{\prime}\right)\right) d s^{\prime} d s \\
& =\int_{0}^{t}(t-s)\left(r\left(1-s^{2}\right)^{-3 / 2}-r^{2} v \cdot\left(\nabla^{2} h(s r v)\right) v\right) d s \\
& \geq \int_{0}^{t}(t-s)\left(r-r^{2}\left|\nabla^{2} h(s r v)\right|\right) d s .
\end{aligned}
$$

The integrand is non-negative if we choose $r \leq \sup \left\{\left|\nabla^{2} h(u)\right|: u \in U_{\delta}\right\}^{-1}$ and $\delta>0$ such that the supremum is finite. In this case both $\tilde{k}-\tilde{h}$ and $\tilde{k}+\tilde{h}$ are non-negative and hence the local ball condition (A.8) is satisfied. This means that $\Sigma \cap B_{\delta}$ satisfies the ball condition at $p_{0}=0$.
(ii.b) Next we prove an estimate of $\sup \left\{\left|\nabla^{2} h(u)\right|: u \in U_{\delta}\right\}$ in terms of the global quantity $\left\|L_{\Sigma}\right\|_{\infty}$. We may again assume that $p_{0}=0$ and $\nu_{\Sigma}\left(p_{0}\right)=e_{n}$. Letting where $p=(u, h(u))$, we want to express the local map $U_{\delta} \ni u \mapsto \nabla^{2} h(u)$ in terms of the global map $\Sigma \ni p \mapsto$ $L_{\Sigma}(p)=-\nabla_{\Sigma} \nu_{\Sigma}(p)$. Let $\nu(u):=\nu_{\Sigma}(u, h(u))$ and $L(u):=L_{\Sigma}(u, h(u))$. From (A.4) we obtain $\nu(u)=\left(\nu(u) \mid e_{n}\right)\left(e_{n}-\nabla h(u)\right)$ and therefore $P \nu(u)=-\left(\nu(u) \mid e_{n}\right) \nabla h(u)$ with $P:=I-e_{n} \otimes e_{n}$. The identities $h(0)=0$ and $\nabla h(0)=0$ yield $\nabla_{\Sigma} \nu_{\Sigma}\left(p_{0}\right)=\nabla \nu(0)=\nabla^{2} h(0)$. Moreover,

$$
-L(u)=\nabla_{\Sigma} \nu_{\Sigma}(p)=\tau_{\Sigma}^{j}(p) \otimes \partial_{u_{j}} \nu_{\Sigma}(u, h(u))=\nabla \nu(u)+\left(\tau_{\Sigma}^{j}(p)-e_{j}\right) \otimes \partial_{j} \nu(u),
$$

where $\tau_{\Sigma}^{j}(p)$ tends to $e^{j}$ as $p \rightarrow p_{0}$. A straightforward computation gives

$$
\begin{equation*}
\nabla^{2} h(u)=\nabla\left(-\frac{P \nu(u)}{\nu(u) \cdot e_{n}}\right)=-\frac{P \nabla \nu(u) P}{\nu(u) \cdot e_{n}}+\frac{P \nu(u) \otimes\left((\nabla \nu(u)) e_{n}\right)}{\left(\nu(u) \mid e_{n}\right)^{2}} . \tag{A.9}
\end{equation*}
$$

Since $\Sigma$ is compact, the quantities $\nu_{\Sigma}$ and $\tau_{\Sigma}^{j}$ are uniformly continuous and thus $\nu(u) \rightarrow \nu(0)$ and $\tau_{\Sigma}^{j}(p) \rightarrow e_{j}$ as $p$ tends to $p_{0}$, uniformly with respect to $p_{0} \in \Sigma$. Hence for some $\delta(\varepsilon)>0$ and all $p_{0}$, we have

$$
\left\|\nabla^{2} h\right\|_{L_{\infty}\left(U_{\delta}\right)} \leq\left\|\nabla_{\Sigma} \nu_{\Sigma}\right\|_{L_{\infty}(\Sigma)}+\varepsilon .
$$

(ii.c) The previous steps imply that for all $r<\left\|\nabla_{\Sigma} \nu_{\Sigma}\right\|_{\infty}^{-1}=\left\|L_{\Sigma}\right\|_{\infty}^{-1}$, there is a number $\delta=\delta(r)>0$ such that every part $\Sigma \cap B_{\delta}\left(p_{0}\right)\left(p_{0} \in \Sigma\right)$ satisfies the uniform ball condition of radius $r$ at the point $p_{0}$. Since $\Sigma$ is compact, there exists $r \in\left(0,\left\|L_{\Sigma}\right\|^{-1}\right)$ such that $\Sigma$ satisfies the uniform ball condition of radius $r$.

Next, we prove the local Lipschitz continuity of $X^{-1}=(\Pi, d): x=p+s \nu_{\Sigma}(p) \mapsto(p, s)$. The signed distance $d$ is Lipschitz with constant 1 and by Proposition A.12, the map $X^{-1}$ is continuous. For every $x_{0} \in B_{r}(\Sigma)$ there exists $\varepsilon>0$ such that $\overline{B_{\varepsilon}}\left(x_{0}\right) \subset B_{r}(\Sigma)$ and $\operatorname{dist}_{\Sigma}\left(\Pi\left(x_{2}\right), \Pi\left(x_{1}\right)\right)<2\left\|L_{\Sigma}\right\|_{\infty}$ for $x_{1}, x_{2} \in B_{\varepsilon}\left(x_{0}\right)$. For $x_{j}=p_{j}+d\left(x_{j}\right) \nu_{\Sigma}\left(p_{j}\right) \in \overline{B_{\varepsilon}}\left(x_{0}\right)$ $(j \in\{1,2\})$, we obtain

$$
\begin{aligned}
\left|x_{2}-x_{1}\right| & \geq\left|p_{2}-p_{1}\right|-\left|d\left(x_{2}\right)\right|\left|\nu_{\Sigma}\left(p_{2}\right)-\nu_{\Sigma}\left(p_{1}\right)\right|-\left|d\left(x_{2}\right)-d\left(x_{1}\right)\right|\left|\nu_{\Sigma}\left(p_{2}\right)\right| \\
& \geq\left|p_{2}-p_{1}\right|-\left|d\left(x_{2}\right)\right| \frac{\left\|L_{\Sigma}\right\|_{\infty}}{\sqrt{2}} \operatorname{dist}_{\Sigma}\left(p_{2}, p_{1}\right)-\left|x_{2}-x_{1}\right| \\
& \geq\left|p_{2}-p_{1}\right|-\left|d\left(x_{2}\right)\right| \frac{\left\|L_{\Sigma}\right\|_{\infty}}{\sqrt{2}} \frac{\left|p_{2}-p_{1}\right|}{2-\left\|L_{\Sigma}\right\|_{\infty} \operatorname{dist}_{\Sigma}\left(p_{2}, p_{1}\right)}-\left|x_{2}-x_{1}\right| .
\end{aligned}
$$

Therefore Proposition A. 12 implies that $X^{-1}$ is locally Lipschitz.
It remains to prove the estimate for the principal curvatures. Fix an arbitrary point $p \in \Sigma$ and a principal curvature direction $v \in T_{p} \Sigma$ so that $L_{\Sigma} v=\kappa v$ and $|\kappa| \leq\left\|L_{\Sigma}\right\|_{\infty}$. By means of a parametrization over $T_{p} \Sigma$ and by reduction to the case $p=0$ and $\nu_{\Sigma}(p)=e_{n}$, the ball condition yields the inequality $|h(t v)| \leq k_{r}(t):=r-\sqrt{r^{2}-t^{2}}$ for some $\delta \in(0, r]$ and all $|t| \leq \delta$. Using $k_{r}(0)=k_{r}^{\prime}(0)=0$ and $k_{r}^{\prime \prime}(0)=1 / r$, we obtain $\left|d^{2} h(t v) / d t^{2}\right|_{t=0} \mid \leq 1 / r$ and therefore $|\kappa|=|\kappa v \cdot v|=\left|\nabla^{2} h(0)(v, v)\right| \leq 1 / r$. Taking sequences $\left(p_{n}\right)_{n},\left(v_{n}\right)_{n}$, and $\left(\kappa_{n}\right)_{n}$ with $v_{n} \in T_{p_{n}} \Sigma$, $L_{\Sigma} v_{n}=\kappa_{n} v_{n}$, and $\left|\kappa_{n}\right| \rightarrow\left\|L_{\Sigma}\right\|_{\infty}$, we obtain the desired inequality $\left\|L_{\Sigma}\right\|_{\infty} \leq 1 / r$.
(iii) Let $\Sigma$ be of class $C^{k+1}(k \geq 1)$. For $\tau \in T_{p} \Sigma$ and $s \in \mathbb{R}$, we obtain $X^{\prime}(p, t)(\tau, s)=$ $\tau+t\left(\nabla_{\Sigma} \nu_{\Sigma}\right)(p) \tau+s \nu_{\Sigma}(p)$. If $X^{\prime}(p, t)(\tau, s)=0$, then $\left(P_{\Sigma}(p)+t\left(\nabla_{\Sigma} \nu_{\Sigma}\right)(p)\right) \tau=0$ and $s \nu_{\Sigma}(p)=0$. Using $\left|P_{\Sigma}(p)+t\left(\nabla_{\Sigma} \nu_{\Sigma}\right)(p)\right| \geq 1-|t|\left|\nabla_{\Sigma} \nu_{\Sigma}\right|>0$ and $\left|\nu_{\Sigma}(p)\right|=1$, this implies $(\tau, s)=0$ and therefore $X^{\prime}(p, t): T_{p} \Sigma \times \mathbb{R} \rightarrow \mathbb{R}^{n}$ is bijective. By the inverse function theorem, the map $X$ is a $C^{k}$-diffeomorphism.
A.18. Lemma (A level function [cf. PS13]). Let $\Sigma \subset \mathbb{R}^{n}$ be a closed (possibly unbounded and possibly disconnected) hypersurface with tubular neighborhood of radius $r>0$. Then $\Sigma$ is a level set $\Sigma=$ $\varphi^{-1}(\{0\})$ with a function $\varphi \in C^{1}\left(\mathbb{R}^{n}\right)$, which has the following properties:
(i) $\left.\nabla \varphi\right|_{\Sigma}$ is a continuous unit normal field on $\Sigma$,
(ii) $\varphi(x) \in\{-1,1\}$ for $x \in \mathbb{R}^{n} \backslash B_{r}(\Sigma)$.

Proof. We extend the construction of [PS13, Section 4.2], which is valid for compact connected hypersurfaces. Let $\Sigma_{j}$ and $\Omega_{k}$ denote the at most countably many connected components of $\Sigma$ and $\mathbb{R}^{n} \backslash \Sigma$, respectively. For every $j$, the component $\Sigma_{j}$ is a closed connected hypersurface and hence there are precisely two domains $\Omega_{k}$ such that $\Sigma_{j} \subset \partial \Omega_{k}$ [cf. Sam69].

In the terminology of graph theory, the vertices $\mathcal{V}:=\left\{\Omega_{k}\right\}$ and the edges $\mathcal{E}=\left\{\Sigma_{j}\right\}$ form a connected graph $(\mathcal{V}, \mathcal{E})$ with two vertices $\Omega_{k}, \Omega_{l}$ being adjacent if and only if $\bar{\Omega}_{k} \cap \bar{\Omega}_{l} \neq \emptyset$. Since there exists precisely one edge $\Sigma_{j} \subset \partial \Omega_{k} \cap \partial \Omega_{l}$ that joins $\Omega_{k}$ to $\Omega_{l}$, the graph is undirected and simple. Suppose that $\left(\Omega_{k_{1}}, \Omega_{k_{2}}, \ldots, \Omega_{k_{m}}\right)$ is a cycle with distinct vertices $\Omega_{k_{l}}(l \in\{1,2, \ldots, m\})$ and corresponding edges $\left(\Sigma_{j_{1}}, \Sigma_{j_{2}}, \ldots, \Sigma_{j_{m}}\right)$. This means $\Sigma_{j_{l}} \subset \partial \Omega_{k_{l}} \cap \partial \Omega_{k_{l+1}}$ for $l<m$ and $\Sigma_{j_{m}} \subset \partial \Omega_{k_{m}} \cap \partial \Omega_{k_{1}}$. Then there exists a closed curve $\gamma:[1 / 2, m+1 / 2] \rightarrow \mathbb{R}^{n}$ such that $\gamma(l) \in \Sigma_{j_{l}}$ for all $l, \gamma(l-t) \in \Omega_{k_{l}}$ for all $l, t \in(0,1)$ and $\gamma(m+t) \in \Omega_{k_{1}}$ for all $t \in(0,1 / 2)$. The component $\Sigma_{j_{1}}$ separates $\mathbb{R}^{n}$ in two components $U_{1}$ and $U_{2}$ such that $\Omega_{k_{1}} \subset U_{1}$. But then $\gamma(t)$ belongs to $U_{2}$ for all $t \in(1, m+1 / 2]$, which is a contradiction. Hence the graph contains no cycles and is therefore bipartite. Consequently, there exists a function $\chi: \mathbb{R}^{n} \rightarrow\{-1,0,1\}$ such that
(i) $\chi(x)=0$ if and only if $x \in \Sigma$,
(ii) $\chi$ is constant in every connected component $\Omega_{k}$ of $\mathbb{R}^{n} \backslash \Sigma$,
(iii) if $\bar{\Omega}_{k} \cap \bar{\Omega}_{l} \neq \emptyset$, then $\left(\chi\left(\Omega_{k}\right), \chi\left(\Omega_{l}\right)\right)=(-1,1)$ or $\left(\chi\left(\Omega_{k}\right), \chi\left(\Omega_{l}\right)\right)=(1,-1)$.

On the connected components $\Sigma_{j}$ of $\Sigma$, we can therefore choose the orientation in such a way that the normal $\left.\nu\right|_{\Sigma_{j}}$ points into $\Omega_{+}:=\left\{x \in \mathbb{R}^{n}: \chi(x)=1\right\}$. We also put $\Omega_{-}:=\{x \in$ $\left.\mathbb{R}^{n}: \chi(x)=-1\right\}$. Then $\Omega_{+} \cup \Omega_{-}=\mathbb{R}^{n} \backslash \Sigma$ and the signed distance satisfies $d(x)>0$ for
$x \in \Omega_{+} \cap B_{r}(\Sigma)$ and $d(x)<0$ for $x \in \Omega_{-} \cap B_{r}(\Sigma)$. We fix some $\psi \in \mathcal{D}(\mathbb{R})$ such that $\psi(t)=1$ for $|t| \leq 1 / 3, \psi(t)=0$ for $|t| \geq 2 / 3$. Then a possible choice for $\varphi$ is

$$
\varphi(x)= \begin{cases}d(x) \psi(d(x) / r)+(1-\psi(d(x) / r)) \operatorname{sign}(d(x)) & \text { for } x \in B_{r}(\Sigma)  \tag{A.10}\\ \chi_{\Omega_{+}}(x)-\chi_{\Omega_{-}}(x) & \text { for } x \notin B_{r}(\Sigma)\end{cases}
$$

In particular, we obtain $\nabla \varphi(x)=\nabla d(x)=\nu_{\Sigma}(\Pi(x))$ for $x \in B_{r / 3}(\Sigma)$.
A.19. Corollary. Every domain $\Omega \subset \mathbb{R}^{n}(n \geq 2)$ that contains a compact $C^{2}$-hypersurface $\Sigma$ can be decomposed into

$$
\Omega=\Omega_{+} \cup \Sigma \cup \Omega_{-} \quad \text { with } \Omega_{ \pm}=\left\{x \in \Omega: \varphi_{\Sigma}(x) \gtrless 0\right\}
$$

where $\varphi_{\Sigma}$ denotes a level function as in Lemma A.18.
The next results deals with important geometric quantities of $C^{2-}$-hypersurfaces.
A.20. Proposition. Let $\Sigma \subset \mathbb{R}^{n}$ be a compact $C^{2-}$-hypersurface with tubular neighborhood $B_{r}(\Sigma)$ and $X:(p, s) \mapsto x=p+s \nu_{\Sigma}(p)$ be the corresponding diffeomorphism with inverse $X^{-1}=(\Pi, d)$. Then the following assertions are valid.
(i) The map

$$
M: x \mapsto\left[I-d(x) L_{\Sigma}(\Pi(x))\right]^{-1}, \quad B_{r}(\Sigma) \rightarrow \mathbb{R}^{n \times n}
$$

is essentially bounded and, for almost all $x \in B_{r}(\Sigma)$ and $p=\Pi(x) \in \Sigma$, the linear map $M(x)$ satisfies $M(x) \nu_{\Sigma}(p)=\nu_{\Sigma}(p)$ and maps $T_{p} \Sigma$ onto itself.
(ii) The map $d$ is of class $C^{2-}$ and satisfies $\nabla d(x)=M(x) \nu_{\Sigma}(p)=\nu_{\Sigma}(p)$.
(iii) The map $\Pi$ is of class $C^{1-}$ and satisfies

$$
\Pi^{\prime}(x)=P_{\Sigma}(\Pi(x)) M(x)=M(x)-\nu_{\Sigma}(p) \otimes \nu_{\Sigma}(p)=\Pi^{\prime}(x)^{\top}
$$

Proof. Let $x \in B_{r}(\Sigma)$ and $p=\Pi(x)$. In view of $\left|d(x) L_{\Sigma}(p)\right|<r \cdot r^{-1}=1$, the matrix $I$ $d(x) L_{\Sigma}(p)$ has maximal rank. It is also symmetric and in view of $L_{\Sigma} \nu_{\Sigma}=0$, it satisfies $[I-$ $\left.d(x) L_{\Sigma}(p)\right] \nu_{\Sigma}(p)=\nu_{\Sigma}(p)$ for all $p \in \Sigma$. Therefore its inverse $M(x)=\left[I-d(x) L_{\Sigma}(p)\right]^{-1}$ satisfies $M(x) \nu_{\Sigma}(p)=\nu_{\Sigma}(p)$ and maps $T_{p} \Sigma$ onto itself.

Let $\phi: U \subset \mathbb{R}^{n-1} \rightarrow \Sigma, u \mapsto p=\phi(u)$ be a chart for $\Sigma$. For each $\lambda \in(-r, r)$, the identities $\Pi(X(\phi(u), \lambda))=\phi(u)$ and $x=X(\phi(u), \lambda)$ imply

$$
\begin{aligned}
& \tau_{i}(p)=\partial_{u_{i}} \phi(u)=\Pi^{\prime}(x)\left(\tau_{i}(p)+\lambda \partial_{i} \nu_{\Sigma}(p)\right)=\Pi^{\prime}(x)\left(I-\lambda L_{\Sigma}(p)\right) \tau_{i}(p) \\
& \Pi^{\prime}(x)\left(I-\lambda L_{\Sigma}(p)\right) \nu_{\Sigma}(p)=\Pi^{\prime}(x) \nu_{\Sigma}(p)=\lim _{s \rightarrow 0} \frac{1}{s}\left(\Pi\left(x+s \nu_{\Sigma}(p)\right)-\Pi(x)\right)=0
\end{aligned}
$$

Hence $\Pi^{\prime}\left[I-d L_{\Sigma} \circ \Pi\right]=P_{\Sigma}$ and this yields $\Pi^{\prime}=P_{\Sigma}\left[I-d L_{\Sigma} \circ \Pi\right]^{-1}=P_{\Sigma} M=M-\nu_{\Sigma} \otimes \nu_{\Sigma}=$ $M P_{\Sigma}=\Pi^{\prime \top}$. Using the relations $d(x)=(x-p) \cdot \nu_{\Sigma}(p)$ and $\left(\partial_{i} \nu_{\Sigma}(p) \mid \nu_{\Sigma}(p)\right)=0$, we obtain

$$
\nabla d(x)=\nu_{\Sigma}(p)-\nu_{\Sigma}(p) \Pi^{\prime}(x)+\Pi^{\prime}(x) \nabla_{\Sigma} \nu_{\Sigma}(p) d(x) \nu_{\Sigma}(p)=\nu_{\Sigma}(p)
$$

As in [PS13], we consider hypersurfaces that are defined in tubular neighborhoods of a given hypersurface.
A.21. Proposition. Let $\Sigma \subset \mathbb{R}^{n}$ be a closed hypersurface of class $C^{k+1}\left[r e s p . C^{k+1-}\right](k \geq 1)$ with tubular neighborhood of radius $r>0$ and let $h \in C^{k}(\Sigma)$ [resp. $\left.h \in C^{k-}(\Sigma)\right]$ satisfy $\|h\|_{\infty}<r$. Then the following assertions are valid.
(i) The image $\Sigma_{h}=\theta_{h}(\Sigma)$ of $\theta_{h}: p \mapsto p+h(p) \nu_{\Sigma}(p)$ is a closed hypersurface of class $C^{k}$ [resp. $C^{k-}$ ] and the map $\theta_{h}: \Sigma \rightarrow \Sigma_{h}$ is a $C^{k}$-diffeomorphism [resp. $C^{k-}$-diffeomorphism if $k \geq 2$ and a homeomorphism with locally Lipschitz continuous inverse if $k=1$ ].
(ii) The normal $\nu_{\Sigma_{h}}$ of $\Sigma_{h}$ is [in the case $k=1$ almost everywhere] given by

$$
\begin{equation*}
\nu_{\Sigma_{h}}\left(p+h(p) \nu_{\Sigma}(p)\right)=\frac{\nu_{\Sigma}(p)-M_{h}(p) \nabla_{\Sigma} h(p)}{\sqrt{1+\left|M_{h}(p) \nabla_{\Sigma} h(p)\right|^{2}}} \quad \text { with } M_{h}(p):=\left(I-h(p) L_{\Sigma}(p)\right)^{-1} \tag{A.11}
\end{equation*}
$$

(iii) Suppose in addition that $\left\|\nabla_{\Sigma} h\right\|_{\infty}<\infty$. Then the normal of $\Sigma_{h}$ satisfies the inequality

$$
\nu_{\Sigma_{h}}\left(p+h(p) \nu_{\Sigma}(p)\right) \cdot \nu_{\Sigma}(p) \geq\left(1+\left\|M_{h} \nabla_{\Sigma} h\right\|_{\infty}^{2}\right)^{-1 / 2}>0 \quad \text { for } p \in \Sigma \text {. }
$$

(iv) The following integral transformation formula is valid for $f \in L_{1}\left(\Sigma_{h}\right)$.

$$
\begin{equation*}
\int_{\Sigma_{h}} f d \Sigma_{h}=\int_{\Sigma} f \circ \theta_{h} \operatorname{det}\left(P_{\Sigma}-h L_{\Sigma}\right) \sqrt{1+\left|M_{h} \nabla_{\Sigma} h\right|^{2}} d \Sigma . \tag{A.12}
\end{equation*}
$$

Proof. (i) Let $\phi: \mathbb{R}^{n-1} \supset U \in u \mapsto p \in \Sigma$ be a parametrization for $\Sigma$ and let $\theta_{h}: \Sigma \rightarrow \Sigma_{h}$, $p \mapsto p+h(p) \nu_{\Sigma}(p)$. Then the derivative $\theta_{h}^{\prime}: T_{p} \Sigma \rightarrow T_{\theta_{h}(p)} \Sigma_{h}$ is given by

$$
\begin{equation*}
\theta_{h}^{\prime}=P_{\Sigma}+\nu_{\Sigma} \otimes \nabla_{\Sigma} h+h \nabla_{\Sigma} \nu_{\Sigma}=P_{\Sigma}+\nu_{\Sigma} \otimes \nabla_{\Sigma} h-h L_{\Sigma} . \tag{A.13}
\end{equation*}
$$

If $\theta_{h}^{\prime}(p) u=0$ for some $u \in T_{p} \Sigma$, then $u-h(p) L_{\Sigma}(p) u=0$. The assumption $\|h\|_{\infty}<r$ implies $\left|h(p)\left(L_{\Sigma}\right)(p)\right|<1$ and this yields $u=0$. Therefore $\theta_{h}^{\prime}$ is bijective for all $p \in \Sigma$ and thereby $\theta_{h}: \Sigma \rightarrow \Sigma_{h}$ is a local diffeomorphism. The map $\theta_{h}$ is also surjective and coincides with a restriction of the map $X: B_{r}(\Sigma) \rightarrow \Sigma \times(-r, r)$ from Proposition A. 17 to $\{(p, t) \in \Sigma \times(-r, r): t=$ $h(p)\}$. Since $X$ is bijective, the map $\theta_{h}$ is a global diffeomorphism and $\theta_{h} \circ \phi$ is a parametrization for $\Sigma_{h}$, which shows that $\Sigma_{h}$ is of class $C^{k}\left[C^{k-}\right]$.
(ii) A derivation of (A.11) can be found in [PS13, Section 3.2]. The inverse $M_{h}(p)$ of $I-$ $h(p) L_{\Sigma}(p)$ is well-defined because of $\|h\|_{\infty}<r$.
(iii) This estimate is a direct consequence of (A.11).
(iv) By means of the parametrization $p=\varphi(u)$ we obtain $d \Sigma(p)=\sqrt{g(u)} d u$ where $g=$ $\operatorname{det} G, G=\varphi^{\prime \top} \varphi^{\prime}$, as well as $d \Sigma_{h}\left(\theta_{h}(p)\right)=\sqrt{g_{h}(u)} d u, g_{h}=\operatorname{det} G_{h}$. Since $\theta_{h} \circ \varphi$ is a parametrization for $\Sigma_{h}$, we have $G_{h}=\left[\theta_{h}^{\prime} \circ \varphi \varphi^{\prime}\right]^{\top} \theta_{h}^{\prime} \circ \varphi \varphi^{\prime}=\varphi^{\prime \top}\left[\theta_{h}^{\prime} \circ \varphi\right]^{\top}\left[\theta_{h}^{\prime} \circ \varphi\right] \varphi^{\prime}$. Hence identity (A.13) yields

$$
\left[\theta_{h}^{\prime} \circ \varphi\right]^{\top}\left[\theta_{h}^{\prime} \circ \varphi\right]=\left(P_{\Sigma}-h L_{\Sigma}\right)^{2}+\nabla_{\Sigma} h \otimes \nabla_{\Sigma} h=M_{h}^{-2}\left(P_{\Sigma}+M_{h}^{2} \nabla_{\Sigma} h \otimes \nabla_{\Sigma} h\right) .
$$

For computing $\operatorname{det} G_{h}$, we recall two facts from linear algebra. First, for any two isomorphisms $A: X \rightarrow Y$ and $B: Y \rightarrow X$ between $n$-dimensional vector spaces $X$ and $Y$ we have $\operatorname{det}_{X}(B A)=\operatorname{det}_{Y}(A B)$, since the determinant $\operatorname{det}_{X}$ in $X$ is given by the identity $\operatorname{det}_{X}(C)=$ $V\left(C x_{1}, \ldots, C x_{n}\right) / V\left(x_{1}, \ldots, x_{n}\right)$ for any $C \in \mathcal{L}(X)$, any volume form $V$ in $X$ and any basis $\left(x_{j}\right)$ of $X$. Second, we have $\operatorname{det}(I+a \otimes b)=1+a \cdot b$ for $a, b \in \mathbb{R}^{n}$. These facts yield

$$
\begin{aligned}
g_{h} & =\operatorname{det}_{\mathbb{R}^{n-1}}\left(\varphi^{\prime \top} \varphi^{\prime}\right) \operatorname{det}_{\mathbb{R}^{n-1}}\left(\varphi^{\prime-1}\left[\theta_{h}^{\prime} \circ \varphi\right]^{\top}\left[\theta_{h}^{\prime} \circ \varphi\right] \varphi^{\prime}\right) \\
& =g \operatorname{det}_{\mathbb{R}^{n-1}}\left(\varphi^{\prime-1} M_{h}^{-2} \varphi^{\prime}\right) \operatorname{det}_{\mathbb{R}^{n-1}}\left(\varphi^{\prime-1} P_{\Sigma} \varphi^{\prime}+\varphi^{\prime-1} M_{h}^{2} \nabla_{\Sigma} h \otimes \varphi^{\prime \top} \nabla_{\Sigma} h\right) \\
& =g \operatorname{det}_{T_{p} \Sigma}\left(P_{\Sigma}-h L_{\Sigma}\right)^{2}\left(1+\left|M_{h} \nabla_{\Sigma} h\right|^{2}\right) .
\end{aligned}
$$

Therefore the asserted equation (A.12) follows.

## A.4. Covariant differentiation

Let $\Gamma$ be a $C^{3}$-hypersurface of $\mathbb{R}^{n}$, equipped with the induced Euclidean metric $(v \mid w)_{g(p)}=v \cdot w$ for $v, w \in T_{p} \Gamma \subset \mathbb{R}^{n}$ and $p \in \Gamma$. We let $\tau_{1}, \ldots, \tau_{n-1}$ be a basis of tangent vectors on $T_{p} \Gamma$ with dual basis $\tau^{1}, \ldots, \tau^{n-1}$ so that $\tau_{j} \cdot \tau^{k}=\delta_{j}^{k}$, we let $\nu$ denote the unit normal on $\Gamma$, and we let $P=I-\nu \otimes \nu$ denote the projection onto the tangent space. Moreover, we let $C^{k}(\Gamma ; T \Gamma)\left(k \in \mathbb{N}_{0}\right)$ denote the Banach space of all tangential vector fields $v=v^{\alpha} \tau_{\alpha}$ of class $C^{k}$ on $\Gamma$.
A.4.1. First order covariant derivatives. We define the (partial) covariant derivative $\widetilde{\nabla}_{j} v$ with respect to the coordinate $x_{j}$ by

$$
v_{; j}:=\widetilde{\nabla}_{j} v:=P \partial_{j}\left(v^{\alpha} \tau_{\alpha}\right)=\left(\partial_{j} v^{\alpha}+\Lambda_{j \beta}^{\alpha} v^{\beta}\right) \tau_{\alpha}=: v^{\alpha}{ }_{; j} \tau_{\alpha} .
$$

Here $\Lambda_{i j}^{k}$ is the Christoffel symbol of the second kind. Moreover, we let

$$
\widetilde{\nabla} v:=\widetilde{\nabla}_{\Gamma} v:=v^{\alpha}{ }_{; \beta} \tau_{\alpha} \otimes \tau^{\beta}=\left(\partial_{\beta} v^{\alpha}+\Lambda_{\beta \gamma}^{\alpha} v^{\gamma}\right) \tau_{\alpha} \otimes \tau^{\beta} \quad \text { for } v \in C^{1}(\Gamma ; T \Gamma),
$$

so that

$$
\widetilde{\nabla}_{u} v:=[\widetilde{\nabla} v] u=\left(\partial_{\beta} v^{\alpha} u^{\beta}+\Lambda_{\beta \gamma}^{\alpha} v^{\gamma} u^{\beta}\right) \tau_{\alpha} \quad \text { for } u \in C(\Gamma ; T \Gamma) .
$$

This definition of $\widetilde{\nabla}_{u} v$ coincides with the Levi-Civita connection on $\Gamma$ and ensures that $\widetilde{\nabla}_{u} v$ is again a tangential vector field.

