# Well-posedness and causality of the non-Newtonian Navier-Stokes equations on 3-dimensional Riemannian $C^{1,1}$-manifolds with respect to strong, local-in-time solutions 

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## Preface

The Navier-Stokes equations have been studied in a variety of cases for their importance in physics and engineering. Yet, it seems, especially non-Newtonian fluids create a lot of problems for the sheer diversity of viscous properties a fluid may have, though the Newtonian case, too, still holds a tight grasp on many interesting questions. Not even well-posedness of strong solutions global in time is known; in fact, this is a Millennium problem of the Clay Mathematics Institute ([5]). Furthermore, causality has never been addressed to my knowledge. In my Dipl. Math. thesis I considered a unified approach for large classes of viscosities in the case of $C^{\infty}$-manifolds without boundary. In these notes, we will expand these findings to include many results interesting for applications, that is, we will consider 3-dimensional $C^{1,1}$-manifolds with or without boundary - the largest class of conceivable Riemannian manifolds. If the manifold has a boundary, then we will consider Dirichlet and Neumann boundary conditions. Fluids are assumed to be incompressible and isotropic.

This approach is largely influenced by Rainer Picard who developed a unified Hilbert space approach for well-posedness and causality of (linear) partial differential equations ([15]). It seems that this unified approach works perfectly for linear partial differential equations encountered in mathematical physics and, hence, he has studied many models as examples; the Stokes equation was one of them. In fact, he observed that it is possible to generalize the viscosity term which I will use as well. The other highly interesting question is, how little regularity of the manifold can we ask for and still obtain local well-posedness of strong solutions. Aside from a Rellich-Kondrachov type condition, $C^{1,1}$-manifolds are as low in regularity as we are able to reduce the problem without having to argue with very special assumptions on the manifold. This is a rather interesting topic in itself and far beyond the scope of these notes. However, it is interesting to keep in mind that this is precisely the lower end of regularity most problems in mathematical physics can support because most problems in mathematical physics contain an operator (here, the Stokes operator) which is a relative of the "mother operator" (cf., [14])

$$
A=\left(\begin{array}{cc}
0 & -\nabla^{*} \\
\nabla & 0
\end{array}\right)
$$

with a suitable domain in the $L_{2}$ space generated by of the set of Lipschitz continuous covariant tensors. Here, $\nabla$ denotes the co-variant derivative $d \otimes$. In other words, the $C^{1,1}$ condition is necessary for the domain of $A$ to be sufficiently rich and, hence, a minimal condition for the problem to be meaningful.

These notes are structured in two parts. In part 1 we will discuss the (functional) analytic background and in part 2 we will use the physical textbook formulation of Navier-Stokes for incompressible fluids "re-modeling" them to find an abstract non-linear Cauchy problem which we are going to solve afterwards.

I will start with chapter 1 which seems rather random but this is a rather intriguing interpretation of the projection theorem and yields many powerful applications. Ever since Rainer Picard has introduced this to me, I have used it quite
extensively and often subtly hidden. Hence, I included this chapter for the reader to see some subtle applications of the projection theorem in proving some important theorems. These methods will be used everywhere.

The content of chapter 2 has also been taught to me by Rainer Picard and it will be used throughout the notes as the spaces we are working on are tensor products and the operators are mostly of the form $1 \otimes A$ or $A \otimes 1$ even though we will only write $A$ in both cases due to the theory explored in chapter 2. More extensive representations of the topic can be found in chapter 1 of $[\mathbf{1 5}]$ and the appendix A of [20].

The $L_{p}$-spaces used in these notes will be introduced in chapter 3 . This is properly standard Lebesgue theory and nothing special; however, for the sake of notation and completeness and since it is not very common to see these $L_{p}$ spaces, I have added this chapter. Furthermore, as we are on a $C^{1,1}$-manifold, it is not at all obvious why Sobolev spaces of higher order should exist. This is subject of the last section of chapter 3 though readers interested in a more detailed account should refer to chapter 2 of [15].

Chapter 4 contains the analytic implicit function theorem. Since the usual approach to the theorem is rather abstract, I chose to adapt a prove that was shown to me by Jürgen Voigt. This proof first proves the implicit function theorem and regularity up to $C^{\infty}$ in a constructive way (which is important since it makes a major difference if we are able to construct solutions of the Navier-Stokes equations or not) such that any second year mathematics student should be able to understand it, if you explain them a few facts about Banach spaces and linear operators. Other than that it is a direct generalization of the finite dimensional theorem. In order to obtain analyticity, we then have to pull out the big guns. The proof shown here is an adaptation of the one shown in [3].

Finally, chapter 5 concludes the analytical background part with some facts about Fredholm operators. These will come in handy as the linearized NavierStokes operator is a Fredholm operator and they will allow a major shortcut in proving that the linearized Navier-Stokes operator is an isomorphism (needed for the implicit function theorem).

Part 2 starts with chapter 6 on modeling Navier-Stokes. Here, we will start from the physical equations of fluid dynamics and "re-model" them into the partial differential equation we are going to solve after identifying the spaces to work in. At this point, the Rellich-Kondrachov condition becomes vital as the equation would be ill-stated otherwise. However, I will not go into detail of the physical implications of the changed viscosity term and, thus, non-Newtonian fluids since this would fill at least a book (cf., e.g., [1]).

Chapter 7 will is a rather short one though important as I think the content should be part of anyone's vocabulary working with non-linear partial differential equations. Chapter 7 contains the framework of the proof, that is, how to construct solutions assuming all theorems are applicable. This will leave us with two holes to fill. First we will have to show that the linearized Navier-Stokes operator is an isomorphism. This will be addressed in chapter 8 by showing that it is an injective Fredholm operator of index zero. This approach is also a standard approach and has previously been applied to Navier-Stokes successfully in multiple special cases (cf., e.g., [2]).

As a corollary we will, furthermore, obtain injectivity of the Navier-Stokes operator which will be used in the causality proof in chapter 9 . This is the second gap to fill in order to make the construction of chapter 7 work. Here, I had to generalize the concept of causality (cf., [15]) to non-linear relations which is not as straight forward as it appears. It turns out there are two slightly different notions
of causality here - weak and strong causality - both having physical meaning. In fact, strong causality is what you want in a classical deterministic theory (such as the Navier-Stokes system) whereas weak causality is the most we can hope for in a quantum system with non-vanishing vacuum fluctuations. For the proof to work, weak causality would suffice but for physics to work strong causality is needed and, as it turns out, physics is fine; we can prove strong causality of the NavierStokes equations. Finally, we can state the well-posedness and causality theorem for Navier-Stokes of non-Newtonian fluids for strong solutions local in time.

At last, I would like to thank Ralph Chill and Rainer Picard for uncountable discussions while supervising my Dipl. Math. thesis, thus, making these notes possible. Furthermore, I thank Ralph Chill, Rainer Picard, and Jürgen Voigt for introducing me to most of theory used in these notes. Finally, I want to thank my parents for their support and patience.

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## Part 1

## Analytic Background

## CHAPTER 1

## Some remarks on the Projection Theorem

We will begin by having a closer look at the projection theorem and some interesting applications as this is used throughout these notes without further mentioning. Let $H_{0}$ and $H_{1}$ be Hilbert spaces over $\mathbb{K} \in\{\mathbb{R}, \mathbb{C}\}$ and $A \subseteq H_{0} \oplus H_{1}$ a closed linear relation. By $-A$ we denote the operational minus

$$
-A:=\left\{(u,-v) \in H_{0} \oplus H_{1} ;(u, v) \in A\right\} .
$$

Then the adjoint relation $A^{*}$ is defined by

$$
\begin{aligned}
A^{*} & :=-\left(A^{-1}\right)^{\perp} \\
& =-\left(\left\{(u, v) \in H_{0} \oplus H_{1} ; \quad(u, v) \in A\right\}^{-1}\right)^{\perp} \\
& =-\left\{(v, u) \in H_{1} \oplus H_{0} ; \quad(u, v) \in A\right\}^{\perp} \\
& =-\left\{(y, x) \in H_{1} \oplus H_{0} ; \forall(u, v) \in A:\langle(y, x),(v, u)\rangle_{H_{1} \oplus H_{0}}=0\right\} \\
& =-\left\{(y, x) \in H_{1} \oplus H_{0} ; \forall(u, v) \in A:\langle y, v\rangle_{H_{1}}+\langle x, u\rangle_{H_{0}}=0\right\} \\
& =\left\{(y,-x) \in H_{1} \oplus H_{0} ; \forall(u, v) \in A:\langle y, v\rangle_{H_{1}}+\langle x, u\rangle_{H_{0}}=0\right\} \\
& =\left\{(y, x) \in H_{1} \oplus H_{0} ; \forall(u, v) \in A:\langle y, v\rangle_{H_{1}}+\langle-x, u\rangle_{H_{0}}=0\right\} \\
& =\left\{(y, x) \in H_{1} \oplus H_{0} ; \forall(u, v) \in A:\langle x, u\rangle_{H_{0}}=\langle y, v\rangle_{H_{1}}\right\} .
\end{aligned}
$$

The last line shows that this is the definition we want, as well as,

$$
A^{*}=-\left(A^{-1}\right)^{\perp}=\left(-A^{-1}\right)^{\perp}=\left((-A)^{-1}\right)^{\perp}=-\left(A^{\perp}\right)^{-1}=\left(-A^{\perp}\right)^{-1}=\left((-A)^{\perp}\right)^{-1}
$$

i.e.,,$-^{\perp}$, and ${ }^{-1}$ commute. Note that $\left({ }^{*},{ }^{*}\right)$ is a Galois connection on the set of linear relations in $H_{0} \oplus H_{1}$ with the inclusion as partial ordering. For $U \subseteq H_{0}$ and $V \subseteq H_{1}$ we will use the notation

$$
\begin{aligned}
& {[V] A:=\{(u, v) \in A ; v \in V\} \text { the pre-set of } V \text { with respect to } A} \\
& A[U]:=\{(u, v) \in A ; u \in U\} \text { the post-set of } U \text { with respect to } A .
\end{aligned}
$$

Note that if $A$ was a function one would call them pre-image and image.
Theorem 1.1 (Projection Theorem). Let $H_{0}$ and $H_{1}$ be Hilbert spaces and $A \subseteq H_{0} \oplus H_{1}$ a closed linear relation. Then we obtain the following orthogonal decompositions.

$$
\begin{aligned}
& H_{0}=[\{0\}] A \oplus \overline{A^{*}\left[H_{1}\right]} \\
& H_{1}=[\{0\}] A^{*} \oplus \overline{A\left[H_{0}\right]}
\end{aligned}
$$

Proof.

$$
\begin{aligned}
y \in \overline{A\left[H_{0}\right]} & \Leftrightarrow y \perp A\left[H_{0}\right] \\
& \Leftrightarrow \forall(u, v) \in A:\langle y, v\rangle_{H_{1}}=0 \\
& \Leftrightarrow \forall(u, v) \in A:\langle y, v\rangle_{H_{1}}+\langle 0, u\rangle_{H_{0}}=0 \\
& \Leftrightarrow \forall(u, v) \in A:\langle(0, y),(u, v)\rangle_{H_{0} \oplus H_{1}}=0 \\
& \Leftrightarrow(0, y) \in A^{\perp}
\end{aligned}
$$

$$
\begin{aligned}
& \Leftrightarrow(y, 0) \in\left(A^{\perp}\right)^{-1} \\
& \Leftrightarrow(y, 0) \in-\left(A^{\perp}\right)^{-1} \\
& \Leftrightarrow y \in[\{0\}] A^{*} .
\end{aligned}
$$

The other identity follows from dualization.

Remark Note that the usual version of the projection theorem reduces to proving that an orthoprojection exists and that it is self-adjoint.

Corollary 1.2. Let $A$ and $A^{*}$ be closed linear operators and $A\left[H_{0}\right]$ closed. Then

$$
A u=f
$$

admits a solution $u$ if and only if $f \perp[\{0\}] A^{*}$. Furthermore, if $u_{0}$ is a solution then the set of solutions is given by $u_{0}+[\{0\}] A$.

Corollary 1.3 (Fredholm Alternative). Let $A$ be a compact operator in $H_{0}$ and $\lambda \in \mathbb{C}$. Considering

$$
\begin{equation*}
(\lambda-A) u=f \tag{*}
\end{equation*}
$$

in $H_{0}$ yields the following cases.
Either (*) admits a unique solution $u$ for every $f \in H_{0}$
or (*) admits a solution $u$ if and only if $f \perp[\{0\}]\left(\lambda^{*}-A^{*}\right)$. In this case, every element of $u+[\{0\}](\lambda-A)$ solves $(*)$ and solutions are unique in $([\{0\}](\lambda-A))^{\perp}$.
Remark The corollary above trivializes the Fredholm alternative to "Either there is a solution or not." However, the Fredholm alternative stated in this form shows that (for any compact operator $A$ ) a non-zero $\lambda \in \mathbb{C}$ is either in the resolvent set or the point spectrum of $A$.

Corollary 1.4. Let $f: H_{0} \rightarrow \mathbb{C}$ be a continuous linear functional. Then $f=0$ or $\operatorname{codim}[\{0\}] f=1$.

Proof. Let $f$ be non-zero. Then $f^{*}$ is non-zero, i.e., $\operatorname{dim} f^{*}[\mathbb{C}]=1$, and $H_{0}=[\{0\}] f \oplus \overline{f^{*}[\mathbb{C}]}$ proves the assertion.

Corollary 1.5 (Riesz' Representation Theorem). Let $f: H_{0} \rightarrow \mathbb{C}$ be a continuous linear functional. Then, there exists $x \in H_{0}$ such that

$$
\forall y \in H_{0}: f(y)=\langle y, x\rangle_{H_{0}}
$$

Proof. If $f=0$ then $x=0 \checkmark$
If $f \neq 0$ then choose $x_{0} \in([\{0\}] f)^{\perp}$ with $\|x\|_{H_{0}}=1$ and define $x:=f\left(x_{0}\right)^{*} x_{0}$. Then $(x)$ is a basis of $([\{0\}] f)^{\perp}$, i.e.,

$$
\begin{aligned}
\forall y \in H_{0}:\langle y, x\rangle_{H_{0}} & =\left\langle y, f\left(x_{0}\right)^{*} x_{0}\right\rangle_{H_{0}} \\
& =f\left(x_{0}\right)\left\langle y, x_{0}\right\rangle_{H_{0}} \\
& =f\left(\left\langle y, x_{0}\right\rangle_{H_{0}} x_{0}\right) \\
& =f(\left\langle y, x_{0}\right\rangle_{H_{0}} x_{0}+\underbrace{y-\left\langle y, x_{0}\right\rangle_{H_{0}} x_{0}}_{\in[\{0\}] f}) \\
& =f(y) .
\end{aligned}
$$

Let $\tilde{x} \in H_{0}$ be such that $\forall y \in H_{0}: f(y)=\langle y, x\rangle_{H_{0}}$ holds, as well. Then

$$
\forall y \in H_{0}: 0=\langle y, x\rangle_{H_{0}}-\langle y, \tilde{x}\rangle_{H_{0}}=\langle y, x-\tilde{x}\rangle_{H_{0}}
$$

holds and we conclude $x=\tilde{x}$, that is, $x$ is unique.

The following example shows that we may also use the projection theorem to solve PDE.
Example Let $\Omega \subseteq \mathbb{R}^{n}$ open and non-empty, $C_{c}^{\infty}(\Omega, \mathbb{K})$ the set of $C^{\infty}\left(\mathbb{R}^{n}, \mathbb{K}\right)$ functions with compact support in $\Omega, \operatorname{grad}_{c}: C_{c}^{\infty}(\Omega, \mathbb{K}) \rightarrow C_{c}^{\infty}\left(\Omega, \mathbb{K}^{n}\right)$ the usual gradient, and $\operatorname{div}_{c}: C_{c}^{\infty}\left(\Omega, \mathbb{K}^{n}\right) \rightarrow C_{c}^{\infty}(\Omega, \mathbb{K})$ the usual divergence. Then it is easy that $\operatorname{grad}_{c}$ and $-\operatorname{div}_{c}$ are formally adjoint (partial integration) in $L_{2}(\Omega, \mathbb{K}) \oplus L_{2}\left(\Omega, \mathbb{K}^{n}\right)$, that is, $\operatorname{grad}_{c} \subseteq-\operatorname{div}_{c}^{*}$. Note that $A \subseteq B^{*}$ implies

$$
\bar{B}=B^{* *} \subseteq A^{*}
$$

which shows that both operators are closable if $A^{*}$ and $B^{*}$ are operators (that is, $A$ and $B$ are densely defined). Hence,

$$
\operatorname{grad}_{0}:=\overline{\operatorname{grad}_{c}}, \operatorname{div}_{0}:=\overline{\operatorname{div}_{c}}, \operatorname{grad}:=-\operatorname{div}_{c}^{*}, \operatorname{div}:=-\operatorname{grad}_{c}^{*}
$$

exist and they are all densely defined closed linear operators.
For $A \in\left\{\operatorname{grad}_{0}, \operatorname{grad}, \operatorname{div}_{0}, \operatorname{div}\right\}$ we define $H(A)$ to be the Hilbert space $D(A)$ endowed with the graph norm $\|x\|_{H(A)}^{2}=\|x\|_{L_{2}}^{2}+\|A x\|_{L_{2}}^{2}$. The projection theorem for inclusion $H\left(\operatorname{grad}_{0}\right) \hookrightarrow H(\operatorname{grad})$ now shows

$$
H(\operatorname{grad})=H\left(\operatorname{grad}_{0}\right) \oplus H\left(\operatorname{grad}_{0}\right)^{\perp}
$$

Let $f \in H\left(\operatorname{grad}_{0}\right)^{\perp} \cap D(\operatorname{div} \operatorname{grad})$. Then

$$
\begin{aligned}
\forall x \in H\left(\operatorname{grad}_{0}\right): 0 & =\langle f, x\rangle_{H(\operatorname{grad})} \\
& =\langle f, x\rangle_{L_{2}(\Omega, \mathbb{R})}+\langle\operatorname{grad} f, \operatorname{grad} x\rangle_{L_{2}\left(\Omega, \mathbb{R}^{n}\right)} \\
& =\langle f, x\rangle_{L_{2}(\Omega, \mathbb{R})}+\left\langle\operatorname{grad} f, \operatorname{grad}_{0} x\right\rangle_{L_{2}\left(\Omega, \mathbb{R}^{n}\right)} \\
& =\langle f, x\rangle_{L_{2}(\Omega, \mathbb{R})}+\langle-\operatorname{div} \operatorname{grad} f, x\rangle_{L_{2}(\Omega, \mathbb{R})} \\
& =\langle(1-\operatorname{div} \operatorname{grad}) f, x\rangle_{L_{2}(\Omega, \mathbb{R})}
\end{aligned}
$$

implies $H\left(\operatorname{grad}_{0}\right)^{\perp}=\overline{[\{0\}](1-\operatorname{div} \operatorname{grad})}=[\{0\}](1-\operatorname{div} \operatorname{grad})$.
We may now use this to solve the inhomogeneous Dirichlet problem

$$
\varphi-\operatorname{div} \operatorname{grad} \varphi=0, \varphi-f \in H\left(\operatorname{grad}_{0}\right), f \in H(\operatorname{grad})
$$

Since $H(\operatorname{grad})=H\left(\operatorname{grad}_{0}\right) \oplus[\{0\}](1-\operatorname{div} \operatorname{grad})$, there are unique $f_{0} \in H\left(\operatorname{grad}_{0}\right)$ and $f_{1} \in[\{0\}](1-\operatorname{div} \operatorname{grad})$ such that $f=f_{0}+f_{1}$ and we obtain

$$
(1-\operatorname{div} \operatorname{grad}) f_{1}=0, f_{1}-f=-f_{0} \in H\left(\operatorname{grad}_{0}\right), f \in H(\operatorname{grad}),
$$

i.e., $\varphi=f_{1}$ solves the inhomogeneous Dirichlet problem by projection.

## CHAPTER 2

## Tensor products of Hilbert spaces

Let $n \in \mathbb{N}$ and $\left(H_{k}\right)_{k \in \mathbb{N}_{\leq n}}$ be a family of real ${ }^{1}$ Hilbert spaces. For $x \in \mathbf{X}_{k=1}^{n} H_{k}$ we define $x_{1} \otimes \ldots \otimes x_{n} \in\left(\mathbf{X}_{k=1}^{n} H_{k}\right)^{*}$ to be the linear functional that suffices

$$
\forall u \in \underset{k=1}{\stackrel{n}{\times}} H_{k}:\left(x_{1} \otimes \ldots \otimes x_{n}\right)(u)=\left\langle x_{1}, u_{1}\right\rangle_{H_{1}} \cdot \ldots \cdot\left\langle x_{n}, u_{n}\right\rangle_{H_{n}}
$$

Let

$$
W_{\otimes}:=\operatorname{lin}\left\{x_{1} \otimes \ldots \otimes x_{n} ; x \in \underset{k=1}{\times} H_{k}\right\}
$$

be equipped with the bilinear continuation of

$$
\left\langle x_{1} \otimes \ldots \otimes x_{n}, u_{1} \otimes \ldots \otimes u_{n}\right\rangle_{H_{1} \otimes \ldots \otimes H_{n}}:=\left\langle x_{1}, u_{1}\right\rangle_{H_{1}} \cdot \ldots \cdot\left\langle x_{n}, u_{n}\right\rangle_{H_{n}}
$$

(i) Symmetry

$$
\begin{aligned}
& \left\langle\sum_{i} \alpha_{i} x_{i, 1} \otimes \ldots \otimes x_{i, n}, \sum_{j} \beta_{j} y_{j, 1} \otimes \ldots \otimes y_{j, n}\right\rangle_{H_{1} \otimes \ldots \otimes H_{n}} \\
= & \sum_{i} \sum_{j} \alpha_{i} \beta_{j}\left\langle x_{i, 1}, y_{j, 1}\right\rangle_{H_{1}} \ldots\left\langle x_{i, n}, y_{j, n}\right\rangle_{H_{n}} \\
= & \sum_{i} \sum_{j} \alpha_{i} \beta_{j}\left\langle y_{j, 1}, x_{i, 1}\right\rangle_{H_{1}} \ldots\left\langle y_{j, n}, x_{i, n}\right\rangle_{H_{n}} \\
= & \left\langle\sum_{j} \beta_{j} y_{j, 1} \otimes \ldots \otimes y_{j, n}, \sum_{i} \alpha_{i} x_{i, 1} \otimes \ldots \otimes x_{i, n}\right\rangle_{H_{1} \otimes \ldots \otimes H_{n}}
\end{aligned}
$$

(ii) Non-negativity

Since the Gramian matrices $G_{k}:=\left(\left\langle x_{i, k}, x_{j, k}\right\rangle_{H_{k}}\right)_{i, j \in \mathbb{N}_{\leq m}}$ are positive semi-definite, the matrices $\left(A_{i j}^{(k)}\right)_{i, j \in \mathbb{N}_{\leq m}}:=\sqrt{G_{k}}$ are positive semi-definite as well. Thus,

$$
\begin{aligned}
& \left\langle\sum_{i=1}^{m} \alpha_{i} x_{i, 1} \otimes \ldots \otimes x_{i, n}, \sum_{j=1}^{m} \alpha_{j} x_{j, 1} \otimes \ldots \otimes x_{j, n}\right\rangle_{H_{1} \otimes \ldots \otimes H_{n}} \\
= & \sum_{i=1}^{m} \sum_{j=1}^{m} \alpha_{i} \alpha_{j}\left\langle x_{i, 1}, x_{j, 1}\right\rangle_{H_{1}} \ldots \cdot\left\langle x_{i, n}, x_{j, n}\right\rangle_{H_{n}} \\
= & \sum_{i=1}^{m} \sum_{j=1}^{m} \sum_{s_{1}} \ldots \sum_{s_{n}} \alpha_{i} \alpha_{j} A_{i s_{1}}^{(1)} A_{s_{1} j}^{(1)} \ldots A_{i s_{n}}^{(n)} A_{s_{n} j}^{(n)} \\
= & \sum_{s_{1}} \ldots \sum_{s_{n}}\left(\sum_{i=1}^{m} A_{i s_{1}}^{(1)} \ldots A_{i s_{n}}^{(n)} \alpha_{i}\right)\left(\sum_{j=1}^{m} A_{s_{1} j}^{(1)} \ldots A_{s_{n} j}^{(n)} \alpha_{j}\right)
\end{aligned}
$$

[^0]$$
=\left\langle\left(\sum_{i=1}^{m} A_{i s_{1}}^{(1)} \ldots A_{i s_{n}}^{(n)} \alpha_{i}\right)_{\left(s_{1}, \ldots, s_{n}\right)},\left(\sum_{i=1}^{m} A_{i s_{1}}^{(1)} \ldots A_{i s_{n}}^{(n)} \alpha_{i}\right)_{\left(s_{1}, \ldots, s_{n}\right)}\right\rangle_{\ell_{2}\left(\mathbb{N}_{\leq m}^{n}\right)} \geq 0
$$
holds.
Hence, $\left(W_{\otimes},\langle\cdot, \cdot\rangle_{H_{1} \otimes \ldots \otimes H_{n}}\right)$ is a semi-scalar product space and called the algebraic tensor product of $\left(H_{k}\right)_{k \in \mathbb{N}_{\leq n}}$. We will also denote algebraic tensor products as $H_{1} \stackrel{a}{\otimes} \ldots \stackrel{a}{\otimes} H_{n}$ or $\stackrel{a}{\otimes}_{k \in \mathbb{N}_{\leq n}} H_{k}$.

Definition 2.1. The completion

$$
H_{1} \otimes \ldots \otimes H_{n}:=\bigotimes_{k=1}^{n} H_{k}:={\overline{W_{\otimes}}}_{\|\cdot\|_{H_{1} \otimes \ldots \otimes H_{n}}}
$$

is called tensor product of $\left(H_{k}\right)_{k \in \mathbb{N}_{s n}}$ where $\|\cdot\|_{H_{1} \otimes \ldots \otimes H_{k}}$ denotes the semi-norm induced by $\langle\cdot, \cdot\rangle_{H_{1} \otimes \ldots \otimes H_{n}}$.

The empty tensor product $\otimes_{\varnothing}$ (sometimes denoted as $\otimes_{\varnothing} H$ with some arbitrary Hilbert space $H$ ) is defined as $\otimes_{\varnothing}:=\mathbb{R}$.

Remark (i) Due to the completion process, elements $x, y \in H_{1} \otimes \ldots \otimes H_{n}$ with $\|x-y\|_{H_{1} \otimes \ldots \otimes H_{n}}=0$ are identified. $H_{1} \otimes \ldots \otimes H_{n}$ is a Hilbert space, thus.
(ii) The choice $\otimes_{\varnothing}:=\mathbb{R}$ is senseful because both $\otimes_{\varnothing}$ and $\mathbb{R}$ act as neutral elements

$$
\bigotimes_{\varnothing}^{\otimes} \otimes \bigotimes_{i \in I} H_{i}=\bigotimes_{i \in I} H_{i} \cong \mathbb{R} \otimes \bigotimes_{i \in I} H_{i} .
$$

(iii) The tensor products introduced here are not tensor products in the algebraic sense as, in general, they fail to have the universal property ${ }^{2}$; cf., [6].

Example Let $H$ be a Hilbert space and $\Omega \subseteq \mathbb{R}$ measurable. The space $L_{2}(\Omega ; H)$ is the completion of

$$
\operatorname{lin}\left\{t \mapsto 1_{I}(t) x ; x \in H, I \subseteq \Omega \text { measurable and with finite measure }\right\}
$$

with respect to the scalar product $(f, g) \mapsto \int_{\Omega}\langle f(t), g(t)\rangle_{H} d t$. For $I \subseteq \Omega$ measurable and $x \in H$ we define

$$
1_{I} \otimes x:=\left(t \mapsto 1_{I}(t) x\right) .
$$

Obviously

$$
\begin{aligned}
\left\langle 1_{I} \otimes x, 1_{J} \otimes y\right\rangle_{L_{2}(\Omega) \otimes H} & =\int_{\Omega} 1_{I}(t) 1_{J}(t) d t\langle x, y\rangle_{H} \\
& =\int_{\Omega}\left\langle 1_{I}(t) x, 1_{J}(t) y\right\rangle_{H} d t \\
& =\left\langle 1_{I} x, 1_{J} y\right\rangle_{L_{2}(\Omega ; H)}
\end{aligned}
$$

holds. Thus, the closure of the linear continuation of $\left(t \mapsto 1_{I}(t) x\right) \mapsto 1_{I} \otimes x$ defines a unitary $\operatorname{map} U: L_{2}(\Omega ; H) \rightarrow L_{2}(\Omega) \otimes H$.

[^1]Theorem 2.2 (Structure of Tensor Products). Let $H_{1}$ and $H_{2}$ be separable, infinite dimensional Hilbert spaces. Then, there exists $T \in L\left(H_{1} \otimes H_{2}, L\left(H_{2}, H_{1}\right)\right)$ satisfying

$$
\forall h_{1} \in H_{1} \forall h_{2}, h_{2}^{\prime} \in H_{2}: T\left(h_{1} \otimes h_{2}\right) h_{2}^{\prime}=\left\langle h_{2}, h_{2}^{\prime}\right\rangle_{H_{2}} h_{1}
$$

The operator $T$ maps $H_{1} \otimes H_{2}$ unitarily to $H S\left(H_{2}, H_{1}\right)$, the set of Hilbert-Schmidtoperators between $H_{2}$ and $H_{1}$.

Furthermore, let $a \in H_{1} \otimes H_{2}$. Then, there exists $\lambda \in \ell_{2}(\mathbb{N})$, an orthonormal basis $\left(\eta_{i}\right)_{i \in \mathbb{N}}$ of $H_{1}$, and and orthonormal basis $\left(\chi_{i}\right)_{i \in \mathbb{N}}$ of $H_{2}$ such that $\|a\|_{H_{1} \otimes H_{2}}=$ $\|\lambda\|_{\ell_{2}(\mathbb{N})}$ and

$$
a=\sum_{n \in \mathbb{N}} \lambda_{n} \eta_{n} \otimes \chi_{n}
$$

hold.
Proof. Let $\left(\varphi_{i}\right)_{i \in \mathbb{N}}$ be an orthonormal basis of $H_{1},\left(\psi_{i}\right)_{i \in \mathbb{N}}$ an orthonormal basis of $H_{2}$, and

$$
\tilde{a}:=\sum_{i=1}^{m} \sum_{j=1}^{n} \tilde{a}_{i j} \varphi_{i} \otimes \varphi_{j} .
$$

Then, $\left(\tilde{a}_{i j}\right)_{(i, j) \in \mathbb{N}^{2}} \in \ell_{2}\left(\mathbb{N}^{2}\right)$ with $\tilde{a}_{i j}:=0$ for $i>m$ or $j>n$, and we obtain $\|\tilde{a}\|_{H_{1} \otimes H_{2}}=$ $\left\|\left(\tilde{a}_{i j}\right)_{(i, j) \in \mathbb{N}^{2}}\right\|_{\ell_{2}\left(\mathbb{N}^{2}\right)}$ by the Pythagorean theorem. Hence, we may decompose any $a \in H_{1} \otimes H_{2}$ as $a=\sum_{i, j \in \mathbb{N}} a_{i j} \varphi_{i} \otimes \varphi_{j}$ with $\|a\|_{H_{1} \otimes H_{2}}=\left\|\left(a_{i j}\right)_{(i, j) \in \mathbb{N}^{2}}\right\|_{\ell_{2}\left(\mathbb{N}^{2}\right)}$.

Let $a \in H_{1} \otimes H_{2}$ satisfy $a=\sum_{i, j \in \mathbb{N}} a_{i j} \varphi_{i} \otimes \varphi_{j}$ and $h_{2} \in H_{2}$. Then, we define

$$
T(a) h_{2}:=\sum_{i, j \in \mathbb{N}} a_{i j}\left\langle\psi_{j}, h_{2}\right\rangle_{H_{2}} \varphi_{i}
$$

and observe for $h_{1}=\sum_{i \in \mathbb{N}} \alpha_{i} \varphi_{i} \in H_{1}, h_{2}=\sum_{j \in \mathbb{N}} \beta_{j} \psi_{j} \in H_{2}$, and $h_{2}^{\prime} \in H_{2}$

$$
T\left(h_{1} \otimes h_{2}\right) h_{2}^{\prime}=\sum_{i, j \in \mathbb{N}} \alpha_{i} \beta_{j}\left\langle\psi_{j}, h_{2}^{\prime}\right\rangle_{H_{2}} \varphi_{i}=\left\langle\sum_{j \in \mathbb{N}} \beta_{j} \psi_{j}, h_{2}^{\prime}\right\rangle \sum_{i \in \mathbb{N}} \alpha_{i} \varphi_{i}=\left\langle h_{2}, h_{2}^{\prime}\right\rangle_{H_{2}} h_{1},
$$

as well as,

$$
\begin{aligned}
\left\|T(a) h_{2}\right\|_{H_{1}}^{2} & =\sum_{i \in \mathbb{N}}\left|\sum_{j \in \mathbb{N}} a_{i j}\left\langle\psi_{j}, h_{2}\right\rangle_{H_{2}}\right|^{2} \\
& =\sum_{i \in \mathbb{N}}\left|\left\langle\sum_{j \in \mathbb{N}} a_{i j} \psi_{j}, h_{2}\right\rangle_{H_{2}}\right|^{2} \\
& \leq \sum_{i \in \mathbb{N}}\left\|\sum_{j \in \mathbb{N}} a_{i j} \psi_{j}\right\|_{H_{2}}^{2}\left\|h_{2}\right\|_{H_{2}}^{2} \\
& \leq \sum_{i \in \mathbb{N}}\left(\sum_{j \in \mathbb{N}}\left|a_{i j}\right|\right)^{2}\left\|h_{2}\right\|_{H_{2}}^{2} \\
& \leq\left(\sum_{i \in \mathbb{N}} \sum_{j \in \mathbb{N}}\left|a_{i j}\right|\right)^{2}\left\|h_{2}\right\|_{H_{2}}^{2} \\
& =\left\|\left(a_{i j}\right)_{(i, j) \in \mathbb{N}^{2}}\right\|_{\ell_{1}\left(\mathbb{N}^{2}\right)}^{2}\left\|h_{2}\right\|_{H_{2}}^{2} .
\end{aligned}
$$

Thus, $T$ extends to a bounded operator on $H_{1} \otimes H_{2}$ and the Hilbert-Schmidt norm $\|T(a)\|_{H S}$ of $T(a)$ satisfies

$$
\left.\|T(a)\|_{H S}^{2}=\sum_{k \in \mathbb{N}}\left\|T(a) \psi_{k}\right\|_{H_{1}}^{2}\right)=\sum_{k \in \mathbb{N}} \sum_{i \in \mathbb{N}}\left|\sum_{j \in \mathbb{N}} a_{i j}\left\langle\psi_{j}, \psi_{k}\right\rangle_{H_{2}}\right|^{2}=\sum_{i, k \in \mathbb{N}}\left|a_{i k}\right|^{2}=\|a\|_{H_{1} \otimes H_{2}}^{2},
$$

i.e., $T: H_{1} \otimes H_{2} \rightarrow H S\left(H_{2}, H_{1}\right)$ is an isometry.

Let $S \in H S\left(H_{2}, H_{1}\right)$. Then, $S^{*} S \in L\left(H_{2}\right)$ is compact ${ }^{3}$, self-adjoint, and nonnegative. Thus, the spectral theorem yields the existence of $N \subseteq \mathbb{N}$, an orthonormal basis $\left(\chi_{n}\right)_{n \in N}$ of $\left([\{0\}] S^{*} S\right)^{\perp}\left([\{0\}] S^{*} S\right.$ is the kernel of $\left.S^{*} S\right)$, and $\lambda \in \mathbb{R}_{>0}^{N}$ such that for every $h_{2} \in H_{2}$

$$
S^{*} S h_{2}=\sum_{n \in N} \lambda_{n}^{2}\left\langle h_{2}, \chi_{n}\right\rangle_{H_{2}} \chi_{n}
$$

holds. Let $\eta_{n}:=\lambda_{n}^{-1} S \chi_{n}$ for $n \in N$. Then,

$$
\left\langle\eta_{n}, \eta_{m}\right\rangle_{H_{1}}=\lambda_{n}^{-1} \lambda_{m}^{-1}\left\langle\chi_{n}, S^{*} S \chi_{m}\right\rangle_{H_{2}}=\lambda_{n}^{-1} \lambda_{m}\left\langle\chi_{n}, \chi_{m}\right\rangle_{H_{2}}=\delta_{m n}
$$

shows that $\left(\eta_{n}\right)_{n \in N}$ is an orthonormal set. Defining $a_{k}:=\sum_{n \in N_{<k}} \lambda_{n} \eta_{n} \otimes \chi_{n}$ for $k \in N$, we observe for $n \in N$

$$
T\left(a_{k}\right) \chi_{n}=\sum_{j \in N_{<k}} \lambda_{j}\left\langle\chi_{j}, \chi_{n}\right\rangle_{H_{2}} \eta_{j}=\left\{\begin{array}{ll}
\lambda_{n} \eta_{n} & , n<k \\
0 & , n \geq k
\end{array}=\left\{\begin{array}{ll}
S \chi_{n} & , n<k \\
0 & , n \geq k
\end{array} .\right.\right.
$$

Hence, $\left\|\left.T\left(a_{k}\right)\right|_{([\{0\}] S)^{\perp}}\right\|_{H S\left(H_{2}, H_{1}\right)} \leq\left\|\left.S\right|_{([\{0\}] S)^{\perp}}\right\|_{H S\left(H_{2}, H_{1}\right)}$. Since $T$ is an isometry, this shows that $a:=\sum_{n \in N} \lambda_{n} \eta_{n} \otimes \chi_{n}$ converges and $\left.T(a)\right|_{([\{0\}] S)^{\perp}}=\left.S\right|_{([\{0\}] S)^{\perp}}$ holds. However, by definition, we have $[\{0\}] T(a)=[\{0\}] S$, i.e., $T(a)=S$, showing surjectivity of $T$. Since isometries are injective, we directly obtain bijectivity, too. Furthermore, setting $\lambda_{n}:=0$ for $n \in \mathbb{N} \backslash N$, we obtain

$$
\sum_{n \in \mathbb{N}}\left|\lambda_{n}\right|^{2}=\left\langle\sum_{n \in N} \lambda_{n} \eta_{n} \otimes \chi_{n}, \sum_{m \in N} \lambda_{m} \eta_{m} \otimes \chi_{m}\right\rangle_{H_{1} \otimes H_{2}}=\|a\|_{H_{1} \otimes H_{2}}^{2}
$$

which shows $\lambda \in \ell_{2}(\mathbb{N})$ and $\|\lambda\|_{\ell_{2}(\mathbb{N})}=\|a\|_{H_{1} \otimes H_{2}}$; thus, completing the proof.

