Ralph Chill

## Skript zur Vorlesung

# Funktionalanalysis 

July 23, 2016
(c) R. Chill

## Vorwort:

Ich habe dieses Skript zur Vorlesung Funktionalanalysis an der Universität Ulm und an der TU Dresden nach meinem besten Wissen und Gewissen geschrieben. Mit Sicherheit schlichen sich jedoch Druckfehler oder gar mathematische Ungenauigkeiten ein, die man beim ersten Schreiben eines Skripts nicht vermeiden kann. Möge man mir diese Fehler verzeihen.

Obwohl die Vorlesung auf Deutsch gehalten wird, habe ich mich entschieden, dieses Skript auf Englisch zu verfassen. Auf diese Weise wird eine Brücke zwischen der Vorlesung und der (meist englischsprachigen) Literatur geschlagen. Mathematik sollte jedenfalls unabhängig von der Sprache sein in der sie präsentiert wird.

Ich danke Johannes Ruf und Manfred Sauter für ihre Kommentare zu einer früheren Version dieses Skripts. Für weitere Kommentare, die zur Verbesserungen beitragen, bin ich sehr dankbar.

## Contents

0 Primer on topology ..... 1
0.1 Metric spaces ..... 1
0.2 Sequences, convergence ..... 4
0.3 Compact spaces ..... 6
0.4 Continuity ..... 7
0.5 Completion of a metric space ..... 8
1 Banach spaces and bounded linear operators ..... 11
1.1 Normed spaces ..... 11
1.2 Product spaces and quotient spaces ..... 17
1.3 Bounded linear operators ..... 20
1.4 The Arzelà-Ascoli theorem ..... 25
2 Hilbert spaces ..... 29
2.1 Inner product spaces ..... 29
2.2 Orthogonal decomposition ..... 33
2.3 * Fourier series ..... 37
2.4 Linear functionals on Hilbert spaces ..... 42
2.5 Weak convergence in Hilbert spaces ..... 43
3 Dual spaces and weak convergence ..... 47
3.1 The theorem of Hahn-Banach ..... 47
3.2 Weak* convergence and the theorem of Banach-Alaoglu ..... 53
3.3 Weak convergence and reflexivity ..... 54
3.4 * Minimization of convex functionals ..... 59
3.5 * The von Neumann minimax theorem ..... 62
4 Uniform boundedness, bounded inverse and closed graph ..... 65
4.1 The lemma of Baire ..... 65
4.2 The uniform boundedness principle ..... 67
4.3 Open mapping theorem, bounded inverse theorem ..... 68
4.4 Closed graph theorem ..... 69
$4.5 *$ Vector-valued analytic functions ..... 72
5 Spectral theory of operators on Banach spaces, compact operators, nuclear operators ..... 75
5.1 Spectrum of closed operators ..... 75
5.2 Compact operators ..... 81
5.3 Nuclear operators ..... 88
5.4 * The mean ergodic theorem ..... 89
6 Banach algebras ..... 97
6.1 Banach algebras and the theorem of Gelfand ..... 97
$6.2 C^{*}$-algebras and the theorem of Gelfand-Naimark ..... 107
7 Operators on Hilbert spaces ..... 111
7.1 Spectral theorem for compact selfadjoint operators ..... 111
7.2 Spectral theorem for bounded, normal operators ..... 118
7.3 Spectral theorem for unbounded selfadjoint operators ..... 123
7.4 Hilbert-Schmidt operators and trace class operators ..... 128
7.5 * Elliptic partial differential equations ..... 128
7.6 * The heat equation ..... 131
7.7 * The wave equation ..... 132
7.8 * The Schrödinger equation ..... 134
8 Calculus on Banach spaces ..... 137
8.1 Differentiable functions between Banach spaces ..... 137
8.2 Local inverse function theorem and implicit function theorem ..... 138
8.3 * Newton's method ..... 142
9 Sobolev spaces ..... 143
9.1 Test functions, convolution and regularization ..... 143
9.2 Sobolev spaces in one dimension ..... 146
9.3 Sobolev spaces in several dimensions ..... 152
9.4 * Elliptic partial differential equations ..... 154
10 Bochner-Lebesgue and Bochner-Sobolev spaces ..... 157
10.1 The Bochner integral ..... 157
10.2 Bochner-Lebesgue spaces ..... 162
10.3 The convolution ..... 164
10.4 Bochner-Sobolev spaces ..... 168
References ..... 171
Index ..... 177

## Chapter 0 Primer on topology

It is the purpose of this introductory chapter to recall some basic facts about metric spaces, sequences in metric spaces, compact metric spaces, and continuous functions between metric spaces. Most of the material should be known, and if it is not known in the context of metric spaces, it has certainly been introduced on $\mathbb{R}^{d}$. The generalization to metric spaces should be straightforward, but it is nevertheless worthwhile to spend some time on the examples.

We also introduce some further notions from topology which may be new; see for example the definitions of density or of completion of a metric space.

### 0.1 Metric spaces

Let $M$ be a set. We call a function $d: M \times M \rightarrow \mathbb{R}_{+}$a metric or a distance on $M$ if for every $x, y, z \in M$
(i) $d(x, y)=0$ if and only if $x=y$,
(ii) $d(x, y)=d(y, x)$ (symmetry), and
(iii) $d(x, y) \leq d(x, z)+d(z, y)$ (triangle inequality).

A pair $(M, d)$ of a set $M$ and a metric $d$ on $M$ is called a metric space.
It will be convenient to write only $M$ instead of $(M, d)$ if the metric $d$ on $M$ is known from the context, and to speak of a metric space $M$.

Example 0.1. 1. Let $M \subseteq \mathbb{R}^{d}$ and

$$
d(x, y):=\sum_{i=1}^{d}\left|x_{i}-y_{i}\right|
$$

or

$$
d(x, y):=\left(\sum_{i=1}^{d}\left|x_{i}-y_{i}\right|^{2}\right)^{\frac{1}{2}}
$$

Then $(M, d)$ is a metric space. The second metric is called the Euclidean metric. Often, if the metric on $\mathbb{R}^{d}$ is not explicitly given, we mean the Euclidean metric.
2. Let $M \subseteq C([0,1])$, the space of all continuous functions on the interval $[0,1]$, and

$$
d(f, g):=\sup _{x \in[0,1]}|f(x)-g(x)| .
$$

Then $(M, d)$ is a metric space.
3. Let $M$ be any set and

$$
d(x, y):= \begin{cases}0 & \text { if } x=y \\ 1 & \text { otherwise }\end{cases}
$$

Then $(M, d)$ is a metric space. The metric $d$ is called the discrete metric.
4. Let $(M, d)$ be a metric space. Then

$$
d_{1}(x, y):=\frac{d(x, y)}{1+d(x, y)}
$$

and

$$
d_{2}(x, y):=\min \{d(x, y), 1\}
$$

define also metrics on $M$.
5. Let $M=C(\mathbb{R})$, the space of all continuous functions on $\mathbb{R}$, and let

$$
d_{n}(f, g):=\sup _{x \in[-n, n]}|f(x)-g(x)| \quad(n \in \mathbb{N})
$$

and

$$
d(f, g):=\sum_{n \in \mathbb{N}} 2^{-n} \frac{d_{n}(f, g)}{1+d_{n}(f, g)} .
$$