For a possibly non-tangential vector field $u=v+w \nu=v^{\alpha} \tau_{\alpha}+w \nu \in C^{1}\left(\Gamma ; \mathbb{R}^{n}\right)$ we define

$$
\begin{equation*}
u_{; k}:=\widetilde{\nabla}_{k} u:=P \partial_{k}\left(v^{\alpha} \tau_{\alpha}+w \nu\right)=\widetilde{\nabla}_{k}\left(v^{\alpha} \tau_{\alpha}\right)-w l_{k \alpha} \tau^{\alpha}, \tag{A.14}
\end{equation*}
$$

where $l_{j k}$ denote the components of the Weingarten tensor $L=l_{\alpha \beta} \tau^{\alpha} \otimes \tau^{\beta}=-\nabla_{\Gamma} \nu$. Then

$$
\partial_{k} u=\widetilde{\nabla}_{k} u+[\nu \otimes \nu] \partial_{k} u=\left(v^{\alpha}{ }_{; k}-w g^{\alpha \beta} l_{k \beta}\right) \tau_{\alpha}+\left(v^{\alpha} l_{k \alpha}+\partial_{k} w\right) \nu .
$$

By abbreviating $v_{\alpha}:=g_{\alpha \gamma} v^{\gamma}$, we rewrite the surface gradient $\nabla_{\Gamma} u=\tau^{\alpha} \otimes \partial_{\alpha} u$ as

$$
\begin{align*}
\nabla_{\Gamma} u & =\tau^{\alpha} \otimes \tau^{\beta}\left(v_{\beta ; \alpha}-w l_{\alpha \beta}\right)+\tau^{\alpha} \otimes \nu\left(v^{\beta} l_{\alpha \beta}+\partial_{\alpha} w\right) \\
& =[\widetilde{\nabla} v]^{\top}-w L+\left(L v+\nabla_{\Gamma} w\right) \otimes \nu . \tag{A.15}
\end{align*}
$$

With the mean curvature $H:=\operatorname{tr} L=-\operatorname{div}_{\Gamma} \nu$, the surface divergence $\operatorname{div}_{\Gamma} u$ satisfies

$$
\begin{equation*}
\operatorname{div}_{\Gamma} u:=\tau^{\alpha} \cdot \partial_{\alpha} u=v^{\alpha}{ }_{; \alpha}-w H=\operatorname{div}_{\Gamma} v-w H . \tag{A.16}
\end{equation*}
$$

The symmetric part $D_{\Gamma}(u)$ of $P\left[\nabla_{\Gamma} u\right] P$ is given by

$$
\begin{equation*}
D_{\Gamma}(u):=\operatorname{sym}\left(P\left[\nabla_{\Gamma} u\right] P\right)=2^{-1} \tau^{\alpha} \otimes \tau^{\beta}\left(v_{\alpha ; \beta}+v_{\beta ; \alpha}\right)-w L=2^{-1}\left(\widetilde{\nabla} v+[\widetilde{\nabla} v]^{\top}\right)-w L . \tag{A.17}
\end{equation*}
$$

We note that $\operatorname{tr} D_{\Gamma}(u)=\operatorname{div}_{\Gamma} u$.
Second order tensors have the form $S^{\alpha \beta} \tau_{\alpha} \otimes \tau_{\beta}, S_{\alpha \beta} \tau^{\alpha} \otimes \tau^{\beta}, S^{\alpha}{ }_{\beta} \tau_{\alpha} \otimes \tau^{\beta}$, or $S_{\alpha}{ }^{\beta} \tau^{\alpha} \otimes \tau_{\beta}$, and their first order covariant derivatives are given by

$$
\begin{aligned}
S^{i j}{ }_{; k}=\partial_{k} S^{i j}+\Lambda_{\alpha k}^{i} S^{\alpha j}+\Lambda_{\alpha k}^{j} S^{i \alpha}, & S^{i}{ }_{j ; k}=\partial_{k} S^{i}{ }_{j}+\Lambda_{\alpha k}^{i} S^{\alpha}{ }_{j}-\Lambda_{j k}^{\alpha} S^{i}{ }_{\alpha}, \\
S_{i j ; k}=\partial_{k} S_{i j}-\Lambda_{i k}^{\alpha} S_{\alpha j}-\Lambda_{j k}^{\alpha} S_{i \alpha}, & S_{i}{ }_{i}{ }_{; k}=\partial_{k} S_{i}{ }^{j}-\Lambda_{i k}^{\alpha} S_{\alpha}{ }^{j}+\Lambda_{\alpha k}^{j} S_{i}{ }^{\alpha} .
\end{aligned}
$$

Then the surface divergence of a second order symmetric tensor $S^{\alpha \beta} \tau_{\alpha} \otimes \tau_{\beta}$ is given by

$$
\begin{equation*}
\operatorname{div}_{\Gamma} S=\operatorname{div}_{\Gamma}\left(S^{\alpha \beta} \tau_{\alpha} \otimes \tau_{\beta}\right):=\left[\partial_{\gamma}\left(S^{\alpha \beta} \tau_{\alpha} \otimes \tau_{\beta}\right)\right] \tau^{\gamma}=S_{; \alpha}^{\alpha \beta} \tau_{\beta}+S^{\alpha \beta} l_{\alpha \beta} \nu_{\Gamma} . \tag{A.18}
\end{equation*}
$$

For symmetric $S=S^{\alpha \beta} \tau_{\alpha} \otimes \tau_{\beta}$ and $u=v^{\alpha} \tau_{\alpha}+w \nu$ we have

$$
\begin{equation*}
\operatorname{div}_{\Gamma}(S u)=\operatorname{div}_{\Gamma} S \cdot u+S: \nabla_{\Gamma} u . \tag{A.19}
\end{equation*}
$$

By using the identity $\partial_{k} \tau_{i} \cdot \tau_{j}=\Lambda_{i k}^{\alpha} g_{\alpha j}$, we can easily deduce the useful identities

$$
\begin{equation*}
g_{i j ; k}=0, \quad g_{; k}^{i j}=0 . \tag{A.20}
\end{equation*}
$$

Thus, the metric tensor $P=g^{i j} \tau_{i} \otimes \tau_{j}=g_{i j} \tau^{i} \otimes \tau^{j}$ satisfies

$$
\begin{equation*}
\operatorname{div}_{\Gamma} P=H \nu . \tag{A.21}
\end{equation*}
$$

The components $l_{i j}=-\tau_{i} \cdot \partial_{j} \nu$ of the Weingarten tensor satisfy the relations

$$
\begin{equation*}
l_{i j ; k}=l_{i k ; j}=l_{j k ; i} . \tag{A.22}
\end{equation*}
$$

More generally, let $T$ be a tensor with the components $T_{j_{1} \cdots j_{b}}^{i_{1} \cdots i_{a}}$, where we agree on not raising or lowering indices when the order between the contravariant and the covariant indices is not indicated. Then the covariant derivative of $T$ with respect to $x_{k}$ is given by

$$
\begin{equation*}
T_{j_{1} \cdots j_{b} ; m}^{i_{1} \cdots i_{a}}=\partial_{m} T_{j_{1} \cdots j_{b}}^{i_{1} \cdots i_{a}}+\sum_{p=1}^{a} \Lambda_{\alpha m}^{i_{p}} T_{j_{1} \cdots j_{b}}^{i_{1} \cdots i_{p-1} \alpha i_{p+1} \cdots i_{a}}-\sum_{p=1}^{b} \Lambda_{j_{i} m}^{\alpha} T_{j_{1} \cdots j_{p-1} \alpha j_{p+1} \cdots j_{b}}^{i_{1} \cdots i_{a}} . \tag{A.23}
\end{equation*}
$$

For two tensors $S$ and $T$ the following product rule is valid.

$$
\begin{equation*}
\left(S_{j_{1} \cdots j_{b}}^{i_{1} \cdots i_{a}} T_{l_{1} \cdots l_{d}}^{k_{1} \cdots k_{c}}\right)_{; m}=S_{j_{1} \cdots j_{b} ; m}^{i_{1} \cdots i_{a}} T_{l_{1} \cdots l_{d}}^{k_{1} \cdots k_{c}}+S_{j_{1} \cdots j_{b}}^{i_{1} \cdots i_{a}} T_{l_{1} \cdots l_{d} ; m}^{k_{1} \cdots k_{c}} . \tag{A.24}
\end{equation*}
$$

A.4.2. Relation to bulk differential operators. Let $\Omega \subset \mathbb{R}^{n}$ be open and let $\Gamma \subset \Omega$ be a $C^{3}$ hypersurface which admits a $C^{1}$-class tubular neighborhood map $(x, s) \mapsto X(x, s)=x+s \nu(x)$ from an open subset $U \subset \Gamma \times \mathbb{R}$ with $U \supset \Gamma \times\{0\}$ onto $V \subset \Omega$. Let $(\Pi, d)=X^{-1}$ so that $\Pi(x+s \nu(x))=x$ and $d(x+s \nu(x))=s$. For a vector field $u: V \rightarrow \mathbb{R}^{n}$ we let

$$
u=v+w \nu \circ \Pi, \quad v:=[P \circ \Pi] u, \quad w:=(\nu \circ \Pi \mid u)
$$

Then we easily find the following identities on $\Gamma$.

$$
\begin{align*}
\nabla u & =\nabla_{\Gamma} u+\nu \otimes \partial_{\nu} u,  \tag{A.25a}\\
\nabla_{\Gamma} u & =\nabla_{\Gamma} v-w L+\nabla_{\Gamma} w \otimes \nu,  \tag{A.25b}\\
\operatorname{div} u & =\operatorname{div}_{\Gamma} u-\left(\nu \mid \partial_{\nu} u\right) . \tag{A.25c}
\end{align*}
$$

A.4.3. Second order covariant derivatives. For a tangential vector field $v=v^{\alpha} \tau_{\alpha}$, we consider the second order covariant derivatives

$$
\widetilde{\nabla}_{j} \widetilde{\nabla}_{k}\left(v^{\alpha} \tau_{\alpha}\right)=\widetilde{\nabla}_{j}\left(v^{\alpha}{ }_{; k} \tau_{\alpha}\right)=v^{\alpha}{ }_{; k j} \tau_{\alpha} .
$$

The operators $\widetilde{\nabla}_{j}$ and $\widetilde{\nabla}_{k}$ do not necessarily commute but satisfy the relations

$$
\begin{equation*}
v^{i}{ }_{; j k}-v^{i}{ }_{; k j}=R_{\alpha j k}^{i} v^{\alpha}, \quad v_{i ; j k}-v_{i ; k j}=-v_{\alpha} R^{\alpha}{ }_{i j k}, \tag{A.26}
\end{equation*}
$$

where $R^{i}{ }_{l j k}=\tau^{i} \cdot R\left(\tau_{j}, \tau_{k}\right) \tau_{l}$ are the components of the Riemann curvature tensor $R$, given by

$$
R^{i}{ }_{l j k}=\partial_{j} \Lambda_{k l}^{i}-\partial_{k} \Lambda_{j l}^{i}+\Lambda_{j \alpha}^{i} \Lambda_{k l}^{\alpha}-\Lambda_{k \alpha}^{i} \Lambda_{j l}^{\alpha}, \quad g_{i m} R^{i}{ }_{l j k}=: R_{m l j k} .
$$

This tensor has the symmetries

$$
R_{i j k l}+R_{i k l j}+R_{i l j k}=0, \quad R_{i j k l}=-R_{i j l k}=-R_{j i k l}=R_{k l i j} .
$$

For a hypersurface $\Gamma$ in $\mathbb{R}^{n}$ we have

$$
\begin{equation*}
R_{i j k l}=l_{i k} l_{j l}-l_{i l} l_{j k} . \tag{A.27}
\end{equation*}
$$

## A.4.4. The tangential Laplace-Beltrami operator. We define

$$
\widetilde{\Delta} v:=\widetilde{\Delta}_{\Gamma} v:=g^{\alpha \beta} \widetilde{\nabla}_{\alpha} \widetilde{\nabla}_{\beta} v=g^{\alpha \beta} v_{; \alpha \beta}^{\gamma} \tau^{\gamma} \quad \text { for } v \in C^{2}(\Gamma ; T \Gamma) .
$$

This definition is consistent with $\widetilde{\Delta} v=-\widetilde{\nabla}^{*} \widetilde{\nabla} v$, where $\widetilde{\nabla}^{*}$ is the formal $L_{2}(\Gamma)$-adjoint of $\widetilde{\nabla}$, which means that $\widetilde{\nabla}^{*} W\left(W \in C^{1}\left(\Gamma ; T \Gamma \otimes T^{*} \Gamma\right)\right)$ is defined by the relation

$$
\left(\widetilde{\nabla}^{*} W \mid v\right)_{L_{2}(\Gamma ; T \Gamma)}=(W \mid \widetilde{\nabla} v)_{L_{2}\left(\Gamma ; T \Gamma \otimes T^{*} \Gamma\right)} \quad \text { for all } v \in C_{c}^{1}(\Gamma ; T \Gamma) .
$$

To check this, we write $W=W^{\alpha}{ }_{\beta} \tau_{\alpha} \otimes \tau^{\beta}$ and $\widetilde{\nabla} v=v^{\alpha}{ }_{; \beta} \tau_{\alpha} \otimes \tau^{\beta}$ and obtain

$$
\begin{aligned}
(W \mid \widetilde{\nabla} v)_{L_{2}\left(\Gamma ; T \Gamma \otimes T^{*} \Gamma\right)} & =\int_{\Gamma} W \tau^{\alpha} \cdot \widetilde{\nabla}_{\alpha} v d \Gamma \\
& =\int_{\Gamma} g^{\alpha \beta} W^{\gamma}{ }_{\beta} \tau_{\gamma} \cdot v_{; \alpha}^{\delta} \tau_{\delta} d \Gamma=\int_{\Gamma} g^{\alpha \beta} W^{\gamma}{ }_{\beta} g_{\gamma \delta} v^{\delta}{ }_{; \alpha} d \Gamma .
\end{aligned}
$$

From $\operatorname{div}_{\Gamma} v=v^{\alpha}{ }_{; \alpha}$, identities (A.20), and the surface divergence theorem, we infer that

$$
(W \mid \widetilde{\nabla} v)_{L_{2}\left(\Gamma ; T \Gamma \otimes T^{*} \Gamma\right)}=-\int_{\Gamma} g^{\alpha \beta} W^{\gamma}{ }_{\beta ; \alpha} g_{\gamma \delta} v^{\delta} d \Gamma=-\left(g^{\alpha \beta} W^{\gamma}{ }_{\beta ; \alpha} \tau_{\gamma} \mid v\right)_{L_{2}(\Gamma ; T \Gamma)},
$$

and thus $\widetilde{\nabla}^{*} W=-g^{\alpha \beta} w^{\gamma}{ }_{\beta ; \alpha} \tau_{\gamma}$. This yields $\widetilde{\Delta} v=-\widetilde{\nabla} * \widetilde{\nabla} v$.
Finally, for $u \in C^{1}(\Gamma ; T \Gamma)$ and $v \in C^{2}(\Gamma ; T \Gamma)$, we calculate

$$
\begin{align*}
(\widetilde{\Delta} v \mid u)_{L_{2}(\Gamma ; T \Gamma)} & =\int_{\Gamma} g^{\alpha \beta} v^{\gamma}{ }_{; \alpha \beta} \tau_{\gamma} \cdot \tau_{\delta} u^{\delta} d \Gamma  \tag{A.28}\\
& =-\int_{\Gamma} v^{\gamma}{ }_{; \alpha} u^{\delta}{ }_{; \beta} g^{\alpha \beta} g_{\gamma \delta} d \Gamma=-\int_{\Gamma} \widetilde{\nabla} v: \widetilde{\nabla} u d \Gamma
\end{align*}
$$

where $S: T=\operatorname{tr}\left(S^{\top} T\right)=\left(S \tau_{\alpha} \mid T \tau^{\alpha}\right)$. Therefore $\widetilde{\Delta}$ is symmetric and negative semi-definite.
For a non-tangential vector field $u=v+w \nu$, equation (A.14) yields

$$
\begin{equation*}
\widetilde{\Delta}(v+w \nu)=\widetilde{\Delta} v-L \nabla_{\Gamma} w-w \nabla_{\Gamma} H . \tag{A.29}
\end{equation*}
$$

## APPENDIX B

## Functional analytic methods

## B.1. Function spaces

B.1.1. Classical function spaces. Let $\Omega$ be an open subset of $\mathbb{R}^{n}(n \in \mathbb{N})$, let $X$ be a Banach space over $\mathbb{K} \in\{\mathbb{R}, \mathbb{C}\}$, and let $k \in \mathbb{N}_{0}$. The vector space $C^{k}(\Omega ; X)$ consists of the $k$ times continuously Fréchet-differentiable functions from $\Omega$ to $X$. We abbreviate $C^{k}(\Omega):=C^{k}(\Omega ; \mathbb{K})$ and $C^{k}(\Omega)^{n}:=C^{k}\left(\Omega ; \mathbb{K}^{n}\right)$, analogously for all subsequent spaces. The subspace $C^{k}(\bar{\Omega} ; X)$ consists of those $u \in C^{k}(\Omega ; X)$ that have a continuous extension onto the closure $\bar{\Omega}$ of $\Omega$, together will all derivatives up to order $k$. The Banach space $B C^{k}(\Omega ; X)$ consists of all bounded functions $u \in C^{k}(\Omega ; X)$ with bounded derivatives up to order $k$, equipped with the norm

$$
\|u\|_{B C^{k}(\Omega ; X)}:=\sup \left\{\left\|\partial_{x}^{\beta} u(x)\right\|_{X}: \beta \in \mathbb{N}_{0}^{n},|\beta| \leq k, x \in \Omega\right\} .
$$

The space $B U C^{k}(\Omega ; X)$ consists of all bounded, uniformly continuous functions $u \in C^{k}(\Omega ; X)$ with bounded, uniformly continuous derivatives up to order $k$. For an interval $J \subset \mathbb{R}$, we let $C_{0}(J ; X)=\left\{u \in C(J ; X):\|u(t)\|_{X} \rightarrow 0\right.$ as $\left.t \rightarrow \infty\right\}$. For given Banach spaces $X$ and $Y$ and an open subset $U \subset X$, the spaces $C^{k}(U ; Y), B C^{k}(U ; Y)$, and $B U C^{k}(U ; Y)$ are defined analogously. For $k \in \mathbb{N}, \alpha \in(0,1]$, and $f: \mathbb{R}^{n} \rightarrow X$, we define the seminorm

$$
\llbracket f \rrbracket_{C^{k, \alpha}}:=\sup _{x \neq y} \frac{\left\|f^{(k)}(x)-f^{(k)}(y)\right\|_{X}}{|x-y|^{\alpha}} .
$$

The space $C^{k, \alpha}\left(\mathbb{R}^{n} ; X\right):=\left\{f \in C^{k}\left(\mathbb{R}^{n} ; X\right): \llbracket f \rrbracket_{C^{k, \alpha}}<\infty\right\}$ is called Hölder space if $\alpha<1$ and Lipschitz space if $\alpha=1$. We also write $C^{k+\alpha}:=C^{k, \alpha}$ if $\alpha \in(0,1)$ and $C^{k-}:=\left\{f \in C^{k-1}\right.$ : $f^{(k-1)}$ is locally Lipschitz $\}$. Rademacher's theorem implies that a function $u \in C\left(\mathbb{R}^{n}\right)$ belongs to $C^{0,1}\left(\mathbb{R}^{n}\right)$ if and only if it is almost everywhere differentiable and its derivative is bounded.

The support supp $u$ of a function $u \in C(\Omega ; X)$ is the closure of the set $\{x \in \Omega: \varphi(x) \neq 0\}$ in $\mathbb{R}^{n}$. The space $C_{c}^{k}(\Omega ; X)$ consists of all $u \in C^{k}(\Omega ; X)$ such that supp $u$ is compact and a subset of $\Omega$. We let $\mathcal{D}(\Omega ; X)=C_{c}^{\infty}(\Omega ; X)$ denote the Fréchet space of test functions. The space of distributions $\mathcal{D}^{\prime}(\Omega ; X)$ consists of all continuous linear maps $\mathcal{D}(\Omega) \rightarrow X$. A function $u \in C^{\infty}\left(\mathbb{R}^{n} ; X\right)$ is called rapidly decreasing, if $x \mapsto|x|^{|\alpha|} \partial_{x}^{\beta} u(x)$ is bounded on $\mathbb{R}^{n}$ for every pair of multi-indices $\alpha, \beta \in \mathbb{N}_{0}^{n}$. Here we let $|\alpha|=\left|\alpha_{1}\right|+\cdots+\left|\alpha_{n}\right|$ and $\partial_{x}^{\beta}=\partial_{1}^{\beta_{1}} \cdots \partial_{n}^{\beta_{n}}$. The Schwartz space of rapidly decreasing functions is denoted by $\mathcal{S}\left(\mathbb{R}^{n} ; X\right)$ and the space of tempered distributions $\mathcal{S}^{\prime}\left(\mathbb{R}^{n} ; X\right)$ consists of all continuous linear maps $\mathcal{S}\left(\mathbb{R}^{n}\right) \rightarrow X$.

Given a measure space $(\Omega, \mathcal{A}, \mu)$, we let $L_{0}(\Omega ; X):=L_{0}(\Omega, \mathcal{A}, \mu ; X)$ denote the vector space of all equivalence classes of strongly $\mu$-measurable functions $\Omega \rightarrow X$. Given $p \in[1, \infty]$ and $m \in \mathbb{N}_{0}$, we employ the usual Bochner-Lebesgue space $L_{p}(\Omega ; X)=L_{p}(\Omega, \mathcal{A}, \mu ; X)$ and the Sobolev space $W_{p}^{m}(\Omega ; X)$. The Bessel potential space of order $s \in \mathbb{R}$ is defined by

$$
H_{p}^{s}\left(\mathbb{R}^{n} ; X\right):=\left\{u \in \mathcal{S}^{\prime}\left(\mathbb{R}^{n} ; X\right):\|f\|_{H_{p}^{s}\left(\mathbb{R}^{n} ; X\right)}=\left\|\mathcal{F}^{-1}\left(\xi \mapsto\left(1+|\xi|^{2}\right)^{s / 2}(\mathcal{F} f)(\xi)\right)\right\|_{L_{p}\left(\mathbb{R}^{n}\right)}<\infty\right\} .
$$

The operator $J^{\sigma}: f \mapsto \mathcal{F}^{-1}\left(\xi \mapsto\left(1+|\xi|^{2}\right)^{\sigma / 2}(\mathcal{F} f)(\xi)\right)$ is called Bessel potential of order $\sigma$ and its realization $J^{\sigma}: H_{p}^{s+\sigma}\left(\mathbb{R}^{n} ; X\right) \rightarrow H_{p}^{s}\left(\mathbb{R}^{n} ; X\right)$ is an isomorphism; that is, a bijective, bounded linear map with bounded inverse.

For $s=m \in \mathbb{N}_{0}$ and $p \in(1, \infty)$, the spaces $H_{p}^{m}\left(\mathbb{R}^{n} ; X\right)$ and $W_{p}^{m}\left(\mathbb{R}^{n} ; X\right)$ coincide with equivalent norms if and only if $X$ is of class $\mathcal{H T}$ [McC84; $\operatorname{Zim} 89]$. For $p \in[1, \infty)$, we have $\left(H_{p}^{s}\left(\mathbb{R}^{n}\right)\right)^{\prime}=H_{p^{\prime}}^{-s}\left(\mathbb{R}^{n}\right)$ [BL76]. The Sobolev-Slobodeckǐ̆ space $W_{p}^{s}(\Omega ; X)$ of order $s \in(0, \infty) \backslash \mathbb{N}_{0}$ with $s=[s]+\{s\},[s] \in \mathbb{N}_{0},\{s\} \in(0,1)$, and $p \in[1, \infty)$ is defined by

$$
W_{p}^{s}(\Omega ; X):=\left\{u \in \mathcal{D}^{\prime}(\Omega ; X):\|u\|_{W_{p}^{s}(\Omega ; X)}:=\|u\|_{W_{p}^{[s]}(\Omega ; X)}+\sum_{|\alpha|=[s]} \llbracket \partial^{\alpha} u \rrbracket_{W_{p}^{\{s\}}(\Omega ; X)}<\infty\right\},
$$

as in [Ama97, p. 10] and [Tri10, Theorem 2.5.7], where the seminorm $\llbracket \cdot \rrbracket_{W_{p}^{\theta}(\Omega ; X)}$ is defined by

$$
\llbracket u \rrbracket_{W_{p}^{\theta}(\Omega ; X)}:=\left(\int_{\Omega} \int_{\Omega} \frac{|u(x)-u(y)|_{X}^{p}}{|x-y|^{n+\theta p}} d x d y\right)^{1 / p} \quad \text { for } \theta \in(0,1) .
$$

We also refer to [Tri95, Theorem 4.2.4] and [Lud14] for some properties of this norm and [KPW13, Section 3.2] for an equivalent norm. Following [Joh95; RS96; Tri10; SSS12], we introduce Besov spaces and Triebel-Lizorkin spaces over $\mathbb{R}^{n}$ in terms of the Fourier transform $\mathcal{F}$ on $\mathcal{S}^{\prime}\left(\mathbb{R}^{n}\right)$ and a partition of unity. Let $\left\{\varphi_{j}\right\}_{j=0}^{\infty} \subset \mathcal{S}\left(\mathbb{R}^{n}\right)$ satisfy the following properties.
(i) There exist $A, B, C \in(0, \infty)$ such that $\operatorname{supp} \varphi_{0} \subset \overline{\mathbb{B}}_{A}$, $\operatorname{supp} \varphi_{j} \subset \overline{\mathbb{B}}_{C^{2 j+1}} \backslash \mathbb{B}_{B 2^{j-1}}$ for $j \in \mathbb{N}$.
(ii) For every $\alpha \in \mathbb{N}_{0}^{n}$ there is $c_{a} \in(0, \infty)$ such that $2^{j|\alpha|}\left|D^{\alpha} \varphi_{j}(x)\right| \leq c_{\alpha}$ for all $x \in \mathbb{R}^{n}, j \in \mathbb{N}_{0}$.
(iii) $\sum_{j=0}^{\infty} \varphi_{j}(x)=1$ for all $x \in \mathbb{R}^{n}$.

The Besov space $B_{p q}^{s}\left(\mathbb{R}^{n} ; X\right)$ and the Triebel-Lizorkin space $F_{p q}^{s}\left(\mathbb{R}^{n} ; X\right)$ of order $s \in \mathbb{R}$, integralexponent $p \in[1, \infty]$, and sum-exponent $q \in[1, \infty]$ are defined by

$$
\begin{aligned}
& B_{p q}^{s}\left(\mathbb{R}^{n} ; X\right):=\left\{u \in \mathcal{S}^{\prime}\left(\mathbb{R}^{n} ; X\right):\|u\|_{B_{p q}^{s}}:=\left\|\left\{2^{s j} \mathcal{F}^{-1}\left[\varphi_{j} \mathcal{F} u\right]\right\}_{j}\right\|_{l_{q}\left(L_{p}\right)}<\infty\right\}, \\
& F_{p q}^{s}\left(\mathbb{R}^{n} ; X\right):=\left\{u \in \mathcal{S}^{\prime}\left(\mathbb{R}^{n} ; X\right):\|u\|_{F_{p q}^{s}}^{s}:=\left\|\left\{2^{s j} \mathcal{F}^{-1}\left[\varphi_{j} \mathcal{F} u\right]\right\}_{j}\right\|_{L_{p}\left(l_{q}\right)}<\infty\right\} .
\end{aligned}
$$

We recall that the identity

$$
W_{p}^{s}\left(\mathbb{R}^{n} ; X\right)=B_{p p}^{s}\left(\mathbb{R}^{n} ; X\right) \quad \text { for } s \in(0, \infty) \backslash \mathbb{N}, p \in(1, \infty),
$$

is valid for every Banach space $X$ [see Ama97, (5.8), (5.9)], whereas

$$
H_{p}^{m}\left(\mathbb{R}^{n} ; X\right)=W_{p}^{m}\left(\mathbb{R}^{n} ; X\right) \quad \text { for } m \in \mathbb{N}_{0}, p \in(1, \infty),
$$

is valid if and only if $X$ is a Banach space of class $\mathcal{H} \mathcal{T}$ [McC84; SSS12]. Moreover,

$$
H_{p}^{s}\left(\mathbb{R}^{n} ; X\right)=F_{p 2}^{s}\left(\mathbb{R}^{n} ; X\right) \quad \text { for } s \in \mathbb{R}, p \in(1, \infty),
$$

is valid if and only if $X$ can be renormed as a Hilbert space [SSS12, Section 2.2].
We collect some properties of interpolation spaces from the monographs [BL76; Lun09; Tri95]. Let $X_{0}$ and $X_{1}$ be Banach spaces with dense embeding $X_{1} \hookrightarrow^{d} X_{0}$ and let also $X, Y$, $Y_{0}$, and $Y_{1}$ be Banach spaces. The space $Y$ is called interpolation space for the couple ( $X_{0}, X_{1}$ ), if $X_{1} \hookrightarrow Y \hookrightarrow X_{0}$ and if every operator $T \in \mathcal{B}\left(X_{0}\right)$ with $\left.T\right|_{X_{1}} \in \mathcal{B}\left(X_{1}\right)$ satisfies $\left.T\right|_{Y} \in \mathcal{B}(Y)$. For $q \in[1, \infty)$ and $\theta \in(0,1)$, we let $X_{\theta, q}=\left(X_{0}, X_{1}\right)_{\theta, q}$ denote the real interpolation space and $X_{\theta}=\left[X_{0}, X_{1}\right]_{\theta}$ denote the complex interpolation space. The following inequalities are valid.

$$
\|\cdot\|_{\left(X_{0}, X_{1}\right)_{\theta, q}} \leq C(\theta, q)\|\cdot\|_{0}^{1-\theta}\|\cdot\|_{1}^{\theta} \text { on } X_{1}, \quad\|\cdot\|_{\left[X_{0}, X_{1}\right]_{\theta}} \leq\|\cdot\|_{0}^{1-\theta}\|\cdot\|_{1}^{\theta} \text { on } X_{1} .
$$

If $X_{1} \hookrightarrow^{d} Y \hookrightarrow^{d} X_{0}$, then

$$
\left(X_{0}, X_{1}\right)_{\theta, q} \hookrightarrow^{d}\left(X_{0}, Y\right)_{\theta, q}, \quad\left[X_{0}, X_{1}\right]_{\theta} \hookrightarrow^{d}\left[X_{0}, Y\right]_{\theta} .
$$

If $\theta \in(0,1), q \in(1, \infty)$ and $r_{j}: X_{j} \rightarrow Y_{j}$ are isomorphisms with $r_{1}=\left.r_{0}\right|_{X_{1}}$, then

$$
\left(r_{0} X_{0}, r_{1} X_{1}\right)_{\theta, q}=r_{0}\left(X_{0}, X_{1}\right)_{\theta, q}, \quad\left[r_{0} X_{0}, r_{1} X_{1}\right]_{\theta}=r_{0}\left[X_{0}, X_{1}\right]_{\theta} .
$$

For $p \in[1, \infty)$ and a $\sigma$-finite measure space $(\Omega, \mathcal{A}, \mu)$, we have

$$
\begin{aligned}
\left(L_{p}\left(\Omega ; X_{0}\right), L_{p}\left(\Omega ; X_{1}\right)\right)_{\theta, q} & =L_{p}\left(\Omega ;\left(X_{0}, X_{1}\right)_{\theta, q}\right) \\
{\left[L_{p}\left(\Omega ; X_{0}\right), L_{p}\left(\Omega ; X_{1}\right)\right]_{\theta} } & =L_{p}\left(\Omega ;\left[X_{0}, X_{1}\right]_{\theta}\right)
\end{aligned}
$$

For two sectorial operators $A$ and $B$ in $X$ with commuting resolvents, we have

$$
(X, D(A) \cap D(B))_{\theta, q}=D_{A}(\theta, q) \cap D_{B}(\theta, q) \quad \text { for } \theta \in(0,1), q \in[1, \infty]
$$

Let us abbreviate $\mathfrak{F}:=\mathfrak{F}\left(\mathbb{R}^{n} ; X\right)$ for $\mathfrak{F} \in\left\{L_{p}, W_{p}^{m}, W_{p}^{s}\right\}$. In terms of real interpolation, [Ama97, (5.8), (5.9)] yields the representation

$$
W_{p}^{s}=B_{p p}^{s}=\left(L_{p}, W_{p}^{m}\right)_{s / m, p} \quad \text { for } p \in[1, \infty), s \in(0, \infty) \backslash \mathbb{N}, m \in \mathbb{N}, s<m
$$

From [Ama97, (5.2)-(5.6), (5.8), (5.15)] we derive the embedding

$$
W_{p}^{s} \hookrightarrow W_{q}^{t} \quad \text { if } s, t \in[0, \infty), s-n / p>t-n / q, \quad 1 \geq 1 / p>1 / q>0
$$

Moreover, from [Ama97, (5.2)-(5.6), (5.8), (5.16)] we obtain

$$
\begin{equation*}
W_{p}^{s} \hookrightarrow B U C^{t} \quad \text { if } s-n / p>t, p \in[1, \infty) \tag{B.1}
\end{equation*}
$$

By [Ama97, (5.2), (5.15)], we further have

$$
\begin{equation*}
B_{p p}^{m+\varepsilon} \hookrightarrow W_{p}^{m} \hookrightarrow B_{p p}^{m-\delta} \quad \text { for } \varepsilon>0, \delta \in(0, m), m \in \mathbb{N}_{0}, p \in[1, \infty) \tag{B.2}
\end{equation*}
$$

By [SSS12, Proposition 2.13, Theorem 2.20] and [Ama97, (5.2)] we have

$$
\begin{equation*}
B_{p p}^{s+\varepsilon} \hookrightarrow H_{p}^{s} \hookrightarrow B_{p p}^{s-\varepsilon} \quad \text { for } \varepsilon>0, s \in \mathbb{R}, p \in[1, \infty) \tag{B.3}
\end{equation*}
$$

## B.1.2. Regularity of domains, embeddings, and extensions.

B.1. Definition (Cone condition). Given $x \in \mathbb{R}^{n}, r>0, \theta>0, v \in \mathbb{R}^{n} \backslash\{0\}$, the set

$$
x+C_{r, \theta, v}=x+\left\{y \in \mathbb{R}^{n}: y=0 \text { or }|y| \in(0, r], \angle(y, v) \leq \theta / 2\right\}
$$

is called finite cone with vertex x , height $r$, direction $v(x)$ and opening angle $\theta$. The angle $\alpha=\angle(y, v) \in[0, \pi]$ between $y, v \in \mathbb{R}^{n} \backslash\{0\}$ is defined by $y \cdot v=|y||v| \cos \alpha$.