Let $\left(H_{0, k}\right)_{k \in \mathbb{N}_{\leq n}}$ and $\left(H_{1, k}\right)_{k \in \mathbb{N}_{\leq n}}$ be families of real Hilbert spaces and for each $k \in \mathbb{N}_{\leq n}$ let $A_{k} \subseteq H_{0, k} \oplus H_{1, k}$ be a linear operator ${ }^{4}$. We define

$$
A_{1} \dot{\otimes} \ldots \dot{\otimes} A_{n}: \bigotimes_{k \in \mathbb{N}_{\leq n}} H_{0, k} \rightarrow \bigotimes_{k \in \mathbb{N}_{\leq n}} H_{1, k}
$$

as linear continuation of $x_{1} \otimes \ldots \otimes x_{n} \mapsto\left(A_{1} x_{1}\right) \otimes \ldots \otimes\left(A_{n} x_{n}\right)$ with

$$
D\left(A_{1} \dot{\otimes} \ldots \dot{\otimes} A_{n}\right):=\bigotimes_{k \in \mathbb{N}_{\leq n}}^{a} D\left(A_{k}\right) .
$$

We will also use the abbreviation $\dot{\otimes}_{k \in \mathbb{N}_{\leq n}} A_{k}$ for $A_{1} \dot{\otimes} \ldots \dot{\otimes} A_{n}$.

[^2]Since $A_{1} \dot{\otimes} \ldots \dot{\otimes} A_{n}$ is a linear continuation, it is a linear subspace of

$$
\bigotimes_{k \in \mathbb{N}_{\leq n}} H_{0, k} \oplus \bigotimes_{k \in \mathbb{N}_{\leq n}} H_{1, k}
$$

thus, we only need to make sure that for all $(0, w) \in A_{1} \dot{\otimes} \ldots \dot{\otimes} A_{n}$

$$
w=0
$$

holds for $A_{1} \dot{\otimes} \ldots \dot{\otimes} A_{n}$ to be an operator. We may express $w$ by

$$
w=\sum_{i} \alpha_{i}\left(A_{1} x_{i, 1}\right) \otimes \ldots \otimes\left(A_{n} x_{i, n}\right)
$$

with

$$
0=\sum_{i} \alpha_{i} x_{i, 1} \otimes \ldots \otimes x_{i, n}
$$

and, thus, we observe

$$
\begin{aligned}
\forall k \in \mathbb{N}_{\leq n} \forall u_{k} \in H_{0, k}: \quad 0 & =\sum_{i} \alpha_{i}\left\langle x_{i, 1}, u_{1}\right\rangle_{H_{0,1}} \cdot \ldots \cdot\left\langle x_{i, n}, u_{n}\right\rangle_{H_{0, n}} \\
& =\left\langle\sum_{i} \alpha_{i}\left\langle x_{i, 2}, u_{2}\right\rangle_{H_{0,2}} \cdot \ldots \cdot\left\langle x_{i, n}, u_{n}\right\rangle_{H_{0, n}} x_{i, 1}, u_{1}\right\rangle_{H_{0,1}}
\end{aligned}
$$

Without loss of generality we may assume that the $x_{i, k}$ are linearly independent in $H_{0, k}$ yielding

$$
0=\alpha_{i}\left\langle x_{i, 2}, u_{2}\right\rangle_{H_{0,2}} \cdot \ldots \cdot\left\langle x_{i, n}, u_{n}\right\rangle_{H_{0, n}}
$$

for every $i$. Since none of the $x_{i, k}$ is zero

$$
\forall i: \alpha_{i}=0
$$

needs to hold. Hence, $w=0 . A_{1} \dot{\otimes} \ldots \dot{\otimes} A_{n}$ is an operator, thus. If it is closable, we will denote the closure by

$$
A_{1} \otimes \ldots \otimes A_{n}
$$

Remark In fact, if all $A_{k}$ are closed operators, then they are Hilbert spaces with respect to the graph norm and the tensor product of the operators is isometrically isomorphic to the tensor product of Hilbert spaces. In particular, $A_{1} \dot{\otimes} \ldots \dot{\otimes} A_{n} \cong$ $A_{1} \stackrel{a}{\otimes} \ldots \stackrel{a}{\otimes} A_{n}$

Lemma 2.3. Let $H_{0}$ and $H_{1}$ be Hilbert spaces, $S_{0} \subseteq H_{0}$ be total, i.e., $\operatorname{lin} S_{0}$ is dense in $H_{0}$, and $S_{1} \subseteq H_{1}$ total. Then, $S_{0} \stackrel{a}{\otimes} S_{1}$ is dense in $H_{0} \otimes H_{1}$.

Proof. Let $x \in H_{0}$ and $y \in H_{1}$. Then, there are sequences $\left(x_{n}\right)_{n \in \mathbb{N}} \in\left(\operatorname{lin} S_{0}\right)^{\mathbb{N}}$ and $\left(y_{n}\right)_{n \in \mathbb{N}} \in\left(\operatorname{lin} S_{1}\right)^{\mathbb{N}}$ with $x_{n} \rightarrow x$ in $H_{0}$ and $y_{n} \rightarrow y$ in $H_{1}$.

Let $n \in \mathbb{N}$. Then, there are $k, m \in \mathbb{N}, s_{1}^{0}, \ldots, s_{k}^{0} \in S_{0}, s_{1}^{1}, \ldots, s_{m}^{1} \in S_{1}$, and $\alpha_{1}, \ldots, \alpha_{k}, \beta_{1}, \ldots, \beta_{m} \in \mathbb{R}$ such that $x_{n}=\sum_{j=1}^{k} \alpha_{j} s_{j}^{0}$ and $y_{n}=\sum_{j=1}^{m} \beta_{j} s_{j}^{1}$. Hence,

$$
x_{n} \otimes y_{n}=\left(\sum_{i=1}^{k} \alpha_{i} s_{i}^{0}\right) \otimes\left(\sum_{j=1}^{m} \beta_{j} s_{j}^{1}\right)=\sum_{i=1}^{k} \sum_{j=1}^{m} \alpha_{i} \beta_{j} s_{i}^{0} \otimes s_{j}^{1} \in S_{0} \stackrel{a}{\otimes} S_{1} .
$$

Furthermore, we obtain

$$
\begin{aligned}
\left\|x_{n} \otimes y_{n}-x \otimes y\right\|_{H_{0} \otimes H_{1}} & \leq\left\|x_{n} \otimes y_{n}-x \otimes y_{n}\right\|_{H_{0} \otimes H_{1}}+\left\|x \otimes y_{n}-x \otimes y\right\|_{H_{0} \otimes H_{1}} \\
& =\left\|\left(x_{n}-x\right) \otimes y_{n}\right\|_{H_{0} \otimes H_{1}}+\left\|x \otimes\left(y_{n}-y\right)\right\|_{H_{0} \otimes H_{1}} \\
& =\underbrace{\left\|x_{n}-x\right\|_{H_{0}}}_{\rightarrow 0} \underbrace{\left\|y_{n}\right\|_{H_{1}}}_{\text {bounded }}+\|x\|_{H_{0}}^{\left\|y_{n}-y\right\|_{H_{1}}} \\
& \rightarrow 0 .
\end{aligned}
$$

Hence, all simple tensors can be approximated by elements of $S_{0} \stackrel{a}{\otimes} S_{1}$.
Let $n \in \mathbb{N}, x_{1}, \ldots, x_{n} \in H_{0}, y_{1}, \ldots, y_{n} \in H_{1}, \alpha_{1}, \ldots, \alpha_{n} \in \mathbb{R}$, and $\varepsilon \in \mathbb{R}_{>0}$. Since simple tensors can be approximated by elements of $S_{0} \stackrel{a}{\otimes} S_{1}$, there are elements $u_{i} \in S_{0} \stackrel{a}{\otimes} S_{1}$ such that

$$
\left\|x_{i} \otimes y_{i}-u_{i}\right\|_{H_{0} \otimes H_{1}}<\left(\sum_{j=1}^{n}\left|\alpha_{j}\right|\right)^{-1} \varepsilon
$$

holds for every $i \in \mathbb{N}_{\leq n}$. Hence,

$$
\left\|\sum_{i=1}^{n} \alpha_{i} x_{i} \otimes y_{i}-\sum_{i=1}^{n} \alpha_{i} u_{i}\right\|_{H_{0} \otimes H_{1}} \leq \sum_{i=1}^{n}\left|\alpha_{i}\right|\left\|x_{i} \otimes y_{i}-u_{i}\right\|_{H_{0} \otimes H_{1}}<\varepsilon
$$

shows density of $S_{0} \stackrel{a}{\otimes} S_{1}$ in $H_{0} \stackrel{a}{\otimes} H_{1}$ and, thus, the assertion as well.

Corollary 2.4. Let $H_{0}$ and $H_{1}$ be Hilbert spaces, and $O_{0} \subseteq H_{0}$ and $O_{1} \subseteq H_{1}$ two complete orthonormal sets. Then,

$$
\left[O_{0}\right] \otimes\left[O_{1}\right]:=\left\{u \otimes v ; u \in O_{0} \wedge v \in O_{1}\right\}
$$

is a complete orthonormal set in $H_{0} \otimes H_{1}$.
Proof. We already know that $O_{0} \stackrel{a}{\otimes} O_{1}$ is dense in $H_{0} \otimes H_{1}$, i.e., $\left[O_{0}\right] \otimes\left[O_{1}\right]$ is total. Hence, it suffices to show that $\left[O_{0}\right] \otimes\left[O_{1}\right]$ is orthonormal. Let $u \otimes v, x \otimes y \in$ $\left[O_{0}\right] \otimes\left[O_{1}\right]$. Then,

$$
\langle u \otimes v, x \otimes y\rangle_{H_{0} \otimes H_{1}}=\langle u, x\rangle_{H_{0}}\langle v, y\rangle_{H_{1}}= \begin{cases}1 & , u=x \wedge v=y \\ 0 & , u \neq x \vee v \neq y\end{cases}
$$

shows the assertion.

Proposition 2.5. Let $H_{00}, H_{01}, H_{10}$, and $H_{11}$ be Hilbert spaces, and $A \subseteq$ $H_{00} \oplus H_{01}$ and $B \subseteq H_{10} \oplus H_{11}$ densely defined closable linear operators. Then, $A \dot{\otimes} B$ is closable and

$$
A \otimes B=\overline{A \dot{\otimes} B} \subseteq\left(A^{*} \dot{\otimes} B^{*}\right)^{*}
$$

holds.

Proof. Let $\xi=\sum_{i=1}^{n} \alpha_{i} x_{i} \otimes y_{i} \in D\left(A^{*}\right) \stackrel{a}{\otimes} D\left(B^{*}\right)$ and $\eta=\sum_{j=1}^{m} \beta_{j} u_{j} \otimes v_{j} \epsilon$ $D(A) \stackrel{a}{\otimes} D(B)$. Then, we observe

$$
\begin{aligned}
\langle A \dot{\otimes} B \eta, \xi\rangle_{H_{01} \otimes H_{11}} & =\sum_{i=1}^{n} \sum_{j=1}^{m} \alpha_{i} \beta_{j}\left\langle A u_{i}, x_{j}\right\rangle_{H_{01}}\left\langle B v_{i}, y_{j}\right\rangle_{H_{11}} \\
& =\sum_{i=1}^{n} \sum_{j=1}^{m} \alpha_{i} \beta_{j}\left\langle u_{i}, A^{*} x_{j}\right\rangle_{H_{00}}\left\langle v_{i}, B^{*} y_{j}\right\rangle_{H_{10}} \\
& =\left\langle\eta, A^{*} \dot{\otimes} B^{*} \xi\right\rangle_{H_{00} \otimes H_{10}}
\end{aligned}
$$

that is, $A^{*} \dot{\otimes} B^{*} \subseteq(A \dot{\otimes} B)^{*}$, which implies

$$
A \dot{\otimes} B \subseteq \bar{A} \dot{\otimes} \bar{B}=A^{* *} \dot{\otimes} B^{* *} \subseteq\left(A^{*} \dot{\otimes} B^{*}\right)^{*}
$$

Since $A$ and $B$ are closable, $A^{*}$ and $B^{*}$ are densely defined, and Lemma 2.3 yields that $A^{*} \dot{\otimes} B^{*}$ is densely defined, i.e., $\left(A^{*} \dot{\otimes} B^{*}\right)^{*}$ is a closed linear operator.

Example Let $H_{1}, H_{2}$ be Hilbert spaces, $A \subseteq H_{2} \oplus H_{2}$ be a closed and densely defined linear operator. The operator $\mathcal{A}$ defined by the $H_{1} \otimes H_{2}$-closure of $x \otimes y \mapsto$ $x \otimes A y$ can be expressed by

$$
\mathcal{A}=1 \otimes A
$$

Proposition 2.6. Let $H_{00}, H_{01}, H_{10}$, and $H_{11}$ be Hilbert spaces, and $A \subseteq$ $H_{00} \oplus H_{01}$ and $B \subseteq H_{10} \oplus H_{11}$ densely defined closable linear operators. Then,

$$
A \otimes B=\bar{A} \otimes \bar{B}
$$

Proof. Clearly,

$$
A \dot{\otimes} B \subseteq \bar{A} \dot{\otimes} \bar{B} \subseteq \bar{A} \otimes \bar{B}
$$

holds and, hence, $A \otimes B \subseteq \bar{A} \otimes \bar{B}$. Let $x=\sum_{i=1}^{n} \alpha_{i} \xi_{i} \otimes \eta_{i} \in D(\bar{A}) \stackrel{a}{\otimes} D(\bar{B})=D(\bar{A} \dot{\otimes} \bar{B})$, $x \neq 0$, and $\varepsilon \in \mathbb{R}_{>0}$. Then, we can find $x_{i} \in D(A)$ and $y_{i} \in D(B)$ such that for every $i \in \mathbb{N}_{\leq n}$

$$
\begin{aligned}
\left\|x_{i}-\xi_{i}\right\|_{H_{00}} & <\frac{\varepsilon}{2 \sum_{j=1}^{n}\left|\alpha_{j}\right|\left\|\eta_{j}\right\|_{H_{10}}}, \\
\left\|A x_{i}-\bar{A} \xi_{i}\right\|_{H_{01}} & <\frac{\varepsilon}{2 \sum_{j=1}^{n}\left|\alpha_{j}\right|\left\|\bar{B} \eta_{j}\right\|_{H_{11}}} \\
\left\|y_{i}-\eta_{i}\right\|_{H_{10}} & <\frac{\varepsilon}{2 \sum_{j=1}^{n}\left|\alpha_{j}\right|\left\|x_{j}\right\|_{H_{00}}},
\end{aligned}
$$

and

$$
\left\|B y_{i}-\bar{B} \eta_{i}\right\|_{H_{11}}<\frac{\varepsilon}{2 \sum_{j=1}^{n}\left|\alpha_{j}\right|\left\|A x_{j}\right\|_{H_{01}}}
$$

hold. Setting $y:=\sum_{j=1}^{n} \alpha_{j} x_{j} \otimes y_{j} \in D(A \dot{\otimes} B)$ yields

$$
\begin{aligned}
\|y-x\|_{H_{00} \otimes H_{10}} & =\left\|\sum_{i=1}^{n} \alpha_{i}\left(x_{i} \otimes y_{i}-\xi_{i} \otimes \eta_{i}\right)\right\|_{H_{00} \otimes H_{10}} \\
& =\left\|\sum_{i=1}^{n} \alpha_{i}\left(x_{i} \otimes\left(y_{i}-\eta_{i}\right)+\left(x_{i}-\xi_{i}\right) \otimes \eta_{i}\right)\right\|_{H_{00} \otimes H_{10}} \\
& \leq \sum_{i=1}^{n}\left|\alpha_{i}\right|\left\|x_{i}\right\|_{H_{00}}\left\|y_{i}-\eta_{i}\right\|_{H_{10}}+\sum_{i=1}^{n}\left|\alpha_{i}\right|\left\|x_{i}-\xi_{i}\right\|_{H_{00}}\left\|e t a_{i}\right\|_{H_{10}}
\end{aligned}
$$

and

$$
\begin{aligned}
&\|(A \dot{\otimes} B) y-(\bar{A} \dot{\otimes} \bar{B}) x\|_{H_{01} \otimes H_{11}}=\left\|\sum_{i=1}^{n} \alpha_{i}\left(A x_{i} \otimes B y_{i}-\bar{A} \xi_{i} \otimes \bar{B} \eta_{i}\right)\right\|_{H_{00} \otimes H_{10}} \\
& \leq \sum_{i=1}^{n}\left|\alpha_{i}\right|\left\|A x_{i}\right\|_{H_{01}}\left\|B y_{i}-\bar{B} \eta_{i}\right\|_{H_{11}} \\
&+\sum_{i=1}^{n}\left|\alpha_{i}\right|\left\|A x_{i}-\bar{A} \xi_{i}\right\|_{H_{00}}\left\|\bar{B} \eta_{i}\right\|_{H_{10}} \\
&<\varepsilon
\end{aligned}
$$

Thus, $x \in D(A \otimes B)$ and $A \otimes B x=\bar{A} \dot{\otimes} \bar{B} x$, i.e., $\bar{A} \dot{\otimes} \bar{B} \subseteq A \otimes B$; thus,

$$
\bar{A} \otimes \bar{B} \subseteq A \otimes B
$$

Proposition 2.7. Let $H_{0}, H_{1}$, and $H_{2}$ be Hilbert spaces. Then,

$$
\left(H_{0} \otimes H_{1}\right) \otimes H_{2}=H_{0} \otimes\left(H_{1} \otimes H_{2}\right)=H_{0} \otimes H_{1} \otimes H_{2}
$$

in the sense of unitary equivalence.
Proof. For $\varphi \in H_{0}, \psi \in H_{1}$, and $\chi \in H_{2}$, we set

$$
U((\varphi \otimes \psi) \otimes \chi):=\varphi \otimes(\psi \otimes \chi)
$$

and extend this mapping to $\left(H_{0} \stackrel{a}{\otimes} H_{1}\right) \stackrel{a}{\otimes} H_{2}$ by

$$
\begin{aligned}
U:\left(H_{0} \stackrel{a}{\otimes} H_{1}\right) \stackrel{a}{\otimes} H_{2} & \rightarrow H_{0} \otimes\left(H_{1} \otimes H_{2}\right) \\
\sum_{j=1}^{m} \beta_{j}\left(\sum_{i=1}^{n_{j}} \alpha_{i}^{j} x_{i}^{j} \otimes y_{i}^{j}\right) \otimes z_{j} & \mapsto \sum_{j=1}^{m} \beta_{j} \sum_{i=1}^{n_{j}} \alpha_{i}^{j} x_{i}^{j} \otimes\left(y_{i}^{j} \otimes z_{j}\right) .
\end{aligned}
$$

First, we will prove that this extension is still right-unique. Let $\varphi \in\left(H_{0} \stackrel{a}{\otimes} H_{1}\right) \stackrel{a}{\otimes} H_{2}$ with

$$
\varphi=\sum_{j=1}^{m} \beta_{j}\left(\sum_{i=1}^{n_{j}} \alpha_{i}^{j} x_{i}^{j} \otimes y_{i}^{j}\right) \otimes z_{j}=\sum_{j=1}^{p} \delta_{j}\left(\sum_{i=1}^{k_{j}} \gamma_{i}^{j} u_{i}^{j} \otimes v_{i}^{j}\right) \otimes w_{j}
$$

Then, we observe for all $a \in H_{0}, b \in H_{1}$, and $c \in H_{2}$,

$$
\begin{aligned}
U\left(\sum_{j=1}^{m} \beta_{j}\left(\sum_{i=1}^{n_{j}} \alpha_{i}^{j} x_{i}^{j} \otimes y_{i}^{j}\right) \otimes z_{j}\right)(a, b \otimes c) & =\left(\sum_{j=1}^{m} \beta_{j} \sum_{i=1}^{n_{j}} \alpha_{i}^{j} x_{i}^{j} \otimes\left(y_{i}^{j} \otimes z_{j}\right)\right)(a, b \otimes c) \\
& =\sum_{j=1}^{m} \beta_{j} \sum_{i=1}^{n_{j}} \alpha_{i}^{j}\left\langle x_{i}^{j}, a\right\rangle_{H_{0}}\left\langle y_{i}^{j} \otimes z_{j}, b \otimes c\right\rangle_{H_{1} \otimes H_{2}} \\
& =\sum_{j=1}^{m} \beta_{j} \sum_{i=1}^{n_{j}} \alpha_{i}^{j}\left\langle x_{i}^{j}, a\right\rangle_{H_{0}}\left\langle y_{i}^{j}, b\right\rangle_{H_{1}}\left\langle z_{j}, c\right\rangle_{H_{2}} \\
& =\sum_{j=1}^{m} \beta_{j} \sum_{i=1}^{n_{j}} \alpha_{i}^{j}\left\langle x_{i}^{j} \otimes y_{i}^{j}, a \otimes b\right\rangle_{H_{0} \otimes H_{1}}\left\langle z_{j}, c\right\rangle_{H_{2}} \\
& =\sum_{j=1}^{m} \beta_{j}\left\langle\sum_{i=1}^{n_{j}} \alpha_{i}^{j} x_{i}^{j} \otimes y_{i}^{j}, a \otimes b\right\rangle_{H_{0} \otimes H_{1}}\left\langle z_{j}, c\right\rangle_{H_{2}} \\
& =\varphi(a \otimes b, c)
\end{aligned}
$$

The same calculation also shows

$$
U\left(\sum_{j=1}^{p} \delta_{j}\left(\sum_{i=1}^{k_{j}} \gamma_{i}^{j} u_{i}^{j} \otimes v_{i}^{j}\right) \otimes w_{j}\right)(a, b \otimes c)=\varphi(a \otimes b, c)
$$

Since these are continuous bi-linear functionals and $H_{0} \stackrel{a}{\otimes}\left(H_{1} \stackrel{a}{\otimes} H_{2}\right)$ is dense in $H_{0} \otimes$ ( $H_{1} \otimes H_{2}$ ), we conclude right-uniqueness of $U$.

Furthermore, $U$ is linear since, for $\kappa \in \mathbb{R}$ and $\varphi, \psi \in H_{0} \stackrel{a}{\otimes}\left(H_{1} \stackrel{a}{\otimes} H_{2}\right)$ with

$$
\varphi=\sum_{j=1}^{m} \beta_{j}\left(\sum_{i=1}^{n_{j}} \alpha_{i}^{j} x_{i}^{j} \otimes y_{i}^{j}\right) \otimes z_{j}
$$

and

$$
\psi=\sum_{j=1}^{p} \delta_{j}\left(\sum_{i=1}^{k_{j}} \gamma_{i}^{j} u_{i}^{j} \otimes v_{i}^{j}\right) \otimes w_{j}
$$

we obtain

$$
\kappa \varphi+\psi=\sum_{j=1}^{m+p} \zeta_{j}\left(\sum_{i=1}^{l_{j}} \eta_{i}^{j} \vartheta_{i}^{j} \otimes \lambda_{i}^{j}\right) \otimes \nu_{j}
$$

with

$$
\begin{aligned}
& \zeta_{j}= \begin{cases}\kappa \beta_{j} & , j \in \mathbb{N}_{\leq m} \\
\delta_{j-m} & , j \in[m+1, m+p] \cap \mathbb{N},\end{cases} \\
& l_{j}= \begin{cases}n_{j} & , j \in \mathbb{N}_{\leq m} \\
k_{j-m} & , j \in[m+1, m+p] \cap \mathbb{N},\end{cases} \\
& \eta_{i}^{j}=\left\{\begin{array}{ll}
\alpha_{i}^{j} & , j \in \mathbb{N}_{\leq m} \\
\gamma_{i}^{j-m} & , j \in[m+1, m+p] \cap \mathbb{N}
\end{array},\right. \\
& \vartheta_{i}^{j}=\left\{\begin{array}{ll}
x_{i}^{j} & , j \in \mathbb{N}_{\leq m} \\
u_{i}^{j-m} & , j \in[m+1, m+p] \cap \mathbb{N}
\end{array},\right. \\
& \lambda_{i}^{j}=\left\{\begin{array}{ll}
y_{i}^{j} & , j \in \mathbb{N}_{\leq m} \\
v_{i}^{j-m} & , j \in[m+1, m+p] \cap \mathbb{N}
\end{array},\right.
\end{aligned}
$$

and

$$
\nu_{j}= \begin{cases}z_{j} & , j \in \mathbb{N}_{\leq m} \\ w_{j-m} & , j \in[m+1, m+p] \cap \mathbb{N}\end{cases}
$$

Hence,

$$
\begin{aligned}
U(\kappa \varphi+\psi) & =\sum_{j=1}^{m+p} \zeta_{j} \sum_{i=1}^{l_{j}} \eta_{i}^{j} \vartheta_{i}^{j} \otimes\left(\lambda_{i}^{j} \otimes \nu_{j}\right) \\
& =\sum_{j=1}^{m} \zeta_{j} \sum_{i=1}^{l_{j}} \eta_{i}^{j} \vartheta_{i}^{j} \otimes\left(\lambda_{i}^{j} \otimes \nu_{j}\right)+\sum_{j=m+1}^{m+p} \zeta_{j} \sum_{i=1}^{l_{j}} \eta_{i}^{j} \vartheta_{i}^{j} \otimes\left(\lambda_{i}^{j} \otimes \nu_{j}\right) \\
& =\kappa \sum_{j=1}^{m} \beta_{j} \sum_{i=1}^{n_{j}} \alpha_{i}^{j} x_{i}^{j} \otimes\left(y_{i}^{j} \otimes z_{j}\right)+\sum_{j=1}^{p} \delta_{j} \sum_{i=1}^{k_{j}} \gamma_{i}^{j} u_{i}^{j} \otimes\left(v_{i}^{j} \otimes w_{j}\right) \\
& =\kappa U(\varphi)+U(\psi) .
\end{aligned}
$$

Now, we can show that $U$ is an isometry. This follows from
$\|U \varphi\|_{H_{0} \otimes\left(H_{1} \otimes H_{2}\right)}^{2}=\sum_{j, k=1}^{m} \beta_{j} \beta_{k} \sum_{i=1}^{n_{j}} \sum_{l=1}^{n_{k}} \alpha_{i}^{j} \alpha_{l}^{k}\left\langle x_{i}^{j} \otimes\left(y_{i}^{j} \otimes z_{j}\right), x_{l}^{k} \otimes\left(y_{l}^{k} \otimes z_{k}\right)\right\rangle_{H_{0} \otimes\left(H_{1} \otimes H_{2}\right)}$

$$
\begin{aligned}
& =\sum_{j, k=1}^{m} \beta_{j} \beta_{k} \sum_{i=1}^{n_{j}} \sum_{l=1}^{n_{k}} \alpha_{i}^{j} \alpha_{l}^{k}\left\langle x_{i}^{j}, x_{l}^{k}\right\rangle_{H_{0}}\left\langle y_{i}^{j}, y_{l}^{k}\right\rangle_{H_{1}}\left\langle z_{j}, z_{k}\right\rangle_{H_{2}} \\
& =\sum_{j, k=1}^{m} \beta_{j} \beta_{k}\left|\left(\sum_{i=1}^{n_{j}} \alpha_{i}^{j} x_{i}^{j} \otimes y_{i}^{j}\right) \otimes z_{j},\left(\sum_{l=1}^{n_{k}} \alpha_{l}^{k} x_{l}^{k} \otimes y_{l}^{k}\right) \otimes z_{k}\right|_{\left(H_{0} \otimes H_{1}\right) \otimes H_{2}} \\
& =\|\varphi\|_{\left(H_{0} \otimes H_{1}\right) \otimes H_{2}}^{2}
\end{aligned}
$$

Finally, if we show that $U$ has dense range, then we can extend $U$ to a unitary operator. Since $H_{0} \stackrel{a}{\otimes}\left(H_{1} \stackrel{a}{\otimes} H_{2}\right)$ is dense, it suffices to show that every

$$
\psi=: \sum_{j=1}^{m} \beta_{j} x_{j} \otimes\left(\sum_{i=1}^{n_{j}} \alpha_{i}^{j} y_{i}^{j} \otimes z_{i}^{j}\right) \in H_{0} \stackrel{a}{\otimes}\left(H_{1} \stackrel{a}{\otimes} H_{2}\right)
$$

is an image of $U$. Let

$$
\varphi:=\sum_{j=1}^{m} \beta_{j} \sum_{i=1}^{n_{j}} \alpha_{i}^{j}\left(x_{i} \otimes y_{i}^{j}\right) \otimes z_{i}^{j} \in\left(H_{0} \stackrel{a}{\otimes} H_{1}\right) \stackrel{a}{\otimes} H_{2} .
$$

Then, linearity of $U$ implies

$$
\begin{aligned}
U \varphi & =\sum_{j=1}^{m} \beta_{j} \sum_{i=1}^{n_{j}} \alpha_{i}^{j} U\left(\left(x_{i} \otimes y_{i}^{j}\right) \otimes z_{i}^{j}\right) \\
& =\sum_{j=1}^{m} \beta_{j} \sum_{i=1}^{n_{j}} \alpha_{i}^{j} x_{i} \otimes\left(y_{i}^{j} \otimes z_{i}^{j}\right) \\
& =\psi
\end{aligned}
$$

The other assertion, $H_{0} \otimes\left(H_{1} \otimes H_{2}\right)=H_{0} \otimes H_{1} \otimes H_{2}$, follows similarly.

Corollary 2.8. For $i \in\{0,1,2\}$ and $j \in\{0,1\}$, let $H_{i j}$ be a Hilbert space and $A_{0} \subseteq H_{00} \oplus H_{01}, A_{1} \subseteq H_{10} \oplus H_{11}$, and $A_{2} \subseteq H_{20} \oplus H_{21}$ densely defined closable linear operators. Then,

$$
\left(A_{0} \otimes A_{1}\right) \otimes A_{2}=A_{0} \otimes\left(A_{1} \otimes A_{2}\right)=A_{0} \otimes A_{1} \otimes A_{2}
$$

Proposition 2.9. For $i, j \in\{0,1\}$, let $H_{i j}$ be a Hilbert space, $A_{0} \in L\left(H_{00}, H_{01}\right)$, and $A_{1} \in L\left(H_{10}, H_{11}\right)$. Then, $A_{0} \otimes A_{1} \in L\left(H_{00} \otimes H_{10}, H_{01} \otimes H_{11}\right)$ with

$$
\left\|A_{0} \otimes A_{1}\right\|_{L\left(H_{00} \otimes H_{10}, H_{01} \otimes H_{11}\right)}=\left\|A_{0}\right\|_{L\left(H_{00}, H_{01}\right)}\left\|A_{1}\right\|_{L\left(H_{10}, H_{11}\right)}
$$

Proof. Let $S_{i j}$ be a complete orthonormal set in $H_{i j}$ and $x=\sum_{j=1}^{l} \kappa_{j} \varphi_{j} \otimes \psi_{j} \epsilon$ $H_{00} \stackrel{a}{\otimes} H_{10}$. Then, we can find sequences $\alpha^{j}, \beta^{j} \in \mathbb{R}^{\mathbb{N}}, \zeta \in S_{00}^{\mathbb{N}}$, and $\xi \in S_{10}^{\mathbb{N}}$ such that for every $j \in \mathbb{N}$

$$
\begin{aligned}
\kappa_{j} \varphi_{j} & =\sum_{n \in \mathbb{N}} \alpha_{n}^{j} \zeta_{n} \\
\psi_{j} & =\sum_{n \in \mathbb{N}} \beta_{n}^{j} \xi_{n}
\end{aligned}
$$

and, hence,

$$
x=\sum_{j=1}^{l}\left(\kappa_{j} \varphi_{j}\right) \otimes \psi_{j}=\sum_{m, n \in \mathbb{N}} \underbrace{\sum_{j=1}^{l} \alpha_{n}^{j} \beta_{m}^{j}}_{=: \gamma_{n m}} \zeta_{n} \otimes \xi_{m}
$$

hold. Since $A_{0}$ and $A_{1}$ are continuous, we observe

$$
A_{0} \otimes A_{1} x=\sum_{j=1}^{l} A_{0} \kappa_{j} \varphi_{j} \otimes A_{1} \psi_{j}
$$

$$
\begin{aligned}
& =\sum_{m, n \in \mathbb{N}} \sum_{j=1}^{n} \alpha_{n}^{j} \beta_{m}^{j} A_{0} \zeta_{n} \otimes A_{1} \xi_{m} \\
& =\sum_{m, n \in \mathbb{N}} \gamma_{n m} A_{0} \zeta_{n} \otimes A_{1} \xi_{m} .
\end{aligned}
$$

Let $y=\sum_{i=1}^{k} \lambda_{i} \sigma_{i} \otimes \tau_{i} \in H_{01} \stackrel{a}{\otimes} H_{11}$. Then, we can find sequences $\nu^{j}, \varrho^{j} \in \mathbb{R}^{\mathbb{N}}, \eta \in S_{01}^{\mathbb{N}}$, and $\vartheta \in S_{11}^{\mathbb{N}}$ such that for every $j \in \mathbb{N}$

$$
\begin{aligned}
\lambda_{j} \sigma_{j} & =\sum_{n \in \mathbb{N}} \nu_{n}^{j} \eta_{n} \\
\tau_{j} & =\sum_{n \in \mathbb{N}} \varrho_{n}^{j} \vartheta_{n}
\end{aligned}
$$

and, hence,

$$
y=\sum_{j=1}^{l}\left(\lambda_{j} \sigma_{j}\right) \otimes \tau_{j}=\sum_{m, n \in \mathbb{N}} \underbrace{\sum_{j=1}^{l} \nu_{n}^{j} \varrho_{m}^{j}}_{=: \delta_{n m}} \eta_{n} \otimes \vartheta_{m}
$$

hold. We, thus, observe

$$
\begin{aligned}
\left\langle A_{0} \otimes A_{1} x, y\right\rangle_{H_{01} \otimes H_{11}} & =\sum_{m, n, s, t \in \mathbb{N}} \gamma_{n m} \delta_{s t}\left\langle A_{0} \zeta_{n}, \eta_{s}\right\rangle_{H_{01}}\left\langle A_{1} \xi_{m}, \vartheta_{t}\right\rangle_{H_{11}} \\
& =\sum_{m, n, s, t \in \mathbb{N}} \gamma_{n m} \delta_{s t}\left\langle A_{0} \zeta_{n}, \eta_{s}\right\rangle_{H_{01}}\left\langle\xi_{m}, A_{1}^{*} \vartheta_{t}\right\rangle_{H_{11}} \\
& =\sum_{m, s \in \mathbb{N}}\left\langle A_{0} \sum_{n \in \mathbb{N}} \gamma_{n m} \zeta_{n}, \eta_{s}\right\rangle_{H_{01}}\left\langle\xi_{m}, A_{1}^{*} \sum_{t \in \mathbb{N}} \delta_{s t} \vartheta_{t}\right\rangle_{H_{11}}
\end{aligned}
$$

and, by Cauchy-Schwarz,

$$
\begin{aligned}
& \left|\left\langle A_{0} \otimes A_{1} x, y\right\rangle_{H_{01} \otimes H_{11}}\right|^{2} \\
\leq & \left(\sum_{m, s \in \mathbb{N}}\left\langle A_{0} \sum_{n \in \mathbb{N}} \gamma_{n m} \zeta_{n}, \eta_{s}\right\rangle_{H_{01}}\right)^{2}\left(\sum_{m, s \in \mathbb{N}}\left\langle\xi_{m}, A_{1}^{*} \sum_{t \in \mathbb{N}} \delta_{s t} \vartheta_{t}\right\rangle_{H_{11}}\right)^{2}
\end{aligned}
$$

which yields (using Bessel's inequality and orthonormality of $\zeta$ and $\vartheta$ )

$$
\begin{aligned}
& \left|\left\langle A_{0} \otimes A_{1} x, y\right\rangle_{H_{01} \otimes H_{11}}\right|^{2} \\
\leq & \left(\sum_{m \in \mathbb{N}}\left\|A_{0} \sum_{n \in \mathbb{N}} \gamma_{n m} \zeta_{n}\right\|_{H_{01}}^{2}\right)\left(\sum_{s \in \mathbb{N}}\left\|A_{1}^{*} \sum_{t \in \mathbb{N}} \delta_{s t} \vartheta_{t}\right\|_{H_{11}}^{2}\right) \\
\leq & \left\|A_{0}\right\|_{L\left(H_{00}, H_{01}\right)}^{2}\left\|A_{1}\right\|_{L\left(H_{10}, H_{11}\right)}^{2}\left(\sum_{m \in \mathbb{N}}\left\|\sum_{n \in \mathbb{N}} \gamma_{n m} \zeta_{n}\right\|_{H_{01}}^{2}\right)\left(\sum_{s \in \mathbb{N}}\left\|\sum_{t \in \mathbb{N}} \delta_{s t} \vartheta_{t}\right\|_{H_{11}}^{2}\right) \\
\leq & \left\|A_{0}\right\|_{L\left(H_{00}, H_{01}\right)}^{2}\left\|A_{1}\right\|_{L\left(H_{10}, H_{11}\right)}^{2}\left(\sum_{m, n \in \mathbb{N}}\left|\gamma_{n m}\right|^{2}\right)\left(\sum_{s, t \in \mathbb{N}}\left|\delta_{s t}\right|^{2}\right) \\
= & \left\|A_{0}\right\|_{L\left(H_{00}, H_{01}\right)}^{2}\left\|A_{1}\right\|_{L\left(H_{10}, H_{11}\right)}^{2}\|x\|_{H_{00} \otimes H_{10}}^{2}\|y\|_{H_{01} \otimes H_{11}}^{2} .
\end{aligned}
$$

For $y=A_{0} \otimes A_{1} x$, this implies

$$
\left\|A_{0} \otimes A_{1} x\right\|_{H_{01} \otimes H_{11}} \leq\left\|A_{0}\right\|_{L\left(H_{00}, H_{01}\right)}\left\|A_{1}\right\|_{L\left(H_{10}, H_{11}\right)}\|x\|_{H_{00} \otimes H_{10}}
$$

i.e.,

$$
\left\|A_{0} \otimes A_{1}\right\|_{L\left(H_{00} \otimes H_{10}, H_{01} \otimes H_{11}\right)} \leq\left\|A_{0}\right\|_{L\left(H_{00}, H_{01}\right)}\left\|A_{1}\right\|_{L\left(H_{10}, H_{11}\right)}
$$

On the other hand, let $x \in B_{H_{00}}^{\mathbb{N}}$ and $y \in B_{H_{10}}^{\mathbb{N}}$ with $\left\|A_{0} x_{n}\right\|_{H_{01}} \rightarrow\left\|A_{0}\right\|_{L\left(H_{00}, H_{01}\right)}$ and $\left\|A_{1} y_{n}\right\|_{H_{11}} \rightarrow\left\|A_{1}\right\|_{L\left(H_{10}, H_{11}\right)}$ for $n \rightarrow \infty$. Then, $\left(x_{n} \otimes y_{n}\right)_{n \in \mathbb{N}} \in B_{H_{00} \otimes H_{10}}^{\mathbb{N}}$ and $\left\|A_{0} \otimes A_{1} x_{n} \otimes y_{n}\right\|_{H_{01} \otimes H_{11}}=\left\|A_{0} x_{n}\right\|_{H_{01}}\left\|A_{1} y_{n}\right\|_{H_{11}} \rightarrow\left\|A_{0}\right\|_{L\left(H_{00}, H_{01}\right)}\left\|A_{1}\right\|_{L\left(H_{10}, H_{11}\right)}$
completes the proof.