Then $(M, d)$ is a metric space. Note that the functions $d_{n}$ are not metrics for any $n \in \mathbb{N}$ !
6. Let $(M, d)$ be a metric space. Then any subset $\tilde{M} \subseteq M$ is a metric space for the induced metric

$$
\tilde{d}(x, y)=d(x, y), \quad x, y \in \tilde{M} .
$$

We may sometimes say that $\tilde{M}$ is a subspace of $M$, that is, a subset and a metric space, but certainly this is not to be understood in the sense of linear subspaces of vector spaces ( $M$ need not be a vector space).
7. Let $\left(M_{n}, d_{n}\right)$ be metric spaces $(n \in \mathbb{N})$. Then the cartesian product $M:=\bigotimes_{n \in \mathbb{N}} M_{n}$ is a metric space for the metric

$$
d(x, y):=\sum_{n \in \mathbb{N}} 2^{-n} \min \left\{d_{n}\left(x_{n}, y_{n}\right), 1\right\} .
$$

Clearly, in a similar way, every finite cartesian product of metric spaces is a metric space.

Let $(M, d)$ be a metric space. For every $x \in M$ and every $r>0$ we define the open ball $B(x, r):=\{y \in M: d(x, y)<r\}$ with center $x$ and radius $r$. A set $O \subseteq M$ is called open if for every $x \in O$ there exists some $r>0$ such that $B(x, r) \subseteq O$. A set $A \subseteq M$ is called closed if its complement $A^{c}=M \backslash A$ is open. A set $U \subseteq M$ is called a neighbourhood of $x \in M$ if there exists $r>0$ such that $B(x, r) \subseteq U$.

Remark 0.2. (a) The notions open, closed, neighbourhood depend on the set M!! For example, $M$ is always closed and open in $M$. The set $\mathbb{Q}$ is not closed in $\mathbb{R}$ (for the Euclidean metric), but it is closed in $\mathbb{Q}$ for the induced metric! Therefore, one should always say in which metric space some given set is open or closed.
(b) Clearly, a set $O \subseteq M$ is open (in $M$ ) if and only if it is a neighbourhood of every of its elements.

Lemma 0.3. Let $(M, d)$ be a metric space. The following are true:
a) Arbitrary unions of open sets are open. That means: if $\left(O_{i}\right)_{i \in I}$ is an arbitrary family of open sets (no restrictions on the index set I), then $\bigcup_{i \in I} O_{i}$ is open.
b) Arbitrary intersections of closed sets are closed. That means: if $\left(A_{i}\right)_{i \in I}$ is an arbitrary family of closed sets, then $\bigcap_{i \in I} A_{i}$ is closed.
c) Finite intersections of open sets are open.
d) Finite unions of closed sets are closed.

Proof. (a) Let $\left(O_{i}\right)_{i \in I}$ be an arbitrary family of open sets and let $O:=\bigcup_{i \in I} O_{i}$. If $x \in O$, then $x \in O_{i}$ for some $i \in I$, and since $O_{i}$ is open, $B(x, r) \subseteq O_{i}$ for some $r>0$. This implies that $B(x, r) \subseteq O$, and therefore $O$ is open.
(c) Next let $\left(O_{i}\right)_{i \in I}$ be a finite family of open sets and let $O:=\bigcap_{i \in I} O_{i}$. If $x \in O$, then $x \in O_{i}$ for every $i \in I$. Since the $O_{i}$ are open, there exist $r_{i}$ such that $B\left(x, r_{i}\right) \subseteq O_{i}$. Let $r:=\min _{i \in I} r_{i}$ which is positive since $I$ is finite. By construction, $B(x, r) \subseteq O_{i}$ for every $i \in I$, and therefore $B(x, r) \subseteq O$, that is, $O$ is open.

The proofs for closed sets are similar or follow just from the definition of closed sets and the above two assertions.

Exercise 0.4 Determine all open sets (respectively, all closed sets) of a metric space $(M, d)$, where $d$ is the discrete metric.

Exercise 0.5 Show that a ball $B(x, r)$ in a metric space $M$ is always open. Show also that

$$
\bar{B}(x, r):=\{y \in M: d(x, y) \leq r\}
$$

is always closed.
Let $(M, d)$ be a metric space and let $S \subseteq M$ be a subset. Then the set $\bar{S}:=\bigcap\{A: A \subseteq M$ is closed and $S \subseteq A\}$ is called the closure of $S$. The set $S^{\circ}:=\bigcup\{O: O \subseteq M$ is open and $O \subseteq S\}$ is called the interior of $S$. Finally, we call $\partial S:=\left\{x \in M: \forall \varepsilon>0 B(x, \varepsilon) \cap S \neq \emptyset\right.$ and $\left.B(x, \varepsilon) \cap S^{c} \neq \emptyset\right\}$ the boundary of $S$.

By Lemma 0.3, the closure of a set $S$ is always closed (arbitrary intersections of closed sets are closed). By definition, $\bar{S}$ is the smallest closed set which contains $S$.

Similarly, the interior of a set $S$ is always open, and by definition it is the largest open set which is contained in $S$. Note that the interior might be empty.

Exercise 0.6 Give an example of a metric space $M$ and some $x \in M, r>0$, to show that $\bar{B}(x, r)$ need not coincide with the closure of $B(x, r)$.

Exercise 0.7 Let $(M, d)$ be a metric space and consider the metrics $d_{1}$ and $d_{2}$ from Example 0.1 (4). Show that the set of all open subsets, closed subsets or neighbourhoods of $M$ is the same for the three given metrics.

The set of all open subsets is also called the topology of M. The three metrics $d$, $d_{1}$ and $d_{2}$ thus induce the same topology. Sometimes it is good to know that one can pass from a given metric $d$ to a finite metric ( $d_{1}$ and $d_{2}$ take only values between 0 and 1) without changing the topology.

### 0.2 Sequences, convergence

Throughout the following, sequences will be denoted by $\left(x_{n}\right)$. Only when it is necessary, we make precise the index $n$; usually, $n \geq 0$ or $n \geq 1$, but sometimes we will also consider finite sequences or sequences indexed by $\mathbb{Z}$.

Let $(M, d)$ be a metric space. We call a sequence $\left(x_{n}\right) \subseteq M$ a Cauchy sequence if for every $\varepsilon>0$ there exists $n_{0}$ such that for every $n, m \geq n_{0}$ one has $d\left(x_{n}, x_{m}\right)<\varepsilon$. We say that a sequence $\left(x_{n}\right) \subseteq M$ converges to some element $x \in M$ if for every $\varepsilon>0$ there exists $n_{0}$ such that for every $n \geq n_{0}$ one has $d\left(x_{n}, x\right)<\varepsilon$. If $\left(x_{n}\right)$ converges to $x$, we also write $\lim _{n \rightarrow \infty} x_{n}=x$ or $x_{n} \rightarrow x$ as $n \rightarrow \infty$.