A domain $\Omega \subset \mathbb{R}^{n},(n \in \mathbb{N})$, satisfies the cone condition if there exist $r>0, \theta>0$ such that each $x \in \Omega$ is the vertex of a finite cone $x+C_{r, \theta, v(x)} \subset \Omega$, for some $v(x) \in \mathbb{R}^{n} \backslash\{0\}$.
B.2. Definition (Local Lipschitz Condition). A bounded domain $\Omega \subset \mathbb{R}^{n}$ satisfies the local Lipschitz condition, if each $x \in \partial \Omega$ has a neighborhood $U_{x} \subset \mathbb{R}^{n}$ such that $U_{x} \cap \partial \Omega$ is the graph of a Lipschitz continuous function; that is, there is $V \subset \mathbb{R}^{n-1}, f \in C^{0,1}(V ; \mathbb{R})$ and an orthogonal transformation $Q$ such that $U_{x} \cap \partial \Omega=x+Q$ graph $f=\{x+Q(v, f(v)): v \in V\}$.
B.3. Theorem. Let $\Omega \subset \mathbb{R}^{n}$ be a domain, $X$ be a Banach space, $p \in[1, \infty), q \in[1, \infty], s \in \mathbb{R}, k \in \mathbb{N}_{0}$. If $\Omega$ satisfies the cone condition, then

$$
B_{p q}^{s+k}(\Omega ; X) \hookrightarrow B C^{k}(\Omega ; X), \quad \text { if } s-n / p>0
$$

If $\Omega$ satisfies the strong local Lipschitz condition, then

$$
B_{p q}^{s+k}(\Omega ; X) \hookrightarrow B U C^{k}(\Omega ; X), \quad \text { if } s-n / p>0
$$

Proof. The assertions for the scalar-valued case $X=\mathbb{K}$ are known [AF03, Theorem 7.34, Theorem 7.37]. To obtain the vector-valued result we consider $u \in W_{p}^{s+k}(\Omega ; X)$ and $x^{\prime} \in X^{\prime}$. Then $x^{\prime} \circ u$ belongs to $W_{p}^{s+k}(\Omega)$ with $\left\|x^{\prime} \circ u\right\|_{W_{p}^{s+k}} \leq\left\|x^{\prime}\right\|_{X^{\prime}}\|u\|_{W_{p}^{s+k}(X)}$. The scalar-valued embedding implies $x^{\prime} \circ u \in B C^{k}(\Omega)$ with $\left\|x^{\prime} \circ u\right\|_{B C^{k}} \leq C\left\|x^{\prime} \circ u\right\|_{W_{p}^{s+k}}$, where $C$ denotes the embedding constant for $W_{p}^{s+k}(\Omega) \hookrightarrow B C^{k}(\Omega)$. Assume in addition that $u \in \mathcal{S}\left(\mathbb{R}^{n} ; X\right)$. Then

$$
\|u\|_{B C^{k}(X)}=\sup _{\left\|x^{\prime}\right\| \leq 1}\left\|x^{\prime}(u)\right\|_{B C^{k}} \leq \sup _{\left\|x^{\prime}\right\| \leq 1} C\left\|x^{\prime}(u)\right\|_{W_{p}^{s+k}} \leq C\|u\|_{W_{p}^{s+k}(X)}
$$

Hence the identity is bounded from a dense subset of $W_{p}^{s+k}(\Omega ; X)$ into $B C^{k}(\Omega ; X)$. Approximation yields $W_{p}^{s+k}(\Omega ; X) \hookrightarrow B C^{k}(\Omega ; X)$. The second embedding can be shown analogously.
B.4. Lemma (Hardy's inequality [cf. Dur70, Appendix B]). Let $p \in[1, \infty), r \in(-1+1 / p, \infty)$, $T \in(0, \infty]$ and let $X$ be a Banach space. If $\left[t \mapsto t^{-r} g(t)\right] \in L_{p}(0, T ; X)$, then $\left[x \mapsto x^{-r-1} \int_{0}^{x} g(t) d t\right] \in$ $L_{p}(0, T ; X)$ and the following inequality is valid.

$$
\begin{equation*}
\left(\int_{0}^{T}\left\|\frac{1}{x^{1+r}} \int_{0}^{x} g(t) d t\right\|^{p} d x\right)^{1 / p} \leq \frac{1}{1+r-1 / p}\left(\int_{0}^{T} \frac{1}{t^{r p}}\|g(t)\|^{p} d t\right)^{1 / p} \tag{B.4}
\end{equation*}
$$

Proof. The result can be proved similarly as in [Dur70, Appendix B], where the case $r=0$ is considered. First let $T<\infty$. We employ the substitution $t=x s / T$ and the continuous version

$$
\left\|\int_{0}^{T} f(t, \cdot) d t\right\|_{L_{p}(\mu)} \leq \int_{0}^{T}\|f(t, \cdot)\|_{L_{p}(\mu)} d t
$$

of Minkowski's inequality with respect to the measure $d \mu(x)=d x / x^{r p}$. Then

$$
\begin{aligned}
\left(\int_{0}^{T}\left(\frac{1}{x^{1+r}} \int_{0}^{x}\|g(t)\|_{X} d t\right)^{p} d x\right)^{1 / p} & =\left(\int_{0}^{T}\left(\frac{1}{T x^{r}} \int_{0}^{T}\left\|g\left(\frac{x s}{T}\right)\right\|_{X} d s\right)^{p} d x\right)^{1 / p} \\
& \leq \frac{1}{T} \int_{0}^{T}\left(\int_{0}^{T} \frac{1}{x^{r p}}\left\|g\left(\frac{x s}{T}\right)\right\|_{X}^{p} d x\right)^{1 / p} d s
\end{aligned}
$$

provided that the right-hand side is finite. But this follows with the substitution $x s / T=u$,

$$
\begin{aligned}
& \frac{1}{T} \int_{0}^{T}\left(\int_{0}^{T} \frac{1}{x^{r p}}\left\|g\left(\frac{x s}{T}\right)\right\|_{X}^{p} d x\right)^{1 / p} d s \\
& \quad \leq \frac{1}{T^{1+r-1 / p}} \int_{0}^{T} t^{r-1 / p}\left(\int_{0}^{t} \frac{1}{u^{r p}}\|g(u)\|_{X}^{p} d u\right)^{1 / p} d t \\
& \quad \leq \frac{1}{1+r-1 / p}\left(\int_{0}^{T} \frac{1}{u^{r p}}\|g(u)\|_{X}^{p} d u\right)^{1 / p} .
\end{aligned}
$$

By Fubini's theorem and the finiteness of the right-hand side, the left-hand side is also finite and this proves Hardy's inequality for the case $T<\infty$. The assertion for $T=\infty$ follows by taking limits as $T \rightarrow \infty$.
B.5. Lemma. Let $X$ be a Banach space, $p \in[1, \infty), T \in(0, \infty), \alpha \in(1 / p, \infty)$. Then the following inequality is valid for every $u \in L_{0}(0, T ; X)$ with $\left[t \mapsto t^{-\alpha} u(t)\right] \in L_{p}(0, T ; X)$.

$$
\left(\int_{0}^{T} \frac{1}{t^{\alpha p}}\|u(t)\|_{X}^{p} d t\right)^{1 / p} \leq \frac{1}{2^{p}} \frac{1+\alpha-1 / p}{\alpha-1 / p}\left(\int_{0}^{T} \int_{0}^{T} \frac{\|u(t)-u(s)\|_{X}^{p}}{|t-s|^{1+\alpha p}} d s d t\right)^{1 / p} .
$$

Proof. This inequality can be checked by an inspection of the proof of [PSS07, (6.8)].
The spaces $L_{p}(\Omega ; X), W_{p}^{m}(\Omega ; X), W_{p}^{s}(\Omega ; X)$ were defined intrinsically; that is, by using only the values of functions at points in $\Omega$. Alternatively, we consider the corresponding spaces of restrictions to $\Omega$ of functions on $\mathbb{R}^{n}$, defined by

$$
\begin{aligned}
& \mathfrak{F}(\Omega ; X):=\left.\mathfrak{F}\left(\mathbb{R}^{n} ; X\right)\right|_{\Omega}:=\left\{\left.u\right|_{\Omega}: u \in \mathfrak{F}\left(\mathbb{R}^{n} ; X\right)\right\}, \\
&\|u\|_{\left.\mathfrak{F}\left(\mathbb{R}^{n} ; X\right)\right|_{\Omega}}:=\inf \left\{\|v\|_{\mathfrak{F}\left(\mathbb{R}^{n} ; X\right)}: v \in \mathfrak{F}\left(\mathbb{R}^{n} ; X\right),\left.v\right|_{\Omega}=u\right\},
\end{aligned}
$$

where $\mathfrak{F} \in\left\{L_{p}, W_{p}^{m}, W_{p}^{s}\right\}$. Then we obtain the embeddings

$$
\left.L_{p}\left(\mathbb{R}^{n} ; X\right)\right|_{\Omega} \hookrightarrow L_{p}(\Omega ; X),\left.\quad W_{p}^{m}\left(\mathbb{R}^{n} ; X\right)\right|_{\Omega} \hookrightarrow W_{p}^{m}(\Omega ; X),\left.\quad W_{p}^{s}\left(\mathbb{R}^{n} ; X\right)\right|_{\Omega} \hookrightarrow W_{p}^{s}(\Omega ; X) .
$$

If $\Omega \subset \mathbb{R}^{n}$ is a bounded domain of cone-type (see [Tri95, Definition 4.2.3]), then

$$
\left.F_{p, 2}^{m}\left(\mathbb{R}^{n}\right)\right|_{\Omega}=\left.H_{p}^{m}\left(\mathbb{R}^{n}\right)\right|_{\Omega}=W_{p}^{m}(\Omega)
$$

for $m \in \mathbb{N}, p \in(1, \infty)$ [Tri95, Theorem 4.2.4]. The Besov space $B_{p q}^{s}(\Omega)$ is also given as the real interpolation space

$$
B_{p q}^{s}(\Omega)=\left(L_{p}(\Omega), H_{p}^{m}(\Omega)\right)_{s / m, q}, \quad s \in(0, \infty), p \in[1, \infty), q \in[1, \infty],
$$

where $m$ is the smallest integer larger than $s$ [AF03, p. 7.32]. If $\Omega=\mathbb{R}^{n}$, we can choose any $m \in \mathbb{N}$ with $m>s$ [Tri95, 2.4.2 Remark 2].

The trivial extension by zero is bounded from $L_{p}(\Omega ; X)$ to $L_{p}\left(\mathbb{R}^{n}\right)$, thus the spaces $L_{p}(\Omega ; X)$ and $\left.L_{p}\left(\mathbb{R}^{n}\right)\right|_{\Omega}$ coincide with equal norms. However, this operator does not map continuous functions on $\Omega$ to continuous functions on $\mathbb{R}^{n}$ and is hence not necessarily bounded from $W_{p}^{m}(\Omega ; X)$ to $W_{p}^{m}\left(\mathbb{R}^{n}\right)$. In fact, function spaces on domains defined via restriction maybe smaller then those defined intrinsically, by nonexistence of extension operators [AF03, Paragraphs 3.20, 6.47.1, 7.32].

Extension theorems guarantee the existence of bounded extension operators, if the boundary of $\Omega$ is sufficiently regular. Then it follows immediately, that the space of restrictions coincides with the intrinsically defined space and the corresponding norms are equivalent, see Corollary B. 8 for an example. This is very useful to transfer properties of function spaces on $\mathbb{R}^{n}$ to those on domains.

We will employ the following extension operators from $\Omega=\mathbb{R}_{+}^{n}$ to $\mathbb{R}^{n}(n \in \mathbb{N})$, which are defined in [AF03, Theorem 5.19] by higher order reflections.
B.6. Theorem. Let $k \in \mathbb{N}_{0}$. We define extension operators $E^{k}$ and $E_{\alpha}^{k}\left(\alpha \in \mathbb{N}_{0}^{n},|\alpha| \leq k\right)$ from $\mathbb{R}_{+}^{n}$ to $\mathbb{R}^{n}$ by (the sum over $1 \leq j \leq 0$ is considered as zero)

$$
\begin{aligned}
& E^{k} u\left(x^{\prime},-x_{n}\right):=\sum_{j=1}^{k} \lambda_{j, k} u\left(x^{\prime}, j x_{n}\right), \\
& E_{\alpha}^{k} u\left(x^{\prime},-x_{n}\right):=\sum_{j=1}^{k}(-j)^{\alpha_{n}} \lambda_{j, k} u\left(x^{\prime}, j x_{n}\right),
\end{aligned}
$$

where $u \in L_{1, \text { loc }}\left(\overline{\mathbb{R}_{+}^{n}}\right), x^{\prime} \in \mathbb{R}^{n-1}, x_{n} \in \mathbb{R}_{+}$, and the numbers $\lambda_{j, k}$ solve the linear system

$$
\sum_{j=1}^{k}(-j)^{l} \lambda_{j, k}=1 \quad \text { for all } l \in\{0,1, \ldots, k-1\}
$$

Then

$$
E^{k} \in \mathcal{B}\left(H_{p}^{l}\left(\mathbb{R}_{+}^{n}\right) ; H_{p}^{l}\left(\mathbb{R}^{n}\right)\right), \quad E_{\alpha}^{k} \in \mathcal{B}\left(H_{p}^{l-|\alpha|}\left(\mathbb{R}_{+}^{n}\right) ; H_{p}^{l-|\alpha|}\left(\mathbb{R}^{n}\right)\right), \quad \partial_{x}^{\alpha} E=E_{\alpha} \partial_{x}^{\alpha},
$$

for all $p \in[1, \infty), l \in\{0,1, \ldots, k\}, \alpha \in \mathbb{N}_{0}^{n}$ with $|\alpha| \leq l$.
B.7. Theorem (Stein's extension theorem [Ste70], [AF03, Theorem 5.24]). If $\Omega$ is a domain in $\mathbb{R}^{n}$ that satisfies the strong local Lipschitz condition, then there exists a linear extension operator, which is bounded from $W_{p}^{m}(\Omega)$ to $W_{p}^{m}\left(\mathbb{R}^{n}\right)$ for all $m \in \mathbb{N}_{0}$ and all $p \in[1, \infty)$.
B.8. Corollary. Let $p \in[1, \infty)$ and suppose that the domain $\Omega \subset \mathbb{R}^{n}(n \in \mathbb{N})$ satisfies the strong local Lipschitz condition. Then the following norms on $W_{p}^{m}(\Omega ; X)$ are equivalent:

$$
\|u\|_{1, \Omega}=\left(\sum_{|\alpha| \leq m}\left\|\partial^{\alpha} u\right\|_{L_{p}(\Omega)}^{p}\right)^{1 / p}, \quad\|u\|_{2, \Omega}=\inf \left\{\|v\|_{1, \mathbb{R}^{n}}: v \in W_{p}^{m}\left(\mathbb{R}^{n} ; X\right),\left.v\right|_{\Omega}=u\right\} .
$$

For a bounded interval $(0, T)$ and a fixed order of differentiability $k \in \mathbb{N}_{0}$ or $s \in[0, \infty)$, it is possible to construct an extension operator with a uniform norm bound with respect to $T \in(0, \infty)$ and power $p \in[1, \infty)$. We also refer to [PSS07, Proposition 6.1].

For $s \in[0, \infty)$ with $s-1 / p \notin \mathbb{N}_{0}$ and a Banach space $X$, we define the space

$$
{ }_{0} W_{p}^{s}(0, T ; X):=\overline{C_{c}^{\infty}((0, T] ; X)}{ }^{\|\cdot\|_{W_{p}^{s}}}=\left\{u \in W_{p}^{s}(0, T ; X):\left.\partial_{t}^{j} u\right|_{t=0}=0 \text { for } j \leq[s-1 / p]\right\} .
$$

Here $[s-1 / p]:=\min \{k \in \mathbb{Z}: k \leq s-1 / p\}$ denotes the integer part of $s-1 / p \in \mathbb{R} \backslash \mathbb{N}_{0}$ and the above characterization of ${ }_{0} W_{p}^{s}(0, T ; X)$ follows from [Ama09, Theorem 4.6.2].
B.9. Lemma ([MS12, Lemma 2.5]). Let $J=(0, T)$ be finite, $p \in(1, \infty), \mu \in(1 / p, 1]$ and $X$ be a Banach space of class $\mathcal{H} \mathcal{T}$. Given $k \in \mathbb{N}$, there is an extension operator $\mathcal{E}_{J}$ from $J$ to $\mathbb{R}_{+}$with

$$
\mathcal{E}_{J} \in \mathcal{B}\left(W_{p, \mu}^{s}(J ; X) ; W_{p, \mu}^{s}\left(\mathbb{R}_{+} ; X\right)\right) \cap \mathcal{B}\left(H_{p, \mu}^{s}(J ; X) ; H_{p, \mu}^{s}\left(\mathbb{R}_{+} ; X\right)\right), \quad \text { for all } s \in[0, k]
$$

Here we can replace $W$ by ${ }_{0} W$ and $H$ by ${ }_{0} H$. There is further an extension operator

$$
\mathcal{E}_{J}^{0} \in \mathcal{B}\left({ }_{0} W_{p, \mu}^{s}(J ; X) ;{ }_{0} W_{p, \mu}^{s}\left(\mathbb{R}_{+} ; X\right)\right) \cap \mathcal{B}\left({ }_{0} H_{p, \mu}^{s}(J ; X) ;{ }_{0} H_{p, \mu}^{s}\left(\mathbb{R}_{+} ; X\right)\right), \quad \text { for all } s \in[0,2]
$$

which is independent of the space $X$ and whose operator norm has a uniform bound with respect to $T \in(0, \infty)$. Moreover,

$$
\mathcal{E}_{J}, \mathcal{E}_{J}^{0} \in \mathcal{B}\left(L_{\infty}(J ; X) ; L_{\infty}\left(\mathbb{R}_{+} ; X\right)\right)
$$

where the operator norms have a uniform bound with respect to $T \in(0, \infty)$.
B.1.3. Intrinsic spaces on hypersurfaces. Let $\Sigma \subset \mathbb{R}^{n+1}(n \in \mathbb{N})$ be a compact smooth hypersurface (without boundary) and let $p \in[1, \infty]$ and $s \in[0, \infty)$. There are two approaches to define the vector-valued Sobolev-Slobodeckiĭ spaces $W_{p}^{s}(\Sigma ; X)$ for a Banach space $X$ over $\mathbb{K} \in\{\mathbb{R}, \mathbb{C}\}$. For the intrinsic approach we define it as the closure of $C^{\infty}(\Sigma ; X)$ in the norm

$$
\begin{equation*}
\|u\|_{W_{p}^{s}}:=\|u\|_{W_{p}^{\lfloor s\rfloor}}+\llbracket \partial^{\lfloor s\rfloor} u \rrbracket_{W_{p}^{\{s\}}}, \quad \llbracket v \rrbracket_{W_{p}^{\{s\}}}:=\left(\int_{\Sigma \times \Sigma} \frac{|v(x)-v(y)|^{p}}{\operatorname{dist}_{\Sigma}(x, y)^{n+\{s\} p}} d \Sigma^{2}(x, y)\right)^{1 / p}, \tag{B.5}
\end{equation*}
$$

where we require that $p<\infty$ if $s \notin \mathbb{N}_{0}$ and in the case $s \in \mathbb{N}_{0}$, the seminorm $\llbracket \partial^{\lfloor s\rfloor} u \rrbracket_{W_{p}^{\{s\}}}$ is omitted. The intrinsic distance dist $\Sigma_{\Sigma}$ is studied on page 133. Note that for a function $u: \Sigma \rightarrow X$, the derivative $\partial_{\Sigma}^{j} u(p)$ of order $j$ belongs to $\mathcal{B}^{j}\left(\left(T_{p} \Sigma\right)^{j} ; X\right)$ and can be identified with some element of $\mathcal{B}^{j}\left(\left(\mathbb{R}^{n+1}\right)^{j} ; X\right)$.

It is useful to relate the space $W_{p}^{s}(\Sigma ; X)$ to the corresponding spaces on the whole space $\mathbb{R}^{n}$, for the purpose of using the known embedding and interpolation properties of the latter spaces. To this end we consider the extrinsic definition of $W_{p}^{s}(\Sigma ; X)$ as a retract of $W_{p}^{s}\left(\mathbb{R}^{n} ; X\right)^{N}$ with some $N \in \mathbb{N}$. A bounded linear operator $r: X \rightarrow Y$ between normed vector spaces $X$ and $Y$ is called retraction if there exists a bounded linear operator $r^{c}: Y \rightarrow X$ such that $r r^{c}=I_{Y}$ In this case we say that $r^{c}$ is a co-retraction for $r$ and $Y$ is a retract of $X$. As in [Tri10, Definition 3.2.2/2], we shall show that the map $r$ defined by $r(u)=\left(\left(\chi_{j} u\right) \circ \varphi_{j}^{-1}\right)_{j}$ is a retraction, where $\left(\varphi_{j}, U_{j}\right)_{j=1}^{N}$ is an atlas and $\left(\chi_{j}\right)_{j=1}^{N}$ is a finite partition of unity for $\Sigma$, subordinate to $\left(U_{j}\right)_{j=1}^{N}$.
B.10. Lemma. Let $n \in \mathbb{N}, p \in[1, \infty), s \in(0,1)$ and let $X, Y, Z$ be Banach spaces with continuous multiplication $X \times Y \rightarrow Z,(x, y) \mapsto x y$. Let $\omega_{n}$ denote the $(n-1)$-dimensional area of $\left\{x \in \mathbb{R}^{n}\right.$ : $|x|=1\}$.
(i) For $u \in W_{\infty}^{1}\left(\mathbb{R}^{n} ; X\right)$ and $v \in W_{p}^{s}\left(\mathbb{R}^{n} ; Y\right)$ we have

$$
\begin{equation*}
\llbracket u v \rrbracket_{W_{p}^{s}} \leq\|u\|_{\infty} \llbracket v \rrbracket_{W_{p}^{s}}+2^{1-s}\left(\frac{\omega_{n}}{s(1-s) p}\right)^{1 / p}\|u\|_{\infty}^{1-s}\|\nabla u\|_{\infty}^{s}\|v\|_{p} \tag{B.6}
\end{equation*}
$$

(ii) For $n=1, T \in(0, \infty)$, $u \in W_{\infty}^{1}(0, T ; X)$, and $v \in W_{p}^{s}(0, T ; Y)$, we also have (B.6).
(iii) Let $\Sigma \subset \mathbb{R}^{n+1}$ be a $C^{1-}$-hypersurface such that the numbers

$$
C_{1}(R)=\sup _{x \in \Sigma}\left(\int_{B_{R}^{\Sigma}(x)} \frac{\operatorname{dist}_{\Sigma}(x, y)^{p} d \Sigma(y)}{\operatorname{dist}_{\Sigma}(x, y)^{n+s p}}\right)^{1 / p}, \quad C_{2}(R)=2 \sup _{x \in \Sigma}\left(\int_{\Sigma \backslash B_{R}^{\Sigma}(x)} \frac{d \Sigma(y)}{\operatorname{dist}_{\Sigma}(x, y)^{n+s p}}\right)^{1 / p}
$$

are finite for some $R>0$. Then for all $u \in W_{\infty}^{1}(\Sigma ; X)$ and $v \in W_{p}^{s}(\Sigma ; Y)$ we have

$$
\llbracket u v \rrbracket_{W_{p}^{s}(\Sigma)} \leq\|u\|_{L_{\infty}(\Sigma)} \llbracket v \rrbracket_{W_{p}^{s}(\Sigma)}+\left(C_{1}(R)^{p}\left\|\nabla_{\Sigma} u\right\|_{L_{\infty}(\Sigma)}^{p}+C_{2}(R)^{p}\|u\|_{L_{\infty}(\Sigma)}^{p}\right)^{1 / p}\|v\|_{L_{p}(\Sigma)}
$$

(iv) Let $\Sigma \subset \mathbb{R}^{n+1}$ be a $C^{3-}$-hypersurface. If $L_{\Sigma}=-\nabla_{\Sigma} \nu_{\Sigma}$ is bounded, then for $R_{*}:=\sqrt{2}\left\|L_{\Sigma}\right\|_{\infty}^{-1}$, $R \in\left(0, R_{*}\right)$, and $\delta=\delta(R):=1-R^{2}\left\|L_{\Sigma}\right\|_{\infty}^{2} / 2 \in(0,1]$, we have

$$
C_{1}(R) \leq \frac{\omega_{n}^{1 / p} R^{1-s}}{\delta^{1-s-1 / p}((1-s) p)^{1 / p}}<\infty .
$$

If $\Sigma$ is compact, then

$$
C_{2}(R) \leq \frac{2|\Sigma|^{1 / p}}{R^{s+n / p}}<\infty .
$$

If $\Sigma$ is a perturbed hyperplane $\left\{(x, h(x)): x \in \mathbb{R}^{n}\right\}$ with $h \in C_{c}^{3-}\left(\mathbb{R}^{n}\right)$, then

$$
C_{2}(R) \leq \frac{2 \omega_{n}^{1 / p}\left(1+\|\nabla h\|_{\infty}^{2}\right)^{1 / 2 p}}{(s p)^{1 / p} R^{s}}<\infty
$$

Proof. (i) First, Minkowski's inequality yields

$$
\begin{aligned}
\llbracket u v \rrbracket_{W_{p}^{s}\left(\mathbb{R}^{n}\right)} & =\left(\iint_{\mathbb{R}^{2 n}} \frac{|u(x+y)(v(x+y)-v(x))+(u(x+y)-u(x)) v(x)|^{p}}{|y|^{n+s p}} d(x, y)\right)^{\frac{1}{p}} \\
& \leq\|u\|_{L_{\infty}\left(\mathbb{R}^{n}\right)} \llbracket v \rrbracket_{W_{p}^{s}\left(\mathbb{R}^{n}\right)}+\left(\iint_{\mathbb{R}^{2 n}} \frac{|u(x+y)-u(x)|^{p}|v(x)|^{p}}{|y|^{n+s p}} d(x, y)\right)^{\frac{1}{p}} .
\end{aligned}
$$

Next, we may consider the case $\nabla u \neq 0$ and let $R:=2\|u\|_{\infty} /\|\nabla u\|_{\infty}$. Then

$$
\begin{aligned}
& \iint_{\mathbb{R}^{2 n}} \frac{|u(x+y)-u(x)|^{p}|v(x)|^{p}}{|y|^{n+s p}} d(x, y) \\
& \quad \leq \int_{\mathbb{R}^{n}}\left(\int_{|y| \leq R} \frac{\|\nabla u\|_{L_{\infty}\left(\mathbb{R}^{n}\right)}^{p}}{|y|^{n-(1-s) p}} d x+\int_{|y|>R} \frac{2^{p}\|u\|_{L_{\infty}\left(\mathbb{R}^{n}\right)}^{p}}{|y|^{n+s p}} d x\right)|v(x)|^{p} d x \\
& \quad=\frac{2^{(1-s) p} \omega_{n}}{s(1-s) p}\|u\|_{L_{\infty}\left(\mathbb{R}^{n}\right)}^{(1-s) p}\|\nabla u\|_{L_{\infty}\left(\mathbb{R}^{n}\right)}^{s p}\|v\|_{L_{p}\left(\mathbb{R}^{n}\right)}^{p} .
\end{aligned}
$$

Combining these estimates, we obtain inequality (B.6).
(ii) For $R=2\|u\|_{\infty} /\left\|u^{\prime}\right\|_{\infty}$ we obtain

$$
\begin{aligned}
& \iint_{(0, T)^{2}} \frac{|u(x)-u(y)|^{p}|v(x)|^{p}}{|x-y|^{1+s p}} d(x, y) \\
& \quad \leq \int_{0}^{T}\left(\int_{x-R}^{x+R} \frac{\left\|u^{\prime}\right\|_{\infty}^{p}}{|x-y|^{s p}} d y+2^{p}\|u\|_{\infty}^{p} 2 \int_{R}^{\infty} y^{-1-s p} d y\right)|v(x)|^{p} d x .
\end{aligned}
$$

Hence, with $2=\omega_{1}$, the assertion follows analogously as above.
(iii) From Minkowski's inequality and Fubini's theorem we infer that

$$
\begin{aligned}
\llbracket u v \rrbracket_{W_{p}^{s}} & \leq\|u\|_{\infty} \llbracket v \rrbracket_{W_{p}^{s}}+\left(\iint_{\Sigma^{2}} \frac{|u(x+y)-u(x)|^{p}|v(x)|^{p}}{\operatorname{dist}_{\Sigma}(x, y)^{n+s p}} d \Sigma^{2}(x, y)\right)^{1 / p} \\
& \leq\|u\|_{\infty} \llbracket v \rrbracket_{W_{p}^{s}}+\sup _{x \in \Sigma}\left(\int_{\Sigma} \frac{|u(x+y)-u(x)|^{p}}{\operatorname{dist}_{\Sigma}(x, y)^{n+s p}} d \Sigma(y)\right)^{1 / p}\|v\|_{p} .
\end{aligned}
$$

Clearly,

$$
|u(x+y)-u(x)| \leq \min \left\{\left\|\nabla_{\Sigma} u\right\|_{\infty} \operatorname{dist}_{\Sigma}(x, y), 2\|u\|_{\infty}\right\}
$$

and therefore

$$
\int_{\Sigma} \frac{|u(x+y)-u(x)|^{p}}{\operatorname{dist}_{\Sigma}(x, y)^{n+s p}} d \Sigma(y) \leq C_{1}(R)^{p}\left\|\nabla_{\Sigma} u\right\|_{\infty}^{p}+C_{2}(R)^{p}\|u\|_{\infty}^{p},
$$

which yields the asserted estimate.
(iv) With Proposition A. 13 we can parametrize every geodesic ball $B_{R}^{\Sigma}(x)$ by

$$
\varphi_{x}: U_{x} \rightarrow B_{R}^{\Sigma}(x), \quad \varphi_{x}(u)=x+Q_{x} u+h_{x}(u) \nu_{\Sigma}(x),
$$

where $U_{x} \subset \mathbb{R}^{n}$ is a neighborhood of the origin, $Q_{x} \in \mathbb{R}^{(n+1) \times(n+1)}$ is an orthogonal matrix with $Q_{x} e_{n+1}=\nu_{\Sigma}(x)$, and $h_{x}: \overline{U_{x}} \rightarrow \mathbb{R}$ is a $C^{3-}$-function with $h_{x}(0)=\left|\nabla h_{x}(0)\right|=0$. We further have $B_{R \delta(R)} \subset U_{x} \subset B_{R}$ and

$$
\left(\nu_{\Sigma}(x) \mid \nu_{\Sigma}\left(\varphi_{x}(u)\right)\right) \geq \delta, \quad\left|\nabla h_{x}(u)\right|^{2} \leq \frac{1-\delta^{2}}{\delta^{2}} \quad \text { for all } u \in \overline{U_{x}}, x \in \Sigma .
$$

Proposition A. 12 yields $|u| \leq \operatorname{dist}_{\Sigma}\left(x, \varphi_{x}(u)\right) \leq\left(1+\left|\nabla h_{x}(u)\right|^{2}\right)^{1 / 2}|u| \leq \delta^{-1}|u|$ and thus

$$
\begin{aligned}
\int_{B_{R}^{\Sigma}(x)} \operatorname{dist} \Sigma(x, y)^{(1-s) p-n} d \Sigma(y) & =\int_{U_{x}} \operatorname{dist}_{\Sigma}\left(x, \varphi_{x}(u)\right)^{(1-s) p-n} \sqrt{1+\left|\nabla h_{x}(u)\right|^{2}} d u \\
& \leq \frac{1}{\delta} \int_{\partial B_{1}} \int_{0}^{R} \chi_{U_{x}}(t \zeta)|t \zeta|^{(1-s) p}|t \zeta|^{-n} t^{n-1} d t d \zeta \\
& \leq \frac{\omega_{n} R^{(1-s) p}}{\delta^{(1-s) p+1}(1-s) p} .
\end{aligned}
$$

This yields the estimate for $C_{1}(R)$. The other estimates follows easily.
The previous estimates allow pointwise multiplication with test functions; for instance,

$$
\begin{equation*}
\|u v\|_{W_{p}^{s}\left(\mathbb{R}^{n}\right)} \leq C(n, p, s)\left(\|u\|_{\infty} \llbracket v \rrbracket_{W_{p}^{s}}+\|u\|_{W_{\infty}^{1}}\|v\|_{p}\right) \quad \text { for } u \in W_{\infty}^{1}\left(\mathbb{R}^{n}\right), v \in W_{p}^{s}\left(\mathbb{R}^{n}\right) . \tag{B.7}
\end{equation*}
$$

This in turn yields the equivalence of intrinsic and extrinsic spaces.
B.11. Lemma. Let $\Sigma \subset \mathbb{R}^{n+1}(n \in \mathbb{N})$ be a smooth bounded hypersurface with smooth compact boundary $\partial \Sigma$, let $s \in[0, \infty), p \in[1, \infty)$, and let $X$ be a Banach space. Then the space $W_{p}^{s}(\Sigma ; X)$ endowed with the intrinsic norm (B.5) is a retract of $W_{p}^{s}\left(\mathbb{R}^{n} ; X\right)^{N}$ for some $N \in \mathbb{N}$.
Proof. By Lemma B. 10 we can find a smooth atlas $\left(\varphi_{j}\left(U_{j}\right), \varphi_{j}^{-1}\right)_{j=1}^{N}$ for $\Sigma$ and a smooth partition of unity $\left(\chi_{j}\right)_{j=1}^{N}$ subordinate to $\left(U_{j}\right)_{j=1}^{N}$ where $B_{R \delta(R)} \subset U_{j} \subset B_{R} \subset \mathbb{R}^{n}$ and $\varphi_{j}(u)=x_{j}+$ $u+h_{j}(u) \nu_{\Sigma}\left(x_{j}\right)$ with $x_{j} \in \Sigma$ and $h_{j} \in C^{\infty}\left(\overline{U_{j}}\right)$. Then Lemma B.10, the chain rule (B.19), the transformation formula (A.12), and Proposition A. 12 imply that

$$
r: W_{p}^{s}(\Sigma ; X) \rightarrow W_{p}^{s}\left(\mathbb{R}^{n} ; X\right)^{N}, \quad u \mapsto\left(\left(\chi_{j} u\right) \circ \varphi_{j}\right)_{j=1}^{N}
$$

is well-defined and bounded.
Let further $\left(\psi_{j}\right)_{j=1}^{N}$ be a collection of smooth functions $\psi_{j} \in \mathcal{D}\left(\varphi_{j}\left(U_{j}\right)\right)$ with $\psi_{j}=1$ on $\operatorname{supp} \varphi_{j}\left(U_{j}\right)$. Then it is also straightforward to check that

$$
r^{c}: W_{p}^{s}\left(\mathbb{R}^{n} ; X\right)^{N} \rightarrow W_{p}^{s}(\Sigma ; X), \quad\left(v_{j}\right)_{j=1}^{N} \mapsto \sum_{j=1}^{N} \psi_{j}\left(\left.v_{j}\right|_{U_{j}} \circ \varphi_{j}^{-1}\right)
$$

is a co-retraction for $r$.
B.1.4. Homogeneous function spaces. We define the homogeneous spaces $\dot{H}_{p}^{s}(\Omega), \dot{B}_{p q}^{s}(\Omega)$, and $\dot{F}_{p q}^{s}(\Omega)$ and collect some of their properties. Further information on homogeneous spaces is given by Bergh and Löfström [BL76], Kozono and Sohr [KS91], Simader and Sohr [SS96], Maz'ya [Maz11], and Triebel [Tri10].

There are two main approaches to define the homogeneous spaces on a domain $\Omega \subset \mathbb{R}^{n}$ or on a possibly disconnected open subset $\Omega \subset \mathbb{R}^{n}$. In the case $\Omega=\mathbb{R}^{n}$, these spaces can be defined as subspaces of $\mathcal{S}_{0}^{\prime}\left(\mathbb{R}^{n}\right)$, see Definition B.13. These spaces consist of distributions modulo polynomials. Then the spaces on domains can be defined extrinsically as spaces of restrictions of functions over $\mathbb{R}^{n}$. For instance we can define $\dot{H}_{p}^{s}(\Omega):=\left.\dot{H}_{p}^{s}\left(\mathbb{R}^{n}\right)\right|_{\Omega}:=\{u: \Omega \rightarrow$ $\left.X: \exists v \in \dot{H}_{p}^{s}\left(\mathbb{R}^{n}\right):\left.v\right|_{\Omega}=u\right\}$, equipped with the norm $\|u\|_{\dot{H}_{p}^{s}(\Omega)}:=\inf \left\{\|v\|_{\dot{H}_{p}^{s}\left(\mathbb{R}^{n}\right)}:\left.v\right|_{\Omega}=u\right\}$. The
extrinsic approach allows to transfer known results on embeddings, lifting properties (Theorem B.15) and traces (Theorem B.31, Theorem B.28) to the spaces over domains.