Observation 2.10. For $i, j \in\{0,1\}$, let $H_{i j}$ be a Hilbert space and $A_{i} \in$ $L\left(H_{i 0}, H_{i 1}\right)$. Then, $A_{0} \otimes 1$ and $1 \otimes A_{1}$ commute. Furthermore, $\left(A_{0} \otimes 1\right)\left(1 \otimes A_{1}\right)$ and $\left(1 \otimes A_{1}\right)\left(A_{0} \otimes 1\right)$ are bounded operators.

Proof. Boundedness of $\left(A_{0} \otimes 1\right)\left(1 \otimes A_{1}\right)$ and $\left(1 \otimes A_{1}\right)\left(A_{0} \otimes 1\right)$ follows directly from boundedness of $A_{0} \otimes 1$ and $1 \otimes A_{1}$.

Let $x \in H_{00}$ and $y \in H_{01}$. Then,

$$
\begin{aligned}
\left(A_{0} \otimes 1\right)\left(1 \otimes A_{1}\right) x \otimes y & =\left(A_{0} \otimes 1\right) x \otimes A_{1} y \\
& =A_{0} x \otimes A_{1} y \\
& =\left(1 \otimes A_{1}\right) A_{0} x \otimes y \\
& =\left(1 \otimes A_{1}\right)\left(A_{0} \otimes 1\right) x \otimes y
\end{aligned}
$$

shows that the commutator [ $A_{0} \otimes 1,1 \otimes A_{1}$ ] vanishes on all algebraic tensors, i.e., $\left[A_{0} \otimes 1,1 \otimes A_{1}\right]=0$ by boundedness.

ObSERVATION 2.11. Let $H_{0}, H_{1}$, and $H_{2}$ be Hilbert spaces, and $A \subseteq H_{1} \oplus H_{2}$ a densely defined closable linear operator. Then,

$$
(1 \otimes A)^{*}=1 \otimes A^{*}
$$

and

$$
(A \otimes 1)^{*}=A^{*} \otimes 1
$$

Proof. So far, we know

$$
1 \otimes A^{*} \subseteq(1 \otimes \bar{A})^{*}=(1 \otimes A)^{*}
$$

To show the missing inclusion let $x \in D\left((1 \otimes A)^{*}\right) \subseteq H_{0} \otimes H_{2}$ and $S_{i} \subseteq H_{i}$ an orthonormal basis for $i \in\{0,1,2\}$. Then, $\left[S_{0}\right] \otimes\left[S_{j}\right]$ is an orthonormal basis of $H_{0} \otimes H_{j}$ for $j \in\{1,2\}$. Hence, there are sequences $\xi \in S_{0}^{\mathbb{N}}, \zeta \in S_{1}^{\mathbb{N}}$, and $\eta \in S_{2}^{\mathbb{N}}$ such that

$$
x=\sum_{n \in \mathbb{N}}\left\langle\xi_{n} \otimes \eta_{n}, x\right\rangle_{H_{0} \otimes H_{2}} \xi_{n} \otimes \eta_{n}
$$

and

$$
(1 \otimes A)^{*} x=\sum_{n \in \mathbb{N}}\left\langle\xi_{n} \otimes \zeta_{n},(1 \otimes A)^{*} x\right\rangle_{H_{0} \otimes H_{1}} \xi_{n} \otimes \zeta_{n}
$$

For $s \in S_{0}$ and $u \in D(A)$, we obtain

$$
\langle(1 \otimes A)(s \otimes u), x\rangle_{H_{0} \otimes H_{2}}=\sum_{n \in \mathbb{N}}\left\langle\xi_{n} \otimes \zeta_{n},(1 \otimes A)^{*} x\right\rangle_{H_{0} \otimes H_{1}}\left\langle s \otimes u, \xi_{n} \otimes \zeta_{n}\right\rangle_{H_{0} \otimes H_{1}}
$$

which is equivalent to

$$
\begin{aligned}
& \sum_{n \in \mathbb{N}}\left\langle\xi_{n} \otimes \eta_{n}, x\right\rangle_{H_{0} \otimes H_{2}}\left\langle s \otimes A u, \xi_{n} \otimes \eta_{n}\right\rangle_{H_{0} \otimes H_{2}} \\
= & \sum_{n \in \mathbb{N}}\left\langle\xi_{n} \otimes \zeta_{n},(1 \otimes A)^{*} x\right\rangle_{H_{0} \otimes H_{1}}\left\langle s \otimes u, \xi_{n} \otimes \zeta_{n}\right\rangle_{H_{0} \otimes H_{1}} .
\end{aligned}
$$

With $s=\xi_{i}$, this implies

$$
\begin{aligned}
\left\langle A u,\left\langle\xi_{i} \otimes \eta_{i}, x\right\rangle_{H_{0} \otimes H_{2}} \eta_{i}\right\rangle_{H_{2}} & =\left\langle\xi_{i} \otimes \eta_{i}, x\right\rangle_{H_{0} \otimes H_{2}}\left\langle A u, \eta_{i}\right\rangle_{H_{2}} \\
& =\left\langle\xi_{i} \otimes \zeta_{i},(1 \otimes A)^{*} x\right\rangle_{H_{0} \otimes H_{1}}\left\langle u, \xi_{i} \otimes \zeta_{i}\right\rangle_{H_{1}} \\
& =\left\langle u,\left\langle\xi_{i} \otimes \zeta_{i},(1 \otimes A)^{*} x\right\rangle_{H_{0} \otimes H_{1}} \xi_{i} \otimes \zeta_{i}\right\rangle_{H_{1}}
\end{aligned}
$$

for all $u \in D(A)$, that is, $\left\langle\xi_{i} \otimes \eta_{i}, x\right\rangle_{H_{0} \otimes H_{2}} \eta_{i} \in D\left(A^{*}\right)$ and

$$
A^{*}\left\langle\xi_{i} \otimes \eta_{i}, x\right\rangle_{H_{0} \otimes H_{2}} \eta_{i}=\left\langle\xi_{i} \otimes \zeta_{i},(1 \otimes A)^{*} x\right\rangle_{H_{0} \otimes H_{1}} \zeta_{i}
$$

for every $i \in \mathbb{N}$. Thus,

$$
\sum_{i=1}^{m}\left\langle\xi_{i} \otimes \eta_{i}, x\right\rangle_{H_{0} \otimes H_{2}} \xi_{i} \otimes \eta_{i} \in H_{0} \otimes D\left(A^{*}\right) \subseteq D\left(1 \otimes A^{*}\right)
$$

and

$$
\left(1 \otimes A^{*}\right) \sum_{i=1}^{m}\left\langle\xi_{i} \otimes \eta_{i}, x\right\rangle_{H_{0} \otimes H_{2}} \xi_{i} \otimes \eta_{i}=\left\langle\xi_{i} \otimes \zeta_{i},(1 \otimes A)^{*} x\right\rangle_{H_{0} \otimes H_{1}} \xi_{i} \otimes \zeta_{i}
$$

holds for all $m \in \mathbb{N}$. Since $1 \otimes A^{*}$ is closed, we conclude that $x \in D\left(1 \otimes A^{*}\right)$ and $\left(1 \otimes A^{*}\right) x=(1 \otimes A)^{*} x$.

The other identity follows similarly.

Observation 2.12. Let $H_{0} \neq\{0\}, H_{1}$, and $H_{2}$ be Hilbert spaces, and $A \subseteq$ $H_{1} \oplus H_{2}$ a densely defined closable operator. Then, $1 \otimes A$ is continuously invertible if and only if $A$ is continuously invertible. In that case,

$$
(1 \otimes A)^{-1}=1 \otimes A^{-1}
$$

holds.
Proof. Let $A$ be continuously invertible. Then, $1 \otimes A^{-1}$ is a bounded operator. Let $x \in H_{0}$ and $y \in H_{1}$. Then,

$$
\left(1 \otimes A^{-1}\right)(1 \otimes A) x \otimes y=\left(1 \otimes A^{-1}\right) x \otimes A y=x \otimes A^{-1} A y=x \otimes y
$$

shows

$$
\left.\left(1 \otimes A^{-1}\right)(1 \otimes A)\right|_{D(1 \dot{\otimes} A)}=\left.1\right|_{D(1 \dot{\otimes} A)} .
$$

Let $x \in D(1 \otimes A)$ and $\left(x_{n}\right)_{n \in \mathbb{N}} \in D(1 \dot{\otimes} A)$ such that $x_{n} \rightarrow x$ in $H_{0} \otimes H_{1}$ and $(1 \otimes$ $A) x_{n} \rightarrow(1 \otimes A) x$ in $H_{0} \otimes H_{2}$. Then, continuity of $1 \otimes A^{-1}$ implies

$$
x_{n}=\left(1 \otimes A^{-1}\right)(1 \otimes A) x_{n} \rightarrow\left(1 \otimes A^{-1}\right)(1 \otimes A) x,
$$

i.e.,

$$
\left.\left(1 \otimes A^{-1}\right)(1 \otimes A)\right|_{D(1 \otimes A)}=\left.1\right|_{D(1 \otimes A)} .
$$

Furthermore, for $y \in H_{2}$,

$$
(1 \otimes A)\left(1 \otimes A^{-1}\right) x \otimes y=(1 \otimes A) x \otimes A^{-1} y=x \otimes A A^{-1} y=x \otimes y
$$

shows

$$
\left.(1 \otimes A)\left(1 \otimes A^{-1}\right)\right|_{H_{0} \otimes H_{2}}=\left.1\right|_{H_{0} \otimes H_{2}} .
$$

Let $y \in H_{0} \otimes H_{2}$ and $\left(y_{n}\right)_{n \in \mathbb{N}} \in\left(H_{0}{ }^{\otimes} H_{2}\right)$ with $y_{n} \rightarrow y$ in $H_{0} \otimes H_{2}$. Then, $(1 \otimes$ $\left.A^{-1}\right) y_{n} \rightarrow\left(1 \otimes A^{-1}\right) y$ in $H_{0} \otimes H_{1}$ by continuity, and

$$
\left(1 \otimes A^{-1}\right) y_{n} \in D(1 \otimes A)
$$

and

$$
(1 \otimes A)\left(1 \otimes A^{-1}\right) y_{n}=y_{n}
$$

hold. Since $1 \otimes A$ is closed, it follows

$$
\left(1 \otimes A^{-1}\right) y \in D(1 \otimes A)
$$

and

$$
(1 \otimes A)\left(1 \otimes A^{-1}\right) y=y .
$$

Hence, $\left(1 \otimes A^{-1}\right)$ is bounded left- and right-inverse of $1 \otimes A$, i.e.,

$$
1 \otimes A^{-1}=(1 \otimes A)^{-1} \in L\left(H_{0} \otimes H_{2}, H_{0} \otimes H_{1}\right)
$$

Let us now assume that $1 \otimes A$ is continuously invertible. Let $x \in H_{0} \backslash\{0\}$ and $y \in[\{0\}] A$. Then, $1 \otimes A x \otimes y=x \otimes A y=x \otimes 0=0$. Since $1 \otimes A$ is injective, this implies $x \otimes y=0$, i.e.,

$$
\forall(\varphi, \psi) \in H_{0} \times H_{1}:\langle x, \varphi\rangle_{H_{0}}\langle y, \psi\rangle_{H_{1}}=x \otimes y(\varphi, \psi)=0
$$

In particular, $\varphi=x$ and $\psi=y$ implies

$$
\|x\|_{H_{0}}^{2}\|y\|_{H_{1}}^{2}=0
$$

and, hence, $y=0$ since we assumed $x \neq 0$. In other words, $A$ is injective.
Let $y \in A\left[H_{1}\right]^{\perp}$. Then, for all $z \in D(A)$, we obtain $\langle A z, y\rangle_{H_{2}}=0$ and, therefore,

$$
\forall \tilde{x} \in H_{0} \forall z \in D(A):\langle(1 \otimes A) \tilde{x} \otimes z, x \otimes y\rangle=0
$$

which implies

$$
\forall \xi \in H_{0} \stackrel{a}{\otimes} D(A):\langle(1 \otimes A) \xi, x \otimes y\rangle=0
$$

and, by continuity of the inner product,

$$
\forall \xi \in D(1 \otimes A):\langle(1 \otimes A) \xi, x \otimes y\rangle=0
$$

Hence, $x \otimes y \in(1 \otimes A)\left[H_{0} \otimes H_{1}\right]^{\perp}=\{0\}$. Since $x$ was assumed non-zero, this implies $y=0$, i.e., $A$ has dense range. Thus, it suffices to show continuity of $A^{-1}$ to prove the assertion. Since

$$
\left.(1 \otimes A)^{-1}\right|_{\left[H_{0}\right] \otimes\left[A\left[H_{1}\right]\right]}=\left.\left(1 \otimes A^{-1}\right)\right|_{\left[H_{0}\right] \otimes\left[A\left[H_{1}\right]\right]}
$$

we obtain, for $y \in A\left[H_{1}\right]$,

$$
\begin{aligned}
\left\|A^{-1} y\right\|_{H_{1}} & =\frac{\|x\|_{H_{0}}\left\|A^{-1} y\right\|_{H_{1}}}{\|x\|_{H_{0}}} \\
& =\frac{1}{\|x\|_{H_{0}}}\left\|x \otimes A^{-1} y\right\|_{H_{0} \otimes H_{1}} \\
& =\frac{1}{\|x\|_{H_{0}}}\left\|(1 \otimes A)^{-1} x \otimes y\right\|_{H_{0} \otimes H_{1}} \\
& \leq \frac{\left\|(1 \otimes A)^{-1}\right\|_{L\left(H_{0} \otimes H_{2}, H_{0} \otimes H_{1}\right)}\|x \otimes y\|_{H_{0} \otimes H_{2}}}{\|x\|_{H_{0}}} \\
& \leq\left\|(1 \otimes A)^{-1}\right\|_{L\left(H_{0} \otimes H_{2}, H_{0} \otimes H_{1}\right)}\|y\|_{H_{2}} .
\end{aligned}
$$

Remark We will use the notation of tensor products in cases where the spaces involved are not Hilbert spaces themselves; e.g., $C^{k}\left(M ; H_{1}\right) \otimes C^{k}\left(M ; H_{2}\right)$ with $H_{1}, H_{2}$ Hilbert spaces. By writing this we mean to consider $C^{k}\left(M ; H_{1} \otimes H_{2}\right)$.

## CHAPTER 3

## $L_{p}$ spaces on $C^{1,1}$-manifolds

Throughout these notes, unless explicitly stated otherwise, let $(M, g)$ be an orientable real 3-dimensional Riemannian $C^{1,1}$-manifold ${ }^{1}$ endowed with a connection $\nabla$. Then, the tangent bundle $T M$ is a Riemannian ( $2 \operatorname{dim} M$ )-manifold and a Hausdorff space itself. Furthermore, $\left(g_{i}(x)\right)_{i \in \mathbb{N}_{\leq \operatorname{dim} M}}$ will always be a local basis of $T_{x} M$ and $\left(g^{i}(x)\right)_{i \in \mathbb{N}_{\leq \operatorname{dim} M}}$ the corresponding dual basis in $T_{x} M^{*} . d \mathrm{vol}_{M}$ will denote the volume form on $M, G$ the Gramian matrix and $\gamma:=\sqrt{\operatorname{det} G}$.

Let $k \in \mathbb{N}_{0} \cup\{\infty, \omega\}, j, N \in \mathbb{N}$ and $\alpha, \beta \in \mathbb{N}_{0}^{N}$. Then, using $f \in C^{\omega}(A ; B): \Leftrightarrow$ $f: A \rightarrow B$ analytic wherever this makes any kind of sense, we define the following spaces

$$
\begin{aligned}
& \mathfrak{X}_{k}(M):=\left\{f \in C^{k, 1}(M ; T M) ; \forall x \in M: f(x) \in T_{x} M\right\} \text { locally Lipschitz } \\
& \text { vector fields } \\
& \mathfrak{X}_{k, c}(M):=\left\{f \in X_{k}(M) ; \operatorname{spt} f \text { compact in int } M\right\} \\
& \mathfrak{T}_{x}(\alpha, \beta):=\bigotimes_{i=1}^{N}\left(\left(\otimes_{j=1}^{\alpha_{i}} T_{x} M\right) \otimes\left(\otimes_{l=1}^{\beta_{i}} T_{x} M^{*}\right)\right) \\
& \mathfrak{T}(\alpha, \beta):=\cup_{x \in M} \mathfrak{T}_{x}(\alpha, \beta) \cong \cup_{x \in M}\{x\} \times \mathfrak{T}_{x}(\alpha, \beta) \\
& \mathfrak{T}^{*}(\alpha, \beta):=\cup_{x \in M} \mathfrak{T}_{x}(\alpha, \beta)^{*} \cong \bigcup_{x \in M}\{x\} \times \mathfrak{T}_{x}(\alpha, \beta)^{*} \\
& \mathfrak{M}_{k}^{(\alpha, \beta)}(M):=\left\{f \in C^{k, 1}\left(M ; \mathfrak{T}^{*}(\alpha, \beta)\right) ; \forall x \in M: \quad f(x) \in \mathfrak{T}_{x}(\alpha, \beta)^{*}\right\} \\
& (\alpha, \beta) \text {-tensor fields } \\
& \mathfrak{M}_{k, c}^{(\alpha, \beta)}(M):=\left\{f \in M_{k}^{(\alpha, \beta)}(M) ; \operatorname{spt} f \text { compact in int } M\right\} \\
& S^{j}:=\left\{\sigma ; \sigma: \mathbb{N}_{\leq j} \rightarrow \mathbb{N}_{\leq j} \text { injective }\right\} \text { permutations } \\
& \mathfrak{A}^{j}(x):=\left\{f \in\left(T_{x} M^{j}\right)^{*} ; \forall X \in T_{x} M^{j} \forall \sigma \in S^{j}: f(X)=\operatorname{sgn} \sigma f(X \circ \sigma)\right\} \\
& \text { alternating linear forms } \\
& \mathfrak{A}^{j}(M):=\cup_{x \in M} \mathfrak{A}^{j}(x) \\
& \Lambda_{k}^{j}(M):=\left\{f \in C^{k, 1}\left(M ; \mathfrak{A}^{j}(M)\right) ; \forall x \in M: f(x) \in \mathfrak{A}^{j}(x)\right\} C^{k}-j \text {-forms }
\end{aligned}
$$

Remark The tensor bundles $\mathfrak{T}(\alpha, \beta)$ and $\mathfrak{T}^{*}(\alpha, \beta)$ are topological spaces. Unfortunately, the topologies are far from adorable. But it is possible to show that if $U \subseteq M$ is open and contractible ${ }^{2}$, then there is a diffeomorphism from $\cup_{x \in U} \mathfrak{T}_{x}(\alpha, \beta)$ (or $\cup_{x \in U} \mathfrak{T}_{x}(\alpha, \beta)^{*}$, respectively) to $U \times \mathbb{R}^{m}$ with $m=(\operatorname{dim} M)^{\sum_{i=1}^{N} \alpha_{i}+\beta_{i}}$. This is strongly liked to local trivializations of $\mathfrak{T}(\alpha, \beta)$ and $\mathfrak{T}^{*}(\alpha, \beta)$. For notational simplicity we will only consider the case $\mathfrak{T}(m, 0)$ as all other cases can be constructed using Riesz identifications. Then, the bundle projection $\pi: \mathfrak{T}(m, 0) \rightarrow M$ is defined by

$$
\forall(x, v) \in\{x\} \times \mathfrak{T}_{x}(m, 0): \pi(x, v)=x
$$

and, given an atlas $\left(U_{i}, \psi_{i}\right)_{i \in I}$, we define

$$
\forall i \in I: \varphi_{i}:\left[U_{i}\right] \pi \rightarrow U_{i} \times \mathbb{R}^{m} ;\left(x, v_{i} g^{i}\right) \mapsto\left(x,\left(v_{1}, \ldots, v_{m}\right)\right)
$$

[^3]These $\varphi_{i}$ are vector space isomorphisms and the $\left(U_{i}, \varphi_{i}\right)$ locally trivialize $\mathfrak{T}(m, 0)$. In fact, this property is very important as for a vector bundle to be locally trivializable ensures the existence of global cross sections with maximal regularity, i.e., $\mathfrak{M}_{k}^{(\alpha, \beta)}(M)$ is non-trivial if $M$ is a $C^{k}$-manifold.

LEmMA 3.1. $\mathfrak{M}_{k}^{(0,0)}(M) \cong C^{k, 1}(M ; \mathbb{R}), \mathfrak{M}_{k}^{(1,0)}(M) \cong \Lambda_{k}^{1}(M), \mathfrak{M}_{k}^{(0,1)}(M) \cong$ $\mathfrak{X}_{k}(M)$ and $\mathfrak{M}_{k}^{(\alpha, \beta)}(M) \otimes \mathfrak{M}_{k}^{\left(\alpha^{\prime}, \beta^{\prime}\right)}(M)=\mathfrak{M}_{k}^{\left(\alpha \oplus \alpha^{\prime}, \beta \oplus \beta^{\prime}\right)}(M)$

Proof.
(i)

$$
\begin{aligned}
\mathfrak{M}_{k}^{(0,0)}(M) & =\left\{f \in C^{k, 1}\left(M ; \mathfrak{T}(0,0)^{*}\right) ; \forall x \in M: f(x) \in \mathfrak{T}_{x}(0,0)^{*}\right\} \\
& =\left\{f \in C^{k, 1}\left(M ; \mathfrak{T}(0,0)^{*}\right) ; \forall x \in M: f(x) \in\left(\bigotimes_{\varnothing} \mathfrak{T}_{x} M \otimes \bigotimes_{\varnothing} T_{x} M^{*}\right)^{*}\right\} \\
& \cong\left\{f \in C^{k, 1}\left(M ; \mathfrak{T}(0,0)^{*}\right) ; \forall x \in M: f(x) \in(\mathbb{R} \otimes \mathbb{R})^{*}\right\} \\
& \cong C^{k, 1}(M ; \mathbb{R})
\end{aligned}
$$

(ii)

$$
\begin{aligned}
\mathfrak{M}_{k}^{(1,0)}(M) & =\left\{f \in C^{k, 1}\left(M ; \mathfrak{T}(1,0)^{*}\right) ; \forall x \in M: f(x) \in \mathfrak{T}_{x}(1,0)^{*}\right\} \\
& \cong\left\{f \in C^{k, 1}\left(M ; \mathfrak{T}(1,0)^{*}\right) ; \forall x \in M: f(x) \in \mathfrak{T}_{x} M^{*}\right\} \\
& =\left\{f \in C^{k, 1}\left(M ; \mathfrak{A}^{1}(M)\right) ; \forall x \in M: f(x) \in \mathfrak{A}^{1}(x)\right\} \\
& =\Lambda_{k}^{1}(M)
\end{aligned}
$$

(iii)

$$
\begin{aligned}
\mathfrak{M}_{k}^{(0,1)}(M) & =\left\{f \in C^{k, 1}\left(M ; \mathfrak{T}(0,1)^{*}\right) ; \forall x \in M: f(x) \in \mathfrak{T}_{x}(0,1)^{*}\right\} \\
& \cong\left\{f \in C^{k, 1}\left(M ; \mathfrak{T}(0,1)^{*}\right) ; \forall x \in M: f(x) \in \mathfrak{T}_{x} M^{* *}\right\} \\
& \cong\left\{f \in C^{k, 1}(M ; T M) ; \forall x \in M: f(x) \in \mathfrak{T}_{x} M\right\} \\
& =\mathfrak{X}_{k}(M)
\end{aligned}
$$

(iv)

$$
\begin{aligned}
\mathfrak{M}_{k}^{(\alpha, \beta)}(M) \otimes \mathfrak{M}_{k}^{\left(\alpha^{\prime}, \beta^{\prime}\right)}(M)= & \left\{f \in C^{k, 1}\left(M ; \mathfrak{T}(\alpha, \beta)^{*} \otimes \mathfrak{T}\left(\alpha^{\prime}, \beta^{\prime}\right)^{*}\right) ;\right. \\
& \left.\forall x \in M: f(x) \in \mathfrak{T}_{x}(\alpha, \beta)^{*} \otimes \mathfrak{T}_{x}\left(\alpha^{\prime}, \beta^{\prime}\right)^{*}\right\} \\
= & \left\{f \in C^{k, 1}\left(M ; \mathfrak{T}(\alpha, \beta)^{*} \otimes \mathfrak{T}\left(\alpha^{\prime}, \beta^{\prime}\right)^{*}\right) ;\right. \\
& \left.\forall x \in M: f(x) \in \mathfrak{T}_{x}\left(\alpha \oplus \alpha^{\prime}, \beta \oplus \beta^{\prime}\right)^{*}\right\} \\
= & \left\{f \in C^{k, 1}\left(M ; \mathfrak{T}\left(\alpha \oplus \alpha^{\prime}, \beta \oplus \beta^{\prime}\right)^{*}\right) ;\right. \\
& \left.\forall x \in M: f(x) \in \mathfrak{T}_{x}\left(\alpha \oplus \alpha^{\prime}, \beta \oplus \beta^{\prime}\right)^{*}\right\} \\
= & \mathfrak{M}_{k}^{\left(\alpha \oplus \alpha^{\prime}, \beta \oplus \beta^{\prime}\right)}(M)
\end{aligned}
$$

Recall that the volume form $d \operatorname{vol}_{M}$ defines a measure on the Borel sets $\mathcal{B}(M)$ by

$$
\forall B \in \mathcal{B}(M): \mu(B):=\int_{B} d \operatorname{vol}_{M}
$$

Using this interpretation we are in the realm of Lebesgue-integrals.

Definition 3.2. Let $p \in \mathbb{R}_{\geq 1}, n \in \mathbb{N}, \alpha, \beta \in \mathbb{N}_{0}^{n}$ and $\langle\cdot, \cdot\rangle_{(\alpha, \beta)}$ be the canonical scalar form on $\mathfrak{M}_{0}^{(\alpha, \beta)}(M)$, i.e. for $x, y \in \mathfrak{M}_{0}^{(\alpha, \beta)}(M)$

$$
\forall p \in M:\langle x, y\rangle_{(\alpha, \beta)}(p)=\langle x(p), y(p)\rangle_{\mathfrak{T}_{p}(\alpha, \beta)^{*}}
$$

holds. We define

$$
\|\cdot\|_{p, \alpha, \beta}: \mathfrak{M}_{0}^{(\alpha, \beta)}(M) \rightarrow \mathbb{R} ; x \mapsto\left(\int_{M}\left|\langle x, x\rangle_{(\alpha, \beta)}\right|^{\frac{p}{2}} d \operatorname{vol}_{M}\right)^{\frac{1}{p}}
$$

and

$$
L_{p}^{(\alpha, \beta)}(M):=\overline{\mathfrak{M}_{0}^{(\alpha, \beta)}(M)}\|\cdot\|_{p, \alpha, \beta}
$$

We will denote measurable and p-integrable functions, i.e. those functions being $\|\cdot\|_{p, \alpha, \beta}$-limits of continuous functions, by $\mathcal{L}_{p}^{(\alpha, \beta)}(M)$.
Remark Fischer-Riesz's theorem (theorem 3.3 below) allows us identify elements of $L_{p}^{(\alpha, \beta)}(M)$ with a functions.

Obviously all $L_{p}^{(\alpha, \beta)}(M)$ are Banach spaces and $L_{2}^{(\alpha, \beta)}(M)$ are Hilbert spaces, since, $\langle x, x\rangle_{(\alpha, \beta)}$ is non-negative and $\langle x, y\rangle_{L_{2}^{(\alpha, \beta)}(M)}=\int_{M}\langle x, y\rangle_{(\alpha, \beta)} d \mathrm{vol}_{M}$ is a scalar product.

ThEOREM 3.3 (Fischer-Riesz). Let $p \in \mathbb{R}_{\geq 1}$ and $\left(f_{n}\right)_{n \in \mathbb{N}} \in \mathcal{L}_{p}^{(\alpha, \beta)}(M)^{\mathbb{N}}$ converging to $f \in \mathcal{L}_{p}^{(\alpha, \beta)}(M)$ in $\mathcal{L}_{p}^{(\alpha, \beta)}(M)$. Then there is $g \in \mathcal{L}_{p}\left(\operatorname{vol}_{M}\right)$ and a subsequence $\left(f_{n_{j}}\right)_{j \in \mathbb{N}}$ of $\left(f_{n}\right)_{n \in \mathbb{N}}$, such that
(i) $f_{n_{j}} \rightarrow f \mu$-almost everywhere
and
(ii) $\forall j \in \mathbb{N}:\left|\left\langle f_{n_{j}}, f_{n_{j}}\right\rangle_{(\alpha, \beta)}\right|^{\frac{1}{2}} \leq g$
hold.
Proof. Choose any subsequence $\left(f_{n_{j}}\right)_{j \in \mathbb{N}}$ of $\left(f_{n}\right)_{n \in \mathbb{N}}$ satisfying

$$
\forall j \in \mathbb{N}:\left\|f_{n_{j+1}}-f_{n_{j}}\right\|_{\mathcal{L}_{p}^{(\alpha, \beta)}(M)} \leq 2^{-j}
$$

For $j \in \mathbb{N}$ let $\tilde{f}_{j}:=f_{n_{j+1}}-f_{n_{j}}$. Then, for $k \in \mathbb{N}$,

$$
\begin{aligned}
\left(\int\left(\sum_{j=1}^{k}\left|\left\langle\tilde{f}_{j}, \tilde{f}_{j}\right\rangle_{(\alpha, \beta)}\right|^{\frac{1}{2}}\right)^{p} d \operatorname{vol}_{M}\right)^{\frac{1}{p}} & =\left\|\sum_{j=1}^{k}\left|\left\langle\tilde{f}_{j}, \tilde{f}_{j}\right\rangle_{(\alpha, \beta)}\right|^{\frac{1}{2}}\right\|_{L_{p}\left(\operatorname{vol}_{M}\right)} \\
& \leq \sum_{j=1}^{k}\left\|\left|\left\langle\tilde{f}_{j}, \tilde{f}_{j}\right\rangle_{(\alpha, \beta)}\right|^{\frac{1}{2}}\right\|_{L_{p}\left(\operatorname{vol}_{M}\right)} \\
& =\sum_{j=1}^{k}\left\|f_{n_{j+1}}-f_{n_{j}}\right\|_{\mathcal{L}_{p}^{(\alpha, \beta)}(M)} \\
& \leq \sum_{j \in \mathbb{N}} 2^{-j} \\
& =1
\end{aligned}
$$

and

$$
\left(\sum_{j=1}^{k}\left|\left\langle\tilde{f}_{j}, \tilde{f}_{j}\right\rangle_{(\alpha, \beta)}\right|^{\frac{1}{2}}\right)^{p} \nearrow\left(\sum_{j \in \mathbb{N}}\left|\left\langle\tilde{f}_{j}, \tilde{f}_{j}\right\rangle_{(\alpha, \beta)}\right|^{\frac{1}{2}}\right)^{p}
$$

hold. Thus, using dominated convergence, we find

$$
\int\left(\sum_{j \in \mathbb{N}}\left|\left\langle\tilde{f}_{j}, \tilde{f}_{j}\right\rangle_{(\alpha, \beta)}\right|^{\frac{1}{2}}\right)^{p} d \operatorname{vol}_{M}<\infty
$$

Hence, $\tilde{g}:=\lim _{k \rightarrow \infty} \sum_{j=1}^{k}\left|\left\langle\tilde{f}_{j}, \tilde{f}_{j}\right\rangle_{(\alpha, \beta)}\right|^{\frac{1}{2}} \in \mathcal{L}_{p}\left(\operatorname{vol}_{M}\right)$ exists with $\tilde{g}<\infty \operatorname{vol}_{M}$-almost everywhere. Since $\mathcal{L}_{p}^{(\alpha, \beta)}(M)$ is a Banach space, $\sum_{j \in \mathbb{N}} \tilde{f}_{j}$ converges $\operatorname{vol}_{M}$-almost everywhere absolutely; $\sum_{j \in \mathbb{N}} \tilde{f}_{j}=: \tilde{f}$. By definition of $\left(\tilde{f}_{j}\right)_{j \in \mathbb{N}}$ we find

$$
\tilde{f} \leftarrow \sum_{j=1}^{k} \tilde{f}_{j}=\sum_{j=1}^{k}\left(f_{n_{j+1}}-f_{n_{j}}\right)=f_{n_{k+1}}-f_{n_{1}} \rightarrow f-f_{n_{1}}
$$

and, hence, $f_{n_{k+1}}=f_{n_{1}}+\sum_{j=1}^{k} \tilde{f}_{j} \rightarrow f \operatorname{vol}_{M}$-almost everywhere. Furthermore, for $j \in \mathbb{N}$

$$
\tilde{g}^{2} \geq\left\langle\tilde{f}_{j}, \tilde{f}_{j}\right\rangle_{(\alpha, \beta)}=\left\langle f_{n_{j+1}}, f_{n_{j+1}}\right\rangle_{(\alpha, \beta)}-2\left\langle f_{n_{j+1}}, f_{n_{j}}\right\rangle_{(\alpha, \beta)}+\left\langle f_{n_{j}}, f_{n_{j}}\right\rangle_{(\alpha, \beta)}
$$

holds. Hence,

$$
0 \leq\left\langle f_{n_{j}}, f_{n_{j}}\right\rangle_{(\alpha, \beta)} \leq \underbrace{\tilde{g}^{2}}_{\in \mathcal{L}_{\frac{p}{2}}\left(\operatorname{vol}_{M}\right)}-\underbrace{\left\langle f_{n_{j+1}}, f_{n_{j+1}}\right\rangle_{(\alpha, \beta)}}_{\epsilon \mathcal{L}_{\frac{p}{2}}\left(\operatorname{vol}_{M}\right)}+\underbrace{2\left\langle f_{n_{j+1}}, f_{n_{j}}\right\rangle_{(\alpha, \beta)}}_{\epsilon \mathcal{L}_{\frac{p}{2}}\left(\operatorname{vol}_{M}\right)} \in \mathcal{L}_{\frac{p}{2}}\left(\operatorname{vol}_{M}\right)
$$

yields (ii) with $g:=\left(\tilde{g}^{2}-\left\langle f_{n_{j+1}}, f_{n_{j+1}}\right\rangle_{(\alpha, \beta)}+2\left\langle f_{n_{j+1}}, f_{n_{j}}\right\rangle_{(\alpha, \beta)}\right)^{\frac{1}{2}} \in \mathcal{L}_{p}\left(\operatorname{vol}_{M}\right)$.
Remark Let $f \in L_{p}^{(\alpha, \beta)}(M)$. Then, there is a representative $g$ such that $g(x) \in$ $\mathfrak{T}_{x}(\alpha, \beta)^{*}$ holds for every $x \in M$ since it holds where a subsequence converges. Since the complement of this set is a null-set, we may choose $g$ to be zero there. Furthermore, $L_{2}^{(\alpha, \beta)}(M)$ may be isometrically embedded in $L_{2}(M) \otimes \mathfrak{T}(\alpha, \beta)^{*}$.

Remark Note that we may restrict all considerations to $\beta=0$, for all spaces with $\beta \neq 0$ can be generated using Riesz identifications.