Exercise 0.8 Let $C([0,1])$ be the metric space from Example 0.1 (2). Show that a sequence $\left(f_{n}\right) \subseteq C([0,1])$ converges to some $f$ for the metric $d$ if and only if it converges uniformly. We say that the metric $d$ induces the topology of uniform convergence.

Show also that a sequence $\left(f_{n}\right) \subseteq C(\mathbb{R})$ (Example 0.1 (5)) converges to some $f$ for the metric $d$ if and only if it converges uniformly on compact subsets of $\mathbb{R}$. In this example, we say that the metric $d$ induces the topology of local uniform convergence.

Exercise 0.9 Determine all Cauchy sequences and all convergent sequences in a discrete metric space.

Lemma 0.10. Let $M$ be a metric space and $\left(x_{n}\right) \subseteq M$ be a sequence. Then:
a) $\lim _{n \rightarrow \infty} x_{n}=x$ for some element $x \in M$ if and only if for every neighbourhood $U$ of $x$ there exists $n_{0}$ such that for every $n \geq n_{0}$ one has $x_{n} \in U$.
b) (Uniqueness of the limit) If $\lim _{n \rightarrow \infty} x_{n}=x$ and $\lim _{n \rightarrow \infty} x_{n}=y$, then $x=y$.

Lemma 0.11. A set $A \subseteq M$ is closed if and only it is sequentially closed, that is, if for every sequence $\left(x_{n}\right) \subseteq A$ which converges to some $x \in M$ one has $x \in A$.

Proof. Assume first that $A$ is closed and let $\left(x_{n}\right) \subseteq A$ be convergent to $x \in M$. If $x$ does not belong to $A$, then it belongs to $A^{c}$ which is open. By definition, there exists $\varepsilon>0$ such that $B(x, \varepsilon) \subseteq A^{c}$. Given this $\varepsilon$, there exists $n_{0}$ such that $x_{n} \in B(x, \varepsilon)$ for every $n \geq n_{0}$, a contradiction to the assumption that $x_{n} \in A$. Hence, $x \in A$.

On the other hand, assume that $\lim _{n \rightarrow \infty} x_{n}=x \in A$ for every convergent $\left(x_{n}\right) \subseteq A$ and assume in addition that $A$ is not closed or, equivalently, that $A^{c}$ is not open. Then there exists $x \in A^{c}$ such that for every $n \in \mathbb{N}$ the set $B\left(x, \frac{1}{n}\right) \cap A$ is nonempty. From this one can construct a sequence $\left(x_{n}\right) \subseteq A$ which converges to $x$, which is a contradiction because $x \in A^{c}$.

Lemma 0.12. Let $(M, d)$ be a metric space, and let $S \subseteq M$ be a subset. Then

$$
\begin{aligned}
\bar{S} & =\left\{x \in M: \exists\left(x_{n}\right) \subseteq S \text { s.t. } \lim _{n \rightarrow \infty} x_{n}=x\right\} \\
& =\left\{x \in M: d(x, S):=\inf _{y \in S} d(x, y)=0\right\} .
\end{aligned}
$$

Proof. Let

$$
A:=\left\{x \in M: \exists\left(x_{n}\right) \subseteq S \text { s.t. } \lim _{n \rightarrow \infty} x_{n}=x\right\}
$$

and

$$
B:=\left\{x \in M: d(x, S):=\inf _{y \in S} d(x, y)=0\right\} .
$$

These two sets are clearly equal by the definition of the inf and the definition of convergence. Moreover, the set $B$ is closed by the following argument. Assume that $\left(x_{n}\right) \subseteq B$ is convergent to $x \in M$. By definition of $B$, for every $n$ there exists $y \in S$ such that $d\left(x_{n}, y_{n}\right) \leq 1 / n$. Hence,

$$
\limsup _{n \rightarrow \infty} d\left(x, y_{n}\right) \leq \limsup _{n \rightarrow \infty} d\left(x, x_{n}\right)+\underset{n \rightarrow \infty}{\limsup } d\left(x_{n}, y_{n}\right)=0,
$$

so that $x \in B$.
Clearly, $B$ contains $S$, and since $B$ is closed, $B$ contains $\bar{S}$. It remains to show that $B \subseteq \bar{S}$. If this is not true, then there exists $x \in B \backslash \bar{S}$. Since the complement of $\bar{S}$ is open in $M$, there exists $r>0$ such that $B(x, r) \cap \bar{S}=\emptyset$, a contradiction to the definition of $B$.

A metric space $(M, d)$ is called complete if every Cauchy sequence converges.
Exercise 0.13 Show that the spaces $\mathbb{R}^{d}, C([0,1])$ and $C(\mathbb{R})$ are complete. Show also that any discrete metric space is complete.

Lemma 0.14. A subspace $N \subseteq M$ of a complete metric space is complete if and only if it is closed in $M$.

Proof. Assume that $N \subseteq M$ is closed, and let $\left(x_{n}\right)$ be a Cauchy sequence in $N$. By the assumption that $M$ is complete, $\left(x_{n}\right)$ is convergent to some element $x \in M$. Since $N$ is closed, $x \in N$.

Assume on the other hand that $N$ is complete, and let $\left(x_{n}\right) \subseteq N$ be convergent to some element $x \in M$. Clearly, every convergent sequence is also a Cauchy sequence,
and since $N$ is complete, $\left(x_{n}\right)$ converges to some element $y \in N$. By uniqueness of the limit, $x=y \in N$. Hence, $N$ is closed.

### 0.3 Compact spaces

We say that a metric space $(M, d)$ is compact if for every open covering there exists a finite subcovering, that is, whenever $\left(O_{i}\right)_{i \in I}$ is a family of open sets (no restrictions on the index set $I$ ) such that $M=\bigcup_{i \in I} O_{i}$, then there exists a finite subset $I_{0} \subseteq I$ such that $M=\bigcup_{i \in I_{0}} O_{i}$.

Lemma 0.15. A metric space $(M, d)$ is compact if and only if it is sequentially compact, that is, if and only if every sequence $\left(x_{n}\right) \subseteq M$ has a convergent subsequence.

Proof. Assume that $M$ is compact and let $\left(x_{n}\right) \subseteq M$. Assume that $\left(x_{n}\right)$ does not have a convergent subsequence. Then for every $x \in M$ there exists $\varepsilon_{x}>0$ such that $B\left(x, \boldsymbol{\varepsilon}_{x}\right)$ contains only finitely many elements of $\left\{x_{n}\right\}$. Note that $\left(B\left(x, \boldsymbol{\varepsilon}_{x}\right)\right)_{x \in M}$ is an open covering of $M$ so that by the compactness of $M$ there exists a finite subset $N \subseteq$ $M$ such that $M=\bigcup_{x \in N} B\left(x, \varepsilon_{x}\right)$. But this means that $\left(x_{n}\right)$ takes only finitely many values, and hence there exists even a constant subsequence which is in particular also convergent; a contradiction to the assumption on $\left(x_{n}\right)$.