The homogeneous spaces can also be defined intrinsically as equivalence classes of functions on $\Omega$ modulo polynomials whose degree does not exceed some $k \in \mathbb{N}_{0}$. The norm only depends on the chosen subset $\Omega \subset \mathbb{R}^{n}$ and is independent of the representative. In the whole space case $\Omega=\mathbb{R}^{n}$, the intrinsic and extrinsic norms are equivalent (Remark B.14). The same is true for a subset $\Omega \subset \mathbb{R}^{n}$ that admits a bounded extension operator from $\Omega$ to $\mathbb{R}^{n}$ for the intrinsic norm.
B.12. Remark (Approximation of $\dot{H}_{p}^{k}$-functions by $C_{c}^{\infty}$-functions). As in [Gal11, Theorem II.7.1] we consider Sobolev's cut-off function (see Sobolev [Sob63])
$\chi_{R}(x)=\chi\left(\frac{\log \log |x|}{\log \log R}\right)$ with $R>e$ and $\chi \in C^{\infty}([0, \infty) ;[0,1]), \chi=1$ on $\left[0, \frac{1}{2}\right], \chi=0$ on $[1, \infty)$. Thus $\chi_{R}(x)=1$ if $|x| \leq e^{\sqrt{\log R}}, \chi_{R}(x)=0$ if $|x| \geq R$ and the support of $\nabla \chi_{R}$ is contained in $\left\{e^{\sqrt{\log R}} \leq|x| \leq R\right\}$. For every $\alpha \in \mathbb{N}_{0}^{n}$ with $|\alpha| \geq 1$ and $R_{0}>e$ we have the estimate

$$
\left|\partial^{\alpha} \chi_{R}(x)\right| \leq \frac{c_{\alpha, R_{0}}}{\log \log R} \frac{1}{|x|^{|\alpha|} \log |x|}, \quad \text { for } e^{\sqrt{\log R}} \leq|x| \leq R, R \geq R_{0}>e
$$

Furthermore, let $\rho_{r}(x)=r^{-n} \rho(x / r)$ denote Friedrichs' mollifiers with some $\rho \in C_{c}^{\infty}\left(\mathbb{R}^{n}\right)$ such that $\operatorname{supp} \rho=\overline{B_{1}(0)}, \rho \geq 0, \rho(x)>0$ for $|x|<1$, and $\int_{\mathbb{R}^{n}} \rho(x) d x=1$.

Let $n \geq 2, p \in[1, \infty), k \in \mathbb{N}_{0}$ and let $u \in \dot{\mathcal{H}}_{p}^{k}\left(\mathbb{R}^{n}\right)$. Then we can find a polynomial $u_{0}$ in $\mathbb{R}^{n}$ of degree $\leq k-1$ such that $u-u_{0}$ can be approximated in the norm $\left\|\nabla^{k} \cdot\right\|_{L_{p}\left(\mathbb{R}^{n}\right)}$ by test functions

$$
u_{k}=\rho_{r_{k}} *\left(\chi_{k} \cdot\left(u-u_{0}\right)\right), \quad k \in \mathbb{N}
$$

with some sequence $\left(r_{k}\right)_{k}$ such that $r_{k} \rightarrow 0$ as $k \rightarrow \infty$.
B.13. Definition (Extrinsic definition of homogeneous spaces). Let $\mathcal{S}_{0}\left(\mathbb{R}^{n}\right):=\left\{\varphi \in \mathcal{S}\left(\mathbb{R}^{n}\right)\right.$ : $\left(\partial^{\alpha} \mathcal{F} \varphi\right)(0)=0$ for all $\left.\alpha \in \mathbb{N}_{0}^{n}\right\}$ The dual space $\mathcal{S}_{0}^{\prime}\left(\mathbb{R}^{n}\right)$ can be identified with $\mathcal{S}^{\prime}\left(\mathbb{R}^{n}\right) / \mathcal{P}$, where $\mathcal{P}=\cup_{k \geq 0} \mathcal{P}_{k}$ and $\mathcal{P}_{k}$ is the linear space of all polynomials of degree not larger than $k$. Then the homogeneous Besov space and the homogeneous Triebel-Lizorkin space are defined by

$$
\begin{aligned}
& \dot{B}_{p q}^{s}\left(\mathbb{R}^{n}\right):=\left\{u \in \mathcal{S}_{0}^{\prime}\left(\mathbb{R}^{n}\right):\|u\|_{\dot{B}_{p q\left(\mathbb{R}^{n}\right)}^{s}}:=\left(\sum_{j \in \mathbb{Z}}\left(2^{j s}\left\|\mathcal{F}^{-1} \varphi_{j} \mathcal{F} u\right\|_{L_{p}\left(\mathbb{R}^{n}\right)}\right)^{q}\right)^{1 / p}<\infty\right\}, \\
& \dot{F}_{p q}^{s}\left(\mathbb{R}^{n}\right):=\left\{u \in \mathcal{S}_{0}^{\prime}\left(\mathbb{R}^{n}\right):\|u\|_{\dot{F}_{p q\left(\mathbb{R}^{n}\right)}}:=\left\|\left(\sum_{j \in \mathbb{Z}}\left|2^{j s} \mathcal{F}^{-1} \varphi_{j} \mathcal{F} u\right|^{q}\right)^{1 / q}\right\|_{L_{p}\left(\mathbb{R}^{n}\right)}<\infty\right\}
\end{aligned}
$$

Furthermore, we define the homogeneous Bessel potential space $H_{p}^{s}\left(\mathbb{R}^{n}\right):=F_{p 2}^{s}\left(\mathbb{R}^{n}\right)$. We refer to [Tri10, Chapter 5], [BL76, Chapter 6] and [RS96, Section 2.6] for further information.
B.14. Remarks (Properties of homogeneous spaces). (i) For $p \in(1, \infty), s \in \mathbb{R}$, the following embeddings are continuous and dense (see [Tri10, Theorem 5.1.5].

$$
\mathcal{S}_{0}\left(\mathbb{R}^{n}\right) \hookrightarrow \dot{H}_{p}^{s}\left(\mathbb{R}^{n}\right) \hookrightarrow \mathcal{S}_{0}^{\prime}\left(\mathbb{R}^{n}\right), \quad \mathcal{S}_{0}\left(\mathbb{R}^{n}\right) \hookrightarrow \dot{B}_{p p}^{s}\left(\mathbb{R}^{n}\right) \hookrightarrow \mathcal{S}_{0}^{\prime}\left(\mathbb{R}^{n}\right)
$$

(ii) We have $\|f\|_{\dot{H}_{s}^{s}}=0$ if and only if $f$ is a polynomial [see BL76, Section 6.3].
(iii) If $s \in \mathbb{R}, p, q \in[1, \infty]$, and $\theta \in(0,1)$, then [see BL76, Theorem 6.3.1],

$$
\left(\dot{H}_{p}^{s_{0}}\left(\mathbb{R}^{n}\right), \dot{H}_{p}^{s_{1}}\left(\mathbb{R}^{n}\right)\right)_{\theta, q}=\dot{B}_{p q}^{s}\left(\mathbb{R}^{n}\right), \quad \text { if } s=(1-\theta) s_{0}+\theta s_{1}, \theta \in(0,1)
$$

(iv) If $s \in(0, \infty)$ and $p, q \in[1, \infty]$, then [see BL76, Theorem 6.3.2]

$$
B_{p q}^{s}\left(\mathbb{R}^{n}\right)=L_{p}\left(\mathbb{R}^{n}\right) \cap \dot{B}_{p q}^{s}\left(\mathbb{R}^{n}\right), \quad H_{p}^{s}\left(\mathbb{R}^{n}\right)=L_{p}\left(\mathbb{R}^{n}\right) \cap \dot{H}_{p}^{s}\left(\mathbb{R}^{n}\right)
$$

(v) If $m \in \mathbb{N}_{0}$ and $p \in(1, \infty)$, then $u \mapsto \sum_{|\alpha|=m}\left\|\partial^{\alpha} u\right\|_{L_{p}\left(\mathbb{R}^{n}\right)}$ the space $\dot{H}_{p}^{m}\left(\mathbb{R}^{n}\right)$ can be identified with the space $\dot{W}_{p}^{m}\left(\mathbb{R}^{n}\right)$ (see [Tri10, Theorem 5.2.3/1]).
(vi) If $s \in(0,1)$ and $p, q \in[1, \infty)$, then the space $\dot{B}_{p p}^{s}\left(\mathbb{R}^{n}\right)$ can be identified with the space $\dot{W}_{p}^{s}\left(\mathbb{R}^{n}\right)$ (see [Tri10, Theorem 5.2.3/2]).
B.15. Theorem ([Tri10, Theorem 5.2.3/1], [Ste70, Section V.1]). Let

$$
\dot{J}_{\sigma} u:=(-\Delta)^{\sigma / 2} u=\mathcal{F}^{-1}\left(\xi \mapsto|\xi|^{\sigma} \mathcal{F} u(\xi)\right) \quad \text { for } \sigma \in \mathbb{R}, u \in \mathcal{S}_{0}^{\prime}\left(\mathbb{R}^{n}\right)
$$

denote the Riesz potential.
(i) If $s \in \mathbb{R}, \sigma \in \mathbb{R}, q \in[1, \infty], p \in[1, \infty]$, then $\dot{J}_{\sigma}: \dot{B}_{p q}^{s}\left(\mathbb{R}^{n}\right) \rightarrow \dot{B}_{p q}^{s-\sigma}\left(\mathbb{R}^{n}\right)$ is an isomorphism.
(ii) If in addition $p \in[1, \infty)$, then $\dot{J}_{\sigma}: \dot{F}_{p q}^{s}\left(\mathbb{R}^{n}\right) \rightarrow \dot{F}_{p q}^{s-\sigma}\left(\mathbb{R}^{n}\right)$ is an isomorphism.

## B.2. Sectorial operators and maximal regularity

B.16. Definition ([cf. DHP03, Definition 3.1]). A family of operators $\mathcal{T} \subset \mathcal{B}(X)$ is called $\mathcal{R}$ bounded, if there are numbers $C>0$ and $p \in[1, \infty)$ such that the inequality

$$
\left\|\sum_{j=1}^{N} \varepsilon_{j} T_{j} x_{j}\right\|_{L_{p}(\Omega ; X)} \leq C\left\|\sum_{j=1}^{N} \varepsilon_{j} x_{j}\right\|_{L_{p}(\Omega ; X)}
$$

is valid for all $N \in \mathbb{N}, T_{j} \in \mathcal{T}, x_{j} \in X$ and for all independent, symmetric $\{-1,1\}$-valued random variables $\varepsilon_{j}$ on a probability space $(\Omega, \mathcal{M}, \mu)$. The smallest such number $C$ is called the $\mathcal{R}$-bound of $\mathcal{T}$, denoted by $\mathcal{R}(\mathcal{T})$.
B.17. Definition ([cf. AHS94; DHP03]). Let $X$ be a complex Banach space and let $\Sigma_{\theta}$ denote the open sector

$$
\Sigma_{\theta}:=\{\lambda \in \mathbb{C} \backslash\{0\}:|\arg \lambda|<\theta\}=\left\{r e^{i \varphi}: r \in(0, \infty), \varphi \in(-\theta, \theta)\right\}, \quad \theta \in(0, \pi]
$$

We write $f \in \mathcal{H}_{0}^{\infty}\left(\Sigma_{\theta}\right)$ if $f: \Sigma_{\theta} \rightarrow \mathbb{C}$ is a bounded holomorphic function such that there exists $s>0$ such that $|f|(\lambda) \leq \frac{c|\lambda|^{s}}{1+|\lambda|^{s}}$ in $\Sigma_{\theta}$ for some $c \geq 0$.
(i) A linear operator $A: D(A) \rightarrow X$ is called sectorial (of type $(K, \vartheta)$ with $K \geq 1, \vartheta \in(0, \pi)$ ) if both $D(A)$ and $R(A)$ are dense in $X$ and

$$
\Sigma_{\vartheta} \subset \rho(-A) \quad \text { and } \quad\left\|\lambda(\lambda+A)^{-1}\right\|_{\mathcal{B}(X)} \leq K \quad \text { for all } \lambda \in \Sigma_{\vartheta}
$$

We call $\phi_{A}:=\inf \{\pi-\vartheta: \exists K \geq 1: A$ is of type $(K, \vartheta)\}$ the spectral angle of $A$.
(ii) A sectorial operator $A: D(A) \rightarrow X$ is called $\mathcal{R}$-sectorial (of type $(K, \vartheta)$ ) if

$$
\Sigma_{\vartheta} \subset \rho(-A) \quad \text { and } \quad \mathcal{R}\left(\left\{\lambda(\lambda+A)^{-1}: \lambda \in \Sigma_{\vartheta}\right\}\right) \leq K
$$

We call $\phi_{A}^{\mathcal{R}}:=\inf \left\{\pi-\vartheta:\left\{\lambda(\lambda+A)^{-1}: \lambda \in \Sigma_{\vartheta}\right\}\right.$ is $\mathcal{R}$-bounded $\}$ the $\mathcal{R}$-angle of $A$.
(iii) A sectorial operator $A$ is said to be of type $(K, \vartheta)$, if

$$
\bar{\Sigma}_{\vartheta}:=\overline{\Sigma_{\vartheta}} \subset \rho(-A) \quad \text { and } \quad(1+|\lambda|)\left\|(\lambda+A)^{-1}\right\|_{\mathcal{B}(X)} \leq K \quad \text { for all } \lambda \in \bar{\Sigma}_{\vartheta}
$$

(iv) A sectorial operator $A$ is said to have bounded imaginary powers (of type $(C, \theta)$ ), if $A^{i t} \in$ $\mathcal{B}(X)$ for all $t \in \mathbb{R}$ and

$$
\left\|A^{i t}\right\|_{\mathcal{B}(X)} \leq C e^{\theta|t|} \quad \text { for } t \in \mathbb{R}
$$

(v) We say that a $A$ has a bounded $\mathcal{H}^{\infty}$-calculus (of type $(M, \vartheta)$ ), if $A$ is sectorial of type $(K, \vartheta)$ with some $K \geq 1$ and

$$
\|f(A)\|_{\mathcal{B}(X)} \leq M\|f\|_{\infty}, \quad \text { for } f \in \mathcal{H}^{\infty}\left(\Sigma_{\pi-\vartheta}\right)
$$

where $f(A)$ is defined by the extended functional calculus, see Remark B.33.
(vi) We say that $A$ has an $\mathcal{R}$-bounded $\mathcal{H}^{\infty}$-calculus (of type $(M, \vartheta)$ ), if

$$
M:=\mathcal{R}\left(\left\{f(A) \in \mathcal{B}(X): f \in \mathcal{H}^{\infty}\left(\Sigma_{\pi-\vartheta}\right),\|f\|_{\infty} \leq 1\right\}\right)<\infty
$$

We introduce the abbreviations

$$
\begin{aligned}
\mathcal{S}(X ; K, \vartheta) & :=\{A: A \text { is sectorial in } X \text { of type }(K, \vartheta)\}, \\
\mathcal{R S}(X ; K, \vartheta): & :\{A: A \text { is } \mathcal{R} \text {-sectorial in } X \text { of type }(K, \vartheta)\}, \\
\mathcal{P}(X ; K, \vartheta) & :=\{A: A \text { is of type }(K ; \vartheta) \text { in } X\}, \\
\mathcal{B I P}(X ; C, \theta): & =\{A: A \text { has bounded imaginary powers of type }(C, \theta) \text { in } X\}, \\
\mathcal{H}^{\infty}(X ; M, \vartheta): & :\left\{A: A \text { has a bounded } \mathcal{H}^{\infty} \text {-calculus in } X \text { of type }(M, \vartheta) \text { in } X\right\}, \\
\mathcal{R} \mathcal{H}^{\infty}(X ; M, \vartheta): & =\left\{A: A \text { has an } \mathcal{R} \text {-bounded } \mathcal{H}^{\infty} \text {-calculus in } X \text { of type }(M, \vartheta) \text { in } X\right\} .
\end{aligned}
$$

Furthermore, we define $\mathcal{S}(X ; \vartheta):=\cup_{K} \mathcal{S}(X ; K, \vartheta)$ and $\mathcal{S}(X):=\cup_{\vartheta} \mathcal{S}(X ; \vartheta)$ and we will write $\mathcal{S}(K, \vartheta):=\mathcal{S}(X ; K, \vartheta)$ and $\mathcal{S}(\vartheta):=\mathcal{S}(X ; \vartheta)$ if no confusion seems likely, analogously for the other classes. Then we define the angles

$$
\begin{aligned}
\phi_{A} & :=\inf \{\phi \geq 0: A \in \mathcal{S}(\pi-\phi)\} \\
\phi_{A}^{\mathcal{R}} & :=\inf \{\phi \geq 0: A \in \mathcal{R} \mathcal{S}(\pi-\phi)\} \\
\theta_{A} & :=\limsup _{|t| \rightarrow \infty} \frac{\log \left\|A^{i t}\right\|}{|t|}=\inf \{\theta \geq 0: A \in \mathcal{B I P} \mathcal{P}(\theta)\}, \\
\phi_{A}^{\infty} & :=\inf \left\{\phi \geq 0: A \in \mathcal{H}^{\infty}(\pi-\phi)\right\}, \\
\phi_{A}^{\mathcal{R} \infty} & :=\inf \left\{\phi \geq 0: A \in \mathcal{R H}^{\infty}(\pi-\phi)\right\}
\end{aligned}
$$

B.2.1. Maximal $L_{p}$-regularity. We collect some material on analytic semigroups and maximal $L_{p}$-regularity from [Dor93], [Ama95], [Lun95], [Wei01], [Prü02], and [DHP03]. We assume that $A: D(A) \rightarrow X$ is a closed linear operator in a complex Banach space $X$ and that $D(A)$ is equipped with the graph norm $\|\cdot\|_{X}+\|A \cdot\|_{X}$.
B.18. Definition (Analytic semigroup). Let $\theta \in(0, \pi / 2]$ and $\Sigma_{\theta}=\{\lambda \in \mathbb{C} \backslash\{0\}:|\arg \lambda|<\theta\}$. A family $T:=\left\{T(t): t \in \Sigma_{\theta} \cup\{0\}\right\} \subset \mathcal{B}(X)$ is called (strongly continuous) analytic semigroup, if
(i) the map $t \mapsto T(t): \Sigma_{\theta} \rightarrow \mathcal{B}(X)$ is analytic,
(ii) $T(0)=I$ and $T(t) T(s)=T(t+s)$ for all $t, s \in \Sigma_{\theta} \cup\{0\}$ (semigroup property),
(iii) $T\left(t_{n}\right) x \rightarrow x$ in $X$ as $\Sigma_{\theta^{\prime}} \ni t_{n} \rightarrow 0$ for all $x \in X, \theta^{\prime} \in(0, \theta)$ (strong continuity).
B.19. Definition. The generator $A: D(A) \rightarrow X$ of an analytic semigroup $T$ is defined by

$$
A x:=\lim _{t \rightarrow 0+} \frac{T(t) x-x}{t}, \quad D(A):=\left\{x \in X: \lim _{t \rightarrow 0+} \frac{T(t) x-x}{t} \text { exists in } X\right\} .
$$

From now on we let $-A$ be the negative generator of the analytic semigroup $e^{-t A}:=T(t)$. For a given function $f \in L_{1, \text { loc }}([0, \infty) ; X)$ we consider the abstract Cauchy problem

$$
\begin{equation*}
\partial_{t} u(t)+A u(t)=f(t), \quad t \in(0, \infty), \quad u(0)=0 . \tag{B.8}
\end{equation*}
$$

It is known [Ama95, Remarks II.2.1.2] that the unique mild solution $u \in C([0, \infty) ; X)$ of (B.8) is given by the variation of parameters formula

$$
u(t)=\int_{0}^{t} e^{-(t-s) A} f(s) d s, \quad t \in[0, \infty) .
$$

We study the solvability of problem (B.8) with respect to the function spaces

$$
{ }_{0} \mathbb{E}(T):={ }_{0} H_{p}^{1}(0, T ; X) \cap L_{p}(0, T ; D(A)), \quad \mathbb{F}(T)=L_{p}(0, T ; X),
$$

where $T \in(0, \infty]$ and $p \in(1, \infty)$.
B.20. Definition. We say that $A$ has maximal $L_{p}(0, T ; X)$-regularity or maximal $L_{p}$-regularity on $(0, T)$ in $X$ if for every $f \in \mathbb{F}(T)$, the mild solution of problem (B.8) belongs to ${ }_{0} \mathbb{E}(T)$. We let $\mathcal{M} \mathcal{R}_{p}(J ; X)$ denote the class of all operators with maximal $L_{p}(0, T ; X)$-regularity.
B.21. Remarks. The following facts are shown in [Dor93], [Ama95], and [Prü02]. (i) If $A \in$ $\mathcal{M R}_{p}(J ; X)$ is valid for some $p \in(1, \infty)$, then it is valid for all $p \in(1, \infty)$. We will therefore simply write $\mathcal{M R}$ instead of $\mathcal{M R}_{p}$ in the following. (ii) If $A \in \mathcal{M R}\left(\left(0, T_{0}\right) ; X\right)$ for some $T_{0} \in$ $(0, \infty]$, then $A \in \mathcal{M R}((0, T) ; X)$ for all $T \in(0, \infty)$. (iii) If $A \in \mathcal{M R}((0,1) ; X)$, then there is $\mu>0$ such that $\mu+A \in \mathcal{M R}\left(\mathbb{R}_{+} ; X\right)$.

The translations $\mu+A$ with large $\mu>0$ can be avoided if the spectrum $\sigma(A)$ of $A$ is contained in the positive right half-plane.
B.22. Theorem (Kato [see Dor93, Theorem 2.4]). If $\mu+A \in \mathcal{M R}\left(\mathbb{R}_{+} ; X\right)$ for some $\mu \in \mathbb{C}$ and if $\sigma(A) \subset\{\lambda \in \mathbb{C}: \operatorname{Re} \lambda>0\}$, then $A \in \mathcal{M R}\left(\mathbb{R}_{+} ; X\right)$.

It is useful to define the larger class ${ }_{0} \mathcal{M R}\left(\mathbb{R}_{+} ; X\right) \supset \mathcal{M} \mathcal{R}\left(\mathbb{R}_{+} ; X\right)$ of all $A \in \mathcal{M R}((0,1) ; X)$ for which the mild solution $u$ to (B.8) satisfies the weaker a priori estimate

$$
\left\|\partial_{t} u\right\|_{L_{p}\left(\mathbb{R}_{+} ; X\right)}+\|A u\|_{L_{p}\left(\mathbb{R}_{+} ; X\right)} \leq C\|f\|_{L_{p}\left(\mathbb{R}_{+} ; X\right)} \quad \text { for all } f \in L_{p}\left(\mathbb{R}_{+} ; X\right) .
$$

We note that $A \in \mathcal{M R}\left(\mathbb{R}_{+} ; X\right)$ if and only if $A \in{ }_{0} \mathcal{M} \mathcal{R}\left(\mathbb{R}_{+} ; X\right)$ and $0 \in \rho(A)$ [see PS15]. The following characterization of maximal $L_{p}$-regularity is very important and useful.
B.23. Theorem (Weis, [Wei01, Theorem 4.2], [cf. DHP03, Theorem 4.4]). Let $X$ be a Banach space of class $\mathcal{H T}$ and let $A$ generate a bounded analytic semigroup in $X$. Then $A$ belongs to ${ }_{0} \mathcal{M R}\left(\mathbb{R}_{+} ; X\right)$ if and only if $\left\{\lambda(\lambda+A)^{-1}: \lambda \in \Sigma_{\theta}\right\}$ is $\mathcal{R}$-bounded for some $\theta>\pi / 2$.

Next, we study exponentially decaying solutions of the abstract initial value problem

$$
\partial_{t} u+A u=f \text { on } J, \quad u(0)=x .
$$

Let $E, \mathbb{X}(J)$ be Banach spaces such that $\mathbb{X}(J) \hookrightarrow L_{1, \text { loc }}(J ; E)$ where $J=(0, T)$ for $T \in(0, \infty]$ and let $\omega \in \mathbb{R}$. We employ the exponentially weighted space

$$
e^{-\omega} \mathbb{X}(J):=\left\{u \in L_{1, \mathrm{loc}}(J ; E):\left[t \mapsto e^{\omega t} u(t)\right] \in \mathbb{X}(J)\right\},
$$

equipped with the norm $\|u\|_{e^{-\omega \mathbb{X}}(J)}:=\left\|\left[t \mapsto e^{\omega t} u(t)\right]\right\|_{\mathbb{X}(J)}$.
B.24. Proposition ([cf. Ama95, Proposition III.1.5.3]). Suppose that $\omega+A: D(A) \rightarrow X$ has maximal $L_{p}\left(\mathbb{R}_{+} ; X\right)$-regularity for some $\omega \in \mathbb{R}$. Then

$$
\left(\partial_{t}+A, \gamma_{0}\right): e^{\omega \cdot}\left[H_{p}^{1}\left(\mathbb{R}_{+} ; X\right) \cap L_{p}\left(\mathbb{R}_{+} ; D(A)\right)\right] \rightarrow e^{\omega \cdot} L_{p}\left(\mathbb{R}_{+} ; X\right) \times D_{A}(1-1 / p, p)
$$

is an isomorphism.
B.2.2. Fractional domains and abstract trace spaces. For two Banach spaces $X_{0}$ and $X_{1}$ with dense embedding $X_{1} \hookrightarrow X_{0}$ and for $M \geq 1$ and $\vartheta \in(\pi / 2, \pi)$, we define the class

$$
\begin{aligned}
\mathcal{P}_{1}\left(X_{1}, X_{0} ; M, \vartheta\right):=\{ & A \in \mathcal{P}\left(X_{0} ; M, \vartheta\right) \cap \mathcal{B}_{\text {isom }}\left(X_{1} ; X_{0}\right):\|A\|_{\mathcal{B}\left(X_{1} ; X_{0}\right)} \leq M, \\
& \left.(1+|\lambda|)^{1-j}\left\|(\lambda+A)^{-1}\right\|_{\mathcal{B}\left(X_{0} ; X_{j}\right)} \leq M \text { for } j \in\{0,1\}, \lambda \in \bar{\Sigma}_{\vartheta}\right\} .
\end{aligned}
$$

If $A$ belongs to $\mathcal{P}_{1}\left(X_{1}, X_{0} ; M, \vartheta\right)$, then $-A$ generates an exponentially stable analytic semigroup $t \mapsto e^{-t A}$. Arguing as in [AHS94, Section 1], it can be shown that there are $\omega_{0}=$ $\omega_{0}\left(X_{1}, X_{0}, M, \vartheta\right)>0$ and $M^{\prime} \geq 1$ such that for all $\omega \in\left(0, \omega_{0}\right)$ we have $A-\omega \in \mathcal{P}_{1}\left(X_{1}, X_{0} ; M^{\prime}, \vartheta\right)$.

For $\alpha \in(0,1)$ and $p \in(1, \infty)$ we define the seminorms

$$
[x]_{D_{A}(\alpha, p)}:=\left(\int_{0}^{\infty}\left|t^{1-\alpha} A e^{-t A} x\right|_{X_{0}}^{p} \frac{d t}{t}\right)^{1 / p}, \quad \llbracket x \rrbracket_{D_{A}(\alpha, p)}:=\left(\int_{0}^{\infty}\left|t^{-\alpha}\left(e^{-t A}-I\right) x\right|_{X_{0}}^{p} \frac{d t}{t}\right)^{1 / p} .
$$

It is shown in [Lun95, Proposition 2.2.4] that these seminorms are equivalent. The fractional domains of $A$ for $\alpha \in(0,1)$ and $p \in(1, \infty)$ are defined by

$$
D_{A}(\alpha, p):=\left\{x \in X_{0}:[x]_{D_{A}(\alpha, p)}<\infty\right\}, \quad|x|_{D_{A}(\alpha, p)}:=|x|_{X_{0}}+[x]_{D_{A}(\alpha, p)} .
$$

We also put $D_{A}(1, p):=D(A)$ with $[x]_{D_{A}(1, p)}:=|A x|_{X_{0}}$ and we let $\left(R_{A} x\right)(t)=e^{-t A} x$.
B.25. Theorem (cf. [Lun95, Section 2.2.1], [Ama95, Proposition III.4.10.3]). Let $X_{0}, X_{1}$ be Banach spaces with dense embedding $X_{1} \hookrightarrow X_{0}$ and let $M \geq 1, \vartheta \in(\pi / 2, \pi), p \in(1, \infty), \alpha \in(1 / p, 1]$ be fixed. Then the following norms are equivalent in $x \in D(A)$ with uniform constants with respect to $A \in \mathcal{P}_{1}\left(X_{1}, X_{0} ; M, \vartheta\right)$ and $T \in(0, \infty]$.

$$
|x|_{D_{A}(\alpha-1 / p, p)}, \quad|x|_{\left(X_{0}, D(A)\right)_{\alpha-1 / p, p}}, \quad\left\|R_{A} x\right\|_{L_{p}\left(0, T ; D_{A}(\alpha, p)\right)}, \quad\left\|R_{A} x\right\|_{W_{p}^{\alpha}\left(0, T ; X_{0}\right)} .
$$

In particular, the operator

$$
R_{A}: x \mapsto\left(t \mapsto e^{-t A} x\right), \quad D_{A}(\alpha-1 / p, p) \rightarrow W_{p}^{\alpha}\left(0, T ; X_{0}\right) \cap L_{p}\left(0, T ; D_{A}(\alpha, p)\right)
$$

is uniformly bounded with respect to $A \in \mathcal{P}_{1}\left(X_{1}, X_{0} ; M, \vartheta\right)$ and $T \in(0, \infty]$.
For the spaces $D_{A}(k+\alpha, p):=\left(D\left(A^{k}\right), D\left(A^{k+1}\right)\right)_{\alpha, p}$ we obtain the following result.
B.26. Corollary. Let $A: D(A) \rightarrow X$ be the negative generator of a bounded analytic semigroup in $X$ such that $A$ is invertible and let $k \in \mathbb{N}_{0}, p \in(1, \infty), \alpha \in(1 / p, 1]$. Then the operator

$$
R_{A}: u \mapsto\left(t \mapsto e^{-t A} u\right), \quad D_{A}(k+\alpha-1 / p, p) \rightarrow W_{p}^{k+\alpha}\left(\mathbb{R}_{+} ; X\right) \cap L_{p}\left(\mathbb{R}_{+} ; D_{A}(k+\alpha, p)\right)
$$

is a bounded right-inverse for trace operator $\left.\cdot\right|_{t=0}$.
Proof. This follows from Theorem B. 25 and the identity $\partial_{t} e^{-t A}=-A e^{-t A}=e^{-t A} A$.
B.27. Theorem ([Dor99]). Let A be invertible and sectorial in $X$ with spectral angle $\phi_{A}$. Then $A$ has a bounded $\mathcal{H}^{\infty}$ functional calculus in $D_{A}(\alpha, p)(\alpha \in(0,1), p \in(1, \infty))$ with $\phi_{A}^{\infty} \leq \phi_{A}$.