Definition 3.4. Let $p \in \mathbb{R}_{\geq 1}, m \in \mathbb{N}$, and $\alpha \in \mathbb{N}_{0}$. Then, we define

$$
\begin{aligned}
& D_{c}\left(\|\cdot\|_{L_{p}^{(\alpha, 0)}(M)}\right):=\left\{x \in \mathfrak{M}_{0, c}^{(\alpha, 0)}(M) ; \int_{M}\left|\langle x, x\rangle_{(\alpha, 0)}\right|^{\frac{p}{2}} d \operatorname{vol}_{M}<\infty\right\}, \\
& D\left(\|\cdot\|_{W_{p}^{1,(\alpha, 0)}(M)}\right):=\left\{x \in \mathfrak{M}_{1}^{(\alpha, 0)}(M) \cap D\left(\|\cdot\|_{L_{p}^{(\alpha, 0)}(M)}\right) ;\right. \\
&\left.\int_{M}\left|\langle\nabla x, \nabla x\rangle_{(\alpha+1,0)}\right|^{\frac{p}{2}} d \operatorname{vol}_{M}<\infty\right\}, \\
& D_{c}\left(\|\cdot\|_{W_{p}^{1,(\alpha, 0)}(M)}\right):=\left\{x \in \mathfrak{M}_{1, c}^{(\alpha, 0)}(M) \cap D\left(\|\cdot\|_{L_{p}^{(\alpha, 0)}(M)}\right) ;\right. \\
&\left.\int_{M}\left|\langle\nabla x, \nabla x\rangle_{(\alpha+1,0)}\right|^{\frac{p}{2}} d \operatorname{vol}_{M}<\infty\right\}, \\
&\|\cdot\|_{W_{p}^{1,(\alpha, 0)}(M)}: D\left(\|\cdot\|_{W_{p}^{1,(\alpha, 0)}(M)}\right) \rightarrow \mathbb{R} ; x \mapsto\left(\|x\|_{L_{p}^{(\alpha, 0)}(M)}^{p}+\|\nabla x\|_{L_{p}^{(\alpha+1,0)}(M)}^{p}\right)^{\frac{1}{p}},
\end{aligned}
$$

as well as,

$$
\begin{aligned}
& L_{p, 0}^{(\alpha, 0)}(M): \overline{D_{c}\left(\|\cdot\|_{L_{p}^{(\alpha, 0)}(M)}\right)} \|^{\|\cdot\|_{L_{p}^{(\alpha, 0)}(M)}},
\end{aligned}
$$

$$
\begin{aligned}
& \left.W_{p, 0}^{1,(\alpha, 0)}(M): \overline{D_{c}\left(\|\cdot\|_{W_{p}^{1,(\alpha, 0)}(M)}\right)}\right)^{\|\cdot\|_{W_{p}^{1,(\alpha, 0)}(M)}} .
\end{aligned}
$$

Remark A priori, we cannot define Sobolev spaces in the same manner as we would define $W_{p}^{k}(\Omega)$ for $\Omega \coprod_{\text {open }} \mathbb{R}^{n}$. However, using Sobolev chains, it is still possible to show that they exists and are dense in $L_{p}(M)$. It is important to keep in mind that this does not imply non-triviality of $C^{k}(M)$ because the Sobolev embedding theorems do not hold, in general.

Definition 3.5. Let $\alpha \in \mathbb{N}_{0}$. Then we define gradient and divergence to be

$$
\begin{gathered}
\operatorname{grad}_{c, \alpha}: \mathfrak{M}_{1, c}^{(\alpha, 0)}(M) \rightarrow \mathfrak{M}_{0, c}^{(\alpha+1,0)}(M) ; x \mapsto \nabla x, \\
\operatorname{div}_{c, \alpha}: \mathfrak{M}_{1, c}^{(\alpha+1,0)}(M) \rightarrow \mathfrak{M}_{0, c}^{(\alpha, 0)}(M) ; x \mapsto \operatorname{tr} \nabla x
\end{gathered}
$$

where $\operatorname{tr}$ denotes the trace acting on the first two components. ${ }^{3}$
Before we can show that the gradient and divergence are formally adjoint, let us recall the Gauss divergence theorem.

Definition 3.6. Let $V$ be a closed subset of $M$. $V$ has smooth boundary if and only if for each $a \in V$ there is an open neighborhood $U \subseteq V$ of a and a function $g \in C^{1}(U ; \mathbb{R})$ such that

$$
V \cap U=\{x \in U ; g(x) \leq 0\}
$$

and $\nabla g(x) \neq 0$ for all $x \in U$ hold. Then we define

$$
\partial V \cap U:=\{x \in U ; g(x)=0\}
$$

and $\partial V$ the union of every such $\partial V \cap U$.
Let a be in $\partial V$. Then we call

$$
\nu(a):=\frac{1}{\|\nabla g(a)\|} \nabla g(a)
$$

the outward-pointing normal at a.
Remark Having smooth boundary as defined above means having $C^{1}$ boundary, i.e., the boundary is locally a $C^{1}$ manifold.

Sketch of proof Let $V \subseteq M$ be a closed subset of $M$ with smooth boundary. Let $p \in \partial V$ and $\varphi$ a chart with $p \in \varphi\left[\mathbb{R}^{\operatorname{dim} M}\right]$. Let $r \in \mathbb{R}_{>0}$ such that $U_{0}:=$ $B_{\mathbb{R}^{\text {dim }} M}\left(\varphi^{-1}(p), r\right) \subseteq[M] \varphi$ and such that there exists $g$ according to the definition with respect to $U:=\varphi\left[U_{0}\right]$. Note that $g \in C^{1}(U)$ means, per definitionem, $\tilde{g}:=$ $g \circ \varphi \in C^{1}\left(U_{0}\right)$ and $\nabla g(x) \neq 0$ is equivalent to $(g \circ \varphi)^{\prime}(x) \neq 0$ since $\nabla_{g_{i}} f=\partial_{i}(g \circ \varphi)$.

[^4]Theorem 3.7 (Level Set Criterion). A set $S \subseteq \mathbb{R}^{n}$ is an m-dimensional $C^{l}$ manifold if and only if for ever $p \in S$ there is an open neighborhood $U_{p}$ of $p$ and a function $g_{p} \in C^{l}\left(U_{p}, \mathbb{R}^{k}\right)$ with $m+k=n$ such that $S \cap U_{p}=[\{0\}] g_{p}$ and $\operatorname{rank} g_{p}^{\prime}=l$ in $U_{p}$.

Since $U_{0}$ is an open neighborhood of $\varphi^{-1}(p)$ and $\tilde{g} \in C^{1}\left(U_{0}, \mathbb{R}\right)$ with $[\partial V] \varphi \cap$ $U_{0}=[\{0\}] \tilde{g}$ and $\operatorname{rank} \tilde{g}^{\prime}=1$, we obtain that $\varphi\left[[\partial V] \varphi \cap U_{0}\right]$ is a $(\operatorname{dim} M-1)$ dimensional $C^{1}$-manifold. Since $p \in \partial V$ was arbitrarily chosen, we conclude that $\partial V$ is a $(\operatorname{dim} M-1)$-dimensional $C^{1}$-manifold.

Theorem 3.8 (Gauss divergence Theorem). Let $V$ be a compact subset of $M$ with smooth boundary and $\nu \in \mathfrak{M}_{0}^{(1,0)}(M)$ such that $\left.\nu\right|_{\partial V}$ is the outward-pointing normal vector field on $\partial V$. Let $d^{\operatorname{vol}} \mathrm{l}_{\partial V}$ be the surface form on $\partial V$. Let $F$ be a continuous vector field on $V$ and continuously differentiable in the interior, i.e., $F \in \mathfrak{M}_{0}^{(1,0)}(V) \cap \mathfrak{M}_{1}^{(1,0)}(V \backslash \partial V)$. Then

$$
\int_{V} \operatorname{tr} \nabla F d \operatorname{vol}_{M}=\int_{\partial V}\langle F, \nu\rangle d \operatorname{vol}_{\partial V}
$$

holds.
ObSERVATION 3.9. $-\operatorname{div}_{c, \alpha} \subseteq\left(\operatorname{grad}_{c, \alpha}\right)^{*}$ holds in $L_{2}^{(\alpha+1,0)}(M) \oplus L_{2}^{(\alpha, 0)}(M)$.
Proof. Let $\varphi \in D\left(\operatorname{grad}_{c, \alpha}\right)$ and $\tau \in D\left(\operatorname{div}_{c, \alpha}\right)$. Then

$$
\begin{aligned}
\langle-\operatorname{tr} \nabla \tau, \varphi\rangle_{(\alpha, 0)}= & \left\langle-\operatorname{tr}\left(\nabla_{g_{i}} \tau_{j_{1} \ldots j_{\alpha+1}} g^{i} \otimes g^{j_{1}} \otimes \ldots \otimes g^{j_{\alpha+1}}\right), \varphi\right\rangle_{(1,0)} \\
= & -\left\langle\nabla_{g_{i}} \tau_{j_{1} \ldots j_{\alpha+1}} g^{i j_{1}} g^{j_{2}} \otimes \ldots \otimes g^{j_{\alpha+1}}, \varphi_{k_{1} \ldots k_{\alpha}} g^{k_{1}} \otimes \ldots \otimes g^{k_{\alpha}}\right\rangle_{(1,0)} \\
= & -\nabla_{g_{i}} \tau_{j_{1} \ldots j_{\alpha+1}} g^{i j_{1}} g^{j_{2} k_{1}} \ldots g^{j_{\alpha+1} k_{\alpha}} \varphi_{k_{1} \ldots k_{\alpha}} \\
= & -\nabla_{g_{i}}\left(\tau_{j_{1} \ldots j_{\alpha+1}} g^{i j_{1}} g^{j_{2} k_{1}} \ldots g^{j_{\alpha+1} k_{\alpha}} \varphi_{k_{1} \ldots k_{\alpha}}\right) \\
& +\tau_{j_{1} \ldots j_{\alpha+1}} g^{i j_{1}} g^{j_{2} k_{1}} \ldots g^{j_{\alpha+1} k_{\alpha}} \nabla_{g_{i}} \varphi_{k_{1} \ldots k_{\alpha}} \\
= & -\nabla_{g_{i}}\left(\tau_{j_{1} \ldots j_{\alpha+1}} g^{i j_{1}} g^{j_{2} k_{1}} \ldots g^{j_{\alpha+1} k_{\alpha}} \varphi_{k_{1} \ldots k_{\alpha}}\right)+\langle\tau, \nabla \varphi\rangle_{(\alpha+1,0)}
\end{aligned}
$$

and, hence, using a $\tilde{M} \subseteq_{\text {compact }} M$ such that $\varphi$ and $\tau$ are compactly supported in $\tilde{M}$,

$$
\begin{aligned}
\int_{\tilde{M}}\langle-\operatorname{tr} \nabla \tau, \varphi\rangle_{(\alpha, 0)} d \operatorname{vol}_{M}= & \int_{\tilde{M}}\langle\tau, \nabla \varphi\rangle_{(\alpha+1,0)} d \operatorname{vol}_{M} \\
& -\int_{\tilde{M}} \nabla g_{i}\left(\tau_{j_{1} \ldots j_{\alpha+1}} g^{i j_{1}} g^{j_{2} k_{1}} \ldots g^{j_{\alpha+1} k_{\alpha}} \varphi_{k_{1} \ldots k_{\alpha}}\right) d \operatorname{vol}_{M} \\
= & \int_{\tilde{M}}\langle\tau, \nabla \varphi\rangle_{(\alpha+1,0)} d \operatorname{vol}_{M} \\
& -\int_{\tilde{M}} \operatorname{tr} \nabla\left(\tau_{j_{1} \ldots j_{\alpha+1}} g^{j_{2} k_{1}} \ldots g^{j_{\alpha+1} k_{\alpha}} \varphi_{k_{1} \ldots k_{\alpha}} g^{j_{1}}\right) d \operatorname{vol}_{M} \\
= & \int_{\tilde{M}}\langle\tau, \nabla \varphi\rangle_{(\alpha+1,0)} d \operatorname{vol}_{M} \\
& -\int_{\partial \tilde{M}}\left\langle\tau_{j_{1} \ldots j_{\alpha+1}} g^{j_{2} k_{1}} \ldots g^{j_{\alpha+1} k_{\alpha}} \varphi_{k_{1} \ldots k_{\alpha}} g^{j_{1}}, \nu\right\rangle d \operatorname{vol}_{\partial \tilde{M}}
\end{aligned}
$$

holds according to the Gauss divergence theorem. However, we have assumed that $\tau$ and $\varphi$ are compactly supported in $\tilde{M} \backslash \partial \tilde{M}$, i.e., the integral over $\partial \tilde{M}$ vanishes which, thence, reduces to

$$
\int_{M}\langle-\operatorname{tr} \nabla \tau, \varphi\rangle_{(\alpha, 0)} d \operatorname{vol}_{M}=\int_{M}\langle\tau, \nabla \varphi\rangle_{(\alpha+1,0)} d \operatorname{vol}_{M}
$$

i.e., the assertion.

Lemma 3.10. Let $X, Y$ be reflexive Banach spaces, and $A \subseteq X \oplus Y, B \subseteq Y^{\prime} \oplus X^{\prime}$ densely defined, linear operators where $X^{\prime}$ and $Y^{\prime}$ denote the dual spaces of $X$ and $Y$, respectively. If $A$ and $B$ are formally adjoint, i.e., $A \subseteq B^{*}$, then $B \subseteq A^{*}$ holds and both operators are closable, where $A^{*}$ and $B^{*}$ denote the respective dual or adjoint operators depending on whether or not $X$ and $Y$ are Hilbert spaces.

Proof. $A^{*}$ and $B^{*}$ are closed operators since $A$ and $B$ are densely defined which directly implies that $A$ is closable. Thus,

$$
B \subseteq \bar{B}=B^{* *} \subseteq A^{*}
$$

shows $B \subseteq A^{*}$ and, therefore, closability of $B$, too.

The lemma above enables us to define

$$
\operatorname{grad}_{0, \alpha}:=\overline{\operatorname{grad}_{c, \alpha}} \quad, \quad \operatorname{div}_{0, \alpha}:=\overline{\operatorname{div}_{c, \alpha}}
$$

as well as,

$$
\operatorname{grad}_{\alpha}:=-\left(\operatorname{div}_{c, \alpha}\right)^{*} \quad, \quad \operatorname{div} \alpha:=-\left(\operatorname{grad}_{c, \alpha}\right)^{*}
$$

From this point on, we will drop the index $\alpha$ as it is uniquely determined by the context.

## Remarks on Sobolev Spaces

Since we are on a $C^{1,1}$-manifold $M$, we only know that the space $C^{1}(M)$ is non-trivial and dense in $C(M)$. For $k \geq 2$ the spaces $C^{k}(M)$ may very well be trivial. Hence, the usual approach to defining Sobolev spaces fails. However, we may use the notion of a Sobolev chain; for further detail, please, refer to [15].

Lemma 3.11 ([15]; Lemma 2.1.3). Let $H$ be a Hilbert space and $A \subseteq H \oplus H$ a closed, densely defined, linear operator with zero in the resolvent set $\varrho(A)$. Then, $A^{n}$ is a closed, densely defined operator for every $n \in \mathbb{N}$ with $0 \in \varrho\left(A^{n}\right)$ and

$$
\forall x \in D\left(A^{n}\right):\left\|A^{n} x\right\|_{H} \geq\left\|A^{-1}\right\|_{L(H)}^{-n}\|x\|_{H}
$$

Let $H_{n}(A)$ be the Hilbert space $D\left(A^{n}\right)$ equipped with the norm $x \mapsto\left\|A^{n} x\right\|_{H}$. Then,

$$
A_{n+1, n}: H_{n+1}(A) \rightarrow H_{n}(A) ; x \mapsto A x
$$

is unitary for every $n \in \mathbb{N}_{0}$.
If $A$ is a closed, densely defined, linear operator with $0 \in \varrho(A)$, then $A^{*}=$ $\left(-A^{-1}\right)^{\perp}$ is closed linear operator. Furthermore, closedness of $A$ implies that $A^{*}$ is densely defined, and $0 \in \varrho\left(A^{*}\right)$ follows from $\varrho\left(A^{*}\right)=\varrho(A)^{*}$. Hence, $H_{n}\left(A^{*}\right)$ is well-defined, as well, and we can extend the family $\left(H_{n}(A)\right)_{n \in \mathbb{N}_{0}}$.

Definition 3.12. Let $H$ be a Hilbert space, $A \subseteq H \oplus H$ a closed, densely defined, linear operator with $0 \in \varrho(A)$, and $n \in \mathbb{Z}$. Then, we define

$$
H_{n}(A):=\left\{\begin{array}{l}
\left(D\left(A^{n}\right),\left\|A^{n} \cdot\right\|_{H}\right) ; n \in \mathbb{N}_{0} \\
H_{-n}\left(A^{*}\right)^{*} ; n \in-\mathbb{N}
\end{array}\right.
$$

where $H_{-n}\left(A^{*}\right)^{*}$ denotes the topological dual of $H_{-n}\left(A^{*}\right)$. Then, we call
(i) $\left(H_{n}(A)\right)_{n \in \mathbb{N}_{0}}$ the positive Sobolev chain associated with $A$,
(ii) $\left(H_{n}(A)\right)_{n \in-\mathbb{N}_{0}}$ the negative Sobolev chain associated with $A$, and
(iii) $\left(H_{n}(A)\right)_{n \in \mathbb{Z}}$ the (long) Sobolev chain associated with $A$.

Lemma 3.13 ([15]; Lemma 2.1.6). Let $\left(H_{n}(A)\right)_{n \in \mathbb{Z}}$ the Sobolev chain associated with the operator $A$. Then, we obtain that the embedding

$$
H_{n+k}(A) \hookrightarrow H_{n}(A)
$$

is dense and continuous (in the sense of canonical embeddings) for every $n \in \mathbb{Z}$ and $k \in \mathbb{N}_{0}$. Furthermore, the operators

$$
D\left(A^{|n|+1}\right) \subseteq H_{n+1}(A) \rightarrow H_{n}(A) ; x \mapsto A x
$$

have unitary closures $A_{n+1, n} \subseteq H_{n+1}(A) \oplus H_{n}(A)$ for every $n \in \mathbb{Z}$.
It is often convenient to define the "closures" of the Sobolev chain

$$
H_{\infty}(A):=\bigcap_{n \in \mathbb{Z}} H_{n}(A)
$$

which is a Fréchet space if equipped with the family of semi-norms $\left(\|\cdot\|_{H_{n}(A)}\right)_{n \in \mathbb{Z}}$ and dense in all $H_{n}(A)$ (in fact, $H_{\infty}(A)$ is a core of $A_{(k)}: H_{k+1}(A) \subseteq H_{k}(A) \xrightarrow{n \in \mathbb{Z}}$ $H_{k}(A) ; x \mapsto A x ;$ cf., Lemma 2.1.15 in [15]) and

$$
H_{-\infty}(A):=\bigcup_{n \in \mathbb{Z}} H_{n}(A)
$$

which is complete if equipped with the topology induced by saying that $x$ is a Cauchy-sequence/convergent in $H_{-\infty}(A)$ if and only if there exists an $n \in \mathbb{Z}$ such that $x$ is a Cauchy-sequence/convergent in $H_{n}(A)$ (cf., Lemma 2.1.11 in [15]).

As for the definition of Sobolev spaces on the $C^{1,1}$-manifold $M$, we know that the gradient grad is a well-defined, closed, densely defined, linear operator on the Hilbert space $L_{2}(M)$. Furthermore, we have the following theorem which can be obtained from the first representation theorem (Theorem VI.2.1 in [11]) applied to the closed, positive, symmetric form $\tau$ with $D(\tau):=D(A)$ and $\forall x, y \in D(\tau)$ : $\tau(x, y):=\langle A x, A y\rangle_{H_{2}}$.

Theorem 3.14 (von Neumann). Let $H_{1}$ and $H_{2}$ be Hilbert spaces and $A \subseteq$ $H_{1} \oplus H_{2}$ a closed and densely defined operator. Then $A^{*} A$ is self-adjoint in $H_{1}$ and its domain is a core of $A$.

Hence, grad* grad is a well-defined, closed, densely defined, linear operator and, additionally, self-adjoint, i.e., so is $|\operatorname{grad}|:=\sqrt{\operatorname{grad}^{*} \operatorname{grad}}$ which allows us to define the Sobolev spaces

$$
\forall k \in \mathbb{Z}: W_{2}^{k}(M):=H_{k}(1+|\operatorname{grad}|)
$$

We may also define $W_{2}^{s}(M)$ for $s \in \mathbb{R}$ to be the closure of $H_{\infty}(1+|\operatorname{grad}|)$ with respect to the scalar product

$$
\forall x, y \in H_{\infty}(1+|\operatorname{grad}|):\langle x, y\rangle_{W_{2}^{s}(M)}:=\left\langle(1+|\operatorname{grad}|)^{s} x,(1+|\operatorname{grad}|)^{s} y\right\rangle_{L_{2}(M)}
$$

Alternatively, we may consider the spaces $\tilde{W}_{2}^{s}(M)$ defined by the closure of $H_{\infty}(1+$ $\mid$ grad|) with respect to the scalar product

$$
\forall x, y \in H_{\infty}(1+|\operatorname{grad}|):\langle x, y\rangle_{\tilde{W}_{2}^{s}(M)}:=\left\langle\left(1+\operatorname{grad}^{*} \operatorname{grad}\right)^{s} x, y\right\rangle_{L_{2}(M)}
$$

for $s \in \mathbb{R}$ which are equivalent but sometimes more suitable. None the less, both families, $\left(W_{2}^{s}(M)\right)_{s \in \mathbb{R}}$ and $\left(\tilde{W}_{2}^{s}(M)\right)_{s \in \mathbb{R}}$, satisfy the interpolation property, that is, for every $s, t \in \mathbb{R}$ and $\vartheta \in[0,1]$

$$
\begin{aligned}
& W_{2}^{(1-\vartheta) s+\vartheta t}(M)=\left[W_{2}^{s}(M), W_{2}^{t}(M)\right]_{\vartheta} \\
& \tilde{W}_{2}^{(1-\vartheta) s+\vartheta t}(M)=\left[\tilde{W}_{2}^{s}(M), \tilde{W}_{2}^{t}(M)\right]_{\vartheta}
\end{aligned}
$$

in the sense of complex interpolation; cf., e.g., section 4.2 in [18]. In fact, if $A$ is strictly positive, then we can extend the Sobolev chain $\left(H_{n}(A)\right)_{n \in \mathbb{Z}}$ to $\left(H_{s}(A)\right)_{s \in \mathbb{R}}$ by setting

$$
H_{s}(A):=\left(D\left(A^{s}\right),\left\|A^{s} \cdot\right\|_{H}\right)
$$

for $s \geq 0$ and by duality for $s<0$. Similarly, we might simply use the interpolation property directly to define the $H_{s}(A)$ for $s \in \mathbb{R} \backslash \mathbb{Z}$ via

$$
\forall s, t \in \mathbb{R}_{\geq 0} \forall \vartheta \in[0,1]: \quad H_{(1-\vartheta) s+\vartheta t}(A)=\left[H_{s}(A), H_{t}(A)\right]_{\vartheta}
$$

Finally, we'd like to note that not all Sobolev embeddings fail to hold. For instance, we still obtain the following theorem.

Theorem 3.15 (Sobolev Embedding). Let $X$ be a Banach space, $S, T \in \mathbb{R}$ and $S<T$. Then

$$
\text { id }: W_{2}^{1}([S, T] ; X) \rightarrow C([S, T] ; X) ; f \mapsto f
$$

is continuous and injective.
Proof. Let $s, t \in[S, T]$ and $f \in C^{\infty}([S, T] ; X)$ (mind that $C^{\infty}([S, T] ; X)$ is a dense subset of $\left.W_{2}^{1}([S, T] ; X)\right)$. Then

$$
\|f(t)\|_{X} \leq\|f(s)\|_{X}+\left|\int_{s}^{t}\left\|f^{\prime}(\tau)\right\|_{X} d \tau\right| \leq\|f(s)\|_{X}+\sqrt{|t-s|}\left|\int_{s}^{t}\left\|f^{\prime}(\tau)\right\|_{X}^{2} d \tau\right|^{\frac{1}{2}}
$$

holds. Integrating $s$ yields

$$
\begin{aligned}
(T-S)\|f(t)\|_{X} & \leq \int_{S}^{T}\|f(s)\|_{X} d s+\int_{S}^{T} \underbrace{\sqrt{|t-s|}}_{\leq \sqrt{T-S}} \underbrace{\left.\int_{s}^{t}\left\|f^{\prime}(\tau)\right\|_{X}^{2} d \tau\right|^{\frac{1}{2}} d s}_{\leq\left|S_{S}^{T}\left\|f^{\prime}(\tau)\right\|_{X}^{2} d \tau\right|^{\frac{1}{2}}} \\
& \leq \sqrt{T-S}\left|\int_{S}^{T}\|f(s)\|_{X}^{2} d s\right|^{\frac{1}{2}}+(T-S)^{\frac{3}{2}}\left|\int_{S}^{T}\left\|f^{\prime}(s)\right\|_{X}^{2} d s\right|^{\frac{1}{2}}
\end{aligned}
$$

Hence,

$$
\|f\|_{C([S, T] ; X)} \leq \max \left\{(T-S)^{\frac{1}{2}},(T-S)^{-\frac{1}{2}}\right\} \sqrt{2}\|f\|_{W_{2}^{1}([S, T] ; X)}
$$

holds, too, where we used

$$
\sqrt{a}+\sqrt{b}=\left\|\binom{\sqrt{a}}{\sqrt{b}}\right\|_{1} \leq\left\|\binom{1}{1}\right\|_{2}\left\|\binom{\sqrt{a}}{\sqrt{b}}\right\|_{2}=\sqrt{2} \sqrt{a+b}
$$

Thus, any $W_{2}^{1}([S, T] ; X)$-Cauchy sequence in $C^{\infty}([S, T] ; X)$ is also a $C([S, T] ; X)$ Cauchy sequence and, therefore, $W_{2}^{1}([S, T] ; X) \subseteq C([S, T] ; X)$.

Furthermore the identities

$$
\begin{aligned}
& \mathrm{id}_{1}: W_{2}^{1}([S, T] ; X) \rightarrow L_{2}([S, T] ; X) ; f \mapsto f \\
& \mathrm{id}_{2}: C([S, T] ; X) \rightarrow L_{2}([S, T] ; X) ; f \mapsto f
\end{aligned}
$$

are injective. Thus, $\mathrm{id}=\mathrm{id}_{2}^{-1} \circ \mathrm{id}_{1}$ is injective.

## CHAPTER 4

## The Analytic Implicit Function Theorem

Before we prove the analytic implicit function theorem, we will recall a few facts about analytic operators. A more extensive account can be found in [3].

Definition 4.1. Let $X$ and $Y$ be Banach spaces and $k \in \mathbb{N}_{0}$. A $k$-linear mapping $m_{k}: X^{k} \rightarrow Y$ is called symmetric if and only if for every permutation $\sigma \in S_{k}$ and $x_{1}, \ldots, x_{k} \in X$

$$
m_{k} x_{1} \cdots x_{k}:=m_{k}\left(x_{\sigma(1)}, \ldots, x_{\sigma(k)}\right)=m_{k}\left(x_{1}, \ldots, x_{k}\right)
$$

holds.
Definition 4.2. Let $X$ and $Y$ be Banach spaces, $U \subseteq X$ open, and $x_{0} \in U$. A mapping $F: U \rightarrow Y$ is called analytic at $x_{0}$ if and only if there exist $r \in \mathbb{R}_{>0}$ and $k$-linear and symmetric operators $m_{k}: X^{k} \rightarrow Y\left(k \in \mathbb{N}_{0}\right)$ such that for every $x \in B_{X}\left(x_{0}, r\right) \subseteq U$

$$
\begin{equation*}
F(x)=\sum_{k \in \mathbb{N}_{0}} m_{k}\left(x-x_{0}\right)^{k} \tag{1}
\end{equation*}
$$

and

$$
\begin{equation*}
\sup _{k \in \mathbb{N}_{0}} r^{k}\left\|m_{k}\right\|_{\text {Lip }}=: M<\infty \tag{2}
\end{equation*}
$$

hold. The series $\sum_{k \in \mathbb{N}_{0}} m_{k}\left(x-x_{0}\right)^{k}$ in (1) is called power series and $F$ is said to be analytic in $U$ if and only if it analytic at every point of $U$.

Due to (2) we observe for $x \in B_{X}\left(x_{0}, r\right)$

$$
\sum_{k \in \mathbb{N}_{0}}\left\|m_{k}\left(x-x_{0}\right)^{k}\right\|_{Y} \leq M \sum_{k \in \mathbb{N}_{0}} \frac{\left\|x-x_{0}\right\|_{X}^{k}}{r^{k}}=M \frac{1}{1-\frac{\left\|x-x_{0}\right\|_{X}}{r}}=\frac{M r}{r-\left\|x-x_{0}\right\|_{X}}<\infty .
$$

Hence, the power series converges absolutely.
Observation 4.3. Let $X, Y_{1}$, and $Y_{2}$ be Banach spaces, $U \subseteq X$ open, and $\left(F_{1}, F_{2}\right): U \rightarrow Y_{1} \times Y_{2}$ analytic. Then, $F_{1}$ and $F_{2}$ are analytic.

Proof. Since $\left(F_{1}, F_{2}\right)$ is analytic, there is a representation

$$
\left(F_{1}, F_{2}\right)(x)=\sum_{k \in \mathbb{N}_{0}} m_{k}\left(x-x_{0}\right)^{k}
$$

with $\sup _{k \in \mathbb{N}_{0}} r^{k}\left\|m_{k}\right\|_{\text {Lip }}=: M<\infty$ for some $r \in \mathbb{R}_{>0}$ and every $x_{0} \in U$. Let $i \in\{1,2\}$. The projection $\mathrm{pr}_{i}: Y_{1} \times Y_{2} \rightarrow Y_{i} ;\left(y_{1}, y_{2}\right) \mapsto y_{i}$ is continuous with norm 1. Hence, we observe

$$
F_{i}(x)=\operatorname{pr}_{i}\left(F_{1}, F_{2}\right)(x)=\sum_{k \in \mathbb{N}_{0}} \operatorname{pr}_{i} m_{k}\left(x-x_{0}\right)^{k}
$$

and

$$
\sup _{k \in \mathbb{N}_{0}} r^{k}\left\|\operatorname{pr}_{i} m_{k}\right\|_{\text {Lip }} \leq \sup _{k \in \mathbb{N}_{0}} r^{k}\left\|\operatorname{pr}_{i}\right\|_{\text {Lip }}\left\|m_{k}\right\|_{\text {Lip }}=M<\infty
$$

Proposition 4.4. Let $F$ be defined by (1) such that (2) holds. Then, $F$ is analytic at every point $x \in B_{X}\left(x_{0}, r\right)=: U_{0}, F \in C^{\infty}\left(U_{0} ; Y\right)$ and for every $k \in \mathbb{N}_{0}$

$$
m_{k}=\frac{\partial^{k} F\left(x_{0}\right)}{k!}
$$

holds. For every $k \in \mathbb{N}_{0}$ and $x \in B_{X}\left(x_{0}, r\right)$ the $k^{\text {th }}$ derivative of $F, \partial^{k} F$, is analytic at $x$. Furthermore for every $x_{1}, \ldots, x_{k} \in X$

$$
\partial^{k} F(x)\left(x_{1}, \ldots, x_{k}\right)=\sum_{j \in \mathbb{N}_{0}} \frac{(j+k)!}{j!} m_{j+k}\left(x-x_{0}\right)^{j} x_{1} x_{2} \cdots x_{k}
$$

holds and there are $C \in \mathbb{R}_{>1}$ and $R \in(0,1)$ such that for every $x \in B_{X}\left(x_{0}, \frac{r}{2}\right)$ and $k \in \mathbb{N}$

$$
\begin{equation*}
\left\|\partial^{k} F(x)\right\|_{\mathrm{Lip}} \leq C \frac{k!}{R^{k}} \tag{3}
\end{equation*}
$$

holds, too. In particular, if $K \subseteq U$ is compact then $C$ and $R$ exist such that (3) holds for every $x \in K$.

Proof. see [3]
Definition 4.5. Let $X, Y$, and $Z$ be Banach spaces, $U \subseteq X \times Y$ open, and $\left(x_{0}, y_{0}\right) \in U$. A mapping $F: U \rightarrow Z$ is called analytic at $\left(x_{0}, y_{0}\right)$ if and only if there exist $r \in \mathbb{R}_{>0}$ and $k$-linear and symmetric operators $m_{k}:(X \times Y)^{k} \rightarrow Z\left(k \in \mathbb{N}_{0}\right)$ such that for every $(x, y) \in B_{X}\left(\left(x_{0}, y_{0}\right), r\right) \subseteq U$

$$
\begin{equation*}
F(x, y)=\sum_{k \in \mathbb{N}_{0}} m_{k}\left(x-x_{0}, y-y_{0}\right)^{k} \tag{4}
\end{equation*}
$$

and

$$
\begin{equation*}
\sup _{k \in \mathbb{N}_{0}} r^{k}\left\|m_{k}\right\|_{\text {Lip }}=: M<\infty \tag{5}
\end{equation*}
$$

hold.
$F$ is said to be analytic in $U$ if and only if it is analytic at every point of $U$.
DEfinition 4.6. $m_{p, q}: X^{p} \times Y^{q} \rightarrow Z$ is $p-q$-linear and symmetric if and only if there is a $k$-linear and symmetric $m_{k}:(X \times Y)^{k} \rightarrow Z$ with $k=p+q$ such that for all $x_{1}, \ldots, x_{p} \in X$ and $y_{1}, \ldots, y_{q} \in Y$

$$
m_{p, q}\left(x_{1}, \ldots, x_{p}, y_{1}, \ldots, y_{q}\right)=m_{k}\left(\left(x_{1}, 0\right), \ldots,\left(x_{p}, 0\right),\left(0, y_{1}\right), \ldots,\left(0, y_{q}\right)\right)
$$

holds.

It is possible to express (cf., e.g., $\S 4.4$ in [3]) $F$ in (4) as

$$
F(x, y)=\sum_{(p, q) \in \mathbb{N}_{0}^{2}} \frac{(p+q)!}{p!q!} m_{p, q}\left(x-x_{0}\right)^{p}\left(y-y_{0}\right)^{q}
$$

with

$$
m_{p, q}=\frac{\partial_{1}^{p} \partial_{2}^{q} F\left(x_{0}, y_{0}\right)}{(p+q)!}
$$

and

$$
\sup _{p, q \in \mathbb{N}_{0}} r^{p+q}\left\|m_{p, q}\right\|<\infty .
$$

In particular, the power series converges absolutely again.
Now we will prove the analytic implicit function theorem. We will start by proving the implicit function theorem for up to $C^{\infty}$ functions with an adaptation of the standard approach in finite dimensional spaces. This has the advantage that it is constructive, i.e., the solutions of Navier-Stokes will be constructable. The prove of analyticity, however, is not "constructable" and, even though it is possible to prove the theorem directly with analyticity, we chose this approach since constructibility of the solution is quite a nice feature.

Proposition 4.7 (Chain Rule). Let $X, Y$, and $Z$ be Banach spaces, $U \subseteq X$ open, $V \subseteq Y$ open, $a \in U, f: U \rightarrow V$ Fréchet-differentiable in $a$, and $g: V \rightarrow Z$ Fréchet-differentiable in $f(a)$. Then, $g \circ f: U \rightarrow Z$ is Fréchet-differentiable in a and satisfies

$$
\left.(g \circ f)^{\prime}(a)=g^{\prime}(f(a)) f^{\prime}(a)\right) \in L(X, Z)
$$

Proof. Let $A:=f^{\prime}(a)$ and $B:=g^{\prime}(f(a))$. Then, we observe for $x \in U$ and $y \in V$

$$
f(x)=f(a)+A(x-a)+\|x-a\|_{X} \varphi(x)
$$

and

$$
g(y)=g(f(a))+B(y-f(a))+\|y-f(a)\|_{Y} \psi(y)
$$

for some $\varphi \in C(U, Y)$ and $\psi \in C(V, Z)$ with $\varphi(x) \rightarrow 0(x \rightarrow a)$ and $\psi(y) \rightarrow 0(y \rightarrow$ $f(a))$. Therefore,

$$
\begin{aligned}
(g \circ f)(x)= & g\left(f(a)+A(x-a)+\|x-a\|_{X} \varphi(x)\right) \\
= & g(f(a))+B\left(A(x-a)+\|x-a\|_{X} \varphi(x)\right) \\
& +\|A(x-a)+\| x-a\left\|_{X} \varphi(x)\right\|_{Y} \psi\left(\left(f(a)+A(x-a)+\|x-a\|_{X} \varphi(x)\right)\right) \\
= & g(f(a))+B A(x-a)+\|x-a\|_{X} \omega(x)
\end{aligned}
$$

with

$$
\omega(x):=B \varphi(x)+\frac{\|A(x-a)+\| x-a\left\|_{X} \varphi(x)\right\|_{Y} \psi(f(x))}{\|x-a\|_{X}}
$$

which satisfies

$$
\|\omega(x)\|_{Z} \leq\|B\|_{L(Y, Z)} \underbrace{\|\varphi(x)\|_{Y}}_{\rightarrow 0}+\left(\|A\|_{L(X, Y)}+\|\varphi(x)\|_{Y}\right) \underbrace{\psi(f(x))}_{\rightarrow \psi(f(a))=0} \rightarrow 0(x \rightarrow a),
$$

thus, showing the assertion.

Proposition 4.8 (Mean Value Inequality). Let $X$ be a Banach space, $a, b \in \mathbb{R}$, $a<b$, and $f \in C([a, b], X)$ differentiable from the right on $(a, b)$. Then, there is $a$ $t \in(a, b)$ such that

$$
\|f(b)-f(a)\|_{X} \leq\left\|f_{r}^{\prime}(t)\right\|_{X}(b-a)
$$

holds where $f_{r}^{\prime}(t)$ denotes the right-hand side derivative of $f$ at $t$.
Proof. (i) Let $\varphi \in C([a, b], \mathbb{R})$ with $\varphi(a)=\varphi(b)=0$. Then, the intermediate value theorem yields the existence of $a_{1}$ and $b_{1}$ with $a<a_{1}<b_{1}<b$ satisfying $\varphi\left(a_{1}\right)=\varphi\left(b_{1}\right)$. By the extreme value theorem, there exists $t \in\left[a_{1}, b_{1}\right)$ such that

$$
\forall r \in\left[a_{1}, b_{1}\right): \varphi(t) \leq \varphi(r)
$$

holds. Hence, there is $h \in \mathbb{R}_{>0}$ such that

$$
\forall s \in[t, t+h]: \varphi(t) \leq \varphi(s)
$$

holds.
(ii) Without loss of generality, let $a=0$ and $f(0)=0$. For $s \in[0, b]$, we define

$$
\varphi(s):=\|f(s)\|_{X}-s\left\|\frac{1}{b} f(b)\right\|_{X}
$$

and observe that $\varphi$ is continuous with $\varphi(0)=\varphi(b)=0$. According to (i), there exists a $t \in(0, b)$ and $h \in(0, b-t)$ such that

$$
\forall s \in[t, t+h]: \varphi(s) \geq \varphi(t)
$$

holds. Hence,

$$
\begin{aligned}
0 & \leq \frac{\varphi(s)-\varphi(t)}{s-t} \\
& =\frac{\|f(s)\|_{X}-\|f(t)\|_{X}}{s-t}-\left\|\frac{1}{b} f(b)\right\|_{X} \\
& \leq \frac{\|f(s)-f(t)\|_{X}}{s-t}-\left\|\frac{1}{b} f(b)\right\|_{X} \\
& =\left\|\frac{f(s)-f(t)}{s-t}\right\|_{X}-\left\|\frac{1}{b} f(b)\right\|_{X} \\
& \rightarrow\left\|f_{r}^{\prime}(t)\right\|_{X}-\left\|\frac{1}{b} f(b)\right\|_{X}
\end{aligned}
$$

holds for $s \searrow t$. In other words,

$$
\|f(b)-f(a)\|_{X}=\|f(b)\|_{X} \leq\left\|f_{r}^{\prime}(t)\right\|_{X} b=\left\|f_{r}^{\prime}(t)\right\|_{X}(b-a)
$$

shows the assertion.