On the other hand, assume that $M$ is sequentially compact and let $\left(O_{i}\right)_{i \in I}$ be an open covering of $M$. We first show that there exists $\varepsilon>0$ such that for every $x \in M$ there exists $i_{x} \in I$ with $B(x, \varepsilon) \subseteq O_{i_{x}}$. If this were not true, then for every $n \in \mathbb{N}$ there exists $x_{n}$ such that $B\left(x_{n}, \frac{1}{n}\right) \nsubseteq O_{i}$ for every $i \in I$. Passing to a subsequence, we may assume that $\left(x_{n}\right)$ is convergent to some $x \in M$. There exists some $i_{0} \in I$ such that $x \in O_{i_{0}}$, and since $O_{i_{0}}$ is open, we find some $\varepsilon>0$ such that $B(x, \varepsilon) \subseteq O_{i_{0}}$. Let $n_{0}$ be such that $\frac{1}{n_{0}}<\frac{\varepsilon}{2}$. By the triangle inequality, for every $n \geq n_{0}$ we have $B\left(x_{n}, \frac{1}{n}\right) \subseteq B(x, \varepsilon) \subseteq O_{i_{0}}$, a contradiction to the construction of the sequence $\left(x_{n}\right)$.

Next we show that $M=\bigcup_{j=1}^{n} B\left(x_{j}, \varepsilon\right)$ for a finite family of $x_{j} \in M$. Choose any $x_{1} \in M$. If $B\left(x_{1}, \varepsilon\right)=M$, then we are already done. Otherwise we find $x_{2} \in M \backslash$ $B\left(x_{1}, \varepsilon\right)$. If $B\left(x_{1}, \varepsilon\right) \cup B\left(x_{2}, \varepsilon\right) \neq M$, then we even find $x_{3} \in M$ which does not belong to $B\left(x_{1}, \varepsilon\right) \cup B\left(x_{2}, \varepsilon\right)$, and so on. If $\bigcup_{j=1}^{n} B\left(x_{j}, \varepsilon\right)$ is never all of $M$, then we find actually a sequence $\left(x_{j}\right)$ such that $d\left(x_{j}, x_{k}\right) \geq \varepsilon$ for all $j \neq k$. This sequence can not have a convergent subsequence, a contradiction to sequential compactness.

Since every of the $B\left(x_{j}, \varepsilon\right)$ is a subset of $O_{i_{x_{j}}}$ for some $i_{x_{j}} \in I$, we have proved that $M=\bigcup_{j=1}^{n} O_{i_{x_{j}}}$, i.e. the open covering $\left(O_{i}\right)$ admits a finite subcovering. The proof is complete.

Lemma 0.16. Any compact metric space is complete.
Proof. Let $\left(x_{n}\right)$ be a Cauchy sequence in $M$. By the preceeding lemma, there exists a subsequence which converges to some $x \in M$. If a subsequence of a Cauchy sequence converges, then the sequence itself converges, too.

### 0.4 Continuity

Let $\left(M_{1}, d_{1}\right),\left(M_{2}, d_{2}\right)$ be two metric spaces, and let $f: M_{1} \rightarrow M_{2}$ be a function. We say that $f$ is continuous at some point $x \in M_{1}$ if

$$
\forall \varepsilon>0 \exists \delta>0 \forall y \in B(x, \delta): d_{2}(f(x), f(y))<\varepsilon .
$$

We say that $f$ is continuous if it is continuous at every point. We say that $f$ is uniformly continuous if

$$
\forall \varepsilon>0 \exists \delta>0 \forall x, y \in M_{1}: d_{1}(x, y)<\delta \Rightarrow d_{2}(f(x), f(y))<\varepsilon .
$$

We say that $f$ is Lipschitz continuous if

$$
\exists L \geq 0 \forall x, y \in M_{1}: d_{2}(f(x), f(y)) \leq L d_{1}(x, y) .
$$

Lemma 0.17. A function $f: M_{1} \rightarrow M_{2}$ between two metric spaces is continuous at some point $x \in M_{1}$ if and only if it is sequentially continuous at $x$, that is, if and only if for every sequence $\left(x_{n}\right) \subseteq M_{1}$ which converges to $x$ one has $\lim _{n \rightarrow \infty} f\left(x_{n}\right)=f(x)$.

Proof. Assume that $f$ is continuous at $x \in M_{1}$ and let $\left(x_{n}\right)$ be convergent to $x$. Let $\varepsilon>0$. There exists $\delta>0$ such that for every $y \in B(x, \delta)$ one has $f(y) \in B(f(x), \varepsilon)$. By definition of convergence, there exists $n_{0}$ such that for every $n \geq n_{0}$ one has $x_{n} \in B(x, \delta)$. For this $n_{0}$ and every $n \geq n_{0}$ one has $f\left(x_{n}\right) \in B(f(x), \varepsilon)$. Hence, $\lim _{n \rightarrow \infty} f\left(x_{n}\right)=f(x)$.

Assume on the other hand that $f$ is sequentially continuous at $x$. If $f$ was not continuous in $x$ then there exists $\varepsilon>0$ such that for every $n \in \mathbb{N}$ there exists $x_{n} \in B\left(x, \frac{1}{n}\right)$ with $f\left(x_{n}\right) \notin B(f(x), \varepsilon)$. By construction, $\lim _{n \rightarrow \infty} x_{n}=x$. Since $f$ is sequentially continuous, $\lim _{n \rightarrow \infty} f\left(x_{n}\right)=f(x)$. But this is a contradiction to $f\left(x_{n}\right) \notin B(f(x), \varepsilon)$, and therefore $f$ is continuous.

Lemma 0.18. A function $f: M_{1} \rightarrow M_{2}$ between two metric spaces is continuous if and only if preimages of open sets are open, that is, if and only if for every open set $O \subseteq M_{2}$ the preimage $f^{-1}(O)$ is open in $M_{1}$.

Proof. Let $f: M_{1} \rightarrow M_{2}$ be continuous and let $O \subseteq M_{2}$ be open. Let $x \in f^{-1}(O)$. Since $O$ is open, there exists $\varepsilon>0$ such that $B(f(x), \varepsilon) \subseteq O$. Since $f$ is continuous, there exists $\delta>0$ such that for every $y \in B(x, \delta)$ one has $f(y) \in B(f(x), \varepsilon)$. Hence, $B(x, \delta) \subseteq f^{-1}(O)$ so that $f^{-1}(O)$ is open.

On the other hand, if the preimage of every open set is open, then for every $x \in M_{1}$ and every $\varepsilon>0$ the preimage $f^{-1}(B(f(x), \varepsilon))$ is open. Clearly, $x$ belongs to this preimage, and therefore there exists $\delta>0$ such that $B(x, \delta) \subseteq f^{-1}(B(f(x), \varepsilon))$. This proves continuity.