## B.2.3. Some concrete trace spaces.

B.28. Theorem (Poisson semigroup [Tri10, Remark 5.2.3/4], [Tri95, p. 2.5.3]). Let $n \in \mathbb{N}$, let

$$
p(x)=\frac{c_{n}}{\left(1+|x|^{2}\right)^{(n+1) / 2}}, \quad \text { with } c_{n}>0 \text { such that } \int_{\mathbb{R}^{n}} p(x) d x=1,
$$

denote the Poisson kernel and put $p_{t}(x)=t^{-n} p(x / t)$. Then the following assertions are valid.
(i) The Poisson semigroup

$$
(P(t) u)(x):=\left(p_{t} * u\right)(x)=\int_{\mathbb{R}^{n}} \frac{c_{n} t u(y) d y}{\left(|x-y|^{2}+t^{2}\right)^{(n+1) / 2}}, \quad u \in L_{p}\left(\mathbb{R}^{n}\right), t>0
$$

is a bounded analytic $C_{0}$-semigroup in $L_{p}\left(\mathbb{R}^{n}\right), p \in(1, \infty)$.
(ii) The identity $P(t) u=\mathcal{F}^{-1}\left(\xi \mapsto e^{-|\xi| t} \mathcal{F} u(\xi)\right)$ is valid for every $u \in \mathcal{S}\left(\mathbb{R}^{n}\right)$.
(iii) Let $\Lambda$ denote the generator of $P$. Then $\Lambda^{2 m}=(-1)^{m} \Delta^{m}, D\left(\Lambda^{2 m}\right)=H_{p}^{2 m}\left(\mathbb{R}^{n}\right)$ for $m \in \mathbb{N}$.
(iv) For $s \in(0, \infty), q \in[1, \infty], m \in \mathbb{N}, m>s$, the following norms are equivalent.

$$
\begin{aligned}
& \|u\|_{B_{p q}^{s}\left(\mathbb{R}^{n}\right)} \sim\|u\|_{L_{p}\left(\mathbb{R}^{n}\right)}+\left(\int_{0}^{\infty} t^{(m-s) q}\left\|\frac{\partial^{m} P(t) u}{\partial t^{m}}\right\|_{L_{p}\left(\mathbb{R}^{n}\right)}^{q} \frac{d t}{t}\right)^{1 / q}, \quad u \in B_{p q}^{s}\left(\mathbb{R}^{n}\right), \\
& \|u\|_{\dot{B}_{p q}^{s}\left(\mathbb{R}^{n}\right)} \sim\left(\int_{0}^{\infty} t^{(m-s) q}\left\|\frac{\partial^{m} P(t) u}{\partial t^{m}}\right\|_{L_{p}\left(\mathbb{R}^{n}\right)}^{q} \frac{d t}{t}\right)^{1 / q}, \quad u \in \dot{B}_{p q}^{s}\left(\mathbb{R}^{n}\right) .
\end{aligned}
$$

B.29. Theorem ([Zac03, Theorem 3.2.1]). Let $X$ be a Banach space of class $\mathcal{H} \mathcal{T}, p \in(1, \infty), \gamma \in$ $[0,1 / p)$ and $s+\gamma>n+1 / p$ with $n \in \mathbb{N}_{0}$. Let further $J=[0, T]$ or $\mathbb{R}_{+}$, and $A$ be an $\mathcal{R}$-sectorial operator in $X$ with $\mathcal{R}$-angle $\phi_{A}^{\mathcal{R}}<\pi / s$. Then for all $0 \leq k \leq n$,

$$
H_{p}^{s+\gamma}(J ; X) \cap H_{p}^{\gamma}\left(J ; D_{A^{s}}\right) \hookrightarrow B U C^{k}\left(J ; D_{A}(s+\gamma-k-1 / p, p)\right)
$$

and

$$
B_{p p}^{s+\gamma}(J ; X) \cap H_{p}^{\gamma}\left(J ; D_{A}(s, p)\right) \hookrightarrow B U C^{k}\left(J ; D_{A}(s+\gamma-k-1 / p, p)\right) .
$$

B.30. Theorem ([SSS12, Theorem 4.19]). Let $X$ be a Banach space and $p \in(1, \infty), m \in \mathbb{N}, s \in$ $(1 / p, \infty)$. For the restriction operator $\left.\varphi \mapsto \varphi\right|_{\mathbb{R}^{n-1}}, C\left(\mathbb{R}^{n} ; X\right) \rightarrow C\left(\mathbb{R}^{n-1} ; X\right)$, the following assertions are valid.
(i) The restriction operator can be extended uniquely to a continuous surjective mapping

$$
\operatorname{tr}: W_{p}^{m}\left(\mathbb{R}^{n} ; X\right) \rightarrow B_{p p}^{m-1 / p}\left(\mathbb{R}^{n-1} ; X\right)
$$

and tr has a continuous right-inverse ext: $B_{p p}^{m-1 / p}\left(\mathbb{R}^{n-1} ; X\right) \rightarrow W_{p}^{m}\left(\mathbb{R}^{n} ; X\right)$.
(ii) The restriction operator can be extended uniquely to a continuous surjective mapping

$$
\operatorname{tr}: H_{p}^{s}\left(\mathbb{R}^{n} ; X\right) \rightarrow B_{p p}^{s-1 / p}\left(\mathbb{R}^{n-1} ; X\right)
$$

and tr has a continuous right-inverse ext: $\left.B_{p p}^{s-1 / p}\left(\mathbb{R}^{n-1} ; X\right) \rightarrow H_{p}^{s}\left(\mathbb{R}^{n} ; X\right)\right)$.
B.31. Theorem ([Jaw77, Theorem 2.1], [Jaw78, Theorem 5.1]). Let $p \in[1, \infty), q \in[1, \infty], s \in$ $(1 / p, \infty), n \in \mathbb{N}, n \geq 2$. For the restriction operator $\left.\varphi \mapsto \varphi\right|_{\mathbb{R}^{n-1}}, \mathcal{S}_{0}\left(\mathbb{R}^{n}\right) \rightarrow \mathcal{S}\left(\mathbb{R}^{n-1}\right)$, the following assertions are valid.
(i) The restriction operator can be extended to a continuous surjective mapping

$$
\operatorname{tr}: \dot{B}_{p q}^{s}\left(\mathbb{R}^{n}\right) \rightarrow \dot{B}_{p q}^{s-1 / p}\left(\mathbb{R}^{n-1}\right)
$$

and there exists a linear operator ext: $\mathcal{S}_{0}^{\prime}\left(\mathbb{R}^{n-1}\right) \rightarrow \mathcal{S}_{0}^{\prime}(\mathbb{R})$ (independent of $\left.p, q, s\right)$ such that the realization ext: $\dot{B}_{p q}^{s-1 / p}\left(\mathbb{R}^{n-1}\right) \rightarrow \dot{B}_{p q}^{s}\left(\mathbb{R}^{n}\right)$ is a continuous right-inverse of tr .
(ii) The restriction operator can be extended to a continuous surjective mapping

$$
\operatorname{tr}: \dot{F}_{p q}^{s}\left(\mathbb{R}^{n}\right) \rightarrow \dot{B}_{p p}^{s-1 / p}\left(\mathbb{R}^{n-1}\right)
$$

and there exists a linear operator ext: $\mathcal{S}_{0}^{\prime}\left(\mathbb{R}^{n-1}\right) \rightarrow \mathcal{S}_{0}^{\prime}(\mathbb{R})$ (independent of $\left.p, q, s\right)$ such that the realization ext: $\dot{B}_{p p}^{s-1 / p}\left(\mathbb{R}^{n-1}\right) \rightarrow \dot{F}_{p q}^{s}\left(\mathbb{R}^{n}\right)$ is a continuous right-inverse of tr .
B.32. Theorem (Spatial trace theorem [cf. MS12, Theorem 4.5]). Let E be a Banach space of class $\mathcal{H T}, J=(0, T)$ be finite or infinite, $p \in(1, \infty), m \in \mathbb{N}, s \in(0,1]$, such that $2 m s \in \mathbb{N}$. Assume that $\Omega \subset \mathbb{R}^{n}$ is a domain with compact smooth boundary, or $\Omega \in\left\{\mathbb{R}^{n}, \mathbb{R}_{+}^{n}\right\}$. Then the trace

$$
\begin{aligned}
\left.u \mapsto u\right|_{\partial \Omega}: & H_{p}^{s}\left(J ; L_{p}(\Omega ; E)\right) \cap L_{p}\left(J ; H_{p}^{2 m s}(\Omega ; E)\right) \\
& \rightarrow W_{p}^{s-1 / 2 m p}\left(J ; L_{p}(\partial \Omega ; E)\right) \cap L_{p}\left(J ; W_{p}^{2 m s-1 / p}(\partial \Omega ; E)\right)
\end{aligned}
$$

is continuous and surjective and has a continuous right-inverse. The restriction of the trace to

$$
{ }_{0} H_{p}^{s}\left(J ; L_{p}(\Omega ; E)\right) \cap L_{p}\left(J ; H_{p}^{2 m s}(\Omega ; E)\right),
$$

is uniformly bounded with respect to the length of $J$.

## B.2.4. Functional calculus for sectorial operators.

B.33. Remark (Functional calculus). Let $A \in \mathcal{S}(X ; \vartheta)$.
(i) The (primary) $\mathcal{H}^{\infty}$-functional calculus $\Phi_{A}: \mathcal{H}_{0}^{\infty}\left(\Sigma_{\pi-\vartheta}\right) \rightarrow \mathcal{B}(X)$ is defined by

$$
\left(\Phi_{A}(f)\right)(x):=f(A) x:=\frac{1}{2 \pi i} \int_{\Gamma} f(\lambda)(\lambda+A)^{-1} d \lambda, \quad \text { for } x \in X,
$$

where the curve $\Gamma=e^{-i \psi}[0, \infty) \cup e^{i \psi}[0, \infty) \subset \rho(-A)$ surrounds $\sigma(-A)$ counterclockwise.
(ii) Extended functional calculus....
B.34. Theorem (Spectral mapping theorem, [Haa06, Theorem 2.7.8]). Let $A \in \mathcal{S}(X), \phi \in\left(\phi_{A}, \pi\right)$ and let $f \in \mathcal{H}_{P}\left(\Sigma_{\phi}\right)$ have polynomial limits at $\{0, \infty\}$. Then

$$
f(\tilde{\sigma}(A))=\tilde{\sigma}(f(A)),
$$

where $\tilde{\sigma}(A):=\sigma(A)$ if $A$ is bounded and $\tilde{\sigma}(A):=\sigma(A) \cup\{\infty\}$ otherwise.
B.2.5. Fractional powers. Set $\rho(\lambda)=\lambda(1+\lambda)^{-2}$. This function belongs to $\mathcal{H}_{0}^{\infty}\left(\Sigma_{\sigma}\right)$ for each $\sigma \in(0, \pi)$ with norm $|\rho|_{L_{\infty}\left(\Sigma_{\sigma}\right)}=(2(1+\cos \sigma))^{-1}$. For given $\sigma \in(0, \pi)$ the function $\lambda \mapsto \lambda^{\alpha}$ is holomorphic and bijective from $\Sigma_{\sigma}$ to $\Sigma_{|\alpha| \sigma}$ for all $\alpha \in \mathbb{R}$ with $|\alpha|<\pi / \sigma$.

Let $A \in \mathcal{S}\left(X, \phi_{A}\right)$ and let

$$
\begin{aligned}
& \Phi_{A}=(f \mapsto f(A)): \mathcal{H}_{0}^{\infty}\left(\Sigma_{\sigma}\right) \rightarrow \mathcal{B}(X), \\
& \Phi_{A}=(f \mapsto f(A)): \mathcal{H}_{p}\left(\Sigma_{\sigma}\right) \rightarrow \mathcal{B}(X)
\end{aligned}
$$

denote the primary and the extended $\mathcal{H}^{\infty}$-calculi of $A$, respectively. Then the fractional powers $A^{\alpha}, \alpha \in \mathbb{C}$, are defined by

$$
A^{\alpha}=\bar{\Phi}_{A}\left(\lambda \mapsto \lambda^{\alpha}\right)=\Phi_{A}(\rho)^{-k} \Phi_{A}\left(\lambda \mapsto \rho(\lambda)^{k} \lambda^{\alpha}\right),
$$

for $k \in \mathbb{N}, k>\alpha$. Their natural domains are given by

$$
D\left(A^{\alpha}\right)=\left\{x \in X: \Phi_{A}\left(\lambda \mapsto \rho(\lambda)^{k} \lambda^{\alpha}\right) x \in D\left(A^{k}\right) \cap R\left(A^{k}\right)\right\} .
$$

B.35. Theorem ([DHP03, Theorem 2.3]). Let $A$ be sectorial in $X$ with spectral angle $\phi_{A}$ and let $\alpha \in\left(-\pi / \phi_{A}, \pi, \phi / A\right)$. Then $A^{\alpha}$ is also sectorial in $X$ with $\phi_{A^{\alpha}} \leq|\alpha| \phi_{A}$.

A sufficient condition for $A^{\alpha} \in \mathcal{H}^{\infty}(X)$ can be derived with the following composition rule.
B.36. Theorem ([cf. Haa06, Theorem 2.4.2]). Let $A \in \mathcal{S}(\omega)(\omega \in[0, \pi))$ be injective and, for some $\phi \in(\omega, \pi)$ and $\omega^{\prime} \in[0, \pi)$, let $g \in \mathcal{H}_{P}\left(\Sigma_{\phi}\right)$ be a function such that $g(A) \in \mathcal{S}\left(\omega^{\prime}\right)$ and such that for every $\phi^{\prime} \in\left(\omega^{\prime}, \pi\right)$ there exists $\phi \in(\omega, \pi)$ such that $g \in \mathcal{H}_{P}\left(\Sigma_{\phi}\right)$ and $g\left(\Sigma_{\phi}\right) \subset \overline{\Sigma_{\phi^{\prime}}}$. Then

$$
(f \circ g)(A)=f(g(A)) \quad \text { for all } \phi^{\prime} \in\left(\omega^{\prime}, \pi\right), f \in \mathcal{H}_{P}\left(\Sigma_{\phi^{\prime}}\right) .
$$

B.37. Corollary. The following implications are valid.

$$
\begin{aligned}
A \in \mathcal{H}^{\infty}(X),-\frac{\pi}{\left|\phi_{A}^{\infty}\right|}<\alpha<\frac{\pi}{\left|\phi_{A}^{\infty}\right|} & \Rightarrow A^{\alpha} \in \mathcal{H}^{\infty}(X), \phi_{A^{\alpha}}^{\infty} \leq|\alpha| \phi_{A}^{\infty}, \\
A \in \mathcal{R} \mathcal{H}^{\infty}(X),-\frac{\pi}{\left|\phi_{A}^{\mathcal{R} \infty}\right|}<\alpha<\frac{\pi}{\left|\phi_{A}^{\mathcal{R}}\right|} & \Rightarrow A^{\alpha} \in \mathcal{R} \mathcal{H}^{\infty}(X), \phi_{A^{\alpha}}^{\mathcal{R} \infty} \leq|\alpha| \phi_{A}^{\mathcal{R} \infty} .
\end{aligned}
$$

B.38. Corollary. For $A \in \mathcal{H}^{\infty}(X ; M, \vartheta)$, the following assertions are valid.
(i) $A \in \mathcal{B I} \mathcal{P}(M, \pi-\vartheta)$.
(ii) For $s \in[0, \pi /(\pi-\vartheta))$, we have $A^{s} \in \mathcal{H}^{\infty}(M, \vartheta+(1-s)(\pi-\vartheta))$.
(iii) For $\varepsilon>0$, we have $\varepsilon+A \in \mathcal{P}\left(M_{1}, \vartheta_{1}\right)$ for every $\vartheta_{1} \in(0, \vartheta)$, where $M_{1}=2 M c\left(1+\varepsilon^{-2}\right)^{1 / 2}$ and $c=1 / \min \left\{1,1+\cos \left(\pi-\left(\vartheta-\vartheta_{1}\right)\right)\right\}$.
Proof. (i) The assertion follows from $\left|z^{i t}\right|=\left|e^{i t(\ln |z|+i \arg z)}\right|=e^{-t \arg z} \leq e^{|t| \arg z}$.
(ii) The function $g_{s}: z \mapsto z^{s}$ maps $\Sigma_{\pi-\vartheta}$ onto $\Sigma_{s(\pi-\vartheta)}$. Hence for $f \in \mathcal{H}^{\infty}\left(\Sigma_{s(\pi-\vartheta)}\right)$ we have $f \circ g_{s} \in \mathcal{H}^{\infty}\left(\Sigma_{\pi-\vartheta}\right)$ with the same $L_{\infty}$-norm. Moreover, the composition rule implies $f\left(A^{s}\right)=\left(f \circ g_{s}\right)(A)$ and this yields the assertion.
(iii) Using $(\lambda+\varepsilon+A)^{-1}=(\lambda+\varepsilon+\cdot)^{-1}(A)$ and (B.14), we obtain

$$
\begin{aligned}
(1+|\lambda|)\left\|(\lambda+\varepsilon+A)^{-1}\right\| & \leq(1+|\lambda|) M\left\|(\lambda+\varepsilon+\cdot)^{-1}\right\|_{L_{\infty}\left(\Sigma_{\pi-\vartheta}\right)} \\
& \leq \frac{\sqrt{2} M}{\min \left\{1,1+\cos \left(\pi-\left(\vartheta-\vartheta_{1}\right)\right)\right\}^{1 / 2}} \frac{1+|\lambda|}{|\lambda+\varepsilon|} \\
& \leq \frac{2 M}{\min \left\{1,1+\cos \left(\pi-\left(\vartheta-\vartheta_{1}\right)\right)\right\}} \sqrt{1+\frac{1}{\varepsilon^{2}}} .
\end{aligned}
$$

B.39. Theorem (cf. [Ama95, (I.2.9.9)] and [Tri95, Theorem 1.15.3]). If $A \in \mathcal{B I P} \mathcal{P}(X)$, then

$$
D\left(A^{\alpha}\right) \cong[X, D(A)]_{\alpha}, \quad \text { for } \alpha \in[0,1],
$$

where $D\left(A^{\alpha}\right)$ is equipped with the norm $x \mapsto\|x\|_{X}+\left\|A^{\alpha} x\right\|_{X}$. Moreover, given $\theta \geq 0, \vartheta \in(0, \pi)$, $M \geq 1, \alpha \in[0,1]$, there exists $C \geq 1$ such that for all $A \in \mathcal{B I P} \mathcal{P}(X ; M, \theta) \cap \mathcal{P}(X ; M, \vartheta)$, we have

$$
C^{-1}\|x\|_{D\left(A^{\alpha}\right)} \leq\|x\|_{[X, D(A)]_{\alpha}} \leq C\|x\|_{D\left(A^{\alpha}\right)} \quad \text { for } x \in D\left(A^{\alpha}\right) \text {. }
$$

B.40. Corollary. Let $M \geq 1, \vartheta \in(0, \pi), s \in(0,1), \varepsilon>0$, let $X_{0}$, $X_{1}$ be Banach spaces with dense embedding $X_{1} \hookrightarrow X_{0}$ and let $0<\phi<\vartheta_{s}:=s \vartheta+(1-s) \pi$. Then there exists $N \geq 1$ such that if

$$
A \in \mathcal{P}_{1}\left(X_{1}, X_{0} ; M, \vartheta\right) \cap \mathcal{H}^{\infty}\left(X_{0} ; M, \vartheta\right),
$$

then

$$
\varepsilon+A^{s} \in \mathcal{P}_{1}\left(\left[X_{0}, X_{1}\right]_{s}, X_{0} ; N, \phi\right) \cap \mathcal{H}^{\infty}\left(X_{0} ; N, \phi\right) .
$$

Proof. First, the norms $\|x\|_{D(A)}=\|x\|_{X_{0}}+\|A x\|_{X_{0}}$ and $\|x\|_{X_{1}}$ are equivalent, uniformly with respect to $A \in \mathcal{P}_{1}\left(X_{1}, X_{0} ; M, \vartheta\right)$, since both $\|A\|_{\mathcal{B}\left(X_{1} ; X_{0}\right)}$ and $\left\|A^{-1}\right\|_{\mathcal{B}\left(X_{0} ; X_{1}\right)}$ are bounded by $M$. Hence, by Theorem B. 39 and Corollary B.38, the norms of $\left[X_{0}, X_{1}\right]_{s},\left[X_{0}, D(A)\right]_{s}$ and $D\left(A^{s}\right)$ are equivalent, uniformly with respect to $A \in \mathcal{P}_{1}\left(X_{1}, X_{0} ; M, \vartheta\right) \cap \mathcal{H}^{\infty}\left(X_{0} ; M, \vartheta\right)$, and this implies

$$
\left\|A^{s}\right\|_{\mathcal{B}\left(\left[X_{0}, X_{1}\right]_{s} ; X_{0}\right)} \sim\left\|A^{s}\right\|_{\mathcal{B}\left(D\left(A^{s}\right) ; X_{0}\right)} \leq 1
$$

Therefore $\left\|\varepsilon+A^{s}\right\|_{\mathcal{B}\left(\left[X_{0}, X_{1}\right]_{s} ; X_{0}\right)}$ is uniformly bounded. Again by Corollary B. 38 and basic resolvent identities like $A^{s}\left(\lambda+A^{s}\right)^{-1}=I-\lambda\left(\lambda+A^{s}\right)^{-1}$ we obtain $\varepsilon+A^{s} \in \mathcal{P}_{1}\left(D\left(A^{s}\right), X_{0} ; M_{1}, \phi\right)$ for all $A \in \mathcal{P}_{1}\left(X_{1}, X_{0} ; M, \vartheta\right) \cap \mathcal{H}^{\infty}\left(X_{0} ; M, \vartheta\right)$. By the uniform equivalence of the norms of $D\left(A^{s}\right)$ and $\left[X_{0}, X_{1}\right]_{s}$, there exists $N \geq 1$ with $\varepsilon+A^{s} \in \mathcal{P}_{1}\left(\left[X_{0}, X_{1}\right]_{s}, X_{0} ; N, \phi\right)$, uniformly in $A$.
B.2.6. Sums and products of sectorial operators. Let $(A, D(A))$ and $(B, D(B))$ be densely defined closed linear operators in a Banach space $X$. We collect several results for the sum $A+B$ and the product $A B$ under the condition that $A$ and $B$ are resolvent commuting. We define the sum $A+B$, the product $A B$ and the commutator $[A, B]$ by

$$
\begin{aligned}
(A+B) x & =A x+B x, \\
(A B) x & =A(B x), \\
{[A, B] x } & =A B x-B A x,
\end{aligned}
$$

$$
\begin{aligned}
D(A+B) & :=D(A) \cap D(B), \\
D(A B) & :=\{x \in D(B): B x \in D(A)\}, \\
D([A, B]) & :=D(A B) \cap D(B A) .
\end{aligned}
$$

B.41. Remark (Commuting operators). (i) A bounded operator $T \in \mathcal{B}(X)$ is said to commute with a closed operator $A: D(A) \subset X \rightarrow X$, if $T A=A T$ on $D(A)$. If $\rho(A) \neq 0$, then this is equivalent to $[T, R(\lambda, A)]=0$ for some (and hence all) $\lambda \in \rho(A)$ [Ama95; Haa06].
(ii) Suppose that $\rho(A) \neq \emptyset, \rho(B) \neq \emptyset$. We say that $A$ and $B$ are resolvent commuting, if $\left[(\lambda-A)^{-1},(\mu-B)^{-1}\right]=0$ for some (and hence all) $\lambda \in \rho(A), \mu \in \rho(B)$. It can be shown that if $A$ and $B$ are resolvent-commuting, then $A B x=B A x$ for all $x \in D(A B) \cap D(B A)$.
(iii) If $A, B \in \mathcal{S}(X)$ are resolvent commuting, then also $f(A), g(B)$ are resolvent commuting for all $f \in \mathcal{H}_{A}\left(\Sigma_{\psi}\right), g \in \mathcal{H}_{B}\left(\Sigma_{\rho}\right), \psi \in\left(\phi_{A}, \pi\right), \rho \in\left(\phi_{B}, \pi\right)$.

Next we state a version of the mixed derivative theorem of Sobolevskiĭ [Sob75, Theorem 6]. A linear operator $A: D(A) \subset X \rightarrow X$ is called positive if it has the properties of a sectorial operator (Definition B.17) except that $R(A)$ does not need to be dense in $X$ and the resolvent estimate is valid in a set $\left\{\lambda=r e^{i \varphi}:|\varphi| \in[\theta, \pi], r \in\left[r_{0}, \infty\right)\right\}$ with some $r_{0} \geq 0$, which may be smaller than the sector $-\Sigma_{\theta}$. In this case $k+A$ for $k \geq r_{0}$ is sectorial and invertible and $D(A)$ is a Banach space for the norm $\|(k+A) \cdot\|_{X}$. We say that two linear operators $A$ and $B$ in $X$ form a coercive pair if for some numbers $M \geq 0$ and $k \in \mathbb{N}_{0}$ we have the estimate

$$
\|(k+A) x\|+\|(k+B) x\| \leq M\|(k+A) x+(k+B) x\| \quad \text { for all } x \in D(A) \cap D(B) .
$$

B.42. Theorem (Mixed derivatives [cf. Sob75, Theorem 6]). Let X be a Banach space and let A and $B$ form a coercive pair of positive operators with commuting resolvents such that their spectral angles satisfy $\phi_{A}+\phi_{B}<\pi$. Then $A+B$ is positive, and for sufficiently large $k$ and arbitrary $0 \leq \alpha \leq 1$ we have the continuous embeddings

$$
D(A+B) \hookrightarrow D\left((k+A)^{\alpha}(k+B)^{1-\alpha}\right) \cap D\left((k+B)^{1-\alpha}(k+A)^{\alpha}\right) .
$$

B.43. Corollary ([MS12, Proposition 1.1]). Let $X$ be a Banach space of class $\mathcal{H T}$ and suppose that the operators $A, B \in \mathcal{B} \mathcal{I} \mathcal{P}(X)$ are resolvent commuting and satisfy $\theta_{A}+\theta_{B}<\pi$. If $A$ or $B$ is invertible, then $A+B$ is invertible, $A+B \in \mathcal{B I P}(X)$ with $\theta_{A+B} \leq \theta_{A}+\theta_{B}$ and $A^{\alpha} B^{1-\alpha}(A+B)^{-1}$ is bounded in $X$ for every $\alpha \in[0,1]$.
B.44. Proposition (Mixed derivative embeddings [cf. MS12, Proposition 3.2]). Let $J=(0, T)$ be finite or infinite, $p \in(1, \infty)$, let $X$ be a Banach space of class $\mathcal{H} \mathcal{T}$ and let $\Omega \subset \mathbb{R}^{n}$ be a domain with compact smooth boundary, or $\Omega \in\left\{\mathbb{R}^{n}, \mathbb{R}_{+}^{n}\right\}$. Let further $t, s \geq 0, \alpha \in(0,2), \beta>0, \rho \in[0,1]$, and set $H_{p}^{t}\left(H_{p}^{s}\right):=H_{p}^{t}\left(J ; H_{p}^{s}(\Omega ; X)\right)$, and analogously for the other anisotropic spaces. Then

$$
H_{p}^{t+\alpha}\left(H_{p}^{s}\right) \cap H_{p}^{t}\left(H_{p}^{s+\beta}\right) \hookrightarrow H_{p}^{t+\rho \alpha}\left(H_{p}^{s+(1-\rho) \beta}\right),
$$

and moreover each of the spaces

$$
H_{p}^{t+\alpha}\left(W_{p}^{s}\right) \cap H_{p}^{t}\left(W_{p}^{s+\beta}\right), W_{p}^{t+\alpha}\left(H_{p}^{s}\right) \cap W_{p}^{t}\left(H_{p}^{s+\beta}\right), W_{p}^{t+\alpha}\left(H_{p}^{s}\right) \cap H_{p}^{t}\left(W_{p}^{s+\beta}\right)
$$

is continuously embedded into

$$
W_{p}^{t+\rho \alpha}\left(H_{p}^{s+(1-\rho) \beta}\right) \cap H_{p}^{t+\rho \alpha}\left(W_{p}^{s+(1-\rho) \beta}\right),
$$

provided that all the occurring $W_{p}$-spaces have a non-integer order of differentiability. Finally, assuming all orders of differentiability to be non-integer, we have

$$
W_{p}^{t+\alpha}\left(W_{p}^{s}\right) \cap W_{p}^{t}\left(W_{p}^{s+\beta}\right) \hookrightarrow W_{p}^{t+\rho \alpha}\left(W_{p}^{s+(1-\rho) \beta}\right) .
$$

These embeddings remain true if $\Omega$ is replaced by its boundary. They are also valid if all $H_{p}$-, $W_{p}$-spaces with respect to time are replaced by ${ }_{0} H_{p^{-}},{ }_{0} W_{p}$-spaces, respectively. Restricting in the latter case to $t+\alpha \leq 2$, the embedding constants have a uniform bound with respect to to the length of $J$.
B.45. Remark. The following mixed derivative embeddings are valid.

$$
\begin{equation*}
H_{p}^{s+\alpha}\left(\mathbb{R}_{+} ; \dot{H}_{p}^{r}\left(\mathbb{R}^{n}\right)\right) \cap H_{p}^{s}\left(\mathbb{R}_{+} ; \dot{H}_{p}^{r+\beta}\left(\mathbb{R}^{n}\right)\right) \hookrightarrow H_{p}^{s+\theta \alpha}\left(\mathbb{R}_{+} ; \dot{H}_{p}^{r+(1-\theta) \beta}\left(\mathbb{R}^{n}\right)\right) \tag{B.9}
\end{equation*}
$$

if $s, \beta \in[0, \infty), \alpha \in[0,2], r \in \mathbb{R}, \theta \in[0,1]$ and

$$
\begin{equation*}
W_{p}^{s+\alpha}\left(\mathbb{R}_{+} ; \dot{H}_{p}^{r}\left(\mathbb{R}^{n}\right)\right) \cap W_{p}^{s}\left(\mathbb{R}_{+} ; \dot{H}_{p}^{r+\beta}\left(\mathbb{R}^{n}\right)\right) \hookrightarrow H_{p}^{s+\theta \alpha}\left(\mathbb{R}_{+} ; \dot{W}_{p}^{r+(1-\theta) \beta}\left(\mathbb{R}^{n}\right)\right) \tag{B.10}
\end{equation*}
$$

if $s, \beta \in(0, \infty), \alpha \in(0,2), r \in \mathbb{R}, \theta \in(0,1)$ and $s, s+\alpha, r+(1-\theta) \beta \notin \mathbb{Z}$ and

$$
\begin{equation*}
H_{p}^{s+\alpha}\left(\mathbb{R}_{+} ; \dot{W}_{p}^{r}\left(\mathbb{R}^{n}\right)\right) \cap H_{p}^{s}\left(\mathbb{R}_{+} ; \dot{W}_{p}^{r+\beta}\left(\mathbb{R}^{n}\right)\right) \hookrightarrow W_{p}^{s+\theta \alpha}\left(\mathbb{R}_{+} ; \dot{H}_{p}^{r+(1-\theta) \beta}\left(\mathbb{R}^{n}\right)\right) \tag{B.11}
\end{equation*}
$$

if $s \in[0, \infty), \alpha \in(0,2), \beta \in(0, \infty), r \in \mathbb{R}, \theta \in(0,1)$ and $r, r+\beta, s+\theta \alpha \notin \mathbb{Z}$.
Proof. The strategy is the same as for the non-homogeneous spaces ([MS12, Proposition 3.2]). We abbreviate $\mathcal{F}(\mathcal{K}):=\mathcal{F}\left(\mathbb{R}_{+} ; \mathcal{K}\left(\mathbb{R}^{n}\right)\right)$ for $\mathcal{F} \in\left\{\dot{H}_{p}^{s}, \dot{W}_{p}^{s}\right\}, \mathcal{K} \in\left\{\dot{H}_{p}^{r}, \dot{W}_{p}^{r}\right\}$. By applying the mixed derivative theorem B. 42 to the $\mathcal{B I P}$-operators $\left(1-\partial_{t}\right)^{\alpha}: H_{p}^{s+\alpha}\left(\dot{H}_{p}^{r}\right) \rightarrow H_{p}^{s}\left(\dot{H}_{p}^{r}\right)$ and $(-\Delta)^{\beta / 2}: H_{p}^{s}\left(\dot{H}_{p}^{r} \cap \dot{H}_{p}^{r+\beta}\right) \rightarrow H_{p}^{s}\left(\dot{H}_{p}^{r}\right)$ (see theorems B.68, B. 15 and Section B.2.5), we obtain

$$
\left\|\left(1-\partial_{t}\right)^{\theta \alpha}(-\Delta)^{(1-\theta) \beta / 2} w\right\|_{H_{p}^{s}\left(\dot{H}_{p}^{r}\right)} \lesssim\|w\|_{H_{p}^{s+\alpha}\left(\dot{H}_{p}^{r}\right) \cap H_{p}^{s}\left(\dot{H}_{p}^{r+\beta}\right)}, \quad \text { for all } \theta \in[0,1] .
$$

By using the invertibility of the operators

$$
\begin{gathered}
\left(1-\partial_{t}\right)^{\theta \alpha}: H_{p}^{s+\theta \alpha}\left(\dot{H}_{p}^{r}\right) \rightarrow H_{p}^{s}\left(\dot{H}_{p}^{r}\right), \\
(-\Delta)^{(1-\theta) \beta / 2}: H_{p}^{s+\theta \alpha}\left(\dot{H}_{p}^{r+(1-\theta) \beta}\right) \rightarrow H_{p}^{s+\theta \alpha}\left(\dot{H}_{p}^{r}\right),
\end{gathered}
$$

we further have

$$
\left\|\left(1-\partial_{t}\right)^{\theta \alpha}(-\Delta)^{(1-\theta) \beta / 2} w\right\|_{H_{p}^{s}\left(\dot{H}_{p}^{r}\right)} \sim\left\|(-\Delta)^{(1-\theta) \beta / 2} w\right\|_{H_{p}^{s+\theta \alpha}\left(\dot{H}_{p}^{r}\right)} \sim\|w\|_{H_{p}^{s+\theta \alpha}\left(\dot{H}_{p}^{r+(1-\theta) \beta}\right)} .
$$

Hence (B.9) is shown.

Next, for proving the embedding (B.10), we choose some sufficiently small $\varepsilon>0$ and put $s_{ \pm}:=s \pm \alpha \varepsilon, \theta_{ \pm}:=\theta \mp \varepsilon$. Then $s+\alpha \pm \alpha \varepsilon=s_{ \pm}+\alpha, s \pm \alpha \varepsilon=s_{ \pm}, s+\theta \alpha=s_{ \pm}+\theta_{ \pm} \alpha$ and $r+(1-\theta) \beta \pm \varepsilon \beta=r+\left(1-\theta_{ \pm}\right) \beta$. We now apply (B.9) with $s_{ \pm}, \theta_{ \pm}$instead of $s, \theta$ and obtain

$$
H_{p}^{s+\alpha \pm \alpha \varepsilon}\left(\dot{H}_{p}^{r}\right) \cap H_{p}^{s \pm \alpha \varepsilon}\left(\dot{H}_{p}^{r+\beta}\right) \hookrightarrow H_{p}^{s+\theta \alpha}\left(\dot{H}_{p}^{r+(1-\theta) \beta \mp \varepsilon \beta}\right) .
$$

Applying the real interpolation functor $(\cdot, \cdot)_{1 / 2, p}$ and the identity $\left(\dot{H}_{p}^{t_{0}}, \dot{H}_{p}^{t_{1}}\right)_{\theta, p}=\dot{W}_{p}^{t}$ for $t=$ $(1-\theta) t_{0}+\theta t_{1} \in \mathbb{R} \backslash \mathbb{Z}$, we obtain (B.10). The embedding (B.11) follows similarly by choosing $r_{ \pm}:=r \mp \beta \varepsilon, \theta_{ \pm}:=\theta \mp \varepsilon$ so that $s+\theta \alpha=s+\theta_{ \pm} \alpha \pm \alpha \varepsilon$ and $r+(1-\theta) \beta=r_{ \pm}+\left(1-\theta_{ \pm}\right) \beta$.
B.46. Theorem ([PS90, Theorem 5]). Let $X$ be a Banach space of class $\mathcal{H T}$, let $A, B \in \mathcal{B I P}(X)$ with $\theta_{A}+\theta_{B}<\pi$ be resolvent commuting, and let $\theta=\max \left(\theta_{A}, \theta_{B}\right), \theta_{A} \neq \theta_{B}$. Then $A+B \in \mathcal{B I P}(X)$ with $\theta_{A+B} \leq \theta$.
B.47. Theorem (Kalton-Weis [KW01, Theorem 6.3]). Suppose $A \in \mathcal{H}^{\infty}(X)$ and $B \in \mathcal{R S}(X)$ are resolvent commuting and $\phi_{A}^{\infty}+\phi_{B}^{\mathcal{R}}<\pi$. Then $A+B$ with domain $D(A) \cap D(B)$ is a closed operator and there is a constant $C$ such that

$$
|A x|+|B x| \leq C|A x+B x|, \quad \text { for all } x \in D(A) \cap D(B) .
$$

Thus, $A+B$ is invertible if either $A$ or $B$ is invertible. Furthermore, if $X$ has property $(\alpha)$, then $A+B \in \mathcal{R S}(X)$ with $\phi_{A+B}^{\mathcal{R}} \leq \max \left(\phi_{A}^{\infty}, \phi_{B}^{\mathcal{R}}\right)$.
B.48. Corollary ([PS07]). Suppose $A \in \mathcal{H}^{\infty}(X)$ and $B \in \mathcal{R} \mathcal{H}^{\infty}(X)$ are commuting such that $\phi_{A}^{\infty}+$ $\phi_{B}^{\text {R }}<\pi$. Then $A+B \in \mathcal{H}^{\infty}(X)$.