Corollary 4.9. Let $X$ and $Y$ be Banach spaces, $U \subseteq X$ open, $f: U \rightarrow Y$ Fréchet-differentiable, and $a, b \in U$ such that their convex hull $\operatorname{conv}\{a, b\}$ is a subset of $U$. Then, there is $a t \in(0, a)$ such that

$$
\|f(b)-f(a)\|_{Y} \leq\left\|f^{\prime}((1-t) a+t b)\right\|_{L(X, Y)}\|b-a\|_{X}
$$

holds.
Proof. Let

$$
g:[0,1] \rightarrow Y ; t \mapsto f((1-t) a+t b)
$$

Then, $g$ is differentiable and the chain rule yields

$$
g^{\prime}(t)=f^{\prime}((1-t) a+t b)(b-a) .
$$

Furthermore, there is $t \in(0,1)$ such that

$$
\|f(b)-f(a)\|_{Y}=\|g(1)-g(0)\|_{Y} \leq\left\|g^{\prime}(t)\right\|_{Y} \leq\left\|f^{\prime}((1-t) a+t b)\right\|_{L(X, Y)}\|b-a\|_{X} .
$$

Proposition 4.10 (Modified Newton's Method). Let $X$ and $Y$ be Banach spaces, $U \subseteq X$ open, $G \subseteq U$ convex and closed in $X, f \in C^{1}(U, Y)$, and $B \in L(Y, X)$. Let $\Phi: U \rightarrow X ; x \mapsto x-B f(x)$ be such that $G$ is $\Phi$-invariant, i.e., $\Phi[G] \subseteq G$, and

$$
k:=\sup _{x \in G}\left\|1-B f^{\prime}(x)\right\|_{L(X)}<1 .
$$

Then, $\Phi$ is a strict contraction on $G$ and its unique fixed point $x^{*} \in G$ satisfies $B f\left(x^{*}\right)=0$.

Proof. Clearly, $\Phi^{\prime}(x)=1-B f^{\prime}(x)$. Hence, for $a, b \in G$, we obtain

$$
\|\Phi(x)-\Phi(a)\|_{X} \leq \sup _{t \in[0,1]}\left\|\Phi^{\prime}((1-t) a+t b)\right\|_{L(X)}\|b-a\|_{X} \leq k\|b-a\|_{X} .
$$

Hence, $\Phi$ is a strict contraction and Banach's fixed point theorem implies the other assertions.

Theorem 4.11 (Implicit Function Theorem). Let $M$ be a topological space, $Y$ and $Z$ Banach spaces, $U \subseteq Y$ open, $F: M \times U \rightarrow Z$ continuous and Fréchetdifferentiable with respect to the second variable, as well as, $(a, b) \in M \times U$ such that $f(a, b)=0, \partial_{2} F$ is continuous at $(a, b)$, and $\partial_{2} F(a, b)$ is an isomorphism.

Then, there are open neighborhoods $V_{1} \subseteq M$ of a and $V_{2} \subseteq U$ of b such that there exists a unique function $g: V_{1} \rightarrow V_{2}$ with $\forall x \in V_{1}: F(x, g(x))=0$. Furthermore, $g$ is continuous.

Proof. Let $B:=\partial_{2} F(a, b)^{-1} \in L(Z, Y)$ and

$$
\Phi: M \times U \rightarrow Y ;(x, y) \mapsto y-B F(x, y) .
$$

Then, clearly, $1-B \partial_{2} F(a, b)=0$ holds and, since $\partial_{2} F$ is continuous at $(a, b)$, there are $W_{1} \subseteq_{\text {open }} M$ and $W_{2} \subseteq_{\text {open }} U$ with $a \in W_{1}$ and $b \in W_{2}$ such that

$$
\forall(x, y) \in W_{1} \times W_{2}:\left\|1-B \partial_{2} F(x, y)\right\|_{L(Y)}<\frac{1}{2}
$$

holds. Let $r \in \mathbb{R}_{>0}$ such that $B_{Y}[b, r] \subseteq W_{2}$. Since $F(a, b)=0$ and $F$ is continuous, there is an open neighborhood $V_{1} \subseteq W_{1}$ of $a$ such that

$$
\sup _{x \in V_{1}}\|B F(x, b)\|_{Y}<\frac{r}{2} .
$$

For $x \in V_{1}$ and $y \in B_{Y}[b, r]$ we, thus, observe

$$
\begin{aligned}
\|\Phi(x, y)-b\|_{Y} & \leq\|\Phi(x, y)-\Phi(x, b)\|_{Y}+\|\Phi(x, b)-b\|_{Y} \\
& \leq \underbrace{\sup _{t \in[0,1]}\left\|1-B \partial_{2} F(x,(1-t) y+t b)\right\|_{L(Y)}}_{<\frac{1}{2}} \underbrace{\|y-b\|_{Y}}_{\leq r}+\underbrace{\|B F(x, b)\|_{Y}}_{<\frac{r}{2}} \\
& <r, \quad
\end{aligned}
$$

i.e., $\Phi(x, \cdot)\left[B_{Y}[b, r]\right] \subseteq B_{Y}(b, r)=: V_{2}$. Hence, $\Phi(x, \cdot)$ has a unique fixed point $g(x) \in V_{2}$ for every $x \in V_{1}$.

Concerning continuity of $g$, we observe for $x, x^{\prime} \in V_{1}$ sufficiently close

$$
\begin{aligned}
\left\|g(x)-g\left(x^{\prime}\right)\right\|_{Y}= & \left\|\Phi(x, g(x))-\Phi\left(x^{\prime}, g\left(x^{\prime}\right)\right)\right\|_{Y} \\
\leq & \left\|\Phi(x, g(x))-\Phi\left(x^{\prime}, g(x)\right)\right\|_{Y}+\left\|\Phi\left(x^{\prime}, g(x)\right)-\Phi\left(x^{\prime}, g\left(x^{\prime}\right)\right)\right\|_{Y} \\
\leq & \left\|\Phi(x, g(x))-\Phi\left(x^{\prime}, g(x)\right)\right\|_{Y} \\
& +\underbrace{\sup _{t \in[0,1]}\left\|1-B \partial_{2} F\left(x^{\prime},(1-t) g(x)+t g\left(x^{\prime}\right)\right)\right\|_{L(Y)}}_{<\frac{1}{2}}\left\|g(x)-g\left(x^{\prime}\right)\right\|_{Y} \\
& <\left\|\Phi(x, g(x))-\Phi\left(x^{\prime}, g(x)\right)\right\|_{Y}+\frac{1}{2}\left\|g(x)-g\left(x^{\prime}\right)\right\|_{Y}
\end{aligned}
$$

and, therefore,
$\left\|g(x)-g\left(x^{\prime}\right)\right\|_{Y}<2\left\|\Phi(x, g(x))-\Phi\left(x^{\prime}, g(x)\right)\right\|_{Y}=2\left\|B\left(F\left(x^{\prime}, g(x)\right)-F(x, g(x))\right)\right\|_{Y}$ which converges to zero as $x^{\prime} \rightarrow x$ because $F$ and $B$ are continuous.

Corollary 4.12. Using the notation of Theorem 4.11, let $g_{0}:=b$ and

$$
\forall n \in \mathbb{N} \forall x \in V_{1}: g_{n}(x):=g_{n-1}(x)-\partial_{2} F(a, b)^{-1} F\left(x, g_{n-1}(x)\right) .
$$

Then, $g_{n}$ converges to the implicit function $g$ pointwise in $Y$.
Proof. In the proof of Theorem 4.11 we constructed $g$ to be the unique fixed point of

$$
g(x)=\Phi(x, g(x))=g(x)-\partial_{2} F(a, b)^{-1} F(x, g(x))
$$

using Banach's fixed point theorem. Now defined $g_{n+1}(x)=\Phi\left(x, g_{n}(x)\right)$. Thus, $b=g_{0}(x) \in V_{2}=B_{Y}(b, r)$ for every $x \in V_{1}$ implies pointwise convergence of $\left(g_{n}\right)_{n \in \mathbb{N}}$ to $g$.

Now, that we can construct implicit functions, the remainder of the chapter will show that the solutions are sufficiently smooth if the function $F$ is and we will state the inverse function theorem since this is the theorem we will end up using.

Proposition 4.13. Let $X, Y$, and $Z$ be Banach spaces, $U_{1} \subseteq X$ open, $U_{2} \subseteq Y$ open, $F: U_{1} \times U_{2} \rightarrow Z,(a, b) \in U_{1}, \times U_{2}, F(a, b)=0, F$ Fréchet-differentiable at $(a, b)$, and $\partial_{2} F(a, b)$ an isomorphism. Let $g: U_{1} \rightarrow U_{2}$ be continuous at $a, g(a)=b$ and $\forall x \in U_{1}: \quad F(x, g(x))=0$.

Then, $g$ is Fréchet-differentiable at a satisfying

$$
g^{\prime}(a)=-\partial_{2} F(a, b)^{-1} \partial_{1} F(a, b) .
$$

Proof. Without loss of generality, let $a=0$ and $b=0$. Let $A:=\partial_{1} F(0,0) \in$ $L(X, Z)$ and $B:=\partial_{2} F(0,0) \in L(Y, Z)$. Then, $B$ is an isomorphism and there exists a function $\varphi: U_{1} \times U_{2} \rightarrow Z$ with $\varphi(x, y) \rightarrow 0((x, y) \rightarrow 0)$ and

$$
F(x, y)=\underbrace{F(0,0)}_{=0}+A x+B y+\left(\|x\|_{X}+\|y\|_{Y}\right) \varphi(x, y) .
$$

Thus,

$$
0=F(x, g(x))=A x+B g(x)+\left(\|x\|_{X}+\|g(x)\|_{Y}\right) \varphi(x, g(x))
$$

implies

$$
g(x)=-B^{-1} A x-\left(\|x\|_{X}+\|g(x)\|_{Y}\right) B^{-1} \varphi(x, g(x))=-B^{-1} A x+\|x\|_{X} \psi(x)
$$

with

$$
\psi(x):=-\left(1+\frac{\|g(x)\|_{Y}}{\|x\|_{X}}\right) B^{-1} \varphi(x, g(x))
$$

for $x \neq 0$. Hence, we will have to show $\psi(x) \rightarrow 0(x \rightarrow 0)$. Let $\delta \in \mathbb{R}_{>0}$ be such that $B_{X}(0, \delta) \subseteq U_{1}$ and

$$
\forall x \in B_{Y}(0, \delta):\left\|B^{-1} \varphi(x, g(x))\right\|_{Y} \leq \frac{1}{2}
$$

Then,

$$
\|g(x)\|_{Y} \leq\left\|B^{-1} A\right\|_{L(X, Y)}\|x\|_{X}+\frac{\|x\|_{X}+\|g(x)\|_{Y}}{2}
$$

implies

$$
\|g(x)\|_{Y} \leq \underbrace{\left(2\left\|B^{-1} A\right\|_{L(X, Y)}+1\right)}_{=: K}\|x\|_{X},
$$

i.e.,

$$
\|\psi(x)\|_{Y} \leq(1+K)\left\|B^{-1} \varphi(x, g(x))\right\|_{Y} \rightarrow 0 \quad(x \rightarrow 0)
$$

Remark on the Neumann series
Let $X$ be a Banach space, $T \in L(X)$, and $\|T\|_{L(X)}<1$. Then, $\sum_{k \in \mathbb{N}_{0}} T^{k}$ converges absolutely since

$$
\sum_{k \in \mathbb{N}_{0}}\left\|T^{k}\right\|_{L(X)} \leq \sum_{k \in \mathbb{N}_{0}}\|T\|_{L(X)}^{k}=\frac{1}{1-\|T\|_{L(X)}}
$$

Furthermore, we obtain

$$
(1-T) \sum_{k \in \mathbb{N}_{0}} T^{k}=\left(\sum_{k \in \mathbb{N}_{0}} T^{k}\right)(1-T)=1
$$

holds, i.e., $1-T$ is a homeomorphism with $(1-T)^{-1}=\sum_{k \in \mathbb{N}_{0}} T^{k}$.

Lemma 4.14. Let $X$ and $Y$ be Banach spaces, $S, T \in L(X, Y), 0 \in \varrho(T)$, and

$$
\|S-T\|_{L(X, Y)}<\left\|T^{-1}\right\|_{L(Y, X)}^{-1}
$$

Then, $0 \in \varrho(S)$ and

$$
B_{L(X, Y)}\left(T,\left\|T^{-1}\right\|_{L(Y, X)}^{-1}\right) \ni S \mapsto S^{-1} \in L(Y, X)
$$

is continuous. In particular, the set of isomorphism in $L(X, Y)$ is open.
Proof. Since $\left\|T^{-1}(S-T)\right\|_{L(X)} \leq\left\|T^{-1}\right\|_{L(Y, X)}\|S-T\|_{L(X, Y)}<1$, the Neumann series yields that $1+T^{-1}(S-T): X \rightarrow X$ is boundedly invertible. Using

$$
S=T\left(1+T^{-1}(S-T)\right)
$$

we obtain the assertion from

$$
S^{-1}=\left(1+T^{-1}(S-T)\right)^{-1} T^{-1} \in L(Y, X)
$$

Corollary 4.15. With the assumptions of Theorem 4.11, $M$ being an open subset of a Banach space $X$, and $F \in C^{m}(M \times U, Z)$ for some $m \in \mathbb{N} \cup\{\infty\}$, the set $V_{1}$ in Theorem 4.11 can be chosen such that $g: V_{1} \rightarrow V_{2}$ is in $C^{m}\left(V_{1}, V_{2}\right)$.

Proof. It is possible to choose $V_{1}$ such that $\forall x \in V_{1}: \partial_{2} F(x, g(x))$ is an isomorphism. Thus, Proposition 4.13 yields continuity of $g^{\prime}$ with

$$
g^{\prime}(x)=-\partial_{2} F(x, g(x))^{-1} \partial_{1} F(x, g(x))
$$

For $m \geq 2$ the right-hand side is Fréchet-differentiable and $g \in C^{2}\left(V_{1}, V_{2}\right)$, therefore. Inductively, we obtain $g \in C^{m}\left(V_{1}, V_{2}\right)$.

Theorem 4.16 (Inverse Function Theorem). Let $X$ and $Y$ be Banach spaces, $U \subseteq X$ open, $m \in \mathbb{N} \cup\{\infty\}, f \in C^{m}(U, Y), a \in U$, and $f^{\prime}(a)$ an isomorphism. Then, there are open neighborhoods $U_{1}$ of $a$ and $U_{2}$ of $b:=f(a)$ such that $f: U_{1} \rightarrow U_{2}$ is $a C^{m}$-diffeomorphism. Furthermore, $\left(f^{-1}\right)^{\prime}(b)=f^{\prime}(a)^{-1}$.

Proof. Let

$$
F: Y \times U \rightarrow Y ; \quad(y, x) \mapsto f(x)-y
$$

Then, $F \in C^{m}(Y \times U, Y), F(b, a)=0$, and $\partial_{2} F(b, a)=f^{\prime}(a)$ is an isomorphism. Hence, there are open neighborhoods $U_{2} \subseteq Y$ of $b$ and $V \subseteq U$ of $a$ such that $g$ : $U_{2} \rightarrow V$ is uniquely determined by $F(y, g(y))=0$ and $g \in C^{m}\left(U_{2}, V\right)$. Since for $x \in C$ and $y \in U_{2}$

$$
x=g(y) \Leftrightarrow F(y, x)=0 \Leftrightarrow y=f(x)
$$

holds, $U_{1}:=g\left[U_{2}\right]=\left[U_{2}\right] f \cap V$ is open and $f: U_{1} \rightarrow U_{2}$ is bijective with $g=\left(\left.f\right|_{U_{1}}\right)^{-1}$. Hence, $\left.g \circ f\right|_{U_{1}}=\left.\mathrm{id}\right|_{U_{1}}$ implies $\left(g^{\prime} \circ f\right) f^{\prime}=1$, i.e.,

$$
g^{\prime}(b)=f^{\prime}(a)^{-1}
$$

At this point we have shown the implicit function theorem and inverse function theorem for $C^{m}$-function with $m \in \mathbb{N} \cup\{\infty\}$. Now we will show that they are also true for analytic functions $\left(C^{\omega}\right)$.

Let $X$ be a Banach space and $r \in(0,1)$. We define $B_{r}:=B_{X}\left(0, r^{2}\right) \times B_{X}(0, r) \subseteq$ $X^{2}$, as well as, $E_{r}$ to be the set of all $u=\left((x, y) \mapsto \sum_{m, n \in \mathbb{N}_{0}} u_{m, n} x^{m} y^{n}\right) \in C^{\omega}\left(B_{r}, X\right)$ satisfying

$$
\|u\|_{E_{r}}:=\sum_{m, n \in \mathbb{N}_{0}}\left\|u_{m, n}\right\|_{\text {Lip }} r^{2 m+n}<\infty
$$

which itself defines a norm on $E_{r}$.
Lemma 4.17. $\left(E_{r},\|\cdot\|_{E_{r}}\right)$ is a Banach space.
Proof. Let $\left(u^{(n)}\right)_{n \in \mathbb{N}} \in E_{r}^{\mathbb{N}}$ be a Cauchy sequence. Then, all $\left(u_{i, j}^{(n)}\right)_{n \in \mathbb{N}} \in$ $L\left(X^{i+j}, X\right)^{\mathbb{N}}$ are Cauchy and, since $X$ is complete, so is $L\left(X^{i+j}, X\right)$, i.e., $u_{i, j}^{(n)} \rightarrow: u_{i, j}$ in $L\left(X^{i+j}, X\right)$ for every $i, j \in \mathbb{N}_{0}$.

Let $x_{1}, \ldots, x_{i+j} \in B_{r}, \alpha \in \mathbb{R}$, and $y \in B_{r}$ sufficiently small such that $u_{k}+\alpha y \in B_{r}$ for every $k \in \mathbb{N}_{\leq i+j}$. Then,

$$
\left.\begin{array}{rl}
u_{i, j}\left(x_{1}, \ldots, x_{k-1}, x_{k}+\alpha y, x_{k+1}, \ldots, x_{i+j}\right) & \leftarrow
\end{array} u_{i, j}^{(n)}\left(x_{1}, \ldots, x_{k-1}, x_{k}+\alpha y, x_{k+1}, \ldots, x_{i+j}\right)\right)
$$

shows multi-linearity of $u_{i, j}$.

Let

$$
\tilde{u}_{n}: B_{r} \rightarrow X ; \quad(x, y) \mapsto \sum_{i, j=0}^{n} u_{i, j} x^{i} y^{j}
$$

and $m \in \mathbb{N}$ sufficiently large such that $\forall i, j \in \mathbb{N}_{0, \leq n}:\left\|u_{i, j}-u_{i, j}^{(m)}\right\|_{\text {Lip }} \leq \frac{\varepsilon}{n^{2} r^{2 i+j}}$. Then,

$$
\begin{aligned}
& \sum_{i, j=0}^{n}\left\|u_{i, j}\right\|_{\text {Lip }} r^{2 i+j} \leq \sum_{i, j=0}^{n}\left\|u_{i, j}-u_{i, j}^{(m)}\right\|_{\text {Lip }} r^{2 i+j}+\sum_{i, j=0}^{n}\left\|u_{i, j}^{(m)}\right\|_{\text {Lip }} r^{2 i+j} \\
& \leq \varepsilon+\underbrace{\sup _{k \in \mathbb{N}}\left\|u^{(k)}\right\|_{E_{r}}}_{<\infty}
\end{aligned}
$$

Hence, the pointwise limit $\tilde{u}_{n} \rightarrow: u(n \rightarrow \infty)$ exists and is an element of $E_{r}$.
In order to show that $\left\|u-\tilde{u}_{n}\right\|_{E_{r}}=\sum_{i, j \in \mathbb{N}_{>_{n}}}\left\|u_{i, j}-u_{i, j}^{(n)}\right\|_{\text {Lip }} r^{2 i+j}$ converges to zero, let $\varepsilon \in \mathbb{R}_{>0}$.
(i) Choose $m_{1} \in \mathbb{N}$ such that $\forall m \in \mathbb{N}_{\geq m_{1}}:\left\|u-\tilde{u}_{m}\right\|_{E_{r}}<\frac{\varepsilon}{4}$.
(ii) Choose $n_{1} \in \mathbb{N}$ such that $\forall n^{\prime}, n^{\prime \prime} \in \mathbb{N}_{\geq n_{1}}$ : $\left\|u^{\left(n^{\prime}\right)}-u^{\left(n^{\prime \prime}\right)}\right\|_{E_{r}}<\frac{\varepsilon}{4}$.
(iii) Choose $m_{2} \in \mathbb{N}_{\geq m_{1}}$ such that $\sum_{i, j \in \mathbb{N}_{>m_{2}}}\left\|u_{i, j}^{\left(n_{1}\right)}\right\|_{\text {Lip }} r^{2 i+j}<\frac{\varepsilon}{4}$.
(iv) Since all $\left\|u_{i, j}^{(n)}\right\|_{\text {Lip }} r^{2 i+j}$ converge to zero, let $n_{2} \in \mathbb{N}_{\geq n_{1}}$ be such that $\forall n \in$

$$
\mathbb{N}_{\geq n_{2}} \forall i, j \in \mathbb{N}_{0, \leq m_{2}}:\left\|u_{i, j}^{(n)}\right\|_{\text {Lip }} r^{2 i+j}<\frac{\varepsilon}{4\left(m_{2}+1\right)^{2}}
$$

Then, we observe for $n \in \mathbb{N}_{\geq n_{2}}$

$$
\begin{aligned}
\left\|u-u^{(n)}\right\|_{E_{r}} & \leq \underbrace{\left\|u-\tilde{u}_{m_{2}}\right\|_{E_{r}}}_{<\frac{\varepsilon}{4}}+\underbrace{\sum_{i, j=0}^{m_{2}}\left\|u_{i, j}-u_{i, j}^{(n)}\right\|_{\text {Lip }}}_{<\frac{\varepsilon}{4}} r^{2 i+j}+\sum_{i, j \in \mathbb{N}_{>m_{2}}}\left\|u_{i, j}^{(n)}\right\|_{\text {Lip }} r^{2 i+j} \\
& <\frac{\varepsilon}{2}+\underbrace{\sum_{i, j \in \mathbb{N}_{>m_{2}}}\left\|u_{i, j}^{(n)}-u_{i, j}^{\left(n_{1}\right)}\right\|_{\text {Lip }} r^{2 i+j}}_{\leq\left\|u^{(n)}-u^{\left(n_{1}\right)}\right\|_{E_{r}}<\frac{\varepsilon}{4}}+\underbrace{\sum_{i, j \in \mathbb{N}_{>m_{2}}}\left\|u_{i, j}^{\left(n_{1}\right)}\right\|_{L i p} r^{2 i+j}}_{<\frac{\varepsilon}{4}} \\
& <\varepsilon
\end{aligned}
$$

which completes the proof.

Let us also consider the subspace

$$
F_{r}:=\left\{u \in E_{r} ; \forall m \in \mathbb{N}_{0}: u_{m, 0}=0\right\} .
$$

Clearly, $F_{r}$ is a closed subspace, i.e., a Banach space itself. Furthermore, let us define $L \in L\left(F_{r}\right)$ by

$$
\forall(x, y) \in B_{r}: L u(x, y):=\sum_{(m, n) \in \mathbb{N}_{0} \times \mathbb{N}} \frac{1}{n} u_{m, n} x^{m} y^{n}
$$

and for $w \in E_{r}$

$$
\forall(x, y) \in B_{r}: L_{w} u(x, y):=\partial_{2} u(x, y) w(x, y)-\partial_{2} u(x, 0) w(x, 0)
$$

Obviously, we obtain $\|L\|_{L\left(F_{r}\right)}=1$ and, for $w_{0}: B_{r} \rightarrow X ; \quad(x, y) \mapsto y, L_{w_{0}} \circ L=\operatorname{id}_{F_{r}}$.
Lemma 4.18. $L_{w} \circ L$ is in $L\left(F_{r}\right)$ and satisfies $\left\|L_{w} \circ L\right\|_{L\left(F_{r}\right)} \leq \frac{\|w\|_{E_{r}}}{r}$.

Proof. Let $w$ be decomposed as

$$
w(x, y)=\sum_{m, n \in \mathbb{N}_{0}} w_{m, n} x^{m} y^{n}
$$

Then, for $u=\left((x, y) \mapsto \sum_{(m, n) \in \mathbb{N}_{0} \times \mathbb{N}} u_{m, n} x^{m} y^{n}\right) \in F_{r}$,

$$
\begin{aligned}
L_{w} L u(z, 0)= & \left((x, y) \mapsto L_{w} \sum_{(m, n) \in \mathbb{N}_{0} \times \mathbb{N}} \frac{1}{n} u_{m, n} x^{m} y^{n}\right)(z, 0) \\
= & \left.\left(\sum_{m, n \in \mathbb{N}_{0}} u_{m, n+1} x^{m} y^{n}\right)\left(\sum_{m, n \in \mathbb{N}_{0}} w_{m, n} x^{m} y^{n}\right)\right|_{(x, y)=(z, 0)} \\
& -\left.\left(\sum_{m \in \mathbb{N}_{0}} u_{m, 1} x^{m}\right)\left(\sum_{m \in \mathbb{N}_{0}} w_{m, 0} x^{m}\right)\right|_{(x, y)=(z, 0)} \\
= & \left(\sum_{m \in \mathbb{N}_{0}} u_{m, 1} z^{m}\right)\left(\sum_{m \in \mathbb{N}_{0}} w_{m, 0} x^{m}\right)-\left(\sum_{m \in \mathbb{N}_{0}} u_{m, 1} z^{m}\right)\left(\sum_{m \in \mathbb{N}_{0}} w_{m, 0} z^{m}\right) \\
= & 0
\end{aligned}
$$

shows $L_{w} L u \in F_{r}$. Furthermore,

$$
\begin{aligned}
& L_{w} L u(x, y) \\
= & L_{w} \sum_{(m, n) \in \mathbb{N}_{0} \times \mathbb{N}} \frac{1}{n} u_{m, n} x^{m} y^{n} \\
= & \left(\sum_{m, n \in \mathbb{N}_{0}} u_{m, n+1} x^{m} y^{n}\right)\left(\sum_{m, n \in \mathbb{N}_{0}} w_{m, n} x^{m} y^{n}\right)-\left(\sum_{m \in \mathbb{N}_{0}} u_{m, 1} x^{m}\right)\left(\sum_{m \in \mathbb{N}_{0}} w_{m, 0} x^{m}\right) \\
= & \sum_{(M, N) \in \mathbb{N}_{0} \times \mathbb{N}}(m, n) \in \mathbb{N}_{0, \leq(M, N)}
\end{aligned}
$$

implies

$$
\begin{aligned}
& \left\|L_{w} L u\right\|_{E_{r}} \\
\leq & \sum_{(M, N) \in \mathbb{N}_{0} \times \mathbb{N}}\left(\sum_{(m, n) \in \mathbb{N}_{0, \leq(M, N)}^{2}}\left\|u_{m, n+1}\right\|_{\text {Lip }}\left\|w_{M-m, N-n}\right\|_{\text {Lip }}\right) r^{2 M+N} \\
= & \frac{1}{r} \sum_{(M, N) \in \mathbb{N}_{0} \times \mathbb{N}}\left(\sum_{(m, n) \in \mathbb{N}_{0, \leq(M, N)}^{2}}\left\|u_{m, n+1}\right\|_{\text {Lip }} r^{2 m+n+1}\left\|w_{M-m, N-n}\right\|_{\text {Lip }} r^{2(M-m)+(N-n)}\right) \\
= & \frac{1}{r}\left(\sum_{m, n \in \mathbb{N}_{0}}\left\|u_{m, n+1}\right\|_{\text {Lip }} r^{2 m+n+1}\right)\left(\sum_{m, n \in \mathbb{N}_{0}}\left\|w_{m, n}\right\|_{\text {Lip }} r^{2 m+n)}\right) \\
= & \frac{\|w\|_{E_{r}}\|u\|_{E_{r}}}{r} .
\end{aligned}
$$

Lemma 4.19. Let $X$ be a Banach space, $U \subseteq X$ an open neighborhood of zero, $F \in C^{\omega}(U, X), F(0)=0$, and $F^{\prime}(0)=1$. Let $V \subseteq X$ be an open neighborhood of zero and $G: V \rightarrow X$ a local inverse of $F$ at zero. Then, $G$ is analytic in an open neighborhood of zero.

Proof. For $r \in \mathbb{R}_{>0}$ sufficiently small, let $v, w \in E_{r}$ be defined by

$$
\forall(x, y) \in B_{r}: v(x, y):=F(y)-x
$$

and

$$
\forall(x, y) \in B_{r}: w(x, y):=v(x, y)-w_{0}(x, y)=F(y)-x-y
$$

Then,

$$
\forall(x, y) \in B_{r}: w(x, y)=-x+\sum_{n \in \mathbb{N}_{\geq 2}} \frac{1}{n!} \partial^{n} F(0) y^{n}
$$

and

$$
\|w\|_{E_{r}} \leq r^{2}+\sum_{n \in \mathbb{N}_{\geq 2}} \frac{\left\|\partial^{n} F(0)\right\|_{\text {Lip }}}{n!} r^{n} \leq r^{2} C_{F}
$$

holds where $C_{F} \in \mathbb{R}_{>0}$ is a constant solely dependent on $F$. From the definitions of $v$ and $w$, we obtain

$$
L_{v} \circ L-1=\left(L_{v}-L_{w_{0}}\right) \circ L=L_{w} \circ L
$$

and, hence, for $r<\frac{1}{C_{F}}$,

$$
\left\|L_{v} \circ L-1\right\|_{L\left(F_{r}\right)} \leq r C_{F}<1
$$

and, according to Lemma 4.14, $L_{v} \circ L$ is an isomorphism on $F_{r}$. Let $u_{0}:=\left(L_{v} \circ\right.$ $L)^{-1} w_{0}$. Then, we obtain for all $(x, y) \in B_{r}$
(*) $y=w_{0}(x, y)=\left(L_{v} \circ L\right) u_{0}(x, y)=\partial_{2}\left(L u_{0}\right)(x, y) v(x, y)-\partial_{2}\left(L u_{0}\right)(x, 0) v(x, 0)$.
In particular, we observe for $y \in B_{X}(0, r)$ and $t \in(0,1)$

$$
\begin{aligned}
t y= & \partial_{2}\left(L u_{0}\right)(0, t y) v(0, t y)-\partial_{2}\left(L u_{0}\right)(0,0) v(0,0) \\
= & \partial_{2}\left(L u_{0}\right)(0, t y) v(0, t y)-\partial_{2}\left(L u_{0}\right)(0, t y) v(0,0) \\
& +\partial_{2}\left(L u_{0}\right)(0, t y) v(0,0)-\partial_{2}\left(L u_{0}\right)(0,0) v(0,0) \\
= & \partial_{2}\left(L u_{0}\right)(0, t y)(F(t y)-F(0))+\left(\partial_{2}\left(L u_{0}\right)(0, t y)-\partial_{2}\left(L u_{0}\right)(0,0)\right) F(0) \\
= & \partial_{2}\left(L u_{0}\right)(0, t y)(F(t y)-F(0))
\end{aligned}
$$

which (dividing by $t$ and $t \searrow 0$ ) shows

$$
\forall y \in B_{X}(0, r): y=\partial_{2}\left(L u_{0}\right)(0,0) F^{\prime}(0) y=\partial_{2}\left(L u_{0}\right)(0,0) y
$$

i.e., $\partial_{2}\left(L u_{0}\right)(0,0)=1=\mathrm{id}_{X}$. Hence, there exists $\varepsilon \in(0, r)$ such that $\partial_{2}\left(L u_{0}\right)(x, y)$ is a bijection on $X$ for every $(x, y) \in B_{X}\left(0, \varepsilon^{2}\right) \times B_{X}(0, \varepsilon)$.

Defining

$$
\tilde{G}: B_{X}\left(0, \varepsilon^{2}\right) \rightarrow X ; x \mapsto \partial_{2}\left(L u_{0}\right)(x, 0) x
$$

we observe $\tilde{G} \in C^{\omega}\left(B_{X}\left(0, \varepsilon^{2}\right), X\right)$ and

$$
\begin{aligned}
\tilde{G}(x) & =\partial_{2}\left(L u_{0}\right)(x, 0) x \\
& =-\partial_{2}\left(L u_{0}\right)(x, 0)(F(0)-x) \\
& =-\partial_{2}\left(L u_{0}\right)(x, 0) v(x, 0) \\
& \stackrel{(*)}{=} y-\partial_{2}\left(L u_{0}\right)(x, y) v(x, y) \\
& =y-\partial_{2}\left(L u_{0}\right)(x, y)(F(y)-x)
\end{aligned}
$$

i.e., for $y=G(x)$,

$$
\tilde{G}(x)=G(x)-\underbrace{\partial_{2}\left(L u_{0}\right)(x, G(x))}_{\text {linear }} \underbrace{(F(G(x))-x)}_{=0}=G(x) .
$$

Hence, $G$ is analytic on $V \cap B_{X}\left(0, \varepsilon^{2}\right)$ which is an open neighborhood of zero.

Theorem 4.20 (Analytic Inverse Function Theorem). Let $X$ and $Y$ be Banach spaces, $U \subseteq X$ open, $m \in \mathbb{N} \cup\{\infty, \omega\}, f \in C^{m}(U, Y)$, $a \in U$, and $f^{\prime}(a)$ an isomorphism. Then, there are open neighborhoods $U_{1}$ of a and $U_{2}$ of $b:=f(a)$ such that $f: U_{1} \rightarrow U_{2}$ is a $C^{m}$-diffeomorphism. Furthermore, $\left(f^{-1}\right)^{\prime}(b)=f^{\prime}(a)^{-1}$.

Proof. The inverse function theorem (Theorem 4.16) yields the assertion for $m \in \mathbb{N} \cup\{\infty\}$, i.e., it suffices to show the assertion for $m=\omega$ knowing that $f: V_{1} \rightarrow V_{2}$ is a $C^{\infty}$-diffeomorphism for some open neighborhoods $V_{1}$ of $a$ and $V_{2}$ of $b$. Let $\tilde{U}:=U-a$ and

$$
\tilde{f}: \tilde{U} \rightarrow X ; x \mapsto f^{\prime}(a)^{-1}(f(x+a)-f(a))
$$

Then, $\tilde{f}(0)=0$ and $\tilde{f}^{\prime}(0)=1$. Thus, Lemma 4.19 yields that $\tilde{f}$ is a $C^{\omega}\left(\tilde{U}_{1}, \tilde{U}_{2}\right)$ diffeomorphism for some neighborhoods $\tilde{U}_{1}$ and $\tilde{U}_{2}$ of zero. Finally,

$$
\forall x \in\left(a+\tilde{U}_{1}\right) \cap V_{1}: f(x)=f(a)+f^{\prime}(a) \tilde{f}(x-a)
$$

implies the assertion for $U_{1}:=\left(a+\tilde{U}_{1}\right) \cap V_{1}$ and $U_{2}:=f\left[U_{1}\right]$.

Theorem 4.21 (Analytic Implicit Function Theorem). Let $X, Y$, and $Z$ be Banach spaces, $U_{1} \subseteq X$ open, $U_{2} \subseteq Y$ open, $m \in \mathbb{N} \cup\{\infty, \omega\}, F \in C^{m}\left(U_{1} \times U_{2}, Z\right)$, $(a, b) \in U_{1} \times U_{2}, F(a, b)=0$, and $\partial_{2} F(a, b)$ an isomorphism. Then, there are open neighborhoods $V_{1} \subseteq X$ of a and $V_{2} \subseteq Y$ of $b$ such that there is a unique function $g: V_{1} \rightarrow V_{2}$ with $\forall x \in V_{1}: F(x, g(x))=0$. Furthermore, $g \in C^{m}\left(V_{1}, V_{2}\right)$.

Proof. We already know the assertion for $m \in \mathbb{N} \cup\{\infty\}$. Hence, let $m=\omega$ and $g \in C^{\infty}\left(W_{1}, W_{2}\right)$ the implicit function with open neighborhoods $W_{1} \subseteq X$ of $a$ and $W_{2} \subseteq Y$ of $b$. Let

$$
G: U_{1} \times U_{2} \rightarrow Z \times X ; \quad(x, y) \mapsto(F(x, y), x)
$$

Then,

$$
G^{\prime}(a, b)(x, y)=\left(\partial_{1} F(a, b) x+\partial_{2} F(a, b) y, x\right)
$$

holds for every $(x, y) \in X \times Y$. Thus, $G^{\prime}(a, b)$ has the bounded inverse

$$
G^{\prime}(a, b)^{-1}: Z \times X \mapsto X \times Y ;(z, x) \mapsto\left(x, \partial_{2} F(a, b)^{-1}\left(z-\partial_{1} F(a, b) x\right)\right)
$$

and the analytic inverse function theorem (Theorem 4.20) yields open sets $\tilde{U}_{1} \subseteq U_{1}$, $\tilde{U}_{2} \subseteq U_{2}$, and $V_{0} \subseteq Z \times X$ such that $(a, b) \in \tilde{U}_{1} \times \tilde{U}_{2}$ and $G$ is a $C^{\omega}\left(\tilde{U}_{1} \times \tilde{U}_{2}, V_{0}\right)$ diffeomorphism. Furthermore, we observe

$$
g(x)=\operatorname{pr}_{2}(x, g(x))=\left(\operatorname{pr}_{2} \circ G^{-1}\right)(F(x, g(x)), x)=\left(\operatorname{pr}_{2} \circ G^{-1}\right)(0, x)
$$

for every $x \in W_{1} \cap \tilde{U}_{1}=: V_{1}$. Observation 4.3, thus, yields that

$$
V_{1} \ni x \mapsto\left(\operatorname{pr}_{2} \circ G^{-1}\right)(0, x) \in Y
$$

is analytic, i.e., $g$ is a $C^{\omega}\left(V_{1}, V_{2}\right)$-diffeomorphism where $V_{2}:=g\left[V_{1}\right]$.