Lemma 0.19. Let $f: K \rightarrow M$ be a continuous function from a compact metric space $K$ into a metric space $M$. Then:
a) The image $f(K)$ is compact.
b) The function $f$ is uniformly continuous.

Proof. (a) Let $\left(O_{i}\right)_{i \in I}$ be an open covering of $f(K)$. Since $f$ is continuous, $f^{-1}\left(O_{i}\right)$ is open in $K$. Moreover, $\left(f^{-1}\left(O_{i}\right)\right)_{i \in I}$ is an open covering of $K$. Since $K$ is compact, there exists a finite subcovering: $K=\bigcup_{i \in I_{0}} f^{-1}\left(O_{i}\right)$ for some finite $I_{0} \subseteq I$. Hence, $\left(O_{i}\right)_{i \in I_{0}}$ is a finite subcovering of $f(K)$.
(b) Let $\varepsilon>0$. Since $f$ is continuous, for every $x \in K$ there exists $\delta_{x}>0$ such that for all $y \in B\left(x, \delta_{x}\right)$ one has $f(y) \in B(f(x), \varepsilon)$. By compactness, there exists a finite family $\left(x_{i}\right)_{1 \leq i \leq n} \subseteq K$ such that $K=\bigcup_{i=1}^{n} B\left(x_{i}, \delta_{x_{i}} / 2\right)$. Let $\delta=\min \left\{\delta_{x_{i}} / 2: 1 \leq i \leq n\right\}$ and let $x, y \in K$ such that $d(x, y)<\delta$. Since $x \in B\left(x_{i}, \delta_{x_{i}} / 2\right)$ for some $1 \leq i \leq n$, we find that $y \in B\left(x_{i}, \delta_{x_{i}}\right)$. By construction, $f(x), f(y) \in B\left(f\left(x_{i}\right), \varepsilon\right)$ so that the triangle inequality implies $d(f(x), f(y))<2 \varepsilon$.

Lemma 0.20. Any Lipschitz continuous function $f: M_{1} \rightarrow M_{2}$ between two metric spaces is uniformly continuous.

Proof. Let $L>0$ be a Lipschitz constant for $f$ and let $\varepsilon>0$. Define $\delta:=\varepsilon / L$. Then, for every $x, y \in M$ such that $d_{1}(x, y) \leq \delta$ one has

$$
d_{2}(f(x), f(y)) \leq L d_{1}(x, y) \leq \varepsilon
$$

and therefore $f$ is uniformly continuous.

### 0.5 Completion of a metric space

We say that a subset $D \subseteq M$ of a metric space $(M, d)$ is dense in $M$ if $\bar{D}=M$. Equivalently, $D$ is dense in $M$ if for every $x \in M$ there exists $\left(x_{n}\right) \subseteq D$ such that $\lim _{n \rightarrow \infty} x_{n}=x$.

Lemma 0.21 (Completion). Let $(M, d)$ be a metric space. Then there exists a complete metric space $(\hat{M}, \hat{d})$ and a continuous, injective $j: M \rightarrow \hat{M}$ such that

$$
d(x, y)=\hat{d}(j(x), j(y)), \quad x, y \in M
$$

and such that the image $j(M)$ is dense in $\hat{M}$.
Let $(M, d)$ be a metric space. A complete metric space $(\hat{M}, \hat{d})$ fulfilling the properties from Lemma 0.21 is called a completion of $M$.

Proof (Proof of Lemma 0.21). Let

$$
\bar{M}:=\left\{\left(x_{n}\right) \subseteq M:\left(x_{n}\right) \text { is a Cauchy sequence }\right\} .
$$

We say that two Cauchy sequences $\left(x_{n}\right),\left(y_{n}\right) \subseteq \bar{M}$ are equivalent (and we write $\left.\left(x_{n}\right) \sim\left(y_{n}\right)\right)$ if $\lim _{n \rightarrow \infty} d\left(x_{n}, y_{n}\right)=0$. Clearly, $\sim$ is an equivalence relation on $\bar{M}$.

We denote by $\left[\left(x_{n}\right)\right]$ the equivalence class in $\bar{M}$ of a Cauchy sequence $\left(x_{n}\right)$, and we let

$$
\hat{M}:=\bar{M} / \sim=\left\{\left[\left(x_{n}\right)\right]:\left(x_{n}\right) \in \bar{M}\right\}
$$

be the set of all equivalence classes. If we define

$$
\hat{d}\left(\left[\left(x_{n}\right)\right],\left[\left(y_{n}\right)\right]\right):=\lim _{n \rightarrow \infty} d\left(x_{n}, y_{n}\right)
$$

then $\hat{d}$ is well defined (the definition is independent of the choice of representatives) and it is a metric on $\hat{M}$. The fact that $\hat{d}$ is a metric and also that $(\hat{M}, \hat{d})$ is a complete metric space are left as exercises.

One also easily verifies that $j: M \rightarrow \hat{M}$ defined by $j(x)=[(x)]$ (the equivalence class of the constant sequence $(x)$ ) is continuous, injective and in fact isometric, i.e.

$$
d(x, y)=\hat{d}(j(x), j(y))
$$

for every $x, y \in M$. The proof is here complete.
Lemma 0.22. Let $\left(\hat{M}_{i}, \hat{d_{i}}\right)(i=1,2)$ be two completions of a metric space $(M, d)$. Then there exists a bijection $b: \hat{M}_{1} \rightarrow \hat{M}_{2}$ such that for every $x, y \in \hat{M}_{1}$

$$
\hat{d}_{1}(x, y)=\hat{d}_{2}(b(x), b(y)) .
$$

Lemma 0.22 shows that up to isometric bijections there exists only one completion of a given metric space and it allows us to speak of the completion of a metric space.

Lemma 0.23. Let $f: M_{1} \rightarrow M_{2}$ be a uniformly (!) continuous function between two metric spaces. Let $\hat{M}_{1}$ and $\hat{M}_{2}$ be the completions of $M_{1}$ and $M_{2}$, respectively. Then there exists a unique continuous extension $\hat{f}: \hat{M}_{1} \rightarrow \hat{M}_{2}$ of $f$.

Proof. Since $f$ is uniformly continuous, it maps equivalent Cauchy sequences into equivalent Cauchy sequences (equivalence of Cauchy sequences is defined as in the proof of Lemma 0.21$)$. Hence, the function $\hat{f}\left(\left[\left(x_{n}\right)\right]\right):=\left[\left(f\left(x_{n}\right)\right)\right]$ is well defined. It is easy to check that $\hat{f}$ is an extension of $f$ and that $\hat{f}$ is continuous (even uniformly continuous).

The assumption of uniform continuity in Lemma 0.23 is necessary in general. The functions $f(x)=\sin (1 / x)$ and $f(x)=1 / x$ on the open interval $(0,1)$ do not admit continuous extensions to the closed interval $[0,1]$ (which is the completion of $(0,1)$ ).