Next, we consider the product $A B$ of two sectorial operators $A, B$.
B.49. Theorem ([PS90, Corollary 3]). Suppose $X$ is of class $\mathcal{H T}$, let $A, B \in \mathcal{B I P}(X)$ with $0 \leq$ $\theta_{A}+\theta_{B}<\pi$ be resolvent commuting. Then $A B$ is closable and $\overline{A B} \in \mathcal{B I P}(X)$ with $\theta_{A B} \leq \theta_{A}+\theta_{B}$. If in addition $A$ is invertible, then $A B$ is closed.
B.50. Corollary ([HDH05, Corollary 2.2]). Let $X$ be a Banach space and assume that $A \in \mathcal{H}^{\infty}(X)$ and $B \in \mathcal{R S}(X)$ are resolvent commuting such that $0 \in \rho(A)$ and $\phi_{A}^{\infty}+\phi_{B}^{\mathcal{R}}<\pi$. Then $A B \in \mathcal{S}(X)$ and $\phi_{A B} \leq \phi_{A}^{\infty}+\phi_{B}^{\mathcal{R}}$. If in addition $B \in \mathcal{R} \mathcal{H}^{\infty}(X)$ with $\phi_{A}^{\infty}+\phi_{B}^{\mathcal{R}}<\pi$, then $A B \in \mathcal{H}^{\infty}(X)$ with $\phi_{A B}^{\infty} \leq \phi_{A}^{\infty}+\phi_{B}^{\mathcal{R} \infty}$.
B.2.7. Estimates for Fourier-Laplace symbols. In order to obtain the mapping properties of linear pseudo-differential operators, we will establish estimates of their Fourier-Laplacesymbols with respect to the temporal and spatial covariables $\lambda, z$ for $\partial_{t}$ and $\sqrt{-\Delta}$, respectively. B.51. Remark (Laplace transform). Let $\mathbb{R}_{+}=[0, \infty)$ and $\mathbb{C}_{>0}=\{z \in \mathbb{C}: \operatorname{Re} z>0\}$. Let $f \in L_{1, \text { loc }}\left(\mathbb{R}_{+} ; X\right)$ be of exponential growth; that is, the integral $\int_{0}^{\infty} e^{-\omega t}|f(t)|_{X} d t$ is finite for some $\omega \in \mathbb{R}$. Then we define the Laplace transform $\mathcal{L} f: \omega+\mathbb{C}_{>0} \rightarrow X$ of $f$ by

$$
\widehat{f}(\lambda):=(\mathcal{L} f)(\lambda):=\int_{0}^{\infty} e^{-\lambda t} f(t) d t, \quad(\operatorname{Re} \lambda \geq \omega) .
$$

Then $\mathcal{L} f \in B U C\left(\omega+\overline{\mathbb{C}_{>0}} ; X\right) \cap \mathcal{H}\left(\omega+\mathbb{C}_{>0} ; X\right)$. For given $\theta_{0} \in(0, \pi]$, let $\mathcal{H}_{b c}\left(\Sigma_{\theta_{0}}\right)$ denote the vector space of all functions on $\Sigma_{\theta_{0}}$ that are holomorphic in $\Sigma_{\theta_{0}}$ and bounded and continuous on each closed sector $\overline{\Sigma_{\theta}}, \theta \in\left[0, \theta_{0}\right)$. Then Cauchy's theorem leads to

$$
\left\{\mathcal{L} f: f \in \mathcal{H}_{b c}\left(\Sigma_{\theta_{0}}\right)\right\}=\left\{g \in \mathcal{H}\left(\Sigma_{\pi / 2+\theta_{0}}\right): \lambda \mapsto \lambda g(\lambda) \in \mathcal{H}_{b c}\left(\Sigma_{\pi / 2+\theta_{0}}\right)\right\} .
$$

See [Prü93, Theorem 0.1]. Uniqueness of the Laplace transform: The complex inversion formula

$$
f(t)=\lim _{N \rightarrow \infty} \frac{1}{2 \pi} \int_{-N}^{N}\left(1-\frac{|\rho|}{N}\right) e^{(\sigma+i \rho) t} \mathcal{L} f(\sigma+i \rho) d \rho
$$

applies for almost all $t \in \mathbb{R}_{+}$and each $\sigma>\omega$. The real inversion formula

$$
f(t)=\lim _{\sigma \rightarrow \infty} \sum_{n=0}^{\infty} \frac{(-1)^{n}\left(\sigma^{2} t\right)^{n+1}}{n!(n+1)!} \mathcal{L} f^{(n)}(\sigma)
$$

applies for almost all $t>0$. The Laplace transform has the following relation to the Fourier transform $\mathcal{F}$. For each $f \in L_{1}(\mathbb{R} ; X)$ such that $f(t)=0,(t<0)$, it is $\mathcal{F} f(\rho)=\mathcal{L} f(i \rho),(\rho \in \mathbb{R})$.
B.52. Definition $(\alpha \lesssim \beta, \alpha \sim \beta)$. Let $D$ be a set, $(X,|\cdot|)$ be a normed vector space and consider two functions $\alpha, \beta: D \rightarrow X$. We say that $\alpha$ is dominated by $\beta$ (in $D$ ) and write

$$
\alpha \lesssim \beta \text { (in } D \text { ) if and only if } \quad \exists C>0 \forall x \in D:|\alpha(x)| \leq C|\beta(x)| .
$$

The functions $\alpha, \beta$ are said to be equivalent, and we write $\alpha \sim \beta$, if $\alpha \lesssim \beta$ and $\beta \lesssim \alpha$.
B.53. Example. By the concavity of the logarithm and the monotonicity of the exponential function, we obtain the following estimate.

$$
\forall a, b \in(0, \infty), \theta \in[0,1]: a^{\theta} b^{1-\theta} \leq \theta a+(1-\theta) b \leq \max \{\theta, 1-\theta\}(a+b) .
$$

We will use it frequently. In particular, it implies that $a+b \leq a+a^{\theta} b^{1-\theta}+b \leq 2(a+b)$; that is,

$$
\begin{equation*}
a^{\theta} b^{1-\theta} \lesssim a+b \text { in }(0, \infty)^{2}, \quad a+b \sim a+a^{\theta} b^{1-\theta}+b \text { in }(0, \infty)^{2} \tag{B.12}
\end{equation*}
$$

for every $\theta \in[0,1]$.
Let us generalize these estimates to complex numbers.
B.54. Lemma. For $\phi \in(0, \pi)$, let $\Sigma_{\phi}:=\{\lambda \in \mathbb{C} \backslash\{0\}:|\arg \lambda|<\phi\}$ denote the open sector centered at zero with opening angle $2 \phi$. Then

$$
\begin{align*}
\lambda_{1}+\lambda_{2} & \sim\left|\lambda_{1}\right|+\left|\lambda_{2}\right| \quad \text { in } \Sigma_{\phi_{1}} \times \Sigma_{\phi_{2}} \text { if } \phi_{1}+\phi_{2}<\pi,  \tag{B.13}\\
\left|\lambda_{1}+\lambda_{2}\right| & \geq 2^{-1 / 2} \sqrt{1+\min \left\{0, \cos \left(\left|\arg \lambda_{1}\right|+\left|\arg \lambda_{2}\right|\right)\right\}}\left(\left|\lambda_{1}\right|+\left|\lambda_{2}\right|\right) \quad \text { in } \mathbb{C} \times \mathbb{C},  \tag{B.14}\\
\lambda_{1}+\lambda_{2} & \sim \lambda_{1}+\lambda_{1}^{1-\theta} \lambda_{2}^{\theta}+\lambda_{2} \quad \text { in } \Sigma_{\phi_{1}} \times \Sigma_{\phi_{2}} \text { if } \theta \in[0,1], \phi_{j} \in[0, \pi / 2), \phi_{1}+\phi_{2}<\pi / 2 . \tag{B.15}
\end{align*}
$$

Proof. (i) Clearly, the estimate $\lambda_{1}+\lambda_{2} \lesssim\left|\lambda_{1}\right|+\left|\lambda_{2}\right|$ applies in $\mathbb{C}^{2}$ by the triangle inequality. Let $\phi_{1} \in[0, \pi), \phi_{2} \in\left[0, \pi-\phi_{1}\right)$ and $\lambda_{1} \in \Sigma_{\phi_{1}}, \lambda_{2} \in \Sigma_{\phi_{2}}$. Then $\operatorname{Re}\left(\lambda_{1} \lambda_{2}\right) \geq\left|\lambda_{1} \lambda_{2}\right| \cos \left(\phi_{1}+\phi_{2}\right)$ where $\cos \left(\phi_{1}+\phi_{2}\right) \in(-1,1]$. For $a, b, c, s \in \mathbb{R}, s \leq 1$ we obtain $a^{2}+2 a b c+b^{2}=s(a+b)^{2}+(1-s)(a+b)^{2}+2 a b(c-1) \geq s(a+b)^{2}+4 a b(1-s)+2 a b(c-1)$. Choosing $s=\frac{1}{2}(1+c)$ yields $a^{2}+2 a b c+b^{2} \geq \frac{1}{2}(1+c)(a+b)^{2}$. Taking $c=\cos \left(\phi_{1}+\phi_{2}\right)$ yields

$$
\left|\lambda_{1}+\lambda_{2}\right|^{2}=\left|\lambda_{1}\right|^{2}+2 \operatorname{Re}\left(\lambda_{1} \bar{\lambda}_{2}\right)+\left|\lambda_{2}\right|^{2} \geq \frac{1}{2}\left(1+\cos \left(\phi_{1}+\phi_{2}\right)\right)\left(\left|\lambda_{1}\right|+\left|\lambda_{2}\right|\right)^{2}, \quad \text { for } \lambda_{j} \in \Sigma_{\phi_{j}} .
$$

This inequality and the triangle inequality yield the asserted inequalities.
(ii) Let $\lambda_{j} \in \Sigma_{\phi_{j}}$ for $\phi_{j} \in[0, \pi), \phi_{1}+\phi_{2}<\pi$ and $\theta \in[0,1]$. Then (B.12) implies

$$
\lambda_{1}+\lambda_{1}^{1-\theta} \lambda_{2}^{\theta}+\lambda_{2} \lesssim\left|\lambda_{1}\right|+\left|\lambda_{1}\right|^{1-\theta}\left|\lambda_{2}\right|^{\theta}+\left|\lambda_{2}\right| \lesssim\left|\lambda_{1}\right|+\left|\lambda_{2}\right| \lesssim \lambda_{1}+\lambda_{2} \text { in } \Sigma_{\phi_{1}} \times \Sigma_{\phi_{2}} .
$$

To prove the converse, we estimate as follows.

$$
\begin{aligned}
&\left|\lambda_{1}+\lambda_{1}^{1-\theta} \lambda_{2}^{\theta}+\lambda_{2}\right|^{2}=\left|\lambda_{1}\right|^{2}+2 \operatorname{Re}\left(\lambda_{1} \bar{\lambda}_{2}\right)+\left|\lambda_{2}\right|^{2}+\left|\lambda_{1}^{1-\theta} \lambda_{2}^{\theta}\right|^{2}+2 \operatorname{Re}\left(\left(\lambda_{1}+\lambda_{2}\right) \bar{\lambda}_{1}^{1-\theta} \bar{\lambda}_{2}^{\theta}\right), \\
& 2 \operatorname{Re}\left(\left(\lambda_{1}+\lambda_{2}\right) \bar{\lambda}_{1}^{1-\theta} \bar{\lambda}_{2}^{\theta}\right) \geq 2 \cos \left(\theta\left(\phi_{1}+\phi_{2}\right)\right)\left|\lambda_{1}\right|^{2-\theta}\left|\lambda_{2}\right|^{\theta}+2 \cos \left((1-\theta)\left(\phi_{1}+\phi_{2}\right)\right)\left|\lambda_{1}\right|^{1-\theta}\left|\lambda_{2}\right|^{1+\theta} .
\end{aligned}
$$

Let us abbreviate $c_{s}:=\cos \left(s\left(\phi_{1}+\phi_{2}\right)\right)$ for $s \in \mathbb{R}$. Since both $c_{\theta}$ and $c_{1-\theta}$ are non-negative, we have $2 \operatorname{Re}\left(\left(\lambda_{1}+\lambda_{2}\right) \bar{\lambda}_{1}^{1-\theta} \bar{\lambda}_{2}^{\theta}\right) \geq 0$ and (B.15) is established.

Let us derive some inequalities for the elements $z=\left(z_{j}\right)_{j=1}^{n} \in \overline{B \Sigma}_{\delta}^{n}$ of the closed bisector

$$
\overline{B \Sigma}_{\delta}^{n}:=\left(\overline{B \Sigma}_{\delta}\right)^{n}, \quad B \Sigma_{\delta}:=\left\{z=r e^{i \varphi} \in \mathbb{C}: r \in \dot{\mathbb{R}},|\arg \varphi-\pi / 2|<\delta\right\}=i \Sigma_{\delta} \cup-i \Sigma_{\delta} .
$$

B.55. Lemma. Let $n \in \mathbb{N}, \delta \in[0, \pi)$ and define $|z|_{-}:=\sqrt{-z \cdot z}=\sqrt{-\sum_{j} z_{j}^{2}}$ for $z \in{\overline{B \Sigma_{\delta}} \backslash\{0\} \text {. }}_{n}^{n}$.
(i) If $\delta \in[0, \pi / 4)$, then $|z|_{-} \in \bar{\Sigma}_{\delta}$ for every $z \in \overline{B \Sigma}_{\delta}^{n} \backslash\{0\}$.
(ii) If $\delta \in[0, \pi / 4)$ satisfies $1+(n-1) \cos (4 \delta)>0$, then

$$
\left||z|_{-}\right| \geq\left(\frac{1+(n-1) \cos (4 \delta)}{n}\right)^{1 / 4}|z| \quad \text { for all } z \in \overline{B \Sigma}_{\delta}^{n} \backslash\{0\}
$$

(iii) If $\delta \in[0, \pi / 8]$, then every $z \in \overline{B \Sigma}_{\delta}^{n} \backslash\{0\}$ satisfies

$$
\begin{equation*}
n^{-1 / 4}|z| \leq\left||z|_{-}\right| \leq|z|, \quad \operatorname{Re}|z|_{-} \geq n^{-1 / 4} \cos (\delta)|z|, \quad \operatorname{Im}|z|_{-} \leq \sin (\delta)|z| \tag{B.16}
\end{equation*}
$$

Proof. (i) For every $z=\left(z_{j}\right) \in \overline{B \Sigma}_{\delta}^{n} \backslash\{0\}$ we have $-z_{j}^{2} \in \bar{\Sigma}_{2 \delta}$ and hence $-z \cdot z=-\sum_{j} z_{j}^{2} \in$ $\bar{\Sigma}_{2 \delta}$ by (B.13) and therefore $|z|_{-}=\sqrt{-\sum_{j} z_{j}^{2}}$ belongs to $\bar{\Sigma}_{\delta}$.
(ii) Hölder's inequality yields

$$
|z|^{2}=|z|_{2}^{2}=\sum_{j}\left|z_{j}\right|^{2} \leq n^{1 / 2}\left(\sum_{j}\left|z_{j}\right|^{4}\right)^{1 / 2}=n^{1 / 2}|z|_{4}^{2} \quad \text { for } z \in \mathbb{C}^{n}
$$

Then the assertion follows from the following estimate.

$$
\begin{aligned}
\|\left.\left. z\right|_{-}\right|^{4} & =\left|-\sum_{j} z_{j}^{2}\right|^{2}=\sum_{j, k} z_{j}^{2} \overline{z_{k}^{2}}=\sum_{j}\left|z_{j}^{2}\right|^{2}+2 \sum_{j<k} \operatorname{Re}\left(z_{j}^{2} \overline{z_{k}^{2}}\right) \\
& \geq \sum_{j}\left|z_{j}^{2}\right|^{2}+2 \cos (4 \delta) \sum_{j<k}\left|z_{j}^{2}\right|\left|z_{k}^{2}\right| \\
& =(1-\cos (4 \delta)) \sum_{j}\left|z_{j}^{2}\right|^{2}+\cos (4 \delta)\left(\sum_{j}\left|z_{j}^{2}\right|^{2}+2 \sum_{j<k}\left|z_{j}^{2} \| z_{k}^{2}\right|\right) \\
& \geq\left(\frac{1-\cos (4 \delta)}{n}+\cos (4 \delta)\right)|z|_{2}^{4} .
\end{aligned}
$$

(iii) This is a simple consequence of (i) and (ii).
B.56. Example. We consider the parabolic symbol

$$
\omega(\lambda, z)=\sqrt{\rho(\tau+\lambda)-\mu z^{2}}, \quad \lambda \in \Sigma_{\phi}, z \in B \Sigma_{\delta}^{n}
$$

where $\rho>0, \tau>0$ and $\mu>0$ are constants and $\phi \in(\pi / 2, \pi)$ and $\delta \in(0, \pi / 8]$ satisfy $\phi+2 \delta<\pi$. Clearly, $|\omega| \lesssim 1+|\lambda|^{1 / 2}+|z|$. Then from inequality (B.14) and Hölder's inequality $a+b+c \leq$ $\sqrt{3} \sqrt{a^{2}+b^{2}+c^{2}}$ for $a, b, c \geq 0$ we obtain

$$
\begin{aligned}
|\omega(\lambda, z)| & \geq \sqrt{\sqrt{\frac{c_{1}}{2}} \rho \tau+\sqrt{\frac{c_{1}}{2}} \sqrt{\frac{c_{2}}{2}}\left(\rho|\lambda|+\mu\left|z^{2}\right|\right)} \\
& \geq \sqrt{\frac{1}{3} \sqrt{\frac{c_{1}}{2}} \rho \tau}+\sqrt{\frac{\sqrt{c_{1} c_{2}}}{6} \rho|\lambda|}+\sqrt{\frac{\sqrt{c_{1} c_{2}}}{6}} \mu\left|z^{2}\right| \quad \text { for } \lambda \in \Sigma_{\phi}, z \in B \Sigma_{\delta}^{n}
\end{aligned}
$$

where $c_{1}^{2}=1+\cos \phi>0$ and $c_{2}^{2}=1+\cos (\phi+2 n \delta)>0$. From (B.16) we conclude that

$$
\omega(\lambda, z)=\sqrt{\rho(\tau+\lambda)-\mu z^{2}} \sim 1+\lambda^{1 / 2}+\sqrt{-z^{2}} \quad \text { for } \lambda \in \Sigma_{\phi}, z \in B \Sigma_{\delta}^{n}
$$

B.2.8. Elliptic differential operators on manifolds. The subsequent theorem of Amann, Hieber and Simonett [AHS94] guarantees that certain elliptic operators on compact manifolds are $\mathcal{R}$-sectorial and have a bounded $\mathcal{H}^{\infty}$ functional calculus in $L_{p}(M ; G)$. Here $(M, g)$ is a compact $n$-dimensional Riemannian $C^{m}$-manifold without boundary $(m \in \mathbb{N}$ ) and $G:=(G, \pi, M)$ is a $C^{m}$-class vector bundle over $M$ whose fibers $\pi^{-1}(\{x\})(x \in M)$ are isomorphic to a Banach space $E \cong \mathbb{C}^{N}$ of finite dimension $N$. A trivializing coordinate system $\left(\kappa, \chi_{\kappa}\right)$ for $G$ consists
of a chart $\kappa: U_{\kappa} \rightarrow E$ and a trivializing map $\pi^{-1}\left(U_{\kappa}\right) \rightarrow U_{\kappa} \times E, g \mapsto\left(\pi(g), \chi_{\kappa}(g)\right)$. The local representation of a section $u$ of $G$ with respect to $\left(\kappa, \chi_{\kappa}\right)$ is given by $u_{\kappa}:=\chi_{\kappa} \circ u \circ \kappa^{-1}$, $\kappa\left(U_{\kappa}\right) \rightarrow E$.

For given $p \in(1, \infty)$ and $k \in\{0,1, \ldots, m\}$, we define the Sobolev spaces $H_{p}^{k}(M ; G)$ of all sections $u$ of $G$ such that $\varphi u_{\kappa}$ belongs to $H_{p}^{k}\left(\mathbb{R}^{n} ; E\right)$ for every $\varphi \in C_{c}^{k}\left(\kappa\left(U_{\kappa}\right)\right)$ and for every trivializing coordinate system $\left(\kappa, \chi_{\kappa}\right)$. Moreover, $H_{p}^{k}(M ; G)$ is a Banach space with respect to the norm $\|u\|_{H_{p}^{k}(M ; G)}:=\sum_{\kappa}\left\|\left(\varphi_{\kappa} \circ \kappa^{-1}\right) u_{\kappa}\right\|_{H_{p}^{k}\left(\mathbb{R}^{n} ; E\right)}$ where the sum is taken over a finite partition of unity $\left(\varphi_{\kappa}\right)_{\kappa}$ for $M$, subordinate to $\left(U_{\kappa}\right)_{\kappa}$. Hence for $K=|\{\kappa\}|$, the map $r: H_{p}^{k}\left(\mathbb{R}^{n} ; \mathbb{C}^{N}\right)^{K} \rightarrow H_{p}^{k}(M ; G),\left(u_{\kappa}\right)_{\kappa} \mapsto u$ is a retraction and thus $H_{p}^{k}(M ; G)$ inherits many embedding and interpolation properties from the space $H_{p}^{k}\left(\mathbb{R}^{n}\right)$. The Lebesgue spaces $L_{p}(M ; G)$ and the Sobolev-Slobodeckiĭ spaces $W_{p}^{s}(M ; G)$ for $s \in[0, m]$ are defined analogously.

Let $\mathcal{A}: H_{p}^{m}(M ; G) \rightarrow L_{p}(M ; G)$ be differential operator with representation $\mathcal{A}_{\kappa}(y, D)=$ $\sum_{|\alpha| \leq m} a_{\kappa, \alpha}(y) D^{\alpha}$ for $y \in \kappa\left(U_{\kappa}\right)$, where $a_{\kappa, \alpha} \in C\left(\kappa\left(U_{\kappa}\right) ; \mathcal{L}(E)\right)$ and $D_{j}:=-i \partial_{j}$ for $1 \leq j \leq n$. The operator $\mathcal{A}$ is called $\theta$-elliptic $\left(\theta \in[0, \pi)\right.$ ), if its principal symbol $\mathcal{A}_{\pi}$ satisfies

$$
\sigma\left(\mathcal{A}_{\pi}\left(\xi_{x}^{*}\right)\right) \subset\{\lambda \in \mathbb{C} \backslash\{0\}:|\arg \lambda| \leq \theta\}, \quad \text { for all } x \in M, \xi_{x}^{*} \in T_{x}^{*} M \backslash\{0\}
$$

Note that $\mathcal{A}_{\pi}\left(\xi_{x}^{*}\right) \in \mathcal{L}\left(\pi^{-1}(\{x\})\right)$ is a homogeneous polynomial in $\xi_{x}^{*}$.
B.57. Theorem ([AHS94, Theorem 10.1, Theorem 10.3]). Let $M$ be a compact $n$-dimensional $C^{m}$ manifold $\left(m \in \mathbb{N} \text { ) without boundary, let } G \text { be a complex } C^{\min \{m, 2\}}\right\}_{\text {-vector bundle over } M \text { and let } p \in}$ $(1, \infty)$. Let $\mathcal{A}: H_{p}^{m}(M ; G) \rightarrow L_{p}(M ; G)$ be a linear differential operator with continuous coefficients such that $\mathcal{A}$ is $\theta_{0}$-elliptic for some $\theta_{0} \in[0, \pi)$.
(i) For every $\theta \in\left(\theta_{0}, \pi\right)$ there exists $\mu_{\theta}>0$ such that $\mu_{\theta}+\mathcal{A}: H_{p}^{m}(M ; G) \rightarrow L_{p}(M ; G)$ is an isomorphism and $\mathcal{R}$-sectorial with $\mathcal{R}$-angle $\theta$.
(ii) If $\mathcal{A}_{\pi}$ has $C^{\varepsilon}$-coefficients for some $\varepsilon \in(0,1)$, then for every $\theta \in\left(0, \theta_{0}\right)$ there exists $\mu_{\theta}>0$ such that $\mu_{\theta}+\mathcal{A}$ has a bounded $\mathcal{H}^{\infty}$ functional calculus in $L_{p}(M ; G)$ with $\mathcal{H}^{\infty}$-angle $\theta$.
B.58. Corollary. Let $\Sigma \subset \mathbb{R}^{n}(n \geq 2)$ be a compact $C^{2}$-hypersurface, let $\Delta_{\Sigma}=\operatorname{div}_{\Sigma} \nabla_{\Sigma}=g^{i j}\left(\partial_{i} \partial_{j}-\right.$ $\left.\Lambda_{i j}^{k} \partial_{k}\right)$ denote the scalar Laplace-Beltrami operator and let $p \in(1, \infty), \mathbb{K}=\mathbb{C}$.
(i) For every $\mu \in(0, \infty)$, the operator $\mu-\Delta_{\Sigma}: H_{p}^{2}(\Sigma) \rightarrow L_{p}(\Sigma)$ is invertible and $\mathcal{R}$-sectorial with $\mathcal{R}$-angle zero.
(ii) If $\Sigma \in C^{2+\varepsilon}$ for some $\varepsilon \in(0,1)$, then for every $\theta \in(0, \pi)$ there exists $\mu_{\theta} \in(0, \infty)$ such that $\mu_{\theta}-\Delta_{\Sigma}$ has a bounded $\mathcal{H}^{\infty}$ functional calculus in $L_{p}(\Sigma)$ with $\mathcal{H}^{\infty}$-angle $\theta$.
(iii) Let $\lambda_{1}>0$ denote the smallest non-zero eigenvalue of $-\Delta_{\Sigma}$. Then for every $\mu \in\left(-\lambda_{1}, \infty\right)$, the operator $\mu-\Delta_{\Sigma}: H_{p}^{2}(\Sigma) \cap L_{p, 0}(\Sigma) \rightarrow L_{p, 0}(\Sigma)$ is invertible and $\mathcal{R}$-sectorial with $\mathcal{R}$-angle zero.
(iv) Let $s \in[0, \infty)$ and assume that $\Sigma$ be smooth and let $s \in[0, \infty)$. Then for every $\mu \in(0, \infty)$, the operator $\mu-\Delta_{\Sigma}: W_{p}^{s+2}(\Sigma) \rightarrow W_{p}^{s}(\Sigma)$ is invertible and $\mathcal{R}$-sectorial with $\mathcal{R}$-angle zero.

Proof. (i) The domain $H_{p}^{2}(\Sigma)$ of $\Delta_{\Sigma}$ is compactly embedded into the ground space $L_{p}(\Sigma)$ and therefore the spectrum of $\Delta_{\Sigma}$ consists solely of eigenvalues with finite multiplicity. The surface divergence theorem implies that all eigenvalues are non-positive and that zero is an eigenvalue with multiplicity one (the corresponding eigenfunctions are the constant functions). Hence, by considering the operators $-e^{i \psi} \Delta_{\Sigma}(\psi \in(-\pi / 2, \pi / 2))$ and using Theorem B.57.(i), Theorem B. 22 and Theorem B.23, we obtain the assertion.
(ii) Since $g^{i j} \in C^{1+\varepsilon}$ and $\Lambda_{i j}^{k} \in C^{\varepsilon}$, the assertion follows from Theorem B.57.(ii).
(iii) We have the direct decomposition $L_{p}(\Sigma)=L_{p, 0}(\Sigma) \oplus \mathbb{K}$, where we identify $\mathbb{K}$ with the constant functions, which form the eigenspace for the eigenvalue zero. Hence the spectrum of the realization of $-\Delta_{\Sigma}$ in $L_{p, 0}(\Sigma)$ is contained in $\left[\lambda_{1}, \infty\right)$ and the assertion follows as in (i).
(iv) By means of a localization procedure as in Section 2.1 it can be shown that the operator $\mu-\Delta_{\Sigma}: H_{p}^{k+2}(\Sigma) \rightarrow H_{p}^{k}(\Sigma)$ is invertible for large $\mu$. This assertion holds true for all $\mu>0$ by Theorem B.22. By means of retractions $r: W_{p}^{s}\left(\mathbb{R}^{n-1}\right)^{K} \rightarrow W_{p}^{s}(\Sigma)$ and real interpolation, it
follows that $\mu-\Delta_{\Sigma}: W_{p}^{s+2}(\Sigma) \rightarrow W_{p}^{s}(\Sigma)$ is invertible and $\mathcal{R}$-sectorial with $\mathcal{R}$-angle zero for every $s \in[0, \infty)$.
B.59. Corollary. Let $\Sigma \subset \mathbb{R}^{n}(n \geq 2)$ be a compact $C^{2}$-hypersurface, let $p \in(1, \infty)$, let $\widetilde{\nabla}_{\Sigma}$ denote the covariant derivative on $\Sigma$ and let $\widetilde{\Delta}_{\Sigma}: H_{p}^{2}(\Sigma ; T \Sigma) \rightarrow L_{p}(\Sigma ; T \Sigma), \widetilde{\Delta}_{\Sigma} v=g^{\alpha \beta} \widetilde{\nabla}_{\alpha} \widetilde{\nabla}_{\beta} v$ denote the Laplace-Beltrami operator for tangential vector fields on $\Sigma$ (see page 141).
(i) For every $\mu \in(0, \infty)$, the operator $\mu-\widetilde{\Delta}_{\Sigma}: H_{p}^{2}(\Sigma ; T \Sigma) \rightarrow L_{p}(\Sigma ; T \Sigma)$ is invertible.
(ii) Let $(T \Sigma)_{\mathbb{C}}$ denote the complexification of $T \Sigma$. Then for every $\mu \in(0, \infty)$, the operator $\mu-$ $\widetilde{\Delta}_{\Sigma}: H_{p}^{2}\left(\Sigma ;(T \Sigma)_{\mathbb{C}}\right) \rightarrow L_{p}\left(\Sigma ;(T \Sigma)_{\mathbb{C}}\right)$ is an isomorphism and $\mathcal{R}$-sectorial with $\mathcal{R}$-angle zero.
(iii) If $\Sigma \in C^{2+\varepsilon}$ for some $\varepsilon \in(0,1)$, then for every $\theta \in(0, \pi)$ there exists $\mu_{\theta} \in(0, \infty)$ such that $\mu_{\theta}-\widetilde{\Delta}_{\Sigma}$ has a bounded $\mathcal{H}^{\infty}$ functional calculus in $L_{p}\left(\Sigma ;(T \Sigma)_{\mathbb{C}}\right)$ with $\mathcal{H}^{\infty}$-angle $\theta$.
Proof. The proof is similar as for Corollary B.58.

## B.3. Joint functional calculus and mixed-order systems

B.3.1. The joint $\mathcal{H}^{\infty}$-functional calculus for $\left(\partial_{t}, \nabla\right)$. We collect some results of Denk and Kaip [DK13] on the joint functional calculus operator tuples like $\nabla=\left(\partial_{1}, \ldots, \partial_{n}\right)$ and $\left(\partial_{t}, \nabla\right)$.
B.60. Definition ([KW04, p. 4.9]). We say that a Banach space $X$ has property $(\alpha)$, if there exists $C>0$ such that

$$
\int_{0}^{1} \int_{0}^{1}\left\|\sum_{i, j=1}^{n} r_{i}(u) r_{j}(v) \alpha_{i j} x_{i j}\right\|_{X} d u d v \leq C \int_{0}^{1} \int_{0}^{1}\left\|\sum_{i, j=1}^{n} r_{i}(u) r_{j}(v) x_{i j}\right\|_{X} d u d v
$$

for all $n \in \mathbb{N}, \alpha_{i j} \in \mathbb{C}$ with $\left|\alpha_{i j}\right| \leq 1, x_{i j} \in X$.
B.61. Remarks ([KW04, p. 4.10], [DK13, Remark 1.15]). Let $X$ have property ( $\alpha$ ).
(i) If $Y$ is a closed subspace of $X$, then $Y$ has property $(\alpha)$.
(ii) If $Y$ is isomorphic to $X$, then $Y$ has property $(\alpha)$.
(iii) If $(\Omega, \mu)$ is a $\sigma$-finite measure space and $p \in[1, \infty)$, then $L_{p}(\Omega, \mu ; X)$ has property $(\alpha)$.
(iv) Every Hilbert space has property ( $\alpha$ ).
B.62. Definition (Ground space $\mathcal{W}$, [DK13, Definition 1.71], [Kai12, Definition 2.25]). Let $n \in \mathbb{N}$, $p_{0}, p_{1}, q_{0}, q_{1} \in(1, \infty), s, \omega \in[0, \infty), r \in \mathbb{R}$ and $X$ be a Banach space of class $\mathcal{H} \mathcal{T}$ with property $(\alpha)$. Then we let

$$
\mathcal{W}:=e^{\omega \cdot}{ }_{0} \mathcal{F}^{s}\left(\mathbb{R}_{+} ; \mathcal{K}^{r}\left(\mathbb{R}^{n} ; X\right)\right), \quad \mathcal{K} \in\left\{B_{p_{1}, q_{1}}, H_{p_{1}}\right\}, \quad \mathcal{F} \in \begin{cases}\left\{B_{p_{0}, q_{0}}, H_{p_{0}}\right\}, & \text { if } s>0, \\ \left\{H_{p_{0}}\right\}, & \text { if } s=0 .\end{cases}
$$

Here the space $e^{\omega \cdot}{ }_{0} \mathcal{F}^{s}\left(\mathbb{R}_{+} ; Y\right)$ consists of all functions $t \mapsto e^{\omega t} u(t)$ such that $u \in{ }_{0} \mathcal{F}^{s}\left(\mathbb{R}_{+} ; Y\right)$, equipped with the norm $\left\|t \mapsto e^{\omega t} u(t)\right\|_{0 \mathcal{F}^{s}\left(\mathbb{R}_{+} ; Y\right)}$.
B.63. Definition (Sectors, bisectors, curves, cf. [DK13, Definition 1.1]). For $\phi \in(0, \pi)$, let $\Sigma_{\phi} \subset$ $\mathbb{C}$ denote the open sector

$$
\Sigma_{\phi}=\left\{z=r e^{i \varphi}: r \in(0, \infty),|\varphi|<\phi\right\}=\{z \in \mathbb{C} \backslash\{0\}:|\arg z|<\phi\} .
$$

For $\delta \in(0, \pi / 2)$, let $B \Sigma_{\delta} \subset \mathbb{C}$ denote the open bisector

$$
B \Sigma_{\delta}=\left\{z=r e^{i \varphi}: r \in \mathbb{R} \backslash\{0\},|\varphi-\pi / 2|<\delta\right\}=i \Sigma_{\delta} \cup-i \Sigma_{\delta} .
$$

For $\psi \in(0, \pi)$ we define the curve $\Gamma_{\psi} \subset \mathbb{C}$ by means of the parametrization $\mathbb{R} \ni r \mapsto$ $|r| e^{-i \psi \operatorname{sign} r}$. Hence $\Gamma_{\psi}=\partial \Sigma_{\psi}$ is oriented counter-clockwise around $\Sigma_{\psi}$.
B.64. Definition ([cf. DK13, Definition 1.17]). Let $n \in \mathbb{N}$, let $\Omega \subset \mathbb{C}^{n}$ be open and let $Y$ be a Banach space. We define
(i) $\mathcal{H}(\Omega ; Y)$, the vector space of all holomorphic $Y$-valued functions on $\Omega$,
(ii) $\mathcal{H}^{\infty}(\Omega ; Y)$, the vector space of all bounded holomorphic $Y$-valued functions on $\Omega$,
(iii) $\mathcal{H}_{0}^{\infty}(\Omega ; Y)=\left\{f \in \mathcal{H}^{\infty}(\Omega ; Y): \exists C, s>0 \forall z \in \Omega:|f(z)|_{Y} \leq C \prod_{j=1}^{n}\left(\min \left\{\left|z_{j}\right|,\left|z_{j}\right|^{-1}\right\}\right)^{s}\right\}$,
(iv) $\mathcal{H}_{P}(\Omega ; Y)=\left\{f \in \mathcal{H}(\Omega ; Y): \exists C, s>0 \forall z \in \Omega:|f(z)|_{Y} \leq C \prod_{j=1}^{n}\left(\max \left\{\left|z_{j}\right|,\left|z_{j}\right|^{-1}\right\}\right)^{s}\right\}$.