To conclude this chapter we will prove the incredibly handy fact that composition of analytic functions yields an analytic function.

Proposition 4.22. Let $X, Y$, and $Z$ be Banach spaces, $U \subseteq X$ open, $V \subseteq Y$ open, $F \in C^{\omega}(U, V)$, and $G \in C^{\omega}(V, Z)$. Then, $G \circ F \in C^{\omega}(U, Z)$.

Proof. Let $W:=U \times(V \times Z)$ and $H: W \rightarrow Y \times Z ;(x,(y, z)) \mapsto(F(x)-$ $y, G(y)-z)$. Let $x_{0} \in U, y_{0}:=F\left(x_{0}\right)$, and $z_{0}:=G\left(y_{0}\right)$. Then, we observe

$$
H\left(x_{0},\left(y_{0}, z_{0}\right)\right)=0
$$

and the equation

$$
(\hat{y}, \hat{z})=\partial_{2} H\left(x_{0},\left(y_{0}, z_{0}\right)\right)(y, z)=(y, z)\left(\begin{array}{cc}
-1 & G^{\prime}\left(y_{0}\right) \\
0 & -1
\end{array}\right)=\left(-y, G^{\prime}\left(y_{0}\right) y-z\right)
$$

is equivalent to $y=-\hat{y}$ and $z=-\hat{z}-G^{\prime}\left(y_{0}\right) \hat{y}$. Thus, $\partial_{2} H\left(x_{0},\left(y_{0}, z_{0}\right)\right)$ is an isomorphism, and, by the analytic implicit function theorem, there is an analytic implicit function $(\hat{Y}, \hat{Z})$ solving

$$
H(x,(\hat{Y}(x), \hat{Z}(x)))=0
$$

in an open neighborhood of $x_{0}$. But $H(x,(y, z))=0$ implies $G(F(x))=z$ and, thus, $G \circ F=\hat{Z}$. Observation 4.3 yields analyticity of $G \circ F$ at $x_{0}$ and, since $x_{0}$ was arbitrarily chosen in $U, G \circ F$ is analytic.

## CHAPTER 5

## Fredholm Operators

At last, we will state a few facts about Fredholm operators as we will use them quite extensively in the proof of well-posedness of the Navier-Stokes equations.

Definition 5.1. Let $X$ and $Y$ be Banach spaces. $T \in L(X, Y)$ is called a Fredholm operator if and only if $\operatorname{dim}[\{0\}] T$ and $\operatorname{codim} T[X]$ are finite.

The number $\operatorname{ind}(T):=\operatorname{dim}[\{0\}] T-\operatorname{codim} T[X]$ is called the index of $T$.
Definition 5.2. Let $X$ and $Y$ be Banach spaces. A linear operator $T \subseteq X \oplus Y$ has finite rank if and only if $\operatorname{dim} T[X]$ is finite.

Corollary 5.3. Every bounded finite rank operator is compact. In particular, if $X$ and $Y$ are Banach spaces, one of which is finite dimensional, then every $T \in L(X, Y)$ has finite rank, i.e., is compact.

Lemma 5.4. Let $H$ be a Hilbert space, $M \subseteq H$ a closed subspace, and $V \subseteq H a$ finite dimensional subspace. Then, $M+V$ is closed.

In particular, if codim $M \in \mathbb{N}_{0}$ and $W \subseteq H$ is a subspace with $M \subseteq W$, then $W$ is closed and codim $W \in \mathbb{N}_{0}$.

Proof. Let $P: H \rightarrow M$ be the orthogonal projection and $V_{\perp}:=(1-P) V$. Then, $M+V=M \oplus V_{\perp}$ where $M \oplus V_{\perp}$ is an orthogonal direct sum, i.e., $M+V$ is closed since a sequence $\left(\left(x_{n}, y_{n}\right)\right)_{n \in \mathbb{N}} \in\left(M \oplus V_{\perp}\right)^{\mathbb{N}}$ converges if and only if $\left(x_{n}\right)_{n \in \mathbb{N}}$ converges in $M$ and $\left(y_{n}\right)_{n \in \mathbb{N}}$ converges in $V_{\perp}$ and both spaces are closed ( $M$ by assumption and $V_{\perp}$ since it is finite dimensional).

Let $\operatorname{codim} M \in \mathbb{N}_{0}$. Then, there is a finite dimensional subspace $\tilde{V}$ such that $W=M+\tilde{V}$, i.e., $W$ is closed by the previous part of the proof, and codim $W \leq$ $\operatorname{codim} M \in \mathbb{N}_{0}$ is trivial.

Proposition 5.5. Let $X$ and $Y$ be Banach spaces, and $T \in L(X, Y)$ a Fredholm operator.
(i) If $\operatorname{ind}(T)=0$ and $T$ is injective, then $T$ is continuously invertible, i.e., $0 \in \varrho(T)$.
(ii) The range $T[X]$ of $T$ is closed. Furthermore, the equation $T x=y$ has a solution $x \in X$ for given $y \in Y$ if and only if $\forall x^{*} \in[\{0\}] T^{*}:\left\langle x^{*}, y\right\rangle=0$ where $T^{*}$ denotes the dual operator.
(iii) Let $S \in L(X, Y)$ be compact. Then, $T+S$ is a Fredholm operator with $\operatorname{ind}(T+S)=\operatorname{ind}(T)$.
(iv) The dual operator $T^{*}$ is a Fredholm operator with

$$
\operatorname{dim}[\{0\}] T^{*}=\operatorname{codim} T[X] \text { and } \operatorname{codim} T^{*}\left[Y^{\prime}\right]=\operatorname{dim}[\{0\}] T
$$

In particular, $\operatorname{ind}(T)=-\operatorname{ind}\left(T^{*}\right)$ and the equation $T^{*} y^{*}=x^{*}$ has a solution $y^{*} \in Y^{\prime}$ for given $x^{*} \in X^{\prime}$ if and only if $\forall x \in[\{0\}] T:\left\langle x^{*}, x\right\rangle=0$.

Proof. [21] Proposition 8.14

Remark The range of a Fredholm operator being closed is non-trivial. Let $X$ be a Banach space and $X_{0} \mp X$ a dense subspace. Let $x_{0} \in X \backslash X_{0}$ and $V:=\{x \in$ $X ;\left(x, x_{0}\right)$ linearly independent $\} \cup\{0\}$. Then, $\operatorname{codim} V=1$ and $X_{0} \subseteq V \mp X$. Since $X_{0}$ is dense, so is $V$. However, $V$ cannot be closed since $V \neq X$.

Proposition 5.6. Let $H_{1}$ and $H_{2}$ be Hilbert spaces and $F \in L\left(H_{1}, H_{2}\right)$. Then, the following are equivalent.
(i) $F$ is a Fredholm operator.
(ii) There exists $A \in L\left(H_{2}, H_{1}\right)$ such that $A F-1$ and $F A-1$ are both compact.
(iii) There exists $A \in L\left(H_{2}, H_{1}\right)$ such that $A F-1$ and $F A-1$ are both of finite rank.

Proof. "(i) $\Rightarrow\left(\right.$ ii)" Let $F$ be a Fredholm operator, $x, y \in[\{0\}] F^{\perp}$, and $F x=F y$. Then, $F(x-y)=0$, i.e., $x-y \in[\{0\}] F \cap[\{0\}] F^{\perp}=\{0\}$. Thus, $F:[\{0\}] F^{\perp} \rightarrow F\left[H_{1}\right]$ is bijective. Let $G: F\left[H_{1}\right] \rightarrow[\{0\}] F^{\perp}$ be the inverse of $F$ on $F\left[H_{1}\right], P: H_{1} \rightarrow$ $[\{0\}] F^{\perp}$ and $Q: H_{2} \rightarrow F\left[H_{1}\right]$ the orthoprojections, and $A:=G Q$. Then,

$$
A F-1=G Q F-1=G F-1=P-1
$$

and

$$
F A-1=F G Q-1=Q-1
$$

hold. Since $P-1$ and $Q-1$ are of finite rank (they are the orthoprojections on $[\{0\}] F$ and $F\left[H_{1}\right]^{\perp}$ ), they are, in particular, compact.
"(ii) $\Rightarrow$ (iii)" Since $A F-1$ is compact, there are $G_{1} \in L\left(H_{1}\right)$ of finite rank and $\Delta_{1} \in B_{L\left(H_{1}\right)}(0,1)$ such that $A F-1=G_{1}+\Delta_{1}$ because compact operators are limits of finite rank operators. Let $A_{1}:=\left(1-\Delta_{1}\right)^{-1} A$. Then, we observe

$$
A_{1} F=\left(1-\Delta_{1}\right)^{-1} A F=\left(1-\Delta_{1}\right)^{-1}\left(1+G_{1}+\Delta_{1}\right)=1+\left(1-\Delta_{1}\right)^{-1} G_{1}
$$

where $\left(1-\Delta_{1}\right)^{-1} G_{1}=: B_{1}$ is another operator of finite rank. Similarly, we can choose $A_{2}, G_{2}, \Delta_{2}$, and $B_{2}$ (with the same properties as the operators with index 1) such that $F A_{2}=1+B_{2}$. Since

$$
A_{1}+A_{1} B_{2}=A_{1} F A_{2}=A_{2}+B_{1} A_{2}
$$

holds, we may define the finite rank operator

$$
J:=A_{1}-A_{2}=B_{1} A_{2}-A_{1} B_{2}
$$

and observe

$$
F A_{1}-1=F A_{1}-\left(F A_{2}-B_{2}\right)=F J+B_{2}
$$

which is of finite rank, as well as,

$$
A_{1} F-1=1+B_{1}-1=B_{1} .
$$

Hence, the operator $A$ can be modified to $A_{1}$ such that $A_{1} F-1$ and $F A_{1}-1$ are both of finite rank.
"(iii) $\Rightarrow(\mathrm{i})$ " Let $A F-1=: G_{1}$ and $F A-1=: G_{2}$. Since $G_{1}$ and $G_{2}$ have finite rank, $A F=1+G_{1}$ and $F A=1+G_{2}$ are Fredholm operators. Thus,

$$
[\{0\}] F \subseteq[\{0\}] A F,
$$

i.e.,

$$
\operatorname{dim}[\{0\}] F \leq \operatorname{dim}[\{0\}] A F \in \mathbb{N}_{0}
$$

and

$$
F\left[H_{1}\right] \supseteq F A\left[H_{2}\right]
$$

i.e.,

$$
\operatorname{codim} F\left[H_{1}\right] \leq \operatorname{codim} F A\left[H_{2}\right] \in \mathbb{N}_{0}
$$

yield the assertion.

Observation 5.7. Let $H_{1}$ and $H_{2}$ be Hilbert spaces and $F \in L\left(H_{1}, H_{2}\right)$ be a Fredholm operator. Then, $F^{*}$ is a Fredholm operator with $\operatorname{ind}\left(F^{*}\right)=-\operatorname{ind}(F)$.

Proof. The observation follows directly from

$$
[\{0\}] F^{*}=F\left[H_{1}\right]^{\perp} \text { and } F^{*}\left[H_{2}\right]^{\perp}=[\{0\}] F
$$

Proposition 5.8. Let $H_{1}$ and $H_{2}$ be Hilbert spaces and $F \in L\left(H_{1}, H_{2}\right)$. Then, $F$ is a Fredholm operator if and only if there are orthogonal decompositions $H_{1}=$ $H_{11} \oplus H_{12}$ and $H_{2}=H_{21} \oplus H_{22}$ such that

- $H_{11}$ and $H_{21}$ are closed,
- $H_{12}$ and $H_{22}$ are finite dimensional, and
- $F$ has the block decomposition

$$
\left(\begin{array}{ll}
F_{11} & F_{12} \\
F_{21} & F_{22}
\end{array}\right): H_{11} \oplus H_{12} \rightarrow H_{21} \oplus H_{22}
$$

with $F_{11} \in L\left(H_{11}, H_{21}\right)$ boundedly invertible.
Furthermore, given this decomposition, $\operatorname{ind}(F)=\operatorname{dim} H_{12}-\operatorname{dim} H_{22}$.
Proof. [4] Lemma 16.34
Proposition 5.9. Let $H_{1}$ and $H_{2}$ be Hilbert spaces, $F_{1}, C, \Delta \in L\left(H_{1}, H_{2}\right), F_{2} \in$ $L\left(H_{2}, H_{1}\right), F_{1}$ and $F_{2}$ Fredholm operators, $C$ compact, and $\|\Delta\|_{L\left(H_{1}, H_{2}\right)}$ sufficiently small. Then, $F_{1}+\Delta, F_{1}+C$. and $F_{1} F_{2}$ are Fredholm operators with $\operatorname{ind}\left(F_{1}+\Delta\right)=$ $\operatorname{ind}\left(F_{1}+C\right)=\operatorname{ind}\left(F_{1}\right)$ and $\operatorname{ind}\left(F_{1} F_{2}\right)=\operatorname{ind}\left(F_{1}\right)+\operatorname{ind}\left(F_{2}\right)$.

In particular, the set of Fredholm operators in $L\left(H_{1}, H_{2}\right)$ is open.
Proof. [4] Proposition 16.35
Proposition 5.10. Let $H_{1}$ and $H_{2}$ be Hilbert spaces. The index of a Fredholm operator is constant on connected components of the set of Fredholm operators in $L\left(H_{1}, H_{2}\right)$ and is a bijection between $\mathbb{Z}$ and the connected components.

Proof. [7] Theorem 1.4 (b)
Remark In fact, it can be shown that the set of Fredholm operators in $L(H)$ of a given index is path connected where $H$ is a Hilbert space. To prove this, recall that if $A_{1}$ and $A_{2}$ have index $k$, then $A_{1}^{*} A_{2}$ has index zero. If $A_{1}^{*} A_{2}$ can be connected to the identity by $\gamma_{1}$, then $A_{1} \gamma_{1}$ connects $A_{1}$ with $A_{1} A_{1}^{*} A_{2}$. On the other hand, if the operator $A_{1} A_{1}^{*}$, which is also of index zero, can be connected to the identity using $\gamma_{2}$, then $\gamma_{2} A_{2}$ connects $A_{2}$ with $A_{1} A_{1}^{*} A_{2}$. Hence, it suffices to show that the set of operators of index zero are path connected. In that case, $[\{0\}] A$ and $A[H]^{\perp}$ are isomorphic since they are finite dimensional spaces of the same dimension. Thus, for such an isomorphism $I$, which we extend by zero on $[\{0\}] A^{\perp}$, the path $[0,1] \ni t \mapsto A+t I \in L(H)$ connects $A$ with an isomorphism (cf., Observation 5.11 below) and $G L(H)$ is known to be path connected (even more so, Kuiper's Theorem states that $G L(H)$ is contractible).

Observation 5.11. Let $H_{1}$ and $H_{2}$ be Hilbert spaces, $G L\left(H_{1}, H_{2}\right)$ the set of isomorphisms mapping $H_{1}$ to $H_{2}$, and $F_{k}\left(H_{1}, H_{2}\right)$ the set of Fredholm operators of index $k \in \mathbb{Z}$ mapping $H_{1}$ to $H_{2}$. Then, $G L\left(H_{1}, H_{2}\right)$ is dense in $F_{0}\left(H_{1}, H_{2}\right)$.

Proof. Obviously $G L\left(H_{1}, H_{2}\right) \subseteq F_{0}\left(H_{1}, H_{2}\right)$ holds since every isomorphism is bijective. Let $A \in F_{0}\left(H_{1}, H_{2}\right)$. Then, $\operatorname{dim}[\{0\}] A=\operatorname{dim} A\left[H_{1}\right]^{\perp} \in \mathbb{N}_{0}$ holds, i.e., [\{0\}] $A$ and $A\left[H_{1}\right]^{\perp}$ are isomorphic. Let $I$ be such an isomorphism, decompose $H_{1}=[\{0\}] A^{\perp} \oplus[\{0\}] A$, and define for $t \in[0,1]$

$$
A_{t}:[\{0\}] A^{\perp} \oplus[\{0\}] A \rightarrow A\left[H_{1}\right] \oplus A\left[H_{1}\right]^{\perp} ; x+y \mapsto A x+t I y
$$

which, by definition, is bijective for $t>0$ (note that $\left.A\right|_{[\{0\}] A^{\perp}}$ is injective and surjective on to $\left.A\left[H_{1}\right]\right)$. Thus, observing $A_{t} \rightarrow A(t \searrow 0)$ completes the proof.

Observation 5.12. Let $X$ and $Y$ be Banach spaces, $S, T \in L(X, Y), S$ bijective, and $T$ compact. Then, $S+T$ is a Fredholm operator of index zero.

In particular, the following are equivalent.
(i) $S+T$ is injective.
(ii) $S+T$ is surjective.
(iii) $S+T$ is bijective.

Proof. Since $S$ is bounded and bijective, $S$ is Fredholm of index zero. Hence, Proposition 5.9 implies that $S+T$ is a Fredholm operator of index zero, as well.

In particular, we have $\operatorname{dim}[\{0\}](S+T)=\operatorname{codim}(S+T)[X]$, i.e., (i) $\Leftrightarrow($ ii $)$, which implies $(\mathrm{i}) \Rightarrow(\mathrm{ii}) \Rightarrow(\mathrm{i}) \wedge(\mathrm{ii}) \Rightarrow(\mathrm{iii}) \Rightarrow(\mathrm{i})$.

## Part 2

## The Navier-Stokes Equations

## CHAPTER 6

## Modeling Navier-Stokes

From now on, we will require $M$ to satisfy the Rellich-Kondrachov condition and .

Definition 6.1 (Rellich-Kondrachov condition). Let $(\tilde{M}, \tilde{g})$ be a finite dimensional Riemannian $C^{1,1}$-manifold. Then, we say $(\tilde{M}, \tilde{g})$ satisfies the RellichKondrachov condition if and only if $\forall q \in[1, \operatorname{dim} \tilde{M}) \forall \alpha \in \mathbb{N}_{0} \forall p \in\left[1, \frac{q \operatorname{dim} \tilde{M}}{\operatorname{dim} \tilde{M}-q}\right)$ :

$$
W_{q}^{1,(\alpha, 0)}(\tilde{M}) \hookrightarrow_{\text {compact }} L_{p}^{(\alpha, 0)}(\tilde{M})
$$

We will start modeling the Navier-Stokes equations on $[0, \tau] \times M$ with the mass flow $\varrho u$ where $\varrho$ is the density of the fluid and $u$ the velocity field. Until we identify the Hilbert spaces, we will assume that all functions are sufficiently smooth. For $V \subseteq M$ open with smooth boundary, we obtain by the Gauss divergence theorem

$$
\forall t \in[0, \tau]: \int_{V} \operatorname{tr} \nabla(\varrho u(t)) d \mathrm{vol}_{M}=\int_{\partial V}\langle\varrho u(t), \nu\rangle d \mathrm{vol}_{\partial V} .
$$

But, since the right-hand side is nothing else than the mass transported out of $V$, we observe

$$
\forall t \in[0, \tau]: \int_{V} \operatorname{tr} \nabla(\varrho(t) u(t)) d \operatorname{vol}_{M}=\int_{\partial V}\langle\varrho u(t), \nu\rangle d \operatorname{vol}_{\partial V}=-\partial_{t} \int_{V} \varrho(t) d \operatorname{vol}_{M}
$$

and, since this holds for ever $V \subseteq M$ with smooth boundary,

$$
\forall t \in[0, \tau]: \operatorname{tr} \nabla(\varrho(t) u(t))=-\partial_{t} \varrho(t)
$$

We want to consider fluids only, that is, an incompressible medium, i.e., $\partial_{t} \varrho=0$. Hence, this last equation yields the continuity equation (continuity)

$$
\operatorname{tr} \nabla u=0 .
$$

Next, we will have a look at the stress term. Let $\nabla^{\text {sym }}:=\operatorname{sym} \nabla$ be the symmetrized ${ }^{1}$ co-variant derivative on $(1,0)$-tensors and $\eta, \zeta$ positive, bounded, and bounded from below (these are the two scalar dynamic viscosities in hydrodynamics). A fluid is called isotropic if and only if the viscous stress tensor

$$
\sigma:=2 \eta \nabla^{\mathrm{sym}} u+\zeta \operatorname{tr}^{*} \operatorname{tr} \nabla u
$$

satisfies

$$
\operatorname{tr} \sigma=0
$$

In this case, we call $\eta$ the shear viscosity and $\zeta+\frac{2 \eta}{\operatorname{dim} M}$ the bulk viscosity. Using the continuity equation, we obtain

$$
\sigma=2 \eta \nabla^{\mathrm{sym}} u
$$

which we generalize to the non-Newtonian case

$$
\sigma=C \nabla^{\mathrm{sym}} u
$$

where $C$ is a viscosity operator.

[^5]Definition 6.2 (Viscosity Operator). Let $C \in L\left(L_{2}\left([0, \tau] ; L_{2}^{(2,0)}(M)\right)\right)$ be a positive operator that furthermore satisfies
(i) $C$ is an isomorphism on the symmetric tensor fields, i.e.,

$$
0 \in \varrho\left(\left.C\right|_{L_{2}\left([0, \tau] ; \operatorname{sym}\left[L_{2}^{(2,0)}(M)\right]\right)} ^{L_{2}\left([0, \tau] ; \operatorname{sym}\left[L_{2}^{(2,0)}(M)\right]\right)}\right)
$$

To simplify notation, let $C^{-1}$ denote $\left(\left.C\right|_{L_{2}\left([0, \tau] ; \operatorname{sym}\left[L_{2}^{(2,0)}(M)\right]\right)} ^{L_{2}\left([0, \tau] \operatorname{sym}\left[L^{(2,0)}(M)\right]\right.}\right)^{-1}$.
(ii) $C$ vanishes on anti-symmetric tensor fields, i.e.,

$$
N(C)=[\{0\}] C=L_{2}\left([0, \tau] ; \operatorname{asym}\left[L_{2}^{(2,0)}(M)\right]\right) .
$$

(iii) $C$ and $C^{-1}$ preserve differentiability classes, i.e.,

$$
x \in W_{2}^{k}\left([0, \tau] ; \operatorname{sym}\left[W_{2}^{k^{\prime},(2,0)}(M)\right]\right)
$$

implies

$$
C x, C^{-1} x \in W_{2}^{k}\left([0, \tau] ; \operatorname{sym}\left[W_{2}^{k^{\prime},(2,0)}(M)\right]\right)
$$

(iv) $C$ is a (timely) causal operator, i.e.,

$$
\forall x \in L_{2}\left([0, \tau] ; \operatorname{sym}\left[L_{2}^{(2,0)}(M)\right]\right): \inf \operatorname{spt}_{0} x \leq \inf \operatorname{spt}_{0} C x
$$

where $\mathrm{spt}_{0}$ denotes the support in $L_{2}([0, \tau]) .{ }^{2}$ In other words, if $x$ is zero on some interval $\left[0, \tau^{\prime}\right]$ then so is $C x$; viz., the viscosity of the fluid does not depend on the future.
(v) $\operatorname{tr} C^{-1} \operatorname{tr}^{*}$ is boundedly invertible.

Remark The "classical" Navier-Stokes problem, cf. [5],

$$
\partial_{t} u+\langle u, \nabla\rangle u=\nu \Delta u-\nabla p+f, \operatorname{div} u=0, u(0)=u_{0}
$$

can be retrieved choosing the viscosity operator $C=2 \nu$ sym.

In order to obtain the entire stress tensor, we will have to take the pressure $p$ into account. The stress tensor $T$ is, then, defined as

$$
T:=\sigma-\operatorname{tr}^{*} p
$$

ObSERVATION 6.3. Let $t \in L_{2}^{(2,0)}(M)$ be anti-symmetric. Then, $\operatorname{tr} t=0$.
Proof. Let $t \in L_{2}^{(2,0)}(M)$ be anti-symmetric, i.e., $t=\frac{t_{i j}-t_{j i}}{2} g^{i} \otimes g^{j}$. Then,

$$
\begin{aligned}
\operatorname{tr} t & =\operatorname{tr}\left(\frac{1}{2}\left(t_{i j} g^{i} \otimes g^{j}-t_{j i} g^{i} \otimes g^{j}\right)\right) \\
& =\frac{1}{2}\left(t_{i j} g^{i j}-t_{j i} g^{i j}\right) \\
& =\frac{1}{2}\left(t_{i j} g^{i j}-t_{i j} g^{j i}\right) \\
& =\frac{1}{2}\left(t_{i j} g^{i j}-t_{i j} g^{i j}\right) \\
& =0
\end{aligned}
$$

holds.

[^6]Thus, we also obtain the stress equation
(continuity)

$$
\begin{array}{r}
\operatorname{tr} \nabla^{\text {sym }} u=0 \\
C^{-1} T-\nabla^{\mathrm{sym}} u+C^{-1} \operatorname{tr}^{*} p=0
\end{array}
$$

Finally, we add the initial condition $u(0)=u_{0}$ and Cauchy's momentum equation (which is Newton's law of motion written down for fluids)

$$
\varrho\left(\partial_{t} u+\langle u, \nabla\rangle u\right)=\operatorname{tr} \nabla T+f
$$

using $\langle u, \nabla\rangle u:=u_{i} g^{i j} \nabla_{g_{j}} u=\nabla_{u} u$ where $f$ is an external force. Without loss of generality, we may assume $\varrho=1$ since we can replace $C$, $p$, and $f$ by $\frac{1}{\varrho} C, \frac{1}{\varrho} p$, and $\frac{1}{\varrho} f$, respectively.

Now, the classical Navier-Stokes system is
(continuity)
(stress)
(Cauchy)

$$
\begin{aligned}
\operatorname{tr} \nabla^{\mathrm{sym}} u & =0 \text { in }(0, \tau) \times M \\
C^{-1} T-\nabla^{\mathrm{sym}} u+C^{-1} \operatorname{tr}^{*} p & =0 \text { in }(0, \tau) \times M \\
\partial_{t} u+\langle u, \nabla\rangle u-\operatorname{tr} \nabla T & =f \text { in }(0, \tau) \times M, \\
u(0) & =u_{0} \text { in } M
\end{aligned}
$$

(initial condition)
where $u$ is the velocity field, $T$ the stress tensor (symmetric), $p$ the pressure, $C$ a viscosity operator, and $f$ an external force. The objective is to find reasonable conditions for all these symbols to be physically senseful and interpretable in an $L_{2}$-sense.

First, let us observe for $\varphi \in \mathfrak{M}_{1}^{(0,0)}(M)$

$$
\begin{aligned}
\operatorname{tr} \nabla \operatorname{tr}^{*} \varphi & =\operatorname{tr} \nabla\left(\varphi g_{j k} g^{j} \otimes g^{k}\right) \\
& =\operatorname{tr}\left(\nabla_{g_{i}} \varphi g_{j k} g^{i} \otimes g^{j} \otimes g^{k}\right) \\
& =\nabla_{g_{i}} \varphi g_{j k} g^{i j} g^{k} \\
& =\nabla_{g_{k}} \varphi g^{k} \\
& =\nabla \varphi
\end{aligned}
$$

and, therefrom, for $u \in \mathfrak{M}_{1}^{(1,0)}(M)$

$$
\begin{aligned}
\langle u, \nabla\rangle u & =u_{i} g^{i j} \nabla_{g_{j}} u_{k} g^{k} \\
& =u_{i} g^{i j}\left(\nabla_{g_{j}} u_{k} g^{k}+\nabla_{g_{k}} u_{j} g^{k}\right)-u_{i} g^{i j} \nabla_{g_{k}} u_{j} g^{k} \\
& =2 \operatorname{tr}(u \otimes \operatorname{sym} \nabla u)-\frac{1}{2}\left(u_{i} g^{i j} \nabla_{g_{k}} u_{j} g^{k}+\nabla_{g_{k}} u_{i} g^{i j} u_{j} g^{k}\right) \\
& =2 \operatorname{tr}(u \otimes \operatorname{sym} \nabla u)-\frac{1}{2} \nabla\langle u, u\rangle_{(1,0)} \\
& =2 \operatorname{tr}(u \otimes \operatorname{sym} \nabla u)-\frac{1}{2} \operatorname{tr} \nabla \operatorname{tr}^{*}\langle u, u\rangle_{(1,0)}
\end{aligned}
$$

Defining

$$
\tilde{p}:=p-\frac{1}{2}\langle u, u\rangle_{(1,0)}
$$

and

$$
\tilde{T}:=T+\operatorname{tr}^{*} \frac{1}{2}\langle u, u\rangle_{(1,0)}=C \nabla^{\mathrm{sym}} u-\operatorname{tr}^{*} p+\operatorname{tr}^{*} \frac{1}{2}\langle u, u\rangle_{(1,0)}=C \nabla^{\mathrm{sym}} u-\operatorname{tr}^{*} \tilde{p}
$$

we observe

$$
\begin{aligned}
\langle u, \nabla\rangle u-\operatorname{tr} \nabla T & =2 \operatorname{tr}\left(u \otimes \nabla^{\mathrm{sym}} u\right)-\frac{1}{2} \operatorname{tr} \nabla \operatorname{tr}^{*}\langle u, u\rangle_{(1,0)}-\operatorname{tr} \nabla \tilde{T}-\operatorname{tr} \nabla \operatorname{tr}^{*} \frac{1}{2}\langle u, u\rangle_{(1,0)} \\
& =2 \operatorname{tr}\left(u \otimes \nabla^{\mathrm{sym}} u\right)-\operatorname{tr} \nabla \tilde{T}
\end{aligned}
$$

$$
\begin{aligned}
& =2 \operatorname{tr}\left(u \otimes\left(C^{-1} T+C^{-1} \operatorname{tr}^{*} p\right)\right)-\operatorname{tr} \nabla \tilde{T} \\
& =2 \operatorname{tr}\left(u \otimes\left(C^{-1} \tilde{T}+C^{-1} \operatorname{tr}^{*} \tilde{p}\right)\right)-\operatorname{tr} \nabla \tilde{T} .
\end{aligned}
$$

Hence, Cauchy's momentum equation reduces to

$$
\partial_{t} u-\operatorname{tr} \nabla \tilde{T}=f-2 \operatorname{tr}\left(u \otimes\left(C^{-1} \tilde{T}+C^{-1} \operatorname{tr}^{*} \tilde{p}\right)\right)
$$

and the model becomes

$$
\begin{aligned}
\operatorname{tr} \nabla^{\mathrm{sym}} u & =0 \text { in }(0, \tau) \times M, \\
C^{-1} T-\nabla^{\mathrm{sym}} u+C^{-1} \operatorname{tr}^{*} p & =0 \text { in }(0, \tau) \times M, \\
\partial_{t} u-\operatorname{tr} \nabla \tilde{T} & =f-2 \operatorname{tr}\left(u \otimes\left(C^{-1} \tilde{T}+C^{-1} \operatorname{tr}^{*} \tilde{p}\right)\right) \text { in }(0, \tau) \times M, \\
u(0) & =u_{0} \text { in } M
\end{aligned}
$$

which, omitting the initial condition for the moment, is equivalent to

$$
\left(\begin{array}{ccc}
0 & \operatorname{trsym}^{\mathrm{sym}} \nabla & 0 \\
0 & \partial_{t} & -\operatorname{tr} \nabla \\
C^{-1} \operatorname{tr}^{*} & -\nabla^{\mathrm{sym}} & C^{-1}
\end{array}\right)\left(\begin{array}{c}
\tilde{p} \\
u \\
\tilde{T}
\end{array}\right)=\left(\begin{array}{c}
0 \\
f-2 \operatorname{tr}\left(u \otimes C^{-1}\left(\tilde{T}+\operatorname{tr}^{*} \tilde{p}\right)\right) \\
0
\end{array}\right)
$$

and, hence, equivalent to

$$
\left(\begin{array}{ccc}
\operatorname{tr} C^{-1} \operatorname{tr}^{*} & 0 & \operatorname{tr} C^{-1} \\
0 & \partial_{t} & -\operatorname{tr} \nabla \\
C^{-1} \operatorname{tr}^{*} & -\nabla^{\operatorname{sym}} & C^{-1}
\end{array}\right)\left(\begin{array}{c}
\tilde{p} \\
u \\
\tilde{T}
\end{array}\right)=\left(\begin{array}{c}
0 \\
f-2 \operatorname{tr}\left(u \otimes C^{-1}\left(\tilde{T}+\operatorname{tr}^{*} \tilde{p}\right)\right) \\
0
\end{array}\right)
$$

as well. Using the (non-unitary) transformation

$$
U:=\left(\begin{array}{ccc}
1 & 0 & 0 \\
0 & 1 & 0 \\
-C^{-1} \operatorname{tr}^{*}\left(\operatorname{tr} C^{-1} \operatorname{tr}^{*}\right)^{-1} & 0 & 1
\end{array}\right)
$$

with

$$
\begin{aligned}
& U^{*}=\left(\begin{array}{ccc}
1 & 0 & -\left(\operatorname{tr} C^{-1} \operatorname{tr}^{*}\right)^{-1} \operatorname{tr} C^{-1} \\
0 & 1 & 0 \\
0 & 0 & 1
\end{array}\right), \\
& \left(U^{*}\right)^{-1}=\left(\begin{array}{ccc}
1 & 0 & \left(\operatorname{tr} C^{-1} \operatorname{tr}^{*}\right)^{-1} \operatorname{tr} C^{-1} \\
0 & 1 & 0 \\
0 & 0 & 1
\end{array}\right), \\
& U\left(\begin{array}{ccc}
\operatorname{tr} C^{-1} \operatorname{tr}^{*} & 0 & \operatorname{tr} C^{-1} \\
0 & \partial_{t} & -\operatorname{tr} \nabla \\
C^{-1} \operatorname{tr}^{*} & -\nabla^{\text {sym }} & C^{-1}
\end{array}\right) U^{*} \\
& =\left(\begin{array}{ccc}
\operatorname{tr} C^{-1} \operatorname{tr}^{*} & 0 & 0 \\
0 & \partial_{t} & -\operatorname{tr} \nabla \\
0 & -\nabla^{\text {sym }} & C^{-1}-C^{-1} \operatorname{tr}^{*}\left(\operatorname{tr} C^{-1} \operatorname{tr}^{*}\right)^{-1} \operatorname{tr} C^{-1}
\end{array}\right),
\end{aligned}
$$

and

$$
\left(U^{*}\right)^{-1}\left(\begin{array}{c}
\tilde{p} \\
u \\
\tilde{T}
\end{array}\right)=\left(\begin{array}{c}
\tilde{p}+\left(\operatorname{tr} C^{-1} \operatorname{tr}^{*}\right)^{-1} \operatorname{tr} C^{-1} \tilde{T} \\
u \\
\tilde{T}
\end{array}\right)
$$

yields

$$
U\left(\begin{array}{ccc}
\operatorname{tr} C^{-1} \operatorname{tr}^{*} & 0 & \operatorname{tr} C^{-1} \\
0 & \partial_{t} & -\operatorname{tr} \nabla \\
C^{-1} \operatorname{tr}^{*} & -\nabla^{\text {sym }} & C^{-1}
\end{array}\right) U^{*}\left(U^{*}\right)^{-1}\left(\begin{array}{c}
\tilde{p} \\
u \\
\tilde{T}
\end{array}\right)=U\left(f-2 \operatorname{tr}\left(u \otimes C^{-1}\left(\tilde{T}+\operatorname{tr}^{*} \tilde{p}\right)\right)\right)
$$

$$
=\left(\begin{array}{c}
0 \\
f-2 \operatorname{tr}\left(u \otimes C^{-1}\left(\tilde{T}+\operatorname{tr}^{*} \tilde{p}\right)\right) \\
0
\end{array}\right)
$$

allowing us to further reduce the system, since

$$
\tilde{p}=-\left(\operatorname{tr} C^{-1} \operatorname{tr}^{*}\right)^{-1} \operatorname{tr} C^{-1} \tilde{T}
$$

decouples, yielding

$$
\begin{aligned}
& \left(\begin{array}{cc}
\partial_{t} & -\operatorname{tr} \nabla \\
-\nabla^{\text {sym }} & C^{-1}-C^{-1} \operatorname{tr}^{*}\left(\operatorname{tr} C^{-1} \operatorname{tr}^{*}\right)^{-1} \operatorname{tr} C^{-1}
\end{array}\right)\binom{u}{\tilde{T}} \\
= & \binom{f-2 \operatorname{tr}\left(u \otimes\left(1-C^{-1} \operatorname{tr}^{*}\left(\operatorname{tr} C^{-1} \operatorname{tr}^{*}\right)^{-1} \operatorname{tr}\right) C^{-1} \tilde{T}\right)}{0}
\end{aligned}
$$