## Chapter 1 <br> Banach spaces and bounded linear operators

Throughout, let $\mathbb{K} \in\{\mathbb{R}, \mathbb{C}\}$.

### 1.1 Normed spaces

Let $X$ be a vector space over $\mathbb{K}$. A function $\|\cdot\|: X \rightarrow \mathbb{R}_{+}$is called a norm if for every $x, y \in X$ and every $\lambda \in \mathbb{K}$
(i) $\|x\|=0$ if and only if $x=0$,
(ii) $\|\lambda x\|=|\lambda|\|x\|$, and
(iii) $\|x+y\| \leq\|x\|+\|y\|$ (triangle inequality).

A pair $(X,\|\cdot\|)$ of a vector space $X$ and a norm $\|\cdot\|$ is called a normed space.
Often, we will speak of a normed space $X$ if it is clear which norm is given on $X$.
Example 1.1. 1. (Finite dimensional spaces) Let $X=\mathbb{K}^{d}$. Then

$$
\|x\|_{p}:=\left(\sum_{i=1}^{d}\left|x_{i}\right|^{p}\right)^{1 / p}, \quad 1 \leq p<\infty
$$

and

$$
\|x\|_{\infty}:=\sup _{1 \leq i \leq d}\left|x_{i}\right|
$$

are norms on $X$.
2. (Sequence spaces) Let $1 \leq p<\infty$, and let

$$
l^{p}:=\left\{\left(x_{n}\right) \subseteq \mathbb{K}: \sum_{n}\left|x_{n}\right|^{p}<\infty\right\}
$$

with norm

$$
\|x\|_{p}:=\left(\sum_{n}\left|x_{n}\right|^{p}\right)^{1 / p}
$$

Then $\left(l^{p},\|\cdot\|_{p}\right)$ is a normed space.
3. (Sequence spaces) Let $X$ be one of the spaces

$$
\begin{aligned}
l^{\infty} & :=\left\{\left(x_{n}\right) \subseteq \mathbb{K}: \sup _{n}\left|x_{n}\right|<\infty\right\}, \\
c & :=\left\{\left(x_{n}\right) \subseteq \mathbb{K}: \lim _{n \rightarrow \infty} x_{n} \text { exists }\right\}, \text { or } \\
c_{0} & :=\left\{\left(x_{n}\right) \subseteq \mathbb{K}: \lim _{n \rightarrow \infty} x_{n}=0\right\}, \text { or } \\
c_{00} & :=\left\{\left(x_{n}\right) \subseteq \mathbb{K}: \text { the set }\left\{n: x_{n} \neq 0\right\} \text { is finite }\right\},
\end{aligned}
$$

and let

$$
\|x\|_{\infty}:=\sup _{n}\left|x_{n}\right| .
$$

Then $\left(X,\|\cdot\|_{\infty}\right)$ is a normed space.
4. (Function spaces: continuous functions) Let $C([a, b])$ be the space of all continuous, $\mathbb{K}$-valued functions on a compact interval $[a, b] \subset \mathbb{R}$. Then

$$
\|f\|_{p}:=\left(\int_{a}^{b}|f(x)|^{p} \mathrm{~d} x\right)^{1 / p}, \quad 1 \leq p<\infty
$$

and

$$
\|f\|_{\infty}:=\sup _{x \in[a, b]}|f(x)|
$$

are norms on $C([a, b])$.
5. (Function spaces: continuous functions) Let $K$ be a compact metric space and let $C(K)$ be the space of all continuous, $\mathbb{K}$-valued functions on $K$. Then

$$
\|f\|_{\infty}:=\sup _{x \in K}|f(x)|
$$

is a norm on $C(K)$.
6. (Function spaces: integrable functions) Let $(\Omega, \mathscr{A}, \mu)$ be a measure space and let $X_{p}=L^{p}(\Omega)(1 \leq p \leq \infty)$. Let

$$
\|f\|_{p}:=\left(\int_{\Omega}|f|^{p} \mathrm{~d} \mu\right)^{1 / p}, \quad 1 \leq p<\infty
$$

or

$$
\|f\|_{\infty}:=\operatorname{ess} \sup |f(x)|:=\inf \left\{c \in \mathbb{R}_{+}: \mu(\{|f|>c\})=0\right\}
$$

Then $\left(X_{p},\|\cdot\|_{p}\right)$ is a normed space.
7. (Function spaces: differentiable functions) Let

$$
C^{1}([a, b]):=\{f \in C([a, b]): f \text { is continuously differentiable }\} .
$$

Then $\|\cdot\|_{\infty}$ and

$$
\|f\|_{C^{1}}:=\|f\|_{\infty}+\left\|f^{\prime}\right\|_{\infty}
$$

are norms on $C^{1}([a, b])$.
We will see more examples in the sequel.
Lemma 1.2. Every normed space $(X,\|\cdot\|)$ is a metric space for the metric

$$
d(x, y):=\|x-y\|, \quad x, y \in X
$$

By the above lemma, also every subset of a normed space becomes a metric space in a natural way. Moreover, it is natural to speak of closed or open subsets (or linear subspaces!) of normed spaces, or of closures and interiors of subsets.

Exercise 1.3 Show that in a normed space $X$, for every $x \in X$ and every $r>0$ the closed ball $\bar{B}(x, r)$ coincides with closure $\bar{B}(x, r)$ of the open ball.

Also the notion of continuity of functions between normed spaces (or between a metric space and a normed space) makes sense. The following is a first example of a continuous function.

Lemma 1.4. Given a normed space, the norm is a continuous function.
This lemma is a consequence of the following lemma.
Lemma 1.5 (Triangle inequality from below). Let $X$ be a normed space. Then, for every $x, y \in X$,

$$
\|x-y\| \geq|\|x\|-\|y\|| .
$$

Proof. The triangle inequality implies

$$
\begin{aligned}
\|x\| & =\|x-y+y\| \\
& \leq\|x-y\|+\|y\|,
\end{aligned}
$$

so that

$$
\|x\|-\|y\| \leq\|x-y\| .
$$

Changing the role of $x$ and $y$ implies

$$
\|y\|-\|x\| \leq\|y-x\|=\|x-y\|,
$$

and the claim follows.
A notion which can not really be defined in metric spaces but in normed spaces is the following. A subset $B$ of a normed space $X$ is called bounded if

$$
\sup \{\|x\|: x \in B\}<\infty .
$$

It is easy to check that if $X$ is a normed space, and $M$ is a metric space, then the set $C(M ; X)$ of all continuous functions from $M$ into $X$ is a vector space for the obvious
addition and scalar multiplication. If $M$ is in addition compact, then $f(M) \subseteq X$ is also compact for every such function, and hence $f(M)$ is necessarily bounded (every compact subset of a normed space is bounded!). So we can give a new example of a normed space.