Thus the spaces $\mathcal{H}_{0}^{\infty}(\Omega ; Y)$ and $\mathcal{H}_{P}(\Omega ; Y)$ consists of functions that have a polynomial decay or a polynomial growth at zero and at infinity.
B.65. Remarks ([DK13, Definition 1.20]). Let $T_{j}(j \in\{1, \ldots, n\})$ be closed linear operators in a complex Banach space $X$ such that each $T_{j}$ is sectorial or bi-sectorial and such that all resolvents the $\left(\lambda-T_{j}\right)^{-1},\left(\mu-T_{l}\right)^{-1}$ commute $\left(\lambda \in \rho\left(T_{j}\right), \mu \in \rho\left(T_{l}\right), j, l \in\{1, \ldots, n\}\right)$.

For sectorial $T_{j}$ we choose $\Omega_{j}:=\Sigma_{\theta_{j}}$ with some $\theta_{j} \in\left(\varphi_{T_{j}}, \pi\right)$ and $\Gamma_{j}:=\Gamma_{\varphi_{j}} \partial \Sigma \ldots$.. For bi-sectorial $T_{j}$ we choose $\Omega_{j}:=B \Sigma_{\delta_{j}}$ with some $\delta_{j} \in\left(\varphi_{T_{j}}^{\mathrm{bi}}, \pi / 2\right)$ and $\Gamma_{j}:=\partial \Sigma_{\phi_{j}}$ with some $\phi_{j} \in()$. We put $\Omega:=\Omega_{1} \times \cdots \times \Omega_{n}$ and $\Gamma:=\Gamma_{1} \times \cdots \times \Gamma_{n}$ and we let $\mathcal{B}_{T} \subset \mathcal{B}(X)$ denote the commutator algebra of $\left\{\left(\lambda-T_{j}\right)^{-1}: \lambda \in \rho\left(T_{j}\right), j \in\{1, \ldots, n\}\right\}$ in $\mathcal{B}(X)$.
(i) Joint $\mathcal{H}^{\infty}$-functional calculus. For $f \in \mathcal{H}_{0}^{\infty}\left(\Omega ; \mathcal{B}_{T}\right)$ we define

$$
f(T):=\frac{1}{(2 \pi i)^{n}} \int_{\Gamma} f(z) \prod_{j=1}^{n}\left(z_{j}-T_{j}\right)^{-1} d z \in \mathcal{B}(X)
$$

(ii) Joint $\mathcal{H}_{P}$-functional calculus. For $k, n \in \mathbb{N}$, the functions

$$
\psi_{k, n}(z):=\frac{k^{2} z_{1}}{\left(1+k z_{1}\right)\left(k+z_{1}\right)} \cdots \frac{k^{2} z_{n}}{\left(1+k z_{n}\right)\left(k+z_{n}\right)}
$$

belong to $\mathcal{H}_{0}^{\infty}(\Omega)$ and for every $f \in \mathcal{H}_{P}(\Omega ; Y)$ there exists $m \in \mathbb{N}_{0}$ such that $\psi_{k, n}^{m} f \in \mathcal{H}_{0}^{\infty}\left(\Omega ; \mathcal{B}_{T}\right)$ for all $k \in \mathbb{N}$. Hence, for $f \in \mathcal{H}_{P}\left(\Omega ; \mathcal{B}_{T}\right)$ we may choose $m \in \mathbb{N}_{0}$ with $\psi_{1, n}^{m} f \in \mathcal{H}_{0}^{\infty}\left(\Omega ; \mathcal{B}_{T}\right)$ and define

$$
f(T): D(f(T)) \rightarrow X, \quad x \mapsto \psi(T)^{-m}\left(\psi^{m} f\right)(T) x
$$

with domain $D(f(T)):=\left\{x \in X:\left(\psi^{m} f\right)(T) x \in R\left(\psi(T)^{m}\right)\right\}$.
B.66. Definition (Joint $\mathcal{H}^{\infty}$-functional calculus for $\left(\mu+\partial_{t}, \nabla_{x}\right)$, [Kai12, Definition 1.10]). Put

$$
\Omega:=\Sigma_{\phi} \times B \Sigma_{\delta_{1}} \times \cdots B \Sigma_{\delta_{n}}, \quad \text { where } \phi \in(\pi / 2, \pi), \delta_{j} \in(0, \pi / 2)
$$

Let $\mu \in[0, \infty)$ and consider the operators $\partial_{t}, \partial_{x_{1}}, \ldots, \partial_{x_{n}}$ as closed operators in $\mathcal{W}$. For the tuple $\left(\mu+\partial_{t}, \nabla\right)$ we define the joint $\mathcal{H}^{\infty}(\Omega)$-functional calculus

$$
f \mapsto f\left(\mu+\partial_{t}, \nabla\right):=\frac{1}{(2 \pi i)^{1+n}} \int_{\Gamma_{\phi^{\prime}}} \int_{\prod_{j}\left(\Gamma_{\delta_{j}^{\prime}} \cup\left(-\Gamma_{\delta_{j}^{\prime}}\right)\right)} f(\tau, z)\left(\tau-\mu-\partial_{t}\right)^{-1} \prod_{j}\left(z_{j}-\partial_{x_{j}}\right)^{-1} d(\tau, z),
$$

where $f \in \mathcal{H}_{0}^{\infty}(\Omega)$ and $\phi^{\prime} \in(\pi / 2, \phi), \delta_{j}^{\prime} \in\left(0, \delta_{j}\right)$ (the integrals do not depend on $\left.\phi^{\prime}, \delta_{j}^{\prime}\right)$. The resolvents $\left(\tau-\mu-\partial_{t}\right)^{-1}$ and $\left(z_{j}-\partial_{x_{j}}\right)^{-1}$ are considered as bounded linear operators in the same ground space $\mathcal{W}$ according to Definition B.62.
B.67. Theorem (Time derivative, [DK13, Theorems 1.83, 1.84]). Let $r, s, \omega \in[0, \infty)$ and let $\mathcal{F}, \mathcal{K}$, $X$ be as in Definition B.62. Let

$$
\mathcal{D}_{t}: u \mapsto \partial_{t} u, \quad e^{\omega \cdot}{ }_{0} \mathcal{F}^{s+1}\left(\mathbb{R}_{+} ; X\right) \rightarrow e^{\omega \cdot}{ }_{0} \mathcal{F}^{s}\left(\mathbb{R}_{+} ; X\right)
$$

denote the realization in $e^{\omega \cdot}{ }_{0} \mathcal{F}^{s}\left(\mathbb{R}_{+} ; X\right)$ of the time derivative. Then the following assertions are valid.
(i) $\mathcal{D}_{t}$ has an $\mathcal{R}$-bounded $\mathcal{H}^{\infty}$-calculus with $\phi_{\mathcal{D}_{t}}^{\mathcal{R} \infty}=\pi / 2$.
(ii) The operator $\mathcal{D}_{t}$ has
(a) the resolvent set $\rho\left(\mathcal{D}_{t}\right)=\{z \in \mathbb{C}: \operatorname{Re} z<\omega\}$,
(b) the residual spectrum $\sigma_{r}\left(\mathcal{D}_{t}\right)=\{z \in \mathbb{C}: \operatorname{Re} z>\omega\}$,
(c) the point spectrum $\sigma_{p}\left(\mathcal{D}_{t}\right)=\emptyset$,
(d) the continuous spectrum $\sigma_{c}\left(\mathcal{D}_{t}\right)=i \mathbb{R}+\omega$.
B.68. Theorem ([MS12, Proposition 2.9]). Let $p \in(1, \infty), X \in \mathcal{H} \mathcal{T}, s \in[0, \infty), \alpha \in(0,2)$, $\omega \in(0, \infty)$. Then the operators

$$
\begin{array}{ll}
\left(\omega-\partial_{t}\right)^{\alpha} & \text { in } H_{p}^{s}\left(\mathbb{R}_{+} ; X\right) \text { with domain } H_{p}^{s+\alpha}\left(\mathbb{R}_{+} ; X\right), \\
\left(\omega-\partial_{t}\right)^{\alpha} & \text { in } W_{p}^{s}\left(\mathbb{R}_{+} ; X\right) \text { with domain } W_{p}^{s+\alpha}\left(\mathbb{R}_{+} ; X\right), s, s+\alpha \notin \mathbb{N}_{0}
\end{array}
$$

are invertible and have bounded $\mathcal{H}^{\infty}$ functional calculi with angle not larger than $\alpha \pi / 2$.
B.69. Theorem (Space derivatives, [DK13, Theorem 1.81]). Let

$$
\mathcal{D}_{x_{j}}: u \mapsto \partial_{x_{j}} u, \quad e^{\omega \cdot}{ }_{0} \mathcal{F}^{s}\left(\mathbb{R}_{+} ; \mathcal{K}^{r+1}\left(\mathbb{R}^{n} ; X\right)\right) \rightarrow e^{\omega \cdot}{ }_{0} \mathcal{F}^{s}\left(\mathbb{R}_{+} ; \mathcal{K}^{r}\left(\mathbb{R}^{n} ; X\right)\right)
$$

denote the realizations of the partial derivatives in $e^{\omega}{ }_{0} \mathcal{F}^{s}\left(\mathbb{R}_{+} ; \mathcal{K}^{r}\left(\mathbb{R}^{n} ; X\right)\right)$. Let $\delta_{j} \in(0, \pi / 2)$ and $\Omega_{x}=\prod_{j=1}^{n} B \Sigma_{\delta_{j}}$. Then the tuple $\mathcal{D}_{x}=\left(\mathcal{D}_{x_{1}}, \ldots, \mathcal{D}_{x_{n}}\right)$ has a bounded joint $\mathcal{H}^{\infty}\left(\Omega_{x}\right)$-calculus.
B.70. Theorem ([Kai12, Definition 1.15, Theorem 2.47]). Let $\mathcal{W}=e^{\omega}{ }_{0} \mathcal{F}^{s}\left(\mathbb{R}_{+} ; \mathcal{K}^{r}\left(\mathbb{R}^{n} ; X\right)\right.$ ) be as in Definition B.62, let $\Omega=\Sigma_{\phi} \times B \Sigma_{\delta_{1}} \times \cdots \times B \Sigma_{\delta_{n}}$ be as in Definition B. 66 and let $\sigma \geq 0$. Then the tuple $\left(\sigma+\mathcal{D}_{t}, \mathcal{D}_{x}\right)$ has a bounded joint $\mathcal{H}^{\infty}(\Omega)$-functional calculus in $\mathcal{W}$.
B.3.2. Parabolic mixed-order systems. We define order functions and Newton polygons. An example is given in Figure B. 1 on the facing page. Then we consider a class of parameterdependent symbols $S\left(\bar{\Sigma}_{\phi} \times \overline{B \Sigma}_{\delta}^{n} \times K\right)$, which are used in Section 3.1 for solving the FouierLaplace transformed differential equations.
B.71. Remarks ([DK13]). (i) A continuous and piecewise linear function $\mu:[0, \infty) \rightarrow \mathbb{R}$ is called an order function if $\mu$ is convex or concave. In this case there exist $M \in \mathbb{N}$ and $\gamma_{l}>0$, $m_{l}(\mu), b_{l}(\mu) \in \mathbb{R}$ for $l \in\{0, \ldots, M\}$ with $0=: \gamma_{0}<\gamma_{1}<\cdots<\gamma_{M}<\gamma_{M+1}:=\infty$ such that

$$
\mu(\gamma)=b_{l}(\mu)+m_{l}(\mu) \gamma \quad \text { for } \gamma \in\left(\gamma_{l}, \gamma_{l+1}\right),
$$

and we have

$$
m_{l-1}(\mu) \leq m_{l}(\mu), \quad b_{l-1}(\mu) \geq b_{l}(\mu) \quad \text { for } l \in\{1, \ldots, M\}
$$

(that is, $\mu$ is convex) or

$$
m_{l-1}(\mu) \geq m_{l}(\mu), \quad b_{l-1}(\mu) \leq b_{l}(\mu) \quad \text { for } l \in\{1, \ldots, M\}
$$

(that is, $\mu$ is concave). If $\mu$ is convex, then we have

$$
\mu(\gamma)=\max \left\{b_{l}(\mu)+m_{l}(\mu) \gamma: l \in\{0, \ldots, M\}\right\} \quad \text { for } \gamma \in[0, \infty) .
$$

(ii) A convex [concave] order function $\mu$ is increasing [decreasing] if $m_{l}(\mu) \geq 0\left[m_{l}(\mu) \leq 0\right.$ ] for all $l \in\{0, \ldots, M\}$. A convex [concave] order function $\mu$ is increasing [decreasing] if $m_{l}(\mu) \geq 0$ [ $\left.m_{l}(\mu) \leq 0\right]$ for all $l \in\{0, \ldots, M\}$.
(iii) An order function $\mu$ is called strictly positive if $\mu$ is convex and $m_{l}(\mu) \geq 0$ and $b_{l}(\mu) \geq 0$ for all $l \in\{0, \ldots, M\}$. An order function $\mu$ is called strictly negative if $-\mu$ is strictly positive.
(iv) For a given finite set $\nu=\left(r_{j}, s_{j}\right)_{j=0}^{J+1} \subset[0, \infty)^{2}, J \in \mathbb{N}_{0}$, the associated Newton polygon $N(\nu)$ is defined as the convex hull in $\mathbb{R}^{2}$ of the set of vertices $(0,0),\left(0, s_{j}\right),\left(r_{j}, 0\right),\left(r_{j}, s_{j}\right)$, $j \in\{0, \ldots, J+1\}$.
B.72. Definition (Symbol class $S\left(\bar{\Sigma}_{\phi} \times \overline{B \Sigma}_{\delta}^{n} \times K\right)$, cf. [DK13]). Let $K \subset \mathbb{C}^{m}$ be compact, $\phi \in$ $(\pi / 2, \pi), \delta \in(0, \pi / 2)$. Then we let $S\left(\bar{\Sigma}_{\phi} \times \overline{B \Sigma}_{\delta}^{n} \times K\right)$ be the set of all functions

$$
\begin{equation*}
P: \bar{\Sigma}_{\phi} \times \overline{B \Sigma}_{\delta}^{n} \times K \rightarrow \mathbb{C}, \quad(\lambda, z, \vartheta) \mapsto P(\lambda, z ; \vartheta)=\sum_{j \in J} \chi_{j}(\vartheta) \omega_{j}(\lambda, z) \varphi_{j}(\lambda) \psi_{j}(z), \tag{B.17}
\end{equation*}
$$

where $J$ is a finite index set and for all $j \in J$,
(i) $\chi_{j}: K \rightarrow \mathbb{C}$ is continuous and nontrivial,



Figure B.1. $\gamma$-order and Newton polygon of the symbol $\omega(\lambda, z)=\left(\lambda+|z|_{-}^{2}\right)^{1 / 2}$.
(ii) $\omega_{j}$ is holomorphic in $\Sigma_{\phi} \times B \Sigma_{\delta}^{n}$, continuous in $\bar{\Sigma}_{\phi} \times \bar{B}_{\delta}^{n}$ and satisfies

$$
\omega_{j}\left(\eta^{2} \lambda, \eta z\right)=\eta^{N_{j}} \omega_{j}(\lambda, z) \neq 0 \quad \text { for } \eta>0,(\lambda, z) \in \bar{\Sigma}_{\phi} \times \overline{B \Sigma}_{\delta}^{n} \backslash\{(0,0)\}
$$

with some $N_{j} \in[0, \infty)$,
(iii) $\varphi_{j}$ is holomorphic in $\Sigma_{\phi}$, continuous in $\bar{\Sigma}_{\phi}$ and satisfies

$$
\varphi_{j}(\eta \lambda)=\eta^{M_{j}} \varphi_{j}(\lambda) \neq 0 \quad \text { for } \eta>0, \lambda \in \bar{\Sigma}_{\phi} \backslash\{0\}
$$

with some $M_{j} \in[0, \infty)$,
(iv) $\psi_{j}$ is holomorphic in $B \Sigma_{\delta}^{n}$, continuous in $\overline{B \Sigma}_{\delta}^{n}$ and satisfies

$$
\psi_{j}(\eta z)=\eta^{L_{j}} \psi_{j}(z) \neq 0 \quad \text { for } \eta>0, z \in \overline{B \Sigma_{\delta}^{n} \backslash\{0\}}
$$

with some $L_{j} \in[0, \infty)$,
(v) for every $\gamma \in(0, \infty]$, the $\gamma$-principal part $\pi_{\gamma} P$ (see below) is not identical zero.
B.73. Definition ( $\gamma$-order and $\gamma$-principal part, cf. [DK13, cf. Definition 2.10]). Let $P \in S\left(\bar{\Sigma}_{\phi} \times\right.$ $\overline{B \Sigma}{ }_{\delta}^{n} \times K$ ) with representation (B.17). We put $P[\vartheta]:=P(\cdot, \cdot ; \vartheta)$ and $J[\vartheta]:=\left\{j \in J: \chi_{j}(\vartheta) \neq 0\right\}$. For $\vartheta \in K$ with $J[\vartheta] \neq \emptyset$ we define the $\gamma$-order

$$
d_{\gamma}(P[\vartheta]):= \begin{cases}\max _{j \in J[\vartheta]}\left\{\gamma M_{j}+N_{j} \max \{\gamma / 2,1\}+L_{j}\right\} & \text { for } \gamma \in(0, \infty), \\ \max _{j \in J[\vartheta]}\left\{M_{j}+N_{j} / 2\right\} & \text { for } \gamma=\infty .\end{cases}
$$

Let

$$
J_{\gamma}[\vartheta]:= \begin{cases}\left\{j \in J: \gamma M_{j}+N_{j} \max \{\gamma / 2,1\}+L_{j}=d_{\gamma}(P[\vartheta])\right\} & \text { for } \gamma \in(0, \infty), \\ \left\{j \in J: M_{j}+N_{j} / 2=d_{\infty}(P[\vartheta])\right\} & \text { for } \gamma=\infty .\end{cases}
$$

We define the $\gamma$-principal part

$$
\pi_{\gamma} P(\lambda, z ; \vartheta):=\left\{\begin{aligned}
\lim _{\eta \rightarrow \infty} \frac{P\left(\eta^{\gamma} \lambda, \eta z ; \vartheta\right)}{\eta^{d_{\gamma}(P[\vartheta])}}=\sum_{j \in J_{\gamma}[\vartheta]} \chi_{j}(\vartheta) \pi_{\gamma} \omega_{j}(\lambda, z) \varphi_{j}(\lambda) \varphi_{j}(z) & \text { for } \gamma \in(0, \infty), \\
\lim _{\eta \rightarrow \infty} \frac{P(\eta \lambda, z ; \vartheta)}{\eta^{d_{\infty}(P[\vartheta])}}=\sum_{j \in J_{\infty}[\vartheta]} \chi_{j}(\vartheta) \omega_{j}(\lambda, 0) \varphi_{j}(\lambda) \psi_{j}(z) & \text { for } \gamma=\infty,
\end{aligned}\right.
$$

where

$$
\pi_{\gamma} \omega_{j}(\lambda, z):= \begin{cases}\omega_{j}(0, z) & \text { for } \gamma \in(0,2) \\ \omega_{j}(\lambda, z) & \text { for } \gamma=2, \\ \omega_{j}(\lambda, 0) & \text { for } \gamma \in(2, \infty]\end{cases}
$$

B.74. Definition ( $N$-parabolic symbol, cf. [Kai12],[DK13]). The class $S_{N}\left(\bar{\Sigma}_{\phi} \times \overline{B \Sigma}_{\delta}^{n} \times K\right)$ of $N$-parabolic symbols consists of all functions $P \in S\left(\bar{\Sigma}_{\phi} \times \overline{B \Sigma_{\delta}^{n}} \times K\right)$ such that
(i) $\pi_{\gamma} P(\cdot, \cdot, \vartheta)$ is nontrivial for all $\gamma \in(0, \infty], \vartheta \in K$ and all $\vartheta \mapsto d_{\gamma}(P(\cdot, \cdot, \vartheta))$ are constant,
(ii) $P$ satisfies a two-sided estimate
$|P(\lambda, z, \vartheta)| \sim \sum_{(r, s) \in N_{V}}|\lambda|^{s}|z|^{r} \quad$ uniformly with respect to $(\lambda, z, \vartheta) \in \bar{\Sigma}_{\phi} \times{\overline{B \Sigma_{\delta}}}^{n} \times K$,
where $N_{V}$ denotes the set of the vertices of the Newton polygon associated to $d_{\gamma}(P)$.
B.75. Theorem (cf. [DK13, Theorem 2.56, Corollary 2.57]). The symbol $P \in S\left(\bar{\Sigma}_{\phi} \times \overline{B \Sigma}_{\delta}^{n} \times K\right)$ is $N$-parabolic if and only if

$$
\pi_{\gamma} P(\lambda, z, \vartheta) \neq 0 \quad \text { for all } \gamma \in(0, \infty], \lambda \in \bar{\Sigma}_{\phi} \backslash\{0\}, z \in \overline{B \Sigma_{\delta}^{n} \backslash\{0\}, \vartheta \in K . . . ~}
$$

The next result implies that every $N$-parabolic symbol induces a topological linear isomorphism with uniform bounds with respect to a compact parameter set $K$. We let $\mathcal{F}^{s}\left(\mathcal{K}^{r}\right)$ be as in Definition B. 62 and apply the joint functional calculus for $\left(\mu+\partial_{t}, \partial_{x_{1}}, \ldots, \partial_{x_{n}}\right)$ from Remark B. 65 .
B.76. Theorem (cf. [Kai12; DK13]). Let $P \in S_{N}\left(\bar{\Sigma}_{\phi} \times \overline{B \Sigma_{\delta}^{n}} \times K\right), \phi \in(\pi / 2, \pi), \delta>0$. Then there exists $\mu_{0} \geq 0$ such that $\vartheta \mapsto P\left(\mu_{0}+\cdot, \cdot, \vartheta\right)^{-1}, K \rightarrow H^{\infty}\left(\Sigma_{\phi} \times B \Sigma_{\delta}^{n}\right)$ is bounded. Moreover, for every such $\mu_{0}$, there exists $C>0$ such that for all $\mu \in\left[\mu_{0}, \infty\right)$ and $\vartheta \in K$, the realization

$$
P\left(\mu+\partial_{t}, \partial_{x_{1}}, \ldots, \partial_{x_{n}}, \vartheta\right): \bigcap_{(r, s) \in N_{V}(P)}\left(0 \mathcal{F}^{s^{\prime}+s}\left(\mathcal{K}^{r^{\prime}+r}\right)\right) \rightarrow{ }_{0} \mathcal{F}^{s^{\prime}}\left(\mathcal{K}^{r^{\prime}}\right)
$$

is an isomorphism and both $P\left(\mu+\partial_{t}, \partial_{x_{1}}, \ldots, \partial_{x_{n}}, \vartheta\right),\left[P\left(\mu+\partial_{t}, \partial_{x_{1}}, \ldots, \partial_{x_{n}}, \vartheta\right)\right]^{-1}$ are bounded by $C$.
B.77. Definition (cf. [DK13]). Let $\phi \in(\pi / 2, \pi), \delta \in(0, \pi / 2)$ and $K$ be a compact topological space. A function $\mathcal{L}: \bar{\Sigma}_{\phi} \times \overline{B \Sigma_{\delta}^{n}} \times K \rightarrow \mathbb{C}^{m \times m}$ is called an $N$-parabolic mixed-order system if
(i) $\mathcal{L}(\cdot, \cdot, \vartheta)$ is holomorphic and polynomially bounded, uniformly in $\vartheta \in K$,
(ii) $\operatorname{det} \mathcal{L}$ is $N$-parabolic in the sense of Definition B.74,
(iii) there are order functions $s_{j}$ and $t_{i}$ such that $s_{j}+t_{i}$ is an upper order function for $\mathcal{L}_{j, i}$ for all $j, i \in\{1, \ldots, m\}$,
(iv) $d_{\gamma}(\operatorname{det} \mathcal{L})=\sum_{j=1}^{m}\left(s_{j}(\gamma)+t_{i}(\gamma)\right)$ for all $\gamma \in(0, \infty]$.
B.78. Definition (cf. [DK13, Definition 2.78]). Let $\mu_{1}$ and $\mu_{2}$ be convex increasing order functions such that $\mu_{1}-\mu_{2}$ is an order function. Let $p \in(1, \infty)$. The scale

$$
\left(\mathcal{F}_{l}, \mathcal{K}_{l}\right) \in\left\{\left(B_{p p}, H_{p}\right),\left(H_{p}, B_{p p}\right)\right\}, \quad l \in\{0, \ldots, M\},
$$

is called $\left(\mu_{1}, \mu_{2}\right)$-admissible, if that there exists $k \in\{0, \ldots, M-1\}$ such that

$$
\begin{array}{r}
\left(\mathcal{F}_{0}, \mathcal{K}_{0}\right)=\cdots=\left(\mathcal{F}_{k}, \mathcal{K}_{k}\right)=\left(H_{p}, B_{p p}\right), \\
\left(\mathcal{F}_{k+1}, \mathcal{K}_{k+1}\right)=\cdots=\left(\mathcal{F}_{M}, \mathcal{K}_{M}\right)=\left(B_{p p}, H_{p}\right),
\end{array}
$$

and

$$
\begin{array}{ll}
\left(b_{k}\left(\mu_{2}\right), m_{k}\left(\mu_{2}\right)\right) \neq\left(b_{k+1}\left(\mu_{2}\right), m_{k+1}\left(\mu_{2}\right)\right) & \text { if } \mu_{1}-\mu_{2} \text { is convex, } \\
\left(b_{k}\left(\mu_{1}\right), m_{k}\left(\mu_{1}\right)\right) \neq\left(b_{k+1}\left(\mu_{1}\right), m_{k+1}\left(\mu_{1}\right)\right) & \text { if } \mu_{1}-\mu_{2} \text { is concave. }
\end{array}
$$

B.79. Theorem (cf. [DK13, Theorem 2.69, Corollary 2.80]). Let $X$ be a Banach space of class $\mathcal{H} \mathcal{T}$ with property ( $\alpha$ ). Let $\mathcal{L}: \bar{\Sigma}_{\phi} \times \overline{B \Sigma_{\delta}^{n}} \times K \rightarrow \mathbb{C}^{m \times m}$ be an $N$-parabolic mixed order system such that for each $i, j \in\{1, \ldots, m\}$, the order function $s_{j}+t_{i}$ is convex and increasing or concave and decreasing. Let $\rho \geq 0, s_{l}^{\prime} \geq 0, r_{l}^{\prime} \in \mathbb{R}, l \in\{0, \ldots, M\}$, such that

$$
\begin{aligned}
& \mu_{\mathbb{H}_{i}}(\gamma):=\max _{l}\left\{\left[s_{l}^{\prime}+m_{l}\left(t_{i}\right)\right] \gamma+r_{l}^{\prime}+b_{l}\left(t_{i}\right)\right\}, \quad \gamma \geq 0, \\
& \mu_{\mathbb{F}_{j}}(\gamma):=\max _{l}\left\{\left[s_{l}^{\prime}-m_{l}\left(s_{j}\right)\right] \gamma+r_{l}^{\prime}-b_{l}\left(s_{j}\right)\right\}, \quad \gamma \geq 0, \quad i, j \in\{1, \ldots, m\},
\end{aligned}
$$

are convex increasing order functions. Furthermore, let $p \in(1, \infty)$ and let the scale

$$
\left(\mathcal{F}_{l}, \mathcal{K}_{l}\right) \in\left\{\left(B_{p p}, H_{p}\right),\left(H_{p}, B_{p p}\right)\right\}, \quad l \in\{0, \ldots, M\},
$$

be ( $\mu_{\mathbb{H}_{i}}, \mu_{\mathbb{F}_{j}}$ )-admissible for all $i, j \in\{1, \ldots, m\}$ and let

$$
s_{l}^{\prime}>\max \left\{\max \left\{-m_{l}\left(t_{i}\right), m_{l}\left(s_{j}\right)\right\}: i, j \in\{1, \ldots, m\}\right\} \quad \text { for all } l \in\{0, \ldots, k\}
$$

with $k$ from Definition B.78. With ${ }_{0} \mathcal{F}_{l}^{s}\left(\mathcal{K}_{l}^{r}\right):={ }_{0} \mathcal{F}_{l}^{s}\left(\mathbb{R}_{+} ; \mathcal{K}_{l}^{r}\left(\mathbb{R}^{n} ; X\right)\right)$ we define the spaces

$$
\mathbb{H}_{i}:=\bigcap_{l=0}^{M} 0 \mathcal{F}_{l}^{s_{l}^{\prime}+m_{l}\left(t_{i}\right)}\left(\mathcal{K}_{l}^{r_{i}^{\prime}+b_{l}\left(t_{i}\right)}\right), \quad \mathbb{F}_{j}:=\bigcap_{l=0}^{M} 0 \mathcal{F}_{l}^{s_{l}^{\prime}-m_{l}\left(s_{j}\right)}\left(\mathcal{K}_{l}^{r_{l}^{\prime}-b_{l}\left(s_{j}\right)}\right)
$$

Then there exists $\tau_{0}>0$ such that for all $\tau \geq \tau_{0}$,

$$
\mathcal{L}\left(\tau+\mathcal{D}_{t}, \mathcal{D}_{x}, \vartheta\right): \prod_{i=1}^{m} \mathbb{H}_{i} \rightarrow \prod_{j=1}^{m} \mathbb{F}_{j}
$$

is a topological linear isomorphism and its inverse

$$
\left(\mathcal{L}\left(\tau+\mathcal{D}_{t}, \mathcal{D}_{x}, \vartheta\right)\right)^{-1}=\mathcal{L}^{-1}\left(\tau+\mathcal{D}_{t}, \mathcal{D}_{x}, \vartheta\right)
$$

is uniformly bounded with respect to $\vartheta \in K$.

## B.4. Analytic Nemytskiĭ operators

The nonlinear problem (T) contains nonlinear operators $(u, \pi, h, t, x) \mapsto F(u, \pi, h)(t, x)$ where $F(u, \pi, h)(t, x)$ only depends on the values of $(u, \pi, h)$ and its derivatives at $(t, x)$. These socalled Nemytskiř operators are studied in Section B.4. In order to prove the analyticity of a Nemytskiĭ operator, we define it in an open subset of a Banach space $X$ such that
(i) $X \hookrightarrow B U C(M ; \mathbb{K})$ for some metric space $M$,
(ii) $X$ is a Banach algebra with respect to pointwise multiplication,
(iii) we have $u^{-1} \in X$ for every $u \in X$ with $\inf \{|u(x)|: x \in M\}>0$.
B.80. Remark. Let $\Sigma \subset \mathbb{R}^{n}$ be a compact smooth hypersurface and let $\theta \in(0,1), p \in(1, \infty)$. Then we have

$$
\llbracket u v \rrbracket_{\theta, p} \leq\|u\|_{\infty} \llbracket v \rrbracket_{\theta, p}+\llbracket u \rrbracket_{\theta, p}\|v\|_{\infty} \quad \text { for } u, v \in W_{p}^{\theta}(\Sigma) \cap L_{\infty}(\Sigma)
$$

Therefore the spaces $W_{p}^{k+\theta}(\Sigma) \cap W_{\infty}^{k}(\Sigma)\left(k \in \mathbb{N}_{0}, \theta \in[0,1], p \in(1, \infty)\right)$ are multiplication algebras.

One more general result is given in
B.81. Lemma (cf. [Mey10, Lemma 1.3.19]). Let $\Omega \subset \mathbb{R}^{n}$ be a domain with compact smooth boundary, or $\Omega \in\left\{\mathbb{R}^{n}, \mathbb{R}_{+}^{n}\right\}$, or let $\Omega$ be the boundary of such a domain. Let further $X$ be a Banach space of class $\mathcal{H} \mathcal{T}$, let $s \in(0, \infty)$ and $p \in(1, \infty)$. Then there exists $C>0$ such that

$$
\|f g\|_{W_{p}^{s}(\Omega ; X)} \leq C\|f\|_{L_{\infty}(\Omega ; \mathcal{B}(X))}\|g\|_{W_{p}^{s}(\Omega ; X)}+C\|f\|_{W_{p}^{s}(\Omega ; \mathcal{B}(X))}\|g\|_{L_{\infty}(\Omega ; X)}
$$

for all $f \in W_{p}^{s}(\Omega ; \mathcal{B}(X)) \cap L_{\infty}(\Omega ; \mathcal{B}(X))$ and $g \in W_{p}^{s}(\Omega ; X) \cap L_{\infty}(\Omega ; X)$.
Hence $W_{p}^{s}(\Sigma)$ is a multiplication algebra for $s \in(0, \infty), p \in(1, \infty)$ with $s-(n-1) / p>0$.
B.82. Definition. Let $M$ be a measure space, let $X, Y$ be Banach spaces, $U \subset X$ be open and $f: M \times U \rightarrow Y$ be a Carathéodory function; that is,
(i) $u \mapsto f(x, u)$ is continuous for almost all $x \in M$,
(ii) $x \mapsto f(x, u)$ is measurable for all $u \in U$.