Let

$$
\mathbb{E}:=1-C^{-\frac{1}{2}} \operatorname{tr}^{*}\left(\operatorname{tr} C^{-1} \operatorname{tr}^{*}\right)^{-1} \operatorname{tr} C^{-\frac{1}{2}}
$$

and

$$
\Theta:=C^{-\frac{1}{2}} \tilde{T}
$$

Then, we observe

$$
(1-\mathbb{E})^{2}=C^{-\frac{1}{2}} \operatorname{tr}^{*}\left(\operatorname{tr} C^{-1} \operatorname{tr}^{*}\right)^{-1} \operatorname{tr} C^{-1} \operatorname{tr}^{*}\left(\operatorname{tr} C^{-1} \operatorname{tr}^{*}\right)^{-1} \operatorname{tr} C^{-\frac{1}{2}}=1-\mathbb{E}
$$

and

$$
\begin{aligned}
\mathbb{E}^{2} & =1-2 C^{-\frac{1}{2}} \operatorname{tr}^{*}\left(\operatorname{tr} C^{-1} \operatorname{tr}^{*}\right)^{-1} \operatorname{tr} C^{-\frac{1}{2}}+\left(C^{-\frac{1}{2}} \operatorname{tr}^{*}\left(\operatorname{tr} C^{-1} \operatorname{tr}^{*}\right)^{-1} \operatorname{tr} C^{-\frac{1}{2}}\right)^{2} \\
& =1-2(1-\mathbb{E})+(1-\mathbb{E})^{2} \\
& =\mathbb{E}
\end{aligned}
$$

Hence, $\mathbb{E}$ is an orthogonal projection onto $[\{0\}] \operatorname{tr} C^{-\frac{1}{2}}$ and self-adjoint in $L_{2}^{(2,0)}(M)$. Furthermore, the Navier-Stokes system becomes

$$
\left(\begin{array}{cc}
\partial_{t} & -\operatorname{tr} \nabla \\
-\nabla^{\operatorname{sym}} & C^{-\frac{1}{2}} \mathbb{E} C^{-\frac{1}{2}}
\end{array}\right)\binom{u}{\tilde{T}}=\binom{f-2 \operatorname{tr}\left(u \otimes C^{-\frac{1}{2}} \mathbb{E} C^{-\frac{1}{2}} \tilde{T}\right)}{0}
$$

i.e.,

$$
\left(\begin{array}{cc}
\partial_{t} & -\operatorname{tr} \nabla C^{\frac{1}{2}} \\
-C^{\frac{1}{2}} \nabla^{\mathrm{sym}} & \mathbb{E}
\end{array}\right)\binom{u}{\Theta}=\binom{f-2 \operatorname{tr}\left(u \otimes C^{-\frac{1}{2}} \mathbb{E} \Theta\right)}{0}
$$

Since $C^{\frac{1}{2}}$ vanishes on anti-symmetric tensors, this last equation is equivalent to

$$
\left(\begin{array}{cc}
\partial_{t} & -\operatorname{tr} \nabla C^{\frac{1}{2}} \\
-C^{\frac{1}{2}} \nabla & \mathbb{E}
\end{array}\right)\binom{u}{\Theta}=\binom{f-2 \operatorname{tr}\left(u \otimes C^{-\frac{1}{2}} \mathbb{E} \Theta\right)}{0}
$$

However, this equation can be interpreted in an $L_{2}^{(\alpha, \beta)}(M)$ setting using the (partial) time derivative $\partial_{0}$ in $L_{2}([0, \tau])$ which yields

$$
\left(\begin{array}{cc}
\partial_{0} & -\operatorname{div}_{(0)} C^{\frac{1}{2}} \\
-C^{\frac{1}{2}} \operatorname{grad}_{(0)} & \mathbb{E}
\end{array}\right)\binom{u}{\Theta}=\binom{f-2 \operatorname{tr}\left(u \otimes C^{-\frac{1}{2}} \mathbb{E} \Theta\right)}{0}
$$

where

$$
\operatorname{grad}_{(0)}= \begin{cases}\operatorname{grad}_{0} & , \text { Dirichlet case } \\ \operatorname{grad} & , \text { Neumann or no boundary case }\end{cases}
$$

and

$$
\operatorname{div}_{(0)}= \begin{cases}\operatorname{div}_{0} & , \text { Neumann case } \\ \operatorname{div} & , \text { Dirichlet or no boundary case }\end{cases}
$$

Recall that for all $t \in[0, \tau]$

$$
u(t) \in N(\operatorname{tr} \nabla)=[\{0\}] \operatorname{tr} \nabla
$$

shall hold and, hence, $C^{\frac{1}{2}} \nabla u$ takes values in $[\{0\}] \operatorname{tr} C^{-\frac{1}{2}}$ which is nothing else than the range of $\mathbb{E}$. Let

$$
Y:[\{0\}] \operatorname{tr} \operatorname{grad} \subseteq \overline{[\{0\}] \operatorname{tr} \operatorname{grad}} \rightarrow[\{0\}] \operatorname{tr} C^{-\frac{1}{2}} ; x \mapsto-C^{\frac{1}{2}} \operatorname{grad} x
$$

in case of no boundary or Neumann boundary conditions ${ }^{3}$. In case of Dirichlet boundary conditions let

$$
Y:[\{0\}] \operatorname{tr} \operatorname{grad}_{0} \subseteq \overline{[\{0\}] \operatorname{tr} \operatorname{grad}_{0}} \rightarrow[\{0\}] \operatorname{tr} C^{-\frac{1}{2}} ; x \mapsto-C^{\frac{1}{2}} \operatorname{grad}_{0} x
$$

Then

$$
Y=-\operatorname{pr}_{[\{0\}] \operatorname{tr} C^{-\frac{1}{2}}} C^{\frac{1}{2}} \operatorname{grad}_{(0)} \operatorname{pr}_{[\{0\}] \operatorname{tr} \operatorname{grad}(0)}^{*}
$$

embedded in $L_{2}^{(1,0)}(M) \oplus L_{2}^{(2,0)}(M)$ and

$$
Y^{*}=\operatorname{pr}_{[\{0\}] \operatorname{tr} \operatorname{grad}_{(0)}} \operatorname{div}_{(0)} C^{\frac{1}{2}} \operatorname{pr}_{[\{0\}] \operatorname{tr} C^{-\frac{1}{2}}}^{*}
$$

embedded in $L_{2}^{(2,0)}(M) \oplus L_{2}^{(1,0)}(M)$.
Remark Let $\Gamma \subseteq \partial M$ be Borel measurable,

$$
D(a):=\left\{u \in W_{2}^{1}(M) ;\left.u\right|_{\Gamma}=0\right\}
$$

and

$$
a: D(a) \times D(a) \rightarrow \mathbb{R} ; \quad(u, v) \mapsto\langle C \operatorname{grad} u, \operatorname{grad} v\rangle_{L_{2}^{(1,0)}(M)}
$$

Then, $a$ generates a positive operator $Y^{*} Y$ which can be considered as a realization of the mixed boundary condition "Dirichlet on $\Gamma$ and Neumann on $\partial M \backslash \Gamma$ ". Similarly, other boundary conditions can be introduced and the following would be virtually the same up to a few subtle changes which we will not address any further.

Hence, the system reduces to

$$
\left(\begin{array}{cc}
\partial_{0} & -Y^{*} \\
Y & \mathbb{E}
\end{array}\right)\binom{u}{\Theta}=\binom{f-2 \operatorname{tr}\left(u \otimes C^{-\frac{1}{2}} \mathbb{E} \Theta\right)}{0}
$$

which yields

$$
\partial_{0} u+Y^{\star} Y u-Y^{*}(1-\mathbb{E}) \Theta=f+2 \operatorname{tr}\left(u \otimes C^{-\frac{1}{2}} Y u\right)
$$

From $Y^{*}=\operatorname{pr} \overline{[\{0\}] \operatorname{tr} \operatorname{grad}_{(0)}} \operatorname{div}_{(0)} C^{\frac{1}{2}} \operatorname{pr}_{[\{0\}] \operatorname{tr} C^{-\frac{1}{2}}}^{*}$ and, since $\left([\{0\}] \operatorname{tr} C^{-\frac{1}{2}}\right)^{\perp}$ is the range of $(1-\mathbb{E})$ (recall that $\mathbb{E}$ is an orthogonal projection), we directly deduce $Y^{*}(1-\mathbb{E})=0$ and, therefore,

$$
\partial_{0} u+Y^{*} Y u=f+2 \operatorname{tr}\left(u \otimes C^{-\frac{1}{2}} Y u\right)
$$

[^7]Since the left-hand side takes values in $\overline{[\{0\}] \operatorname{tr} \operatorname{grad}_{(0)}}$, so the right-hand side has to. Thus, we may state the system as

$$
\partial_{0} u+Y^{*} Y u-2 \operatorname{pr} \frac{\sum_{[\{0\}] \operatorname{tr} \operatorname{grad}_{(0)}}}{} \operatorname{tr}\left(u \otimes C^{-\frac{1}{2}} Y u\right)=f
$$

where $f$ takes values in $H:=\operatorname{pr}_{[\{0\}] \operatorname{trgrad}(0)}\left[L_{2}^{(1,0)}(M)\right]$.
Remark Note that the operator $Y^{*} Y$ is densely defined due to a theorem by von Neumann.

Theorem 6.4 (von Neumann). Let $H_{1}$ and $H_{2}$ be Hilbert spaces and $A \subseteq$ $H_{1} \oplus H_{2}$ a closed and densely defined operator. Then $A^{*} A$ is self-adjoint in $H_{1}$ and its domain is a core of $A$.

This theorem can be obtained from the first representation theorem (Theorem VI.2.1 in [11]) applied to the closed, positive, symmetric form $\tau$ with $D(\tau):=D(A)$ and $\forall x, y \in D(\tau): \tau(x, y):=\langle A x, A y\rangle_{H_{2}}$.

We define the space of maximal regularity

$$
\mathfrak{M R}_{\tau}:=W_{2}^{1}([0, \tau] ; H) \cap L_{2}\left([0, \tau] ; D\left(Y^{\star} Y\right)\right)
$$

endowed with the norm

$$
\|\cdot\|_{\mathfrak{M R}_{\tau}}: \mathfrak{M R}_{\tau} \rightarrow \mathbb{R} ; x \mapsto\left(\|x\|_{W_{2}^{1}([0, \tau] ; H)}^{2}+\|x\|_{L_{2}\left([0, \tau] ; D\left(Y^{*} Y\right)\right)}^{2}\right)^{\frac{1}{2}}
$$

Then, the Sobolev Embedding Theorem (Theorem 3.15) yields

$$
\mathfrak{M R}_{\tau} \hookrightarrow_{\text {continuous }} C([0, \tau] ; H)
$$

The embedding $\mathfrak{M} \mathfrak{R}_{\tau} \rightarrow_{\text {continuous }} C([0, \tau] ; H)$ is, in fact, compact as can be shown using Arzelà-Ascoli's theorem. We, on the other hand, only need continuity since this ensures that

$$
\mathfrak{M R}_{\tau, 0}:=\left\{x \in \mathfrak{M} \mathfrak{R}_{\tau} ; x(0)=0\right\}
$$

and

$$
\mathfrak{T} \mathfrak{R}:=\mathfrak{M} \mathfrak{R}_{\tau} / \mathfrak{M R}_{\tau, 0} \cong\left\{x(0) \in H ; x \in \mathfrak{M} \mathfrak{R}_{\tau}\right\}
$$

are well-defined Hilbert spaces.
As of now, we have identified the abstract Cauchy problem we would like to consider and the spaces the equation should hold in. The only thing in question is whether the non-linearity behaves nicely. This is where we need the RellichKondrachov condition. The Rellich-Kondrachov condition implies

$$
\forall \alpha \in \mathbb{N}_{0}: W_{2}^{1,(\alpha, 0)}(M) \subseteq L_{4}^{(\alpha, 0)}(M)
$$

which combined with $Y\left[D\left(Y^{*} Y\right)\right] \subseteq W_{2}^{1,(2,0)}(M)$ yields

$$
\forall u \in \mathfrak{M R}_{\tau}: \operatorname{tr}\left(u \otimes C^{-\frac{1}{2}} Y u\right) \in L_{2}\left([0, \tau] ; L_{2}^{(1,0)}(M)\right)
$$

Now, we may actually state the Navier-Stokes problem we want to address.
Problem 6.5 (Navier-Stokes). Let $\tau \in \mathbb{R}_{>0}, f \in L_{2}([0, \tau] ; H)$ and $u_{0} \in \mathfrak{T} \mathfrak{R}$. Find $u \in \mathfrak{M R}_{\tau}$ such that
(Navier-Stokes)

$$
\begin{aligned}
\partial_{0} u+Y^{\star} Y u-2 \operatorname{pr} & \frac{[\{0\}] \operatorname{tr} \operatorname{grad}_{(0)}}{} \operatorname{tr}\left(u \otimes C^{-\frac{1}{2}} Y u\right)
\end{aligned}=f \quad \text { in }(0, \tau) \times M, ~ \begin{aligned}
u(0) & =u_{0} \text { in } M
\end{aligned}
$$

holds.

## CHAPTER 7

## Construction of Solutions and Analytic Dependence

In order to solve the Navier-Stokes problem, let us define

$$
B: \mathfrak{M R}_{\tau} \rightarrow L_{2}([0, \tau] ; H) ; x \mapsto-2 \operatorname{pr}_{[\{0\}] \operatorname{tr} \operatorname{grad}_{(0)}} \operatorname{tr}\left(x \otimes C^{-\frac{1}{2}} Y x\right)
$$

and

$$
F_{\tau}: \mathfrak{M R}_{\tau} \rightarrow L_{2}([0, \tau] ; H) \times \mathfrak{T} \mathfrak{R} ; x \mapsto\left(\partial_{0} x+Y^{*} Y x+B(x), x(0)\right) .
$$

These yield the nice and short notation

$$
F_{\tau}(u)=\left(f, u_{0}\right)
$$

for the Navier-Stokes equations. $F_{\tau}$ will, thus, be called the Navier-Stokes operator and our objective is to continuously invert $F_{\tau}$ locally in time and show that the inverse is an analytic operator on the reduced time interval.

Note that $F_{\tau}$ is an analytic operator since it is a polynomial of degree two and for $u, v \in \mathfrak{M R}_{\tau}$ we observe

$$
F_{\tau}^{\prime}(v) u=\left(d_{0} u+Y^{*} Y u+B^{\prime}(v) u, u(0)\right) .
$$

Hence, if we can find a $v \in \mathfrak{M R}_{\tau}$ such that $F_{\tau}^{\prime}(v)$ is an isomorphism and our data $\left(f, u_{0}\right)$ are sufficiently close to $F_{\tau}(v)$, then the analytic inverse function theorem yields existence and constructibility of solutions and analytic dependence on the data. We are going to achieve this by defining

$$
g_{\tau^{\prime}, v}:=f 1_{\left[0, \tau^{\prime}\right]}+\left(\partial_{0} v+Y^{*} Y v+B(v)\right) 1_{\left(\tau^{\prime}, \tau\right]}
$$

for $\tau^{\prime} \in(0, \tau)$ and $v \in \mathfrak{M R}_{\tau}$. Then we observe for $v \in \mathfrak{M R}_{\tau}$ with $v(0)=u_{0}$

$$
\begin{aligned}
& \left\|\left(g_{\tau^{\prime}, v}, u_{0}\right)-F_{\tau}(v)\right\|_{L_{2}([0, \tau] ; H) \times \mathfrak{T} \mathfrak{R}}^{2} \\
= & \int_{0}^{\tau^{\prime}}\left\|f(s)-\partial_{0} v(s)-Y^{*} Y v(s)-B(v)(s)\right\|_{H}^{2} d s \searrow 0
\end{aligned}
$$

as $\tau^{\prime} \searrow 0$, i.e., for $\tau^{\prime}$ sufficiently small, we can solve a slightly Navier-Stokes system and $\tau^{\prime}$ is even locally constant. In order for us to know that this solution with respect to $\left(g_{\tau^{\prime}, v}, u_{0}\right)$ also solves the Navier-Stokes system with respect to ( $f, u_{0}$ ) on $\left[0, \tau^{\prime}\right]$, we need to make sure that the solution on $\left[0, \tau^{\prime}\right]$ does not depend on the data on $\left(\tau^{\prime}, \tau\right]$. Hence, our to-do-list is:

- Find $v \in \mathfrak{M R}_{\tau}$ with $v(0)=u_{0}$ and $F_{\tau}^{\prime}(v)$ being an isomorphism; in fact, we are going to show that $F_{\tau}^{\prime}(v)$ is always an isomorphism and the existence of a $v$ with $v(0)=u_{0}$ is trivial by definition of $\mathfrak{T} \mathfrak{R}$.
- Show injectivity of $F_{\tau}$.
- Show causality of solutions.


## CHAPTER 8

## Linearized Navier-Stokes

This chapter is devoted to showing that $F_{\tau}^{\prime}(v)$ is an isomorphism but, lucky us, will also yield injectivity of $F_{\tau}$ as a corollary. In order to show that $F_{\tau}^{\prime}(v)$ is an isomorphism, we will prove that the Stokes operator ${ }^{1}$

$$
I: \mathfrak{M}_{\tau} \rightarrow L_{2}([0, \tau] ; H) \times \mathfrak{T} \mathfrak{R} ; x \mapsto\left(\partial_{0} x+Y^{*} Y x, x(0)\right)
$$

is an isomorphism, first, and then we will consider the perturbation

$$
\mathcal{B}_{v}: \mathfrak{M R}_{\tau} \rightarrow L_{2}([0, \tau] ; H) \times \mathfrak{T} \mathfrak{R} ; x \mapsto\left(B^{\prime}(v) x, 0\right)
$$

Lemma 8.1. Let $\lambda \in \mathbb{R}_{>0}$. Then

$$
\left(\partial_{0}+Y^{*} Y+\lambda\right): \mathfrak{M}_{\tau, 0} \rightarrow L_{2}([0, \tau] ; H)
$$

is an isomorphism.
Furthermore, if $\lambda \geq \lambda_{0} \in \mathbb{R}_{>0}$ is uniformly bounded away from zero then

$$
\left\|\left(\partial_{0}+Y^{*} Y+\lambda\right)^{-1}\right\|_{L\left(L_{2}([0, \tau] ; H), \mathfrak{M} \Re_{\tau, 0}\right)} \text { and } \|_{\left.\left(\partial_{0}+Y^{*} Y+\lambda\right)^{-1} \|_{L\left(L_{2}([0, \tau] ; H)\right)}\right)}
$$

are uniformly bounded (varying $\lambda$ ).
Proof. Since

$$
Y^{*} Y: D\left(Y^{*} Y\right) \subseteq \operatorname{pr}_{[\{0\}] \operatorname{trgrad}(0)}\left[W_{2}^{1,(1,0)}(M)\right] \rightarrow H
$$

is self-adjoint and non-negative, the spectral theorem warrants the existence of a measure space $(\Omega, \mathcal{A}, \mu)$ and $a: \Omega \rightarrow \mathbb{R}_{\geq 0}$ measurable such that $Y^{*} Y$ is unitarily equivalent to

$$
a(m): D(a(m)) \subseteq L_{2}(\mu) \rightarrow L_{2}(\mu) ; f \mapsto\left(\Omega \ni x \mapsto a(x) f(x) \in \mathbb{R}_{\geq 0}\right)
$$

with

$$
D(a(m)):=\left\{f \in L_{2}(\mu) ;\left(\Omega \ni x \mapsto a(x) f(x) \in \mathbb{R}_{\geq 0}\right) \in L_{2}(\mu)\right\}
$$

Without loss of generality, we may, hence, assume that $Y^{*} Y=a(m)$.
Let $f \in L_{2}\left([0, \tau] ; L_{2}(\mu)\right)$ and $\tilde{f}$ be a representative. For $(t, x) \in[0, \tau] \times \Omega$ with $\tilde{f}(\cdot, x) \in L_{1}([0, \tau])$ define

$$
S_{\lambda} f(t, x):=\int_{0}^{t} e^{(a(x)+\lambda)(s-t)} \tilde{f}(s, x) d s
$$

Note that $S_{\lambda}$ is unitarily equivalent to $\left(\partial_{0}+Y^{*} Y+\lambda\right)^{-1}$. Let $b: \Omega \rightarrow \mathbb{R}_{\geq 0}$ be measurable and $c \in \mathbb{R}_{>0}$ such that for $\mu$-almost every $x \in \Omega$

$$
0 \leq \frac{b(x)}{a(x)+\lambda} \leq c
$$

[^8]holds. Then we observe
\[

$$
\begin{aligned}
& \left\|b(m) S_{\lambda} f\right\|_{L_{2}\left([0, \tau] ; L_{2}(\mu)\right)}^{2} \\
& =\int_{0}^{\tau}\left\|b(m) S_{\lambda} f(t)\right\|_{L_{2}(\mu)}^{2} d t \\
& =\int_{0}^{\tau} \int_{\Omega}\left|b(x) S_{\lambda} f(t, x)\right|^{2} d \mu(x) d t \\
& =\int_{\Omega} \int_{0}^{\tau}|b(x)|^{2}\left|\int_{0}^{t} e^{(a(x)+\lambda)(s-t)} f(s, x) d s\right|^{2} d t d \mu(x) \\
& =\int_{\Omega}|b(x)|^{2}\left\|t \mapsto \int_{0}^{t} e^{-(a(x)+\lambda)(t-s)} f(s, x) d s\right\|_{L_{2}([0, \tau])}^{2} d \mu(x) \\
& =\int_{\Omega}|b(x)|^{2}\left\|t \mapsto \int_{0}^{t} e^{-(a(x)+\lambda)(t-s)} 1_{[0, \tau]}(t-s) f(s, x) 1_{[0, \tau]}(s) d s\right\|_{L_{2}([0, \tau])}^{2} d \mu(x) \\
& =\int_{\Omega}|b(x)|^{2}\left\|t \mapsto \int_{\mathbb{R}} e^{-(a(x)+\lambda)(t-s)} 1_{[0, \tau]}(t-s) f(s, x) 1_{[0, \tau]}(s) d s\right\|_{L_{2}([0, \tau])}^{2} d \mu(x) \\
& =\int_{\Omega}|b(x)|^{2} \|\left(e^{\left.-(a(x)+\lambda) \cdot 1_{[0, \tau]}\right) *\left(f(\cdot, x) 1_{[0, \tau]}\right) \|_{L_{2}([0, \tau])}^{2} d \mu(x)}\right. \\
& \leq \int_{\Omega}|b(x)|^{2}\left\|\left(e^{-(a(x)+\lambda) \cdot} 1_{[0, \tau]}\right) *\left(f(\cdot, x) 1_{[0, \tau]}\right)\right\|_{L_{2}(\mathbb{R})}^{2} d \mu(x) \\
& \leq \text { Young } \int_{\Omega}|b(x)|^{2}\left\|e^{-(a(x)+\lambda) \cdot} 1_{[0, \tau]}\right\|_{L_{1}(\mathbb{R})}^{2}\left\|f(\cdot, x) 1_{[0, \tau]}\right\|_{L_{2}(\mathbb{R})}^{2} d \mu(x) \\
& =\int_{\Omega}|b(x)|^{2}\left(\int_{0}^{\tau} e^{-(a(x)+\lambda) s} d s\right)^{2}\left\|f(\cdot, x) 1_{[0, \tau]}\right\|_{L_{2}(\mathbb{R})}^{2} d \mu(x) \\
& =\int_{\Omega}|b(x)|^{2}\left(\frac{e^{-(a(x)+\lambda) \tau}-1}{-(a(x)+\lambda)}\right)^{2}\left\|f(\cdot, x) 1_{[0, \tau]}\right\|_{L_{2}(\mathbb{R})}^{2} d \mu(x) \\
& \leq \int_{\Omega}\left(\frac{\left(e^{-(a(x)+\lambda) \tau}-1\right) b(x)}{a(x)+\lambda}\right)^{2}\left\|f(\cdot, x) 1_{[0, \tau]}\right\|_{L_{2}(\mathbb{R})}^{2} d \mu(x) \\
& \leq c^{2}\|f\|_{L_{2}\left([0, \tau] ; L_{2}(\mu)\right)}^{2}
\end{aligned}
$$
\]

For $b=1$ we may choose $c=\frac{1}{\lambda}$ and for $b=a$ we may choose $c=1$. Then we obtain

$$
\begin{aligned}
\left\|S_{\lambda} f\right\|_{L_{2}([0, \tau] ; D(a(m)))} & \leq\left\|S_{\lambda} f\right\|_{L_{2}\left([0, \tau] ; L_{2}(\mu)\right)}+\left\|a(m) S_{\lambda} f\right\|_{L_{2}\left([0, \tau] ; L_{2}(\mu)\right)} \\
& \leq\left(\frac{1}{\lambda}+1\right)\|f\|_{L_{2}\left([0, \tau] ; L_{2}(\mu)\right)}
\end{aligned}
$$

Furthermore, for $f$ continuous, the fundamental theorem of calculus for Bochner integrals implies $S_{\lambda} f \in C^{1}\left([0, \tau] ; L_{2}(\mu)\right)$ and, thus, $S_{\lambda} f \in W_{2}^{1}\left([0, \tau] ; L_{2}(\mu)\right)$ for $f \in L_{2}\left([0, \tau] ; L_{2}(\mu)\right)$. Hence, $\partial_{0} S_{\lambda} f=f-(a(m)+\lambda) S_{\lambda} f$ implies

$$
\begin{aligned}
& \left\|S_{\lambda} f\right\|_{W_{2}^{1}\left([0, \tau] ; L_{2}(\mu)\right)} \\
\leq & \left\|S_{\lambda} f\right\|_{L_{2}\left([0, \tau] ; L_{2}(\mu)\right)}+\left\|\partial_{0} S_{\lambda} f\right\|_{L_{2}\left([0, \tau] ; L_{2}(\mu)\right)} \\
\leq & \left\|S_{\lambda} f\right\|_{L_{2}\left([0, \tau] ; L_{2}(\mu)\right)}+\left\|f-(a(m)+\lambda) S_{\lambda} f\right\|_{L_{2}\left([0, \tau] ; L_{2}(\mu)\right)} \\
\leq & \frac{\|f\|_{L_{2}\left([0, \tau] ; L_{2}(\mu)\right)}}{\lambda}+\|f\|_{L_{2}\left([0, \tau] ; L_{2}(\mu)\right)}+\left\|(a(m)+\lambda) S_{\lambda} f\right\|_{L_{2}\left([0, \tau] ; L_{2}(\mu)\right)} \\
\leq & \left(\frac{1}{\lambda}+1+1\right)\|f\|_{L_{2}\left([0, \tau] ; L_{2}(\mu)\right)} .
\end{aligned}
$$

Remark To show $S_{\lambda}=\left(\partial_{0}+Y^{*} Y+\lambda\right)^{-1}$ in the sense of unitary equivalence, we observe for $f \in L_{2}\left([0, \tau], L_{2}(\mu)\right), g \in \mathfrak{M R}_{\tau, 0}$, and $\tilde{f}$ and $\tilde{g}$ representatives

$$
\begin{aligned}
\left(\partial_{0}+Y^{\star} Y+\lambda\right) S_{\lambda} f(t, x)= & \partial_{0} \int_{0}^{t} e^{(a(x)+\lambda)(s-t)} \tilde{f}(s, x) d s \\
& +\int_{0}^{t}(a(x)+\lambda) e^{(a(x)+\lambda)(s-t)} \tilde{f}(s, x) d s \\
= & \int_{0}^{t}-(a(x)+\lambda) e^{(a(x)+\lambda)(s-t)} \tilde{f}(s, x) d s+\tilde{f}(t, x) \\
& +\int_{0}^{t}(a(x)+\lambda) e^{(a(x)+\lambda)(s-t)} \tilde{f}(s, x) d s \\
= & \tilde{f}(t, x)
\end{aligned}
$$

and (using Hille's theorem ${ }^{2}$ )

$$
\begin{aligned}
S_{\lambda}\left(\partial_{0}+Y^{\star} Y+\lambda\right) g(t, x)= & \int_{0}^{t} e^{(a(x)+\lambda)(s-t)}\left(\partial_{0}+Y^{\star} Y+\lambda\right) \tilde{g}(s, x) d s \\
= & \int_{0}^{t} e^{(a(x)+\lambda)(s-t)} \partial_{0} \tilde{g}(s, x) d s \\
& +\int_{0}^{t} e^{(a(x)+\lambda)(s-t)}\left(Y^{\star} Y+\lambda\right) \tilde{g}(s, x) d s \\
= & \int_{0}^{t} \partial_{0}\left(e^{(a(x)+\lambda)(s-t)} \tilde{g}(s, x)\right) d s \\
& -\int_{0}^{t} \partial_{0} e^{(a(x)+\lambda)(s-t)} \tilde{g}(s, x) d s \\
& +\left(Y^{*} Y+\lambda\right) \int_{0}^{t} e^{(a(x)+\lambda)(s-t)} \tilde{g}(s, x) d s \\
= & \int_{0}^{t} \partial_{0}\left(e^{(a(x)+\lambda)(s-t)} \tilde{g}(s, x)\right) d s \\
& -\left(Y^{*} Y+\lambda\right) \int_{0}^{t} e^{(a(x)+\lambda)(s-t)} \tilde{g}(s, x) d s \\
& +\left(Y^{\star} Y+\lambda\right) \int_{0}^{t} e^{(a(x)+\lambda)(s-t)} \tilde{g}(s, x) d s \\
= & \tilde{g}(t, x)-e^{-(a(x)+\lambda) t} \underbrace{\tilde{g}(0, x)}_{=0} \\
= & \tilde{g}(t, x)
\end{aligned}
$$

almost everywhere. Hence, $S_{\lambda}=\left(\partial_{0}+Y^{*} Y+\lambda\right)^{-1}$ in the sense of unitary equivalence.

Proposition 8.3. Let $B_{0} \in L\left(\mathfrak{M R}_{\tau}, L_{2}([0, \tau] ; H)\right)$ with

$$
\forall \lambda \in \mathbb{R}_{>0}: \quad B_{0} e^{\lambda m_{0}}=e^{\lambda m_{0}} B_{0}
$$

( $m_{0}$ is the multiplication operator with the "time" argument, i.e., in $L_{2}([0, \tau])$, with maximal domain) and $\forall \alpha \in(0,1) \exists C_{\alpha} \in \mathbb{R}_{>0} \forall u \in \mathfrak{M R}_{\tau}$ :

$$
\left\|B_{0} u\right\|_{L_{2}([0, \tau] ; H)} \leq C_{\alpha}\|u\|_{L_{2}([0, \tau] ; H)}+\alpha\|u\|_{\mathfrak{M}_{\tau}}
$$

2
Theorem 8.2 (Hille). Let $I \subseteq \mathbb{R}$ be an interval, $X$ and $Y$ Banach spaces, $A \subseteq X \oplus Y a$ closed linear operator, $f: I \rightarrow X$ Bochner-integrable, $\forall t \in I: f(t) \in D(A)$, and $t \mapsto A f(t)$ Bochner-integrable. Then, $\int_{I} f(t) d t \in D(A)$ and $A \int_{I} f(t) d t=\int_{I} A f(t) d t$ holds.

Proof. see [10]

Then

$$
J: \mathfrak{M R}_{\tau} \rightarrow L_{2}([0, \tau] ; H) \times \mathfrak{T} \mathfrak{R} ; x \mapsto\left(\partial_{0} x+Y^{*} Y x+B_{0} x, x(0)\right)
$$

is an isomorphism.
Proof. For $f \in L_{2}([0, \tau] ; H)$ and $u_{0} \in \mathfrak{T} \mathfrak{R}$ we want to find a solution $u \in \mathfrak{M R}_{\tau}$ of

$$
\begin{aligned}
\left(\partial_{0}+Y^{*} Y+B_{0}\right) u & =f \\
u(0) & =u_{0}
\end{aligned}
$$

Case $u_{0}=0$ : For $\lambda \in \mathbb{R}_{>0}$ consider

$$
\Phi_{\lambda}: L_{2}([0, \tau] ; H) \rightarrow L_{2}([0, \tau] ; H) ; x \mapsto e^{-\lambda m_{0}} f-B_{0}\left(\partial_{0}+Y^{*} Y+\lambda\right)^{-1} x
$$

The lemma above ensures that $\Phi_{\lambda}$ is well-defined and for $x, y \in L_{2}([0, \tau] ; H)$ we observe

$$
\begin{aligned}
& \left\|\Phi_{\lambda}(x)-\Phi_{\lambda}(y)\right\|_{L_{2}([0, \tau] ; H)} \\
= & \left\|B_{0}\left(\partial_{0}+Y^{*} Y+\lambda\right)^{-1}(x-y)\right\|_{L_{2}([0, \tau] ; H)} \\
\leq & C_{\alpha}\left\|\left(\partial_{0}+Y^{*} Y+\lambda\right)^{-1}(x-y)\right\|_{L_{2}([0, \tau] ; H)}+\alpha\left\|\left(\partial_{0}+Y^{*} Y+\lambda\right)^{-1}(x-y)\right\|_{\mathfrak{M}_{\mathfrak{R}_{\tau}}} \\
\leq & \left(\frac{C_{\alpha}}{\lambda}+\alpha\left\|\left(\partial_{0}+Y^{*} Y+\lambda\right)^{-1}\right\|_{L\left(L_{2}([0, \tau] ; H), \mathfrak{M} \mathfrak{R}_{\tau}\right)}\right)\|x-y\|_{L_{2}([0, \tau] ; H)} .
\end{aligned}
$$

For $\alpha$ sufficiently small and subsequently $\lambda$ large, the lemma above implies that

$$
\lambda \mapsto\left\|\left(\partial_{0}+Y^{*} Y+\lambda\right)^{-1}\right\|_{L\left(L_{2}([0, \tau] ; H), \mathfrak{M}_{\tau}\right)}
$$

can be uniformly bounded and, hence, there are choices of $\alpha$ and $\lambda$ such that

$$
\left(\frac{C_{\alpha}}{\lambda}+\alpha\left\|\left(\partial_{0}+Y^{*} Y+\lambda\right)^{-1}\right\|_{L\left(L_{2}([0, \tau] ; H), \mathfrak{M} \mathfrak{R}_{\tau}\right)}\right)<1
$$

i.e., $\Phi_{\lambda}$ a contraction.

Let $x^{*} \in L_{2}([0, \tau] ; H)$ be the unique fixed point of $\Phi_{\lambda}$, i.e.,

$$
x^{*}=e^{-\lambda m_{0}} f-B_{0}\left(\partial_{0}+Y^{*} Y+\lambda\right)^{-1} x^{*}
$$

holds. Considering $u^{*}:=e^{\lambda m_{0}}\left(\partial_{0}+Y^{*} Y+\lambda\right)^{-1} x^{*} \in \mathfrak{M} \mathfrak{R}_{0}$ we observe

$$
\left(\partial_{0}+Y^{*} Y+\lambda\right) e^{-\lambda m_{0}} u^{*}=e^{-\lambda m_{0}} f-B_{0} e^{-\lambda m_{0}} u^{*}
$$

and, therefore,

$$
\left(\partial_{0}+Y^{\star} Y+B_{0}\right) u^{*}=f
$$

Case $u_{0} \neq 0$ : Choose $w \in \mathfrak{M R}_{\tau}$ with $w(0)=u_{0}$ and consider

$$
\begin{equation*}
\left(\partial_{0}+Y^{*} Y+B_{0}\right)(u-w)=f-\left(\partial_{0}+Y^{*} Y+B_{0}\right) w \tag{*}
\end{equation*}
$$

Then the first case yields a solution $v$ of $(*)$ and $u^{*}:=v+w \in \mathfrak{M} \mathfrak{R}_{\tau}$ solves the initial problem.

The two cases above show that $J: \mathfrak{M R}_{\tau} \rightarrow L_{2}([0, \tau] ; H) \times \mathfrak{T} \mathfrak{R}$ is a bijection and the bounded inverse theorem ${ }^{3}$ yields that $J$ is, in fact, an isomorphism.

3
Theorem 8.4 (Bounded Inverse Theorem). Let $X_{1}$ and $X_{2}$ be Banach spaces and $T \in$ $L\left(X_{1}, X_{2}\right)$ bijective. Then $T^{-1} \in L\left(X_{2}, X_{1}\right)$.

Corollary 8.5. The Stokes operator

$$
I: \mathfrak{M R}_{\tau} \rightarrow L_{2}([0, \tau] ; H) \times \mathfrak{T} \mathfrak{R} ; x \mapsto\left(\partial_{0} x+Y^{*} Y x, x(0)\right)
$$

is an isomorphism.
We are now going to prove that $\mathcal{B}_{v}$ is a compact operator. Therefore, we need to have a look at some compact embedding theorems, first.

Lemma 8.6 (Aubin-Lions). Let $X_{0}, X_{1}, X_{2}$ be Banach spaces, $X_{0}$ and $X_{1}$ be reflexive, and

$$
X_{0} \hookrightarrow_{\text {compact }} X \hookrightarrow_{\text {continuous }} X_{1} .
$$

Let $p, q \in \mathbb{R}_{\geq 1}$ and

$$
W:=\left\{f \in L_{p}\left([0, \tau] ; X_{0}\right) ; f^{\prime} \in L_{q}\left([0, \tau] ; X_{1}\right)\right\}
$$

Then

$$
W \rightarrow_{\text {compact }} L_{p}([0, \tau] ; X) .
$$

Proof. see [17]; Proposition III.1.3
We will start by proving two embedding theorems.
Lemma 8.7. $\mathfrak{M R}_{\tau} \rightarrow_{\text {compact }} L_{2}([0, \tau] ; H)$
Proof. Clearly, $\mathfrak{M P}_{\tau}=W_{2}^{1}([0, \tau], H) \cap L_{2}\left([0, \tau], D\left(Y^{\star} Y\right)\right)$ is continuously embedded into

$$
W:=\left\{u \in L_{2}\left([0, \tau], D\left(Y^{*} Y\right)\right) ; u^{\prime} \in L_{2}([0, \tau], H)\right\}
$$

Using Aubin-Lions' Lemma (Lemma 8.6) with $X_{0}:=D\left(Y^{*} Y\right), X:=X_{1}:=H$, and $p:=q:=2$, the assertion reduces to showing

$$
D\left(Y^{*} Y\right) \subseteq \operatorname{pr}_{[\{0\}] \operatorname{tr} \operatorname{grad}_{(0)}}\left[W_{2}^{1,(1,0)}(M)\right] \hookrightarrow_{\text {compact }} H
$$

Let $\left(f_{n}\right)_{n \in \mathbb{N}} \in D\left(Y^{*} Y\right)^{\mathbb{N}}$ be a bounded sequence. Then, $\left(\operatorname{pr}_{[\{0\}] \operatorname{trgrad}(0)}^{*} f_{n}\right)_{n \in \mathbb{N}}$ is a bounded sequence in $W_{2}^{1,(1,0)}(M)$ which is compactly embedded in $L_{2}^{(1,0)}(M)$ by the Rellich-Kondrachov condition. In other words, there exists a subsequence $\left(\operatorname{pr}_{[\{0\}] \operatorname{trgrad}(0)}^{*} f_{n_{k}}\right)_{k \in \mathbb{N}}$ which converges in $L_{2}^{(1,0)}(M)$. Hence,

$$
\left(f_{n_{k}}\right)_{k \in \mathbb{N}}=\left(\operatorname{pr}_{\overline{[\{0\}] \operatorname{tr} \operatorname{grad}_{(0)}}} \operatorname{pr}_{[\{0\}] \operatorname{tr} \operatorname{grad}_{(0)}}^{*} f_{n_{k}}\right)_{k \in \mathbb{N}}
$$

converges in $\operatorname{pr}_{[\{0\}] \operatorname{tr} \operatorname{grad}_{(0)}}\left[L_{2}^{(1,0)}(M)\right]=H$.