Example 1.6. 8. (Function spaces: vector-valued continuous functions) Let ( $X, \| \cdot$ $\|)$ be a normed space and let $K$ be a compact metric space. Let $E=C(K ; X)$ be the space of all $X$-valued continuous functions on $K$. Then

$$
\|f\|_{\infty}:=\sup _{x \in K}\|f(x)\|
$$

is a norm on $C(K ; X)$.
Also the notions of Cauchy sequences and convergent sequences make sense in normed spaces. In particular, one can speak of a complete normed space, that is, a normed space in which every Cauchy sequence converges. A complete normed space is called a Banach space.

Example 1.7. The finite dimensional spaces, the sequence spaces $l^{p}(1 \leq p \leq \infty), c$, and $c_{0}$, and the function spaces $\left(C([a, b]),\|\cdot\|_{\infty}\right),\left(L^{p}(\Omega),\|\cdot\|_{p}\right)$ are Banach spaces.

The spaces $\left(c_{00},\|\cdot\|_{\infty}\right),\left(C([a, b]),\|\cdot\|_{p}\right)(1 \leq p<\infty)$ are not Banach spaces.
If $X$ is a Banach space, then also $\left(C(K ; X),\|\cdot\|_{\infty}\right)$ is a Banach space.
We say that two norms $\|\cdot\|_{1}$ and $\|\cdot\|_{2}$ on a real or complex vector space $X$ are equivalent if there exist two constants $c, C>0$ such that for every $x \in X$

$$
c\|x\|_{1} \leq\|x\|_{2} \leq C\|x\|_{1} .
$$

Lemma 1.8. Let $\|\cdot\|_{1},\|\cdot\|_{2}$ be two norms on a vector space $X$ (over $\mathbb{K}$ ). The following are equivalent:
(i) The norms $\|\cdot\|_{1},\|\cdot\|_{2}$ are equivalent.
(ii) A set $O \subseteq X$ is open for the norm $\|\cdot\|_{1}$ if and only if it is open for the norm $\|\cdot\|_{2}$ (and similarly for closed sets).
(iii) A sequence $\left(x_{n}\right) \subseteq X$ converges to 0 for the norm $\|\cdot\|_{1}$ if and only if it converges to 0 for the norm $\|\cdot\|_{2}$.

In other words, if two norms $\|\cdot\|_{1},\|\cdot\|_{2}$ on a vector space $X$ are equivalent, then the open sets, the closed sets and the null sequences are the same. We also say that the two norms define the same topology. In particular, if $X$ is a Banach space for one norm then it is also a Banach space for the other (equivalent) norm.

Exercise 1.9 The norms $\|\cdot\|_{\infty}$ and $\|\cdot\|_{p}$ are not equivalent on $C([0,1])$.
Theorem 1.10. Any two norms on a finite dimensional real or complex vector space are equivalent.

Proof. We may without loss of generality consider $\mathbb{K}^{d}$. Let $\|\cdot\|$ be a norm on $\mathbb{K}^{d}$ and let $\left(e_{i}\right)_{1 \leq i \leq d}$ be the canonical basis of $\mathbb{K}^{d}$. For every $x \in \mathbb{K}^{d}$

$$
\begin{aligned}
\|x\| & =\left\|\sum_{i=1}^{d} x_{i} e_{i}\right\| \\
& \leq \sum_{i=1}^{d}\left|x_{i}\right|\left\|e_{i}\right\| \\
& \leq C\|x\|_{1},
\end{aligned}
$$

where $C:=\sup _{1 \leq i \leq d}\left\|e_{i}\right\|<\infty$ and $\|\cdot\|_{1}$ is the norm from Example 1.1.1. By the triangle inequality from below, for every $x, y \in \mathbb{K}^{d}$,

$$
|\|x\|-\|y\|| \leq\|x-y\| \leq C\|x-y\|_{1} .
$$

Hence, the norm $\|\cdot\|:\left(\mathbb{K}^{d},\|\cdot\|_{1}\right) \rightarrow \mathbb{R}_{+}$is continuous (on $\mathbb{K}^{d}$ equipped with the norm $\|\cdot\|_{1}$ ). If $S:=\left\{x \in \mathbb{K}^{d}:\|x\|_{1}=1\right\}$ denotes the unit sphere for the norm $\|\cdot\|_{1}$, then $S$ is compact. As a consequence

$$
c:=\inf \{\|x\|: x \in S\}>0
$$

since the infimum is attained by the continuity of $\|\cdot\|$. This implies

$$
c\|x\|_{1} \leq\|x\| \quad \text { for every } x \in \mathbb{K}^{d} .
$$

We have proved that every norm on $\mathbb{K}^{d}$ is equivalent to the norm $\|\cdot\|_{1}$. Hence, any two norms on $\mathbb{K}^{d}$ are equivalent.

Corollary 1.11. Any finite dimensional normed space is complete. Any finite dimensional subspace of a normed space is closed.

Proof. The space $\left(\mathbb{K}^{d},\|\cdot\|_{1}\right)$ is complete (exercise!). If $\|\cdot\|$ is a second norm on $\mathbb{K}^{d}$ and if $\left(x_{n}\right)$ is a Cauchy sequence for that norm, then it is also a Cauchy sequence in ( $\mathbb{K}^{d},\|\cdot\|_{1}$ ) (use that the norms $\|\cdot\|_{1}$ and $\|\cdot\|$ are equivalent), and therefore convergent in $\left(\mathbb{K}^{d},\|\cdot\|_{1}\right)$. By equivalence of norms again, the sequence $\left(x_{n}\right)$ is also convergent in $\left(\mathbb{K}^{d},\|\cdot\|\right)$, and therefore $\left(\mathbb{K}^{d},\|\cdot\|\right)$ is complete.

Let $Y$ be a finite dimensional subspace of a normed space $X$, and let $\left(x_{n}\right) \subseteq Y$ be a convergent sequence with $x=\lim _{n \rightarrow \infty} x_{n} \in X$. Since $\left(x_{n}\right)$ is also a Cauchy sequence, and since $Y$ is complete, we find (by uniqueness of the limit) that $x \in Y$, and therefore $Y$ is closed (Lemma 0.11).

Let $\left(x_{n}\right)$ be a sequence in a normed space $X$. We say that the series $\sum_{n} x_{n}$ is convergent if the sequence $\left(\sum_{j \leq n} x_{j}\right)$ of partial sums is convergent. We say that the series $\sum_{n} x_{n}$ is absolutely convergent if $\sum_{n}\left\|x_{n}\right\|<\infty$.

Lemma 1.12. Let $\left(x_{n}\right)$ be a sequence in a normed space $X$. If the series $\sum_{n} x_{n}$ is convergent, then necessarily $\lim _{n \rightarrow \infty} x_{n}=0$.

Note that in a normed space not every absolutely convergent series is convergent. In fact, the following is true.

Lemma 1.13. A normed space $X$ is a Banach space if and only if every absolutely convergent series converges.

Proof. Assume that $X$ is a Banach space, and let $\sum_{n} x_{n}$ be absolutely convergent. It follows easily from the triangle inequality that the corresponding sequence of partial sums is a Cauchy sequence, and since $X$ is complete, the series $\sum_{n} x_{n}$ is convergent.