Then the map

$$
F: U^{M} \rightarrow Y^{M}, \quad u \mapsto[x \mapsto F(u)(x):=f(x, u(x))]
$$

is called the Nemytksǐ operator (of order zero) induced by $f$.
B.83. Definition. Let $M$ be a set and $\mathbb{K} \in\{\mathbb{R}, \mathbb{C}\}$. A Banach space $X \subset \mathbb{K}^{M}$ is called a multiplication algebra if pointwise multiplication

$$
X \times X \rightarrow X, \quad(u, v) \mapsto u v=[x \mapsto u(x) v(x)]
$$

is continuous. In this case there exists $C_{X}>0$ such that $\|u v\|_{X} \leq C_{X}\|u\|_{X}\|v\|_{X}$ for all $u, v \in X$.

We collect some information on analytic operators between open subsets of Banach spaces from Appell and Zabrejko [AZ90], Deimling [Dei10], and Zeidler [Zei86].

Let $X_{1}, \ldots, X_{k}, Y$ be Banach spaces over the same scalar field $\mathbb{K} \in\{\mathbb{R}, \mathbb{C}\}$. We say that a $k$-linear operator $T: X_{1} \times \cdots \times X_{k} \rightarrow Y$ is bounded, if there exists a number $C \geq 0$ such that $\left\|T\left(x_{1}, \ldots, x_{k}\right)\right\|_{Y} \leq C\left\|x_{1}\right\|_{X_{1}} \cdots\left\|x_{k}\right\|_{X_{1}}$ for all tuples $\left(x_{1}, \ldots, x_{k}\right)$. The infimum of such numbers $C$ is the norm of $T$, denoted by $\|T\|$ or $\|T\|_{\mathcal{B}^{k}\left(X_{1} \times \cdots \times X_{k} ; Y\right)}$. We put

$$
\mathcal{B}^{k}\left(X_{1} \times \cdots \times X_{k} ; Y\right):=\left\{T: X_{1} \times \cdots \times X_{k} \rightarrow Y: T \text { is } k \text {-linear and bounded }\right\} .
$$

For $k=0$, we put $X_{1} \times \cdots \times X_{k}:=\{0\}$ and $\mathcal{B}^{0}(\{0\} ; Y):=Y$. For a multi-index $\alpha \in \mathbb{N}_{0}^{k}$ we let $K=\left|\left\{j \in\{1, \ldots, k\}: \alpha_{j} \neq 0\right\}\right|$ and $\phi:\{1, \ldots, K\} \rightarrow\{1, \ldots, k\}$ be strictly increasing such that $\alpha_{\phi(j)} \neq 0$ for all $j$. Then we identify $X_{1}^{\alpha_{1}} \times \cdots \times X_{k}^{\alpha_{k}}$ with $X_{\varphi(1)}^{\alpha_{\varphi(1)}} \times \cdots \times X_{\varphi(K)}^{\alpha_{\varphi(K)}}$ and define

$$
\mathcal{B}^{\alpha}\left(X_{1}^{\alpha_{1}} \times \cdots \times X_{k}^{\alpha_{k}} ; Y\right):=\mathcal{B}^{|\alpha|}\left(X_{1}^{\alpha_{1}} \times \cdots \times X_{k}^{\alpha_{k}} ; Y\right)
$$

A map $T: X^{k}=X \times \cdots \times X \rightarrow Y$ is called symmetric if $T\left(x_{1}, \ldots, x_{k}\right)=T\left(x_{\sigma(1)}, \ldots, x_{\sigma(k)}\right)$ for all tuples $\left(x_{1}, \ldots, x_{k}\right)$ and all permutations $\sigma$ of $\{1,2, \ldots, k\}$. The map $T: X_{1}^{\alpha_{1}} \times \cdots \times X_{k}^{\alpha_{k}}(\alpha \in$ $\left.\mathbb{N}_{0}^{k}\right)$ is called partially symmetric, if it is symmetric with respect to every tuple $\left(x_{j, 1}, \ldots, x_{j, \alpha_{j}}\right) \in$ $X_{j}^{\alpha_{j}}$ when the other variables are fixed. We define

$$
\mathcal{B}_{\mathrm{sym}}^{k}\left(X^{k} ; Y\right):=\left\{T \in \mathcal{B}^{k}\left(X^{k} ; Y\right): T \text { is symmetric }\right\},
$$

$\mathcal{B}_{\text {sym }}^{\alpha}\left(X_{1}^{\alpha_{1}} \times \cdots \times X_{k}^{\alpha_{k}} ; Y\right):=\left\{T \in \mathcal{B}^{|\alpha|}\left(X_{1}^{\alpha_{1}} \times \cdots \times X_{k}^{\alpha_{k}} ; Y\right): T\right.$ is partially symmetric $\}$.
A map $M: X_{1} \times \cdots \times X_{k} \rightarrow Y$ is called monomial (operator) of degree $\alpha \in \mathbb{N}_{0}^{k}$ induced by the multilinear map $T \in \mathcal{B}_{\text {sym }}^{\alpha}\left(X_{1}^{\alpha_{1}} \times \cdots \times X_{k}^{\alpha_{k}} ; Y\right)$ if

$$
M\left(x_{1}, \ldots, x_{k}\right)=T\left(x_{1}^{\alpha_{1}}, \ldots, x_{k}^{\alpha_{k}}\right) \quad \text { for all }\left(x_{1}, \ldots, x_{k}\right) \in X_{1} \times \cdots \times X_{k}
$$

where $x_{j}^{\alpha_{j}}$ denotes the tuple $\left(x_{j}, \ldots, x_{j}\right) \in X_{j}^{\alpha_{j}}$.
A map $P: X_{1} \times \cdots \times X_{k} \rightarrow Y$ is called polynomial (operator) of degree lesser than or equal to $\alpha \in \mathbb{N}_{0}^{k}$, if there exist finitely many monomials $M^{(i)}: X_{1} \times \cdots \times X_{k} \rightarrow Y$ of degree $\alpha^{(i)} \in \mathbb{N}_{0}^{k}$ with $\alpha^{(i)} \leq \alpha$ such that $P=\sum_{i} M^{(i)}$.
B.84. Definition (Analytic operator). Let $X, Y$ be Banach spaces over $\mathbb{K} \in\{\mathbb{R}, \mathbb{C}\}$ and $U \subset X$ be open. We say that $F: U \subset X \rightarrow Y$ is ( $\mathbb{K}-$ ) analytic at $u \in U$, if there exists $r>0$ and symmetric operators $F_{k} \in \mathcal{B}_{\text {sym }}^{k}\left(X^{k} ; Y\right), k \geq 0$, such that

$$
\begin{equation*}
\sum_{k=0}^{\infty}\left\|F_{k}\right\|_{\mathcal{B}^{k}\left(X^{k} ; Y\right)}\|h\|_{X}^{k}<\infty \text { and } F(u+h)=\sum_{k=0}^{\infty} F_{k} h^{k} \quad \text { for all } h \in B_{r}^{X} \tag{B.18}
\end{equation*}
$$

A function is called analytic in $U$, if it is analytic at every point $u_{0} \in U$.
If $F$ is analytic at $u$, then $F$ is $C^{\infty}$ near $u$ and we have $F_{k}=F^{(k)}(u) / k!$. We next define

$$
r_{F}(u):=\min \left\{\operatorname{dist}_{X}(u, \partial U), C_{F}(u)^{-1}\right\}, \quad C_{F}(u):=\lim \sup _{k \rightarrow \infty}\left\|F^{(k)}(u) / k!\right\|_{\mathcal{B}^{k}\left(X^{k} ; Y\right)}^{1 / k} .
$$

Then the Taylor series $\sum_{k=0}^{\infty} F^{(k)}(u) h^{k} / k!$ converges in $Y$ and equals $F(u+h)$ for $\|h\|_{X}<r_{F}(u)$. If $\mathbb{K}=\mathbb{C}$, then a function is analytic in $U$ if and only if it is holomorphic in $U$. We then have

$$
F^{(k)}(u) h^{k}=\frac{k!}{2 \pi i} \int_{|\zeta|=\rho} \frac{F(u+\zeta h)}{\zeta^{k+1}} d \zeta \quad \text { for } 0<\rho\|h\|_{X}<r_{F}(u), k \in \mathbb{N}_{0}
$$

and Cauchy's estimates are valid:

$$
\left\|F^{(k)}(u)\right\|_{\mathcal{B}^{k}\left(X^{k} ; Y\right)} \leq \frac{k!}{\delta^{k}}\|F\|_{L_{\infty}\left(B_{\delta}(u) ; Y\right)} \quad \text { for } 0<\delta<r_{F}(u), k \in \mathbb{N}_{0}
$$

B.85. Remark (Chain rule). Let $U \subset \mathbb{R}^{n}$ and $V \subset \mathbb{R}^{k}$ be open subsets. The following chain rule is valid for sufficiently smooth maps $f: V \rightarrow \mathbb{R}$ and $u: U \rightarrow V$ at $x \in U$ (see [RS96, (5.2.1/6)]).

$$
\partial^{\alpha}(f \circ u)(x)=\sum_{j=1}^{|\alpha|} \sum c_{\alpha, j, \beta^{(1)}, \ldots, \beta^{(j)}} f^{(j)}(u(x)) \partial^{\beta^{(1)}} u(x) \cdots \partial^{\beta^{(j)}} u(x) .
$$

Here the second sum is taken over all multi-indices $\beta^{(1)}, \ldots, \beta^{(j)} \in \mathbb{N}_{0}^{n} \backslash\{0\}$ such that $\beta^{(1)}+$ $\cdots+\beta^{(j)}=\alpha$ and $c_{\alpha, j, \beta^{(1)}, \ldots, \beta^{(j)}}$ are some constants that do not depend on $f$ and $u$.

We next state Fraenkel's chain rule [Fra78, Formula A]. Let $X, Y, Z$ be Banach spaces, $U \subset X, V \subset Y$ be open sets and let $f \in C^{N}(V ; Z), u \in C^{N}(U ; V)$. Then $f \circ u \in C^{N}(U ; Z)$ and for $n \in\{1, \ldots, N\}, x \in U$ and $\left(v_{1}, \ldots, v_{n}\right) \in X^{n}$, we have

$$
\begin{equation*}
(f \circ u)^{(n)} v_{1} \cdots v_{n}=\sum_{j=1}^{n} \sum_{\beta, \sigma} \frac{f^{(j)} \circ u}{j!\beta!}\left(u^{\left(\beta_{1}\right)} v_{\sigma(1)} \cdots v_{\sigma\left(\beta_{1}\right)}\right) \cdots\left(u^{\left(\beta_{j}\right)} v_{\sigma\left(n-\beta_{j}+1\right)} \cdots v_{\sigma(n)}\right), \tag{B.19}
\end{equation*}
$$

where the sum $\sum_{\beta, \sigma}$ is taken over multi-indices $\beta \in \mathbb{N}^{j}$ such that $|\beta|=n$ and all $n$ ! permutations $\sigma$ of $\{1, \ldots, n\}$. It can be shown that

$$
\begin{equation*}
\sum_{\beta \in \mathbb{N}^{j},|\beta|=1} 1=\left|\left\{\beta \in \mathbb{N}^{j}:|\beta|=n\right\}\right|=\binom{n-1}{j-1} \quad \text { for } 1 \leq j \leq n . \tag{B.20}
\end{equation*}
$$

B.86. Corollary. If $F: U \subset X \rightarrow V \subset Y$ and $G: V \subset Y \rightarrow Z$ are analytic, then $G \circ F: U \subset X \rightarrow Z$ is analytic.

Proof. Let $M_{F}:=\sup _{j \in \mathbb{N}}\left\|F^{(j)}(x) / j!\right\|^{1 / j}, M_{G}:=\sup _{j \in \mathbb{N}}\left\|G^{(j)}(F(x)) / j!\right\|^{1 / j}$ for $x \in U$. The chain rule (B.19) yields

$$
\begin{aligned}
\frac{\left\|(G \circ F)^{(n)}(x)\right\|}{n!} & \leq M_{F}^{n} \sum_{j=1}^{n} M_{G}^{j}\left|\left\{\beta \in \mathbb{N}^{j}:|\beta|=n\right\}\right| \\
& \leq M_{F}^{n} \sum_{j=1}^{n} M_{G}^{j}\binom{n-1}{j-1}=M_{F}^{n} M_{G}\left(1+M_{G}\right)^{n-1} .
\end{aligned}
$$

Therefore $\sum_{n \geq 0}(G \circ F)^{(n)}(x) h^{n} / n!$ converges for $\|h\|<\left[M_{F}\left(1+M_{G}\right)\right]^{-1}$. The representation

$$
G(F(x+h))=\sum_{j \geq 0} \frac{G^{(j)}(F(x))}{j!}(F(x+h)-F(x))^{j}=\sum_{j \geq 0} \frac{G^{(j)}(F(x))}{j!}\left(\sum_{l \geq 1} \frac{F^{(l)}(x) h^{l}}{l!}\right)^{j}
$$

is valid for small $h$. As in the proof of [Fra78, Formula A] we rewrite the right-hand side as

$$
G(F(x))+\sum_{n \geq 1} \sum_{j=1}^{n} \sum_{\beta \in \mathbb{N} j,|\beta|=n} \sum_{\sigma} \frac{G^{(j)}(F(x))}{j!\beta!}\left(F^{\left(\beta_{1}\right)}(x) h^{\beta_{1}}\right) \cdots\left(F^{\left(\beta_{j}\right)}(x) h^{\beta_{j}}\right) .
$$

By the chain rule, $G(F(x+h))$ coincides with its Taylor series for small $h$. Therefore $G \circ F$ is analytic at $x$.
B.87. Proposition. Let $M$ be a metric space, $X \hookrightarrow B U C(M ; \mathbb{K})$ be a multiplication algebra, $U \subset \mathbb{K}^{m}$ $(m \in \mathbb{N})$ be open and $f: U \subset \mathbb{K}^{m} \rightarrow \mathbb{K}$ be analytic. Define

$$
\begin{aligned}
\mathcal{U} & :=\left\{u \in X^{m}: u(M) \subset U, \inf r_{f}(u(M))>0, f \circ u \in X, C_{F}(u)<\infty\right\}, \\
C_{F}(u) & =\lim \sup _{j \rightarrow \infty}\left\|\partial^{j} f \circ u / j!\right\|_{\mathcal{B}^{j}\left(\left(X^{m}\right)^{j} ; X\right)}^{1 / j} \text { for } u \in X^{m} \text { with } u(M) \subset U .
\end{aligned}
$$

Then $F: u \mapsto f \circ u, \mathcal{U} \subset X^{m} \rightarrow X$ is analytic.

Proof. For all $u \in X^{m}$ with $u(M) \subset U$ and all $h \in X^{m}$ with $\|h\|_{X^{m}}<C(u)^{-1}$, the Taylor series $\sum_{j \geq 0} F^{(j)}(u) h^{j} / j$ ! converges in $X$. Since $f$ is analytic on $u(M) \subset U$, we obtain the representation

$$
F(u+h)(x)=f(u(x)+h(x))=\sum_{j \geq 0} \partial^{j} f(u(x)) h(x)^{j} / j!\quad \text { for } x \in M,
$$

provided $|h(x)|<r_{f}(u(x))$. From $X \hookrightarrow B U C(M ; \mathbb{K})$ we infer that

$$
|h(x)| \leq\|h\|_{B U C(M)^{m}} \leq\|I\|_{X \rightarrow B U C(M)}\|h\|_{X^{m}}<\inf r_{f}(u(M)) \leq r_{f}(u(x)) \quad \text { for all } x \in M,
$$ for $h \in X^{m}$ with $\|h\|_{X^{m}}<\|I\|_{X \rightarrow B U C(M)}^{-1} \inf r_{f}(u(M))$. Therefore $F$ is analytic at $u$ with

$$
r_{F}(u) \geq \min \left\{\|I\|_{X \rightarrow B U C(M)}^{-1} \inf r_{f}(u(M)), C_{F}(u)^{-1}\right\}>0 .
$$

B.88. Proposition. Let $M$ be a metric space, $X \hookrightarrow B U C(M ; \mathbb{K})$ be a multiplication algebra and $m \in \mathbb{N}$. Then the map $A \mapsto[A(\cdot)]^{-1},\left\{A(\cdot) \in X^{m \times m}:[A(\cdot)]^{-1} \in X^{m \times m}\right\} \rightarrow X^{m \times m}$ is analytic.
Proof. Let $U:=\left\{A \in \mathbb{K}^{m \times m}: A\right.$ is invertible $\}$ and $f: A \mapsto A^{-1}, U \subset \mathbb{K}^{m \times m} \rightarrow \mathbb{K}^{m \times m}$. Then

$$
\begin{equation*}
\partial^{j} f(A)\left(B_{1}, \ldots, B_{k}\right)=(-1)^{j} \sum_{\sigma}\left(\prod_{i=1}^{j}\left(A^{-1} B_{\sigma(i)}\right)\right) A^{-1} \quad \text { for } j \in \mathbb{N}_{0}, B \in \mathbb{K}^{m \times m} \tag{B.21}
\end{equation*}
$$

where the sum is taken over all $j$ ! permutations $\sigma$ of $\{1, \ldots, j\}$. Hence $C_{f}(A)=\left|A^{-1}\right|$ for $A \in U$. For $A \in U$ and $B \in \mathbb{K}^{m \times m}$ with $|B|<\left|A^{-1}\right|^{-1}$ we have $A+B=A\left(I+A^{-1} B\right) \in U$ and thus $\operatorname{dist}_{\mathbb{K}^{m \times m}}(A, \partial U) \geq\left|A^{-1}\right|^{-1}$. Therefore $f$ is analytic with $r_{f}(A)=\left|A^{-1}\right|^{-1}$.

The space $X^{m \times m}$ with norm $\|\cdot\|_{X}:=\|\cdot\|_{X^{m \times m}}$ is a Banach algebra with respect to pointwise matrix multiplication and there exists $C_{X}>0$ such that $\|A B\|_{X} \leq C_{X}\|A\|_{X}\|B\|_{X}$. Hence

$$
\left\|\partial^{j} f(A(\cdot))\right\|_{X} \leq C_{X}^{2 j}\left\|A^{-1}\right\|_{X}^{j+1} j!\quad \text { for } j \in \mathbb{N}_{0}, A \in X^{m \times m} \text { with } A^{-1} \in X^{m \times m} .
$$

Proposition B. 87 with $C_{F}(A)=C_{X}^{2}\left\|A^{-1}\right\|_{X}<\infty$ yields the assertion.
B.89. Proposition. Let $M$ be a metric space and $X \hookrightarrow B U C(M ; \mathbb{K})$ be a multiplication algebra. Then the map $F: u \mapsto u(\cdot)^{1 / 2},\left\{u \in X: \inf _{M} \operatorname{dist}\left(u(\cdot), \mathbb{R}_{-}\right)>0, u^{1 / 2}, u^{-1} \in X\right\} \rightarrow X$ is analytic.
Proof. The map $f: z \mapsto z^{1 / 2}, \mathbb{K} \backslash \mathbb{R}_{-} \rightarrow \mathbb{K}$ is analytic with $C_{f}(z)=|z|$ and $r_{f}(z)=\operatorname{dist}\left(z, \mathbb{R}_{-}\right)$ and

$$
\left\|\frac{\partial^{k} f \circ u}{k!}\right\|_{X}^{1 / k}=\left|\frac{1}{2 \cdot 1} \cdot\left(-\frac{1}{2 \cdot 2}\right) \cdots\left(-\frac{2 k-3}{2 \cdot k}\right)\right|^{1 / k}\left\|u^{-k+1 / 2}\right\|_{X}^{1 / k} \leq c_{k} C_{X}\left\|u^{1 / 2}\right\|_{X}^{1 / k}\left\|u^{-1}\right\|_{X}
$$

with $\lim _{k \rightarrow \infty} c_{k}=1$. Hence $C_{F}(u) \leq C_{X}\left\|u^{-1}\right\|_{X}$ and Proposition B. 87 yields analyticity.
B.90. Lemma. Let $\Sigma \subset \mathbb{R}^{n}(n \geq 2)$ be a compact smooth hypersurface and let $s \in[0, \infty), p \in(1, \infty)$, and $m \in \mathbb{N}$. Then the pointwise matrix inversion operator

$$
A(\cdot) \mapsto A(\cdot)^{-1}, \quad\left\{A \in\left(W_{p}^{s} \cap C\right)\left(\Sigma ; \mathbb{K}^{m \times m}\right):\left\|A(\cdot)^{-1}\right\|_{\infty}<\infty\right\} \rightarrow\left(W_{p}^{s} \cap C\right)\left(\Sigma ; \mathbb{K}^{m \times m}\right)
$$

and the pointwise square root operator

$$
u(\cdot) \mapsto \sqrt{u(\cdot)}, \quad\left\{u \in\left(W_{p}^{s} \cap C\right)(\Sigma): \inf _{\Sigma} \operatorname{dist}\left(u(\cdot), \mathbb{R}_{-}\right)>0\right\} \rightarrow\left(W_{p}^{s} \cap C\right)\left(\Sigma ; \mathbb{K}^{m \times m}\right)
$$

are analytic.
Proof. The matrix inversion operator is well-defined, since we can control $A^{-1}$ in the $W_{p}^{s}$-norm by means of the identity (B.21), Lemma B. 81 and the inequalities

$$
\left\|A^{-1}\right\|_{p} \leq|\Sigma|^{1 / p}\left\|A^{-1}\right\|_{\infty}, \quad \llbracket A^{-1} \rrbracket_{\theta, p} \leq\left\|A^{-1}\right\|_{\infty}^{2} \llbracket A \rrbracket_{\theta, p} .
$$

Then Proposition B. 88 yields analyticity. The second assertion follows from the estimates

$$
\|\sqrt{u}\|_{p} \leq|\Sigma|^{1 / p}\|u\|_{\infty}^{1 / 2}, \quad \llbracket \sqrt{u} \rrbracket_{\theta, p} \leq\left(2 \inf _{\Sigma}|u|^{1 / 2}\right)^{-1} \llbracket u \rrbracket_{\theta, p} .
$$

## B.5. Computation of the boundary symbol

For the derivation of the boundary symbol in Section 3.1 .1 we have employed the identity (3.13) on page 58. This identity can be checked with the software Maxima [Max] with the following source code.

```
gam1: %omega[1]+z;
gam2: %omega[2]+z;
%alpha[1]: %mu[1]*%omega[1]*gam1;
%alpha[2]: %mu[2]*%omega[2]*gam2;
%Omega[p]: %mu[1] *%omega[1]*gam1+%mu[2] *%omega[2] *gam2;
%Omega[s]: %mu[s]*z^2
c[6]*%mu[1]*%omega[1]+c[6] *%mu[2]*%omega[2]+;
Lw1: c[5]*z*%alpha[1]*(%mu[1]-%mu[2])
    + c[6]*%mu[1]*%omega[1]*%alpha[2]
    + c[6]*%mu[2]*%omega[2]*%alpha[1];
Lw2: c[5]*z*%alpha[2]*(%mu[1]-%mu[2])
    - c[6]*%mu[1]*%omega[1]*%alpha[2]
    - c[6]*%mu[2]*%omega[2]*%alpha[1];
Lq: c[5]*z*(%mu[1]-%mu[2])
    + c[6]*%mu[2]*%omega[2] - c[6]*%mu[1]*%omega[1];
ratvars(z,%mu[1],%mu[2],%omega[1],%omega[2],%lambda);
B: matrix([-%omega[2]*%Omega[s]*%Omega[p]
            +z*Lw2-z^2*%Omega[p]*%lambda[s]*%omega[2],
        Lw1*z, -c[1]*z^4, Lq*z^2],
    [%omega[2], %omega[1], 0, 0],
    [-%alpha[2], -%alpha[1], %lambda, -z],
    [(2*%mu[2]-c [2])*%omega[2]*%Omega[p] *%Omega[s]
                -2*%theta[3]*z*Lw2
            +2*%theta[3]*%lambda[s]*%omega[2]*z^ 2*%Omega[p],
        2*%mu[1]*%omega[1]*%Omega[p] *%Omega[s]
            -2*%theta[3]*Lw1*z,
            (c[%sigma]+%theta[4])*%Omega[s]*z^2
                +2*c[1]*%theta[3]*z^4,
        %Omega[s]*%Omega [p]
            -2*%theta[3]*z^2*Lq]);
detB: expand(determinant(B));
factor(detB);
P: expand(divide(detB,%omega[1]*%omega[2]
    * (%mu[2] *%omega [2] *z+%mu[1] *%omega[1]*z
    +%mu[2] *%omega[2]^2+%mu[1]*%omega[1]^2)
    * (%mu[s]*z^2+c[6]*%mu[2]*%omega[2]
    +c[6]*%mu[1]*%omega[1]))[1]);
Q: -scsimp(P,
        (%theta[4]+c[%sigma])*(%mu[s]+%lambda[s])
    +c[1]*c[2]+2*c[1]*%theta[3]=%beta[s]*d,
    %mu[s]+%lambda[s]=%beta[s]);
```


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## List of symbols

## Numbers

$\mathbb{C}$ the complex numbers. 11
$\mathbb{C}_{+}\{z \in \mathbb{C}: \operatorname{Re} z \geq 0\}$, the complex numbers with non-negative real part. 11
$\mathbb{K}$ either $\mathbb{R}$ or $\mathbb{C} .11$
$\mathbb{N}\{1,2, \ldots\}$, the positive integers. 11
$\mathbb{N}_{0}\{0,1,2, \ldots\}$, the non-negative integers. 11
$\mathbb{R}$ the real numbers. 11
$\mathbb{R}_{+}[0, \infty)$, the non-negative real numbers. 11
$\mathbb{R}_{-}(-\infty, 0]$, the non-positive real numbers. 11
$\mathbb{Z}\{\ldots,-1,0,1, \ldots\}$, the integers. 11

## Surface differential operators

$\widetilde{\Delta}_{\Gamma} g^{\alpha \beta} \widetilde{\nabla}_{\alpha} \widetilde{\nabla}_{\beta}$, Laplace-Beltrami operator for tangential vector fields. 141, 164
$D_{\Gamma}(u) \operatorname{sym}\left(P_{\Gamma}\left[\nabla_{\Gamma} u\right] P_{\Gamma}\right)$, surface rate-ofstrain tensor for $u: \Gamma \rightarrow \mathbb{R}^{n}$. 19, 140
$\operatorname{div}_{\Gamma} S\left[\partial_{\alpha} S\right] \tau^{\alpha}$, surface divergence of a symmetric tensor $S^{\alpha \beta} \tau_{\alpha} \otimes \tau_{\beta} .140$
$\operatorname{div}_{\Gamma} u \tau^{\alpha} \cdot \partial_{\alpha} u$, surface divergence of a vector field $u=v^{\alpha} \tau_{\alpha}+w \nu .140$
$\nabla_{\Gamma} \tau_{\Gamma}^{j} \otimes \partial_{j}$, surface gradient of $\Gamma .140$
$\widetilde{\nabla}_{\Gamma}$ covariant derivative of $\Gamma .139,140$
Symbols related to basic function spaces
$B_{p q}^{s}$ Besov space of order $s$ with exponents $p$ and $q .144$
$B U C^{k}$ space of all bounded and uniformly continuous functions with bounded and uniformly continuous derivatives up to order $k .143$
$C(U ; V)$ space of all continuous functions $f: U \subset X \rightarrow V \subset Y .11$
$C_{0}(J ; X)$ Banach space of all $u \in C(J ; X)$ such that $\|u(t)\|_{X} \rightarrow 0$ as $t \rightarrow \infty .143$
$C^{k-}$ space of all $u \in C^{k-1}$ such that $\nabla^{k-1}$ is locally Lipschitz. 143
$C^{k, \alpha}$ Hölder space of all $u \in C^{k}$ such that $\nabla^{k} u \in C^{0, \alpha} .143$
$\mathcal{D}(\Omega ; X)$ space of test functions in $\Omega .143$
$\mathcal{D}^{\prime}(\Omega ; X)$ space of distributions in $\Omega .143$
$F_{p q}^{s}$ Triebel-Lizorkin space of order $s$ with exponents $p, q .144$
$\mathcal{H}(\Omega ; Y)$ space of all holomorphic $Y$-valued functions on $\Omega .164$
$\mathcal{H}^{\infty}(\Omega ; Y)$ space of all bounded holomorphic $Y$-valued functions on $\Omega .164$
$\mathcal{H}_{0}^{\infty}(\Omega ; Y)$ subspace of $\mathcal{H}^{\infty}(\Omega ; Y)$ with polynomial decay near 0 and $\infty .164$
$\mathcal{H}_{P}(\Omega ; Y)$ subspace of $\mathcal{H}(\Omega ; Y)$ with polynomial growth near 0 and $\infty .165$
$\dot{H}_{p}^{k}$ homogeneous Sobolev space of order $k$ with exponent $p .40$
$\dot{\mathcal{H}}_{p}^{k}$ semi-normed version of $\dot{H}_{p}^{k} .23,40$
$\hat{H}_{p}^{-1}$ dual space of $\dot{H}_{p^{\prime}}^{1} .25,40$
$H_{p}^{s}$ Bessel-potential space of order $s$ with exponent $p .143$
$L_{p}$ Lebesgue space with exponent $p .143$
$\mathcal{P}_{k}$ space of all polynomials. 151
$\mathcal{P}_{k}$ space of all polynomials of degree lesser or equal than $k .151$
$\mathcal{S}$ space of rapidly decreasing functions. 143
$\mathcal{S}^{\prime}$ space of tempered distributions. 143
$\mathcal{S}_{0}$ subspace of all $\varphi \in \mathcal{S}\left(\mathbb{R}^{n}\right)$ with $\left(\partial^{\alpha} \mathcal{F} \varphi\right)(0)=0$ for all $\alpha \in \mathbb{N}_{0}^{n} .151$
$\mathcal{S}_{0}^{\prime}$ dual space of $\mathcal{S}_{0}$, equivalence classes of $\mathcal{S}^{\prime}$ modulo polynomials. 151
$\Sigma_{\theta}\left\{r e^{i \varphi}: r \in(0, \infty), \varphi \in(-\theta, \theta)\right\}$, open sector. 152, 161, 164
$W_{p}^{s}$ Sobolev-Slobodeckiĭ space of order $s$ with exponent $p$. 143, 144

## Special function spaces

$\mathbb{E}_{\partial \Theta}$ space of interface regularity for the Jacobian of the normal-preserving map. 99
$\tilde{\mathbb{E}}_{\partial \Theta}$ a larger space than $\mathbb{E}_{\partial \Theta} .113,115$
$\mathbb{E}_{h}$ space of time-dependent height functions $h .99$
$\tilde{\mathbb{E}}_{h}$ a larger space than $\mathbb{E}_{h} .112$
$\mathbb{E}_{\Theta}$ space of interior regularity for the normal-preserving map. 102
$\mathbb{E}_{u, v, w, \llbracket \mu \partial_{\nu} w \rrbracket}$ space of all $u \in \mathbb{E}_{u}$ with $\left.P_{\Sigma} u\right|_{\Sigma} \in \mathbb{E}_{v}, w:=\left.\nu_{\Sigma} \cdot u\right|_{\Sigma} \in \mathbb{E}_{w},\left.\partial_{\nu} w_{ \pm}\right|_{\Sigma} \in$ $\mathbb{G}_{w} .69$
$\mathbb{F}_{d, \Sigma}$ space of all $f_{d} \in \mathbb{F}_{d}$ with $f_{d, \pm} \mid \Sigma \in \mathbb{G}_{w}$. 69, 123
$\mathbb{G}_{v}$ space for the tangential interface stress balance. 115
$\mathbb{G}_{w}$ space for the normal interface stress balance. 115
$\mathcal{P}_{M, T}$ parameter set for perturbed version of problem (MP). 82
$\mathcal{P}_{M, T, \eta, R}$ parameter set for problem (PL). 71
$\mathcal{U}_{h}$ set of certain $h: J \times \Sigma \rightarrow \mathbb{R}$ for which $\Theta_{h}$ is a diffeomorphism. 101, 102
$\mathcal{U}_{h_{0}}$ set of certain $h_{0}: \Sigma \rightarrow \mathbb{R}$ for which $\Theta_{h_{0}}$ is a diffeomorphism. 101, 102
Linear operators
$D_{A}(\alpha, p)$ fractional domain of $A .154$
$\mathcal{H}(X)$ class of linear operators $A: D(A) \subset$ $X \rightarrow X$ with bounded $\mathcal{H}^{\infty}$-calculus. 153
$\mathcal{P}_{1}\left(X_{1}, X_{0} ; M, \vartheta\right)$ class of invertible linear operators $A: X_{1} \rightarrow X_{0}$ satisfying the estimate $(1+|\lambda|)\left\|(\lambda+A)^{-1}\right\| \leq K$ for $|\arg \lambda| \leq$ ข. 154

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## Veröffentlichungen

Optimal regularity and long-time behavior of solutions for the Westervelt equation, Applied Mathematics and Optimization 64, 257-271, 2011 (mit Mathias Wilke)
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Optimal regularity and exponential stability for the Blackstock-Crighton equation in $L_{p}$-spaces with Dirichlet and Neumann boundary conditions, Journal of Evolution Equations, 37 S., 2016 (mit Rainer Brunnhuber)

## Eidesstattliche Erklärung

Hiermit erkläre ich an Eides statt, dass ich die vorliegende Arbeit selbständig und ohne fremde Hilfe verfasst habe, keine anderen als die von mir angegebenen Quellen und Hilfsmittel benutzt habe und die den benutzen Werken wörtlich oder inhaltlich entnommenen Stellen als solche kenntlich gemacht habe.

Halle (Saale), 28. November 2015

## Stefan Meyer