Remark Using the theorem of Arzelà-Ascoli, it is possible to show that the embedding $\mathfrak{M R}_{\tau} \hookrightarrow C([0, \tau], H)$ is compact, as well.

Lemma 8.8. $|Y|\left[\mathfrak{M R}_{\tau}\right] \hookrightarrow_{\text {compact }} L_{2}([0, \tau] ; H)$
Proof. Note that $|Y|:=\sqrt{Y^{*} Y}$ is non-negative, i.e., $-1 \in \varrho(|Y|)$. Let

$$
\forall x \in H:\|x\|_{H_{-1}}:=\left\|(|Y|+1)^{-1} x\right\|_{H}
$$

and

$$
H_{-1}:=\bar{H}^{\|\cdot\|_{H_{-1}}} .
$$

Then, $|Y|+1$ maps $H$ unitarily to $X_{1}$. Furthermore, $\left(\partial_{0} \otimes 1\right)(1 \otimes|Y|)$ and $(1 \otimes$ $|Y|)\left(\partial_{0} \otimes 1\right)$ coincide in $L\left(\mathfrak{M R}_{\tau}, L_{2}\left([0, \tau], H_{-1}\right)\right)$. Thus,

$$
\forall x \in \mathfrak{M R}_{\tau}: \partial_{0}|Y| x=|Y| \partial_{0} x \in L_{2}\left([0, \tau], H_{-1}\right)
$$

implies that $|Y|\left[\mathfrak{M R}_{\tau}\right]$ is continuously embedded in

$$
W:=\left\{x \in L_{2}([0, \tau], D(|Y|)) ; \partial_{0} x \in L_{2}\left([0, \tau], H_{-1}\right)\right\} .
$$

Furthermore, $D(|Y|)=\operatorname{pr} \frac{}{[\{0\}] \operatorname{grad}_{(0)}}\left[W_{2}^{1,(1,0)}(M)\right]$ is compactly embedded in $H$ by the Rellich-Kondrachov condition and the calculation in the proof of Lemma 8.7. Choosing $p:=q:=2, X_{0}:=D(|Y|), X:=H$, and $X_{1}:=H_{-1}$ in Aubin-Lions' Lemma (Lemma 8.6), thus, yields that $W$ is compactly embedded in $L_{2}([0, \tau], H)$ and, hence, the assertion.

Before proving compactness of $\mathcal{B}_{v}$, we will need one last lemma.
Lemma 8.9. Let $H_{1}$ and $H_{2}$ be Hilbert spaces and $T \in L\left(H_{1}, H_{2}\right)$. Then, the following are equivalent.
(i) $T$ is compact.
(ii) $T$ maps weakly-convergent sequences to norm-convergent sequences.

Proof. "(i) $\Rightarrow$ (ii)" Let $\left(x_{n}\right)_{n \in \mathbb{N}} \in H_{1}^{\mathbb{N}}$ be weakly convergent to $x \in H_{1}$. Then, $\left(T x_{n}\right)_{n \in \mathbb{N}} \in H_{2}^{\mathbb{N}}$ converges weakly to $T x$ since

$$
\forall y \in H_{2}:\left\langle T x_{n}, y\right\rangle_{H_{2}}=\left\langle x_{n}, T^{*} y\right\rangle_{H_{1}} \rightarrow\left\langle x, T^{*} y\right\rangle_{H_{1}}=\langle T x, y\rangle_{H_{2}}
$$

Suppose $\left(T x_{n}\right)_{n \in \mathbb{N}}$ does not converge in norm. Then, there exists $\delta \in \mathbb{R}_{>0}$ and a subsequence $\left(T x_{n_{k}}\right)_{k \in \mathbb{N}}$ such that

$$
\forall k \in \mathbb{N}:\left\|T x_{n_{k}}-T x\right\|_{H_{2}}>\delta
$$

The uniform boundedness principle ${ }^{4}$ for $F=\left\{y \mapsto\left\langle x_{n}, y\right\rangle_{H_{1}} ; n \in \mathbb{N}\right\}$ yields that $\left(x_{n}\right)_{n \in \mathbb{N}}$ is a bounded sequence. Therefore, $\left(T x_{n_{k}}\right)_{k \in \mathbb{N}}$ is bounded and, since $T$ is compact, there is a norm-convergent subsequence $\left(T x_{n_{k_{j}}}\right)_{j \in \mathbb{N}}$ with $T x_{n_{k_{j}}} \rightarrow: \hat{y}(j \rightarrow$ $\infty)$. Since norm-convergence implies weak convergence, we obtain that $\left(T x_{n_{k_{j}}}\right)_{j \in \mathbb{N}}$ converges weakly to $\hat{y}$, as well. But the weak limit was $T x$, i.e., $\hat{y}=T x$ by the Highlander principle ${ }^{5}$ which is a contradiction.
"(ii) $\Rightarrow(\mathrm{i}) "$ Since $H_{1}$ is a Hilbert space, the unit ball $B_{H_{1}}$ is weakly compact (Banach-Alaoglu). Let $\left(x_{n}\right)_{n \in \mathbb{N}} \in B_{H_{1}}^{\mathbb{N}}$. Then, $\left(x_{n}\right)_{n \in \mathbb{N}}$ contains a weakly convergent subsequence which, by (ii), is mapped to a norm-convergent subsequence of $\left(T x_{n}\right)_{n \in \mathbb{N}}$; hence, $T$ is compact.

Proposition 8.11. Let $v \in \mathfrak{M R}_{\tau}$. Then $\mathcal{B}_{v}$ is compact.
Proof. Note that it suffices to show that $B^{\prime}(v)$ maps weakly convergent sequences in $\mathfrak{M} \mathfrak{R}_{\tau}$ to norm-convergent sequences in $L_{2}([0, \tau] ; H)$. Let $w \in \mathfrak{M}_{\tau}^{\mathbb{N}}$ be weakly convergent to $w_{0} \in \mathfrak{M R}_{\tau}$.

4
Theorem 8.10 (uniform boundedness principle). Let $X$ be a Banach space and $N a$ normed vector space. Let $F \subseteq L(X, N)$ be such that $\forall x \in X: \sup _{T \in F}\|T x\|_{N}<\infty$. Then, $\sup _{T \in F}\|T\|_{L(X, N)}<\infty$.
${ }^{5}$ There can only be one [limit].
(i) Using polar decomposition $Y=V|Y|$ and observing that

$$
L_{2}^{(1,0)} \ni x \mapsto \operatorname{tr}\left(v(t) \otimes C^{-\frac{1}{2}} V x\right) \in L_{2}^{(1,0)}(M)
$$

is continuous with

$$
\sup _{t \in[0, \tau]}\left\|L_{2}^{(1,0)}(M) \ni x \mapsto \operatorname{tr}\left(v(t) \otimes C^{-\frac{1}{2}} V x\right) \in L_{2}^{(1,0)}(M)\right\|_{L\left(L_{2}^{(1,0)}(M)\right)}<\infty
$$

since $\mathfrak{M}_{\tau} \subseteq C([0, \tau] ; H)$, it remains to show that $\left(|Y| w_{n}\right)_{n \in \mathbb{N}}$ is norm-convergent in $L_{2}([0, \tau] ; H)$ which follows directly from $|Y|\left[\mathfrak{M R}_{\tau}\right] \rightarrow_{\text {compact }} L_{2}([0, \tau] ; H)$.
(ii) Note, $\mathfrak{M} \mathfrak{R}_{\tau} \subseteq C([0, \tau] ; H) \cap L_{2}\left([0, \tau] ; D\left(Y^{*} Y\right)\right)$ also implies that every continuous representative of $v$ takes values in $D\left(Y^{*} Y\right)$ almost everywhere, i.e., for almost every $t$ we obtain

$$
C^{-\frac{1}{2}} Y v(t) \in W_{2}^{1,(2,0)}(M) \subseteq L_{2}^{(2,0)}(M)=: H_{2}
$$

Let $D:=C^{-\frac{1}{2}} Y v \in L_{\infty}\left([0, \tau] ; H_{2}\right)$. Introducing the abbreviations $x_{n}:=w_{n}-w_{0}$ and $E:=D_{\beta \gamma} D_{\varepsilon \delta} g^{\gamma \delta} g^{\beta} \otimes g^{\varepsilon} \in L_{\infty}\left([0, \tau] ; H_{2}\right)$, we observe

$$
\begin{aligned}
& \left\|\operatorname{tr}\left(x_{n} \otimes D\right)\right\|_{L_{2}([0, \tau] ; H)}^{2} \\
= & \int_{0}^{\tau}\left\|x_{n}(t)_{\alpha} g^{\alpha \beta} D(t)_{\beta \gamma} g^{\gamma}\right\|_{H}^{2} d t \\
= & \int_{0}^{\tau} x_{n}(t)_{\alpha} g^{\alpha \beta} D(t)_{\beta \gamma} g^{\gamma \delta} D(t)_{\varepsilon \delta} g^{\zeta \varepsilon} x_{n}(t)_{\zeta} d t \\
= & \int_{0}^{\tau}\left|\left\langle x_{n}(t)_{\alpha} x_{n}(t)_{\zeta} g^{\alpha} \otimes g^{\zeta}, D(t)_{\beta \gamma} D(t)_{\varepsilon \delta} g^{\gamma \delta} g^{\beta} \otimes g^{\varepsilon}\right\rangle_{(2,0)}\right| d t \\
\leq & \int_{0}^{\tau}\left\|\left(x_{n} \otimes x_{n}\right)(t)\right\|_{L_{2}^{(2,0)}(M)}\|E(t)\|_{L_{2}^{(2,0)}(M)} d t \\
\leq & \|E\|_{L_{\infty}\left([0, \tau] ; H_{2}\right)} \int_{0}^{\tau}\left\|x_{n}(t)\right\|_{H}^{2} d t \\
= & \|E\|_{L_{\infty}\left([0, \tau] ; H_{2}\right)}\left\|w_{n}-w_{0}\right\|_{L_{2}([0, \tau] ; H)}^{2}
\end{aligned}
$$

which converges to zero since $\mathfrak{M R}_{\tau} \rightarrow_{\text {compact }} L_{2}([0, \tau] ; H)$.

Hence, $F_{\tau}^{\prime}(v)$ is a Fredholm operator of index zero, i.e., injective if and only if its range is dense. Since the range is also closed we obtain the following corollary.

Corollary 8.12. Let $v \in \mathfrak{M}_{\tau}$. Then, $F_{\tau}^{\prime}(v)$ is an isomorphism if and only if $F_{\tau}^{\prime}(v)$ is injective.

For $x, y, z \in \mathfrak{M R}_{\tau}$ let

$$
\beta(x, y, z):=\left\langle-2 \operatorname{pr}_{[\{0\}] \operatorname{trgrad}(0)} \operatorname{tr}\left(x \otimes C^{-\frac{1}{2}} Y y\right), z\right\rangle_{H}
$$

Note that

$$
\begin{aligned}
\beta(x, y, z) & =-2\left\langle\operatorname{tr}\left(x \otimes \operatorname{sym} \operatorname{grad}_{(0)} y\right), z\right\rangle_{(1,0)} \\
& =\int_{M} x_{i}\left(\nabla_{g_{j}} y_{k}+\nabla_{g_{k}} y_{j}\right) g^{i j} g^{k m} z_{m} d \operatorname{vol}_{M}
\end{aligned}
$$

Thus,

$$
\begin{aligned}
\beta(x, y, y) & =\int_{M} x_{i}\left(\nabla_{g_{j}} y_{k}+\nabla_{g_{k}} y_{j}\right) g^{i j} g^{k m} y_{m} d \mathrm{vol}_{M} \\
& =-\int_{M} x_{i} g^{i j}\left(\nabla_{g_{j}} y_{k} g^{k m} y_{m}+\nabla_{g_{k}} y_{j} g^{k m} y_{m}\right) d \operatorname{vol}_{M} \\
& =-\int_{M} x_{i} g^{i j} \frac{\nabla_{g_{j}}\langle y, y\rangle_{(1,0)}}{2}+x_{i} g^{i j}\left\langle\operatorname{grad}_{(0)} y_{j}, y\right\rangle_{(1,0)} d \operatorname{vol}_{M}
\end{aligned}
$$

$$
\begin{aligned}
& =-\int_{M} \frac{1}{2}\langle\underbrace{\operatorname{div}_{(0)} x}_{=0},\langle y, y\rangle_{(1,0)}\rangle_{(0,0)}+\langle x,\langle y_{j}, \underbrace{\operatorname{div}_{(0)} y}_{=0}\rangle_{(1,0)}\rangle_{(0,0)} d \mathrm{vol}_{M} \\
& =0
\end{aligned}
$$

as well as,

$$
0=\beta(x, y+z, y+z)=\beta(x, y, y)+\beta(x, y, z)+\beta(x, z, y)+\beta(x, z, z)
$$

i.e.,

$$
\beta(x, y, z)=-\beta(x, z, y)
$$

The last ingredient we need to prove injectivity of $F_{\tau}^{\prime}(v)$ is Gronwall's lemma.
Lemma 8.13 (Gronwall's Lemma). Let $f, g, h: \mathbb{R}_{\geq 0} \rightarrow \mathbb{R}_{\geq 0}$ be measurable with

$$
f(t) \leq g(t)+\int_{0}^{t} f(s) h(s) d s
$$

for almost every $t \in \mathbb{R}_{\geq 0}$. Then

$$
f(t) \leq g(t)+\int_{0}^{t} g(s) h(s) \exp \left(\int_{s}^{t} h(r) d r\right) d s
$$

holds for almost every $t \in \mathbb{R}_{\geq 0}$.
Proof. see [16]; Theorem A. 43
Proposition 8.14. Let $v \in \mathfrak{M R}_{\tau}$. Then, $F_{\tau}^{\prime}(v)$ is injective. In particular, $F_{\tau}$ is locally a diffeomorphism.

Proof. Let $x \in \mathfrak{M} \mathfrak{R}_{\tau}$ and $F_{\tau}^{\prime}(v) x=0$. To show: $x=0$. First, note that $F_{\tau}^{\prime}(v) x=0$ is equivalent to

$$
x(0)=0 \wedge \partial_{0} x+Y^{*} Y x+B^{\prime}(v) x=0
$$

Multiplying the latter scalarly with $x$ in $H$ yields

$$
\begin{aligned}
0 & =\left\langle\partial_{0} x, x\right\rangle_{H}+\langle | Y|x,|Y| x\rangle_{H}+\beta(x, v, x)+\beta(v, x, x) \\
& =\frac{1}{2}\left(\|x\|_{H}^{2}\right)^{\prime}+\||Y| x\|_{H}^{2}+\beta(x, v, x) .
\end{aligned}
$$

With $\tilde{v}:=\left(\nabla_{g_{j}} v_{k}+\nabla_{g_{k}} v_{j}\right) g^{k n}\left(\nabla_{g_{m}} v_{n}+\nabla_{g_{n}} v_{m}\right) g^{j} \otimes g^{m}$ this last equation yields

$$
\begin{aligned}
& \left(\|x\|_{H}^{2}\right)^{\prime}+\||Y| x\|_{H}^{2} \\
\leq & 2|\beta(x, v, x)| \\
= & 2\left|\left\langle-2 \operatorname{pr}_{[\{0\}] \operatorname{tr} \operatorname{grad}_{(0)}} \operatorname{tr}\left(x \otimes C^{-\frac{1}{2}} Y v\right), x\right\rangle_{H}\right| \\
\leq & 4\left\|\operatorname{tr}\left(x \otimes \operatorname{symgrad}_{(0)} v\right)\right\|_{H}\|x\|_{H} \\
= & 4\|x\|_{H}\left(\int_{M} x_{i} g^{i j} \frac{1}{2}\left(\nabla_{g_{j}} v_{k}+\nabla_{g_{k}} v_{j}\right) g^{k n} x_{l} g^{l m} \frac{1}{2}\left(\nabla_{g_{m}} v_{n}+\nabla_{g_{n}} v_{m}\right) d \mathrm{vol}_{M}\right)^{\frac{1}{2}} \\
= & 2\|x\|_{H}\left|\left\langle x \otimes x,\left(\nabla_{g_{j}} v_{k}+\nabla_{g_{k}} v_{j}\right) g^{k n}\left(\nabla_{g_{m}} v_{n}+\nabla_{g_{n}} v_{m}\right) g^{j} \otimes g^{m}\right\rangle_{L_{2}^{(2,0)}(M)}\right|^{\frac{1}{2}} \\
\leq & 2\|x\|_{H}\left(\|x \otimes x\|_{L_{2}^{(2,0)}(M)}\|\tilde{v}\|_{L_{2}^{(2,0)}(M)}\right)^{\frac{1}{2}} \\
= & 2\|\tilde{v}\|_{L_{2}^{(2,0)}(M)}^{\frac{1}{2}}\|x\|_{H}^{2} \\
\leq & 2\|\tilde{v}\|_{L_{2}^{(2,0)}(M)}^{\frac{1}{2}}\|x\|_{H}^{2}+\||Y| x\|_{H}^{2}
\end{aligned}
$$

and, hence,

$$
\left(\|x\|_{H}^{2}\right)^{\prime} \leq 2\|\tilde{v}\|_{L_{2}^{(2,0)}(M)}^{\frac{1}{2}}\|x\|_{H}^{2}
$$

Thus, integration yields

$$
\|x(t)\|_{H}^{2} \leq\|x(0)\|_{H}^{2}+\int_{0}^{t} 2\|\tilde{v}(s)\|_{L_{2}^{(2,0)}(M)}^{\frac{1}{2}}\|x(s)\|_{H}^{2} d s
$$

for almost every $t$ and, investing Gronwall's lemma, gives

$$
\|x(t)\|_{H}^{2} \leq \underbrace{\|x(0)\|_{H}^{2}}_{=0}+\int_{0}^{t} 2 \underbrace{\|x(0)\|_{H}^{2}}_{=0}\|\tilde{v}(s)\|_{L_{2}^{(2,0)}(M)}^{\frac{1}{2}} e^{\int_{s}^{t} 2\|\tilde{v}(r)\|_{\left.L_{2}^{2}, 0\right)}^{\frac{1}{2}}(M)} d r d s=0
$$

The very same mechanism also yields the following proposition.
Proposition 8.15. $F_{\tau}$ is injective. In particular, $F_{\tau}$ is a diffeomorphism.
Proof. Let $x, y \in \mathfrak{M R}_{\tau}$ with $F_{\tau}(x)=F_{\tau}(y)$. Then $z:=x-y$ satisfies

$$
\begin{aligned}
\partial_{0} z+Y^{*} Y z & =B(y)-B(x) \text { in }(0, \tau) \times M, \\
z(0) & =0 \text { in } M .
\end{aligned}
$$

Just as before, but now in $z$, we obtain

$$
\begin{aligned}
\left(\|z\|_{H}^{2}\right)^{\prime}+2\||Y| z\|_{H}^{2} & =-2 \beta(x, x, z)+2 \beta(y, y, z) \\
& =-2(z, x, z)-2 \beta(y, x, z)+2 \beta(y, y, z) \\
& =-2 \beta(z, x, z)-2 \underbrace{\beta(y, z, z)}_{=0} \\
& \leq 2|\beta(z, x, z)|
\end{aligned}
$$

Choosing $\tilde{x}:=\left(\nabla_{g_{j}} x_{k}+\nabla_{g_{k}} x_{j}\right) g^{k n}\left(\nabla_{g_{m}} x_{n}+\nabla_{g_{n}} x_{m}\right) g^{j} \otimes g^{m}$ yields

$$
\begin{aligned}
& \left(\|z\|_{H}^{2}\right)^{\prime}+\||Y| z\|_{H}^{2} \\
\leq & 2|\beta(z, x, z)| \\
= & 2\left|\left\langle-2 \operatorname{pr}_{[\{0\}] \operatorname{tr} \operatorname{grad}_{(0)}} \operatorname{tr}\left(z \otimes C^{-\frac{1}{2}} Y x\right), z\right\rangle_{H}\right| \\
\leq & 4\left\|\operatorname{tr}\left(z \otimes{\operatorname{sym} \operatorname{grad}_{(0)}} x\right)\right\|_{H}\|z\|_{H} \\
= & 4\|z\|_{H}\left(\int_{M} z_{i} g^{i j} \frac{1}{2}\left(\nabla_{g_{j}} x_{k}+\nabla_{g_{k}} x_{j}\right) g^{k n} z_{l} g^{l m} \frac{1}{2}\left(\nabla_{g_{m}} x_{n}+\nabla_{g_{n}} x_{m}\right) d \operatorname{vol}_{M}\right)^{\frac{1}{2}} \\
= & 2\|z\|_{H}\left|\left\langle z \otimes z,\left(\nabla_{g_{j}} x_{k}+\nabla_{g_{k}} x_{j}\right) g^{k n}\left(\nabla_{g_{m}} x_{n}+\nabla_{g_{n}} x_{m}\right) g^{j} \otimes g^{m}\right\rangle_{L_{2}^{(2,0)}(M)}\right|^{\frac{1}{2}} \\
\leq & 2\|z\|_{H}\left(\|z \otimes z\|_{L_{2}^{(2,0)}(M)}\|\tilde{x}\|_{L_{2}^{(2,0)}(M)}\right)^{\frac{1}{2}} \\
= & 2\|\tilde{x}\|_{L_{2}^{(2,0)}(M)}^{\frac{1}{2}}\|z\|_{H}^{2} \\
\leq & 2\|\tilde{x}\|_{L_{2}^{2}}^{\frac{1}{2},(0)}(M)
\end{aligned}\|z\|_{H}^{2}+\||Y| z\|_{H}^{2} .
$$

and, hence,

$$
\left(\|z\|_{H}^{2}\right)^{\prime} \leq 2\|\tilde{x}\|_{L_{2}^{(2,0)}(M)}^{\frac{1}{2}}\|z\|_{H}^{2}
$$

Thus, integration yields

$$
\|z(t)\|_{H}^{2} \leq\|z(0)\|_{H}^{2}+\int_{0}^{t} 2\|\tilde{x}(s)\|_{L_{2}^{(2,0)}(M)}^{\frac{1}{2}}\|z(s)\|_{H}^{2} d s
$$

for almost every $t$ and, investing Gronwall's lemma, gives

$$
\|z(t)\|_{H}^{2} \leq \underbrace{\|z(0)\|_{H}^{2}}_{=0}+\int_{0}^{t} 2 \underbrace{\|z(0)\|_{H}^{2}}_{=0}\|\tilde{x}(s)\|_{L_{2}^{(2,0)}(M)}^{\frac{1}{2}} e^{\int_{s}^{t} 2\|\tilde{x}(r)\|_{L_{2}^{(2,0)}(M)}^{\frac{1}{2}} d r} d s=0
$$

## CHAPTER 9

## Causality and Well-posedness

As of the end of chapter 8 we know that $F_{\tau}$ is injective and locally an analytic diffeomorphism, i.e., if we have a solution $u$ of the Navier-Stokes problem then it is unique with respect to the data $F_{\tau}(u)$, changing $F_{\tau}(u)$ slightly does not destroy unique solvability, and the corresponding solutions depend analytically on the data. But for our construction of solutions with arbitrary data to work, the notion of causality is needed.

Definition 9.1. Let $R \subseteq L_{2}\left([0, \tau] ; X_{1}\right) \oplus L_{2}\left([0, \tau] ; X_{2}\right)$ where $X_{1}$ and $X_{2}$ are Banach spaces.
(i) The relation $R$ is called weakly causal if and only if

$$
\forall\left(u, f_{u}\right),\left(v, f_{v}\right) \in R: \inf \operatorname{spt}_{0}(u-v) \leq \inf \operatorname{spt}_{0}\left(f_{u}-f_{v}\right)
$$

where $\operatorname{spt}_{0}$ denotes the support with respect to time, i.e., in $L_{2}([0, \tau])$.
(ii) The relation $R$ is called strongly causal if and only if $R$ is weakly causal and

$$
\forall(u, f) \in R: \inf _{\operatorname{spt}_{0}} u \leq \inf _{\operatorname{spt}_{0} f}
$$

(iii) $R$ is said to have weakly causal solutions if and only if $R^{-1}$ is weakly causal.
(iv) $R$ is said to have strongly causal solutions if and only if $R^{-1}$ is strongly causal.

Additionally, let $R$ be linear.
(v) The linear relation $R$ is called causal if and only if

$$
\forall(u, f) \in R: \inf \operatorname{spt}_{0} u \leq \inf \operatorname{spt}_{0} f .
$$

(vi) The linear relation $R$ is said to have causal solutions if and only if $R^{-1}$ is causal.

Corollary 9.2. Let $X_{1}$ and $X_{2}$ be Banach spaces and $R \subseteq L_{2}\left([0, \tau] ; X_{1}\right) \oplus$ $L_{2}\left([0, \tau] ; X_{2}\right)$ with $0 \in R$. Then, weak causality and strong causality are equivalent.

Proof. For $R=\varnothing$ the assertion is trivial and, since strong causality implies weak causality, there is only one direction to show. Let $(u, f) \in R$. Then,

$$
\forall\left(v, f_{v}\right) \in R: \inf \operatorname{spt}_{0}(u-v) \leq \inf \operatorname{spt}_{0}\left(f-f_{v}\right)
$$

implies

$$
\operatorname{inf~spt}_{0} u=\inf \operatorname{spt}_{0}(u-0) \leq \inf _{\operatorname{spt}_{0}}(f-0)=\inf \operatorname{spt}_{0} f
$$

because $0 \in R$.

Corollary 9.3. Let $X_{1}$ and $X_{2}$ be Banach spaces and $R \subseteq L_{2}\left([0, \tau] ; X_{1}\right) \oplus$ $L_{2}\left([0, \tau] ; X_{2}\right)$ linear. Then, weak causality, strong causality, and "inear" causality are equivalent.

Proof. For $R=\varnothing$ the assertion is trivial and, since linear relations contain zero and strong causality trivially implies "linear" causality, it suffices to show that "linear" causality implies weak causality. Let $\left(u, f_{u}\right),\left(v, f_{v}\right) \in R$. Then, $\left(u-v, f_{u}-\right.$ $\left.f_{v}\right) \in R$ by linearity and, thus,

$$
\inf \operatorname{spt}_{0}(u-v) \leq \inf \operatorname{spt}_{0}\left(f-f_{v}\right)
$$

These are very neat properties as we only need to show weak causality of solutions for our construction to work whereas strong causality of solutions implies that the "vacuum solution" is zero, that is, a motionless fluid will remain at rest as long as no external force acts on it. This property does not hold if we have proper weak causality of solutions but is essential for the system to be physically senseful.

Lemma 9.4. Let $X_{1}$ and $X_{2}$ be Banach spaces, and $\left(R_{t}\right)_{t \in \mathbb{R}_{>0}}$ a family of leftunique ${ }^{1}$ relations $R_{t} \subseteq L_{2}\left([0, t] ; X_{1}\right) \oplus L_{2}\left([0, t] ; X_{2}\right)$ such that

$$
\begin{equation*}
\forall t \in \mathbb{R}_{>0} \forall s \in(0, t) \forall(u, f) \in R_{t}:\left(\left.u\right|_{[0, s]},\left.f\right|_{[0, s]}\right) \in R_{s} \tag{6}
\end{equation*}
$$

holds. Then, all $R_{t}$ have weakly causal solutions.
Proof. Suppose $R_{t}$ does not have weakly causal solutions for some $t \in \mathbb{R}_{>0}$. Then there are $\left(u, f_{u}\right),\left(v, f_{v}\right) \in R_{t}$ with

$$
\inf \sup _{0}(u-v)<\inf \operatorname{spt}_{0}\left(f_{u}-f_{v}\right)
$$

By left-uniqueness, this implies $u \neq v$ because $f_{u}$ and $f_{v}$ must be distinct. Choose $s \in\left(\inf _{\sup _{0}}(u-v), \inf \operatorname{spt}_{0}\left(f_{u}-f_{v}\right)\right)$. Then

$$
\left.u\right|_{[0, s]} \neq v_{[0, s]}
$$

and
(*)

$$
\left.f_{u}\right|_{[0, s]}=\left.f_{v}\right|_{[0, s]}
$$

hold. But from (*) and left-uniqueness of $R_{s}$ we deduce

$$
\left.u\right|_{[0, s]}=v_{[0, s]}
$$

which is a contradiction.

Since $\operatorname{grad}_{(0)}, \operatorname{div}_{(0)}, \partial_{0}, \otimes, \operatorname{tr}$, and sym obviously are causal operators and $C$ was defined to be causal, we conclude that all $F_{\tau}$ are weakly causal (which implies (6)) and, therefore, they all have strongly causal solutions. Choosing $C$ to be local, as well, is not possible in our general setting because many non-Newtonian fluids have viscous memory, that is, $C$ contains delay terms. However, it would break physics to assume the viscosity depended on the future. Hence, $C$ ought to be causal. This was the last missing piece of our jigsaw and we can, now, state our main result.

Theorem 9.5 (Well-posedness and Causality). Let $\tau \in \mathbb{R}_{>0}, u_{0} \in \mathfrak{T} \mathfrak{R}$, and $f \in L_{2}([0, \tau] ; H)$.
(i) There exist $\tau^{\prime} \in(0, \tau)$ and $u \in \mathfrak{M R}_{\tau}$ such that the Navier-Stokes equations

$$
\begin{aligned}
\partial_{0} u+Y^{*} Y u+B(u) & =f \quad \text { in }\left(0, \tau^{\prime}\right) \times M \\
u(0) & =u_{0} \text { in } M
\end{aligned}
$$

are satisfied. Furthermore, solutions are strongly causal and unique in $\mathfrak{M R}_{\tau^{\prime}}$.

[^9](ii) There exists an open neighborhood $U \subseteq L_{2}([0, \tau] ; H) \times \mathfrak{T} \mathfrak{R}$ of $\left(f, u_{0}\right)$ and $\tau^{\prime} \in(0, \tau)$ such that
$$
G: U \rightarrow \mathfrak{M R}_{\tau^{\prime}} ;\left.\quad\left(g, v_{0}\right) \mapsto F_{\tau}^{-1}\left(g, v_{0}\right)\right|_{\left[0, \tau^{\prime}\right]}
$$
is analytic and all $G\left(g, v_{0}\right)$ solve the Navier-Stokes system in $\left(0, \tau^{\prime}\right) \times M$ with respect to the data $\left(g, v_{0}\right) \in U$, i.e., the solutions depend analytically on the data and $\tau^{\prime}$ is locally constant.

## Bibliography

[1] D. J. ACHESON, Elementary Fluid Dynamics, Oxford University Press, New York, NY, 1990.
[2] M. BENEŠ and P. KUČERA, Solutions to the Navier-Stokes Equations with various types of boundary conditions, Arch. Math. 98 (2012), no. 5, 487-497. MR2922692
[3] B. BUFFONI and J. TOLAND, Analytic Theory of Bifurcation, Princeton University Press, Princeton, NJ, 2003.
[4] B. K. DRIVER, Analysis Tools with Applications and PDE Notes, Springer preprint, 2003.
[5] C. L. FEFFERMAN, Existence and Smoothness of the Navier-Stokes Equation, Clay Mathematics Institute (2000).
[6] P. GARRETT, Non-existence of tensor products of Hilbert spaces, Lecture notes, http://www.math.umn.edu/~garrett/m/v (July, 2010).
[7] J. GIL, T. KRAINER, and I WITT, Aspects of Boundary problems in Analysis and Geometry, Birkhäuser, Basel/Boston, MA/Berlin, 2004.
[8] P. HAJŁASZ and P. KOSKELA, Sobolev Met Poincaré, Memoirs of the AMS 145 (2000), no. 688.
[9] E. HEBEY, Nonlinear Analysis on Manifolds: Sobolev Spaces and Inequalities, Courant Institute of Mathematical Sciences/AMS, New York, NY/Providence, RI, 1999.
[10] E. HILLE and R. S. PHILLIPS, Functional Analysis and Semi-groups, AMS Colloquium Publications XXXI (1968).
[11] T. KATO, Perturbation Theory for Linear Operators, 2nd ed., Springer, Berlin/Heidelberg, 1980.
[12] P. C. KUNSTMANN and L. WEIS, Maximal $L_{p}$-regularity for Parabolic Equations, Fourier Multiplier Theorems and $H^{\infty}$-functional Calculus, Lecture Notes in Mathematics Volume 1855 (2004), 65-311.
[13] A. LUNARDI, Analytic Semigroups and Optimal Regularity in Parabolic Problems, Birkhäuser, Basel, 1995.
[14] R. H. PICARD, Mother operators and their descendants, J. Math. Anal. Appl. 403 (2013), no. 1, 54-62. MR3035071
[15] R. H. PICARD and D. F. MCGHEE, Partial Differential Equations: A Unified Hilbert Space Approach, de Gruyter, Berlin/New York, NY, 2011.
[16] R. L. SCHILLING and L. PARTZSCH, Brownian Motion: An Introduction to Stochastic Processes, de Gruyter, Berlin/Boston, MA, 2012.
[17] R. E. SHOWALTER, Monotone Operators in Banach Space and Nonlinear Partial Differential Equations, Mathematical Surveys and Monographs 49 (1997).
[18] M. E. TAYLOR, Partial Differential Equations I, Applied Mathematical Sciences, vol. 115, Springer, Berlin/Heidelberg/New York, NY, 1996.
[19] R. TEMAM, Navier-Stokes Equations: Theory and Numerical Analysis, North-Holland Publishing Company, Amsterdam/New York, NY/Oxford, 1977.
[20] S. TROSTORFF, Well-posedness and causality for a class of evolutionary inclusions, Dr. rer. nat. thesis, Technische Universität Dresden (2011).
[21] E. ZEIDLER, Nonlinear Functional Analysis and its Applications I: Fixed Point Theorems, Springer, New York, NY, 1986.


[^0]:    ${ }^{1}$ It works complex, as well, using the obvious adaptations to obtain sesqui-linearity.

[^1]:    ${ }^{2}$ For two infinite dimensional Hilbert spaces $H_{1}$ and $H_{2}$, there is no Hilbert space $H$ and bounded bi-linear map $j: H_{1} \times H_{2} \rightarrow H$ such that for every Hilbert space $\tilde{H}$ and bounded bi-linear $\operatorname{map} \tilde{j}: H_{1} \times H_{2} \rightarrow \tilde{H}$ there is a bounded linear operator $L: H \rightarrow \tilde{H}$ satisfying $\tilde{j}=L \circ j$.

[^2]:    ${ }^{3}$ Every Hilbert-Schmidt operator is compact.
    ${ }^{4}$ We do not distinguish between an operator (or, more generally, a function) and its graph as a function $f: X \rightarrow Y$ is, by definition, a right-unique, that is, single-valued, relation which is usually considered the graph of the function. Furthermore, we do not assume a function to be left-total since closed unbounded operators in Banach spaces may at most be densely defined.

    Also note that we induce a topology on $X \oplus Y$ if $X$ and $Y$ are Banach spaces. This topology can be defined using the norm $\|(x, y)\|_{X \oplus Y}:=\|x\|_{X}+\|y\|_{Y}$ or $\|(x, y)\|_{X \oplus Y}=\max \left\{\|x\|_{X},\|y\|_{Y}\right\}$. In the Hilbert space case, it is common to choose $\|(x, y)\|_{X \oplus Y}:=\sqrt{\|x\|_{X}^{2}+\|y\|_{Y}^{2}}$ since then $X \oplus Y$ is a Hilbert space assuming $X$ and $Y$ are.

[^3]:    ${ }^{1} f \in C^{1,1}$ means $f$ is Fréchet-differentiable and its derivative $f^{\prime}$ is Hölder continuous with Hölder exponent one, i.e., $f^{\prime}$ is locally Lipschitz.
    ${ }^{2} \mathrm{~A}$ topological space is called contractible if and only if the identity map is null-holomorphic.

[^4]:    $3_{\text {that is, for instance, }} \operatorname{tr}\left(a_{i j k} g^{i} \otimes g^{j} \otimes g^{k}\right)=a_{i j k} g^{i j} g^{k}$

[^5]:    ${ }^{1}$ Let $\tau$ be a $(2,0)$-tensor. Then, $\operatorname{sym} T(x, y):=\frac{1}{2}(T(x, y)+T(y, x))=\frac{T_{i j}+T_{j i}}{2} g^{i} \otimes g^{j}$.

[^6]:    ${ }^{2}$ Mind that $L_{2}([0, \tau] ; H)=L_{2}([0, \tau]) \otimes H$ holds for every Hilbert space $H$.

[^7]:    ${ }^{3}$ Note that these Neumann boundary conditions do not have vanishing normal derivative but con-normal derivative $\langle C \operatorname{grad} u, \nu\rangle=0$ where $\nu$ is the exterior normal on the boundary. Furthermore, this is a generalization of vanishing con-normal derivative which only makes sense if the boundary is sufficiently smooth because for general boundary there is no reason why the trace on $\partial M$ of $C \operatorname{grad} u$ should even exist due to the partial derivatives occurring.

[^8]:    ${ }^{1}$ The name "Stokes operator" is ambiguous here. To be precise, the operator $Y^{*} Y$ should be called Stokes operator whereas $I$ is the operator associated with the Stokes system. However, we will always refer to $Y^{*} Y$ as $Y^{*} Y$ and chose the name "Stokes operator" for $I$ for reasons of brevity.

[^9]:    ${ }^{1}$ left-uniqueness resembles injectivity