On the other hand, assume that every absolutely convergent series is convergent. Let $\left(x_{n}\right)_{n \geq 1} \subseteq X$ be a Cauchy sequence. From this Cauchy sequence, one can extract a subsequence $\left(x_{n_{k}}\right)_{k \geq 1}$ such that $\left\|x_{n_{k+1}}-x_{n_{k}}\right\| \leq 2^{-k}, k \geq 1$. Let $y_{0}=x_{n_{1}}$ and $y_{k}=x_{n_{k+1}}-x_{n_{k}}, k \geq 1$. Then the series $\sum_{k \geq 0} y_{k}$ is absolutely convergent. By assumption, it is also convergent. But by construction, $\left(\sum_{l=0}^{k} y_{l}\right)=\left(x_{n_{k}}\right)$, so that $\left(x_{n_{k}}\right)$ is convergent. Hence, we have extracted a subsequence of the Cauchy sequence $\left(x_{n}\right)$ which converges. As a consequence, $\left(x_{n}\right)$ is convergent, and since $\left(x_{n}\right)$ was an arbitrary Cauchy sequence, $X$ is complete.

Lemma 1.14 (Riesz). Let $X$ be a normed space and let $Y \subseteq X$ be a closed linear subspace. If $Y \neq X$, then for every $\delta>0$ there exists $x \in X \backslash Y$ such that $\|x\|=1$ and

$$
\operatorname{dist}(x, Y)=\inf \{\|x-y\|: y \in Y\} \geq 1-\delta
$$

Proof. Let $z \in X \backslash Y$. Since $Y$ is closed,

$$
d:=\operatorname{dist}(z, Y)>0
$$

Let $\delta>0$. By definition of the infimum, there exists $y \in Y$ such that

$$
\|z-y\| \leq \frac{d}{1-\delta}
$$

Let $x:=\frac{z-y}{\|z-y\|}$. Then $x \in X \backslash Y,\|x\|=1$, and for every $u \in Y$

$$
\begin{aligned}
\|x-u\| & =\|z-y\|^{-1}\|z-(y+\|z-y\| u)\| \\
& \geq\|z-y\|^{-1} d \geq 1-\delta,
\end{aligned}
$$

since $(y+\|z-y\| u) \in Y$.
Theorem 1.15. A normed space is finite dimensional if and only if every closed bounded set is compact.

Proof. If the normed space is finite dimensional, then every closed bounded set is compact by the Theorem of Heine-Borel. Note that by Theorem 1.10 it is not important which norm on the finite dimensional space is considered. By Lemma 1.8 , the closed and bounded sets do not change.

On the other hand, if the normed space is infinite dimensional, then, by the Lemma of Riesz, one can construct inductively a sequence $\left(x_{n}\right) \subseteq X$ such that
$\left\|x_{n}\right\|=1$ and $\operatorname{dist}\left(x_{n+1}, X_{n}\right) \geq \frac{1}{2}$ for every $n \in \mathbb{N}$, where $X_{n}=\operatorname{span}\left\{x_{i}: 1 \leq i \leq n\right\}$ (note that $X_{n}$ is closed by Corollary 1.11). By construction, $\left(x_{n}\right)$ belongs to the closed unit ball, but it can not have a convergent subsequence (even not a Cauchy subsequence). Hence, the closed unit ball is not compact. We state this result separately.

Theorem 1.16. In an infinite dimensional Banach space the closed unit ball is not compact.

Lemma 1.17 (Completion of a normed space). For every normed space $X$ there exists a Banach space $\hat{X}$ and a linear injective $j: X \rightarrow \hat{X}$ such that $\|j(x)\|=\|x\|$ $(x \in X)$ and $j(X)$ is dense in $\hat{X}$. Up to isometry, the Banach space $\hat{X}$ is unique (up to isomorphism). It is called the completion of $X$.

Proof. It suffices to repeat the proof of Lemma 0.21 and to note that the completion $\hat{X}$ of $X$ (considered as a metric space) carries in a natural way a linear structure: addition of - equivalence classes of - Cauchy sequences is their componentwise addition, and also multiplication of - an equivalence class - of a Cauchy sequence and a scalar is done componentwise. Moreover, for every [ $\left.\left(x_{n}\right)\right]$, one defines the norm

$$
\left\|\left[\left(x_{n}\right)\right]\right\|:=\lim _{n \rightarrow \infty}\left\|x_{n}\right\| .
$$

Uniqueness of $\hat{X}$ follows from Lemma 0.22 .

### 1.2 Product spaces and quotient spaces

Lemma 1.18 (Product spaces). Let $\left(X_{i}\right)_{i \in I}$ be a finite (!) family of normed spaces, and let $\mathscr{X}:=\bigotimes_{i \in I} X_{i}$ be the cartesian product. Then

$$
\|x\|_{p}:=\left(\sum_{i \in I}\left\|x_{i}\right\|_{X_{i}}^{p}\right)^{1 / p} \quad(1 \leq p<\infty)
$$

and

$$
\|x\|_{\infty}:=\sup _{i \in I}\left\|x_{i}\right\|_{X_{i}}
$$

define equivalent norms on $\mathscr{X}$. In particular, the cartesian product is a normed space.

Proof. The easy proof is left to the reader.
Lemma 1.19. Let $\left(X_{i}\right)_{i \in I}$ be a finite family of normed spaces, and let $\mathscr{X}:=\bigotimes_{i \in I} X_{i}$ be the cartesian product equipped with one of the equivalent norms $\|\cdot\|_{p}$ from Lemma 1.18. Then a sequence $\left(x^{n}\right)=\left(\left(x_{i}^{n}\right)_{i}\right) \subseteq \mathscr{X}$ converges (is a Cauchy sequence) if and only if $\left(x_{i}^{n}\right) \subseteq X_{i}$ is convergent (is a Cauchy sequence) for every $i \in I$.

As a consequence, $\mathscr{X}$ is a Banach space if and only if all the $X_{i}$ are Banach spaces.

Proposition 1.20 (Quotient space). Let $X$ be a vector space (!) over $\mathbb{K}$, and let $Y \subseteq X$ be a linear subspace. Define, for every $x \in X$, the affine subspace

$$
x+Y:=\{x+y: y \in Y\},
$$

and define the quotient space or factor space

$$
X / Y:=\{x+Y: x \in X\}
$$

Then $X / Y$ is a vector space for the addition

$$
(x+Y)+(z+Y):=(x+z+Y)
$$

and the scalar multiplication

$$
\lambda(x+Y):=(\lambda x+Y)
$$

The neutral element is $Y$.
For the definition of quotient spaces, it is not important that we consider real or complex vector spaces.

Examples of quotient spaces are already known. In fact, $L^{p}$ is such an example. Usually, one defines

$$
\mathscr{L}^{p}(\Omega, \mathscr{A}, \mu)
$$

to be the space of all measurable functions $f: \Omega \rightarrow \mathbb{K}$ such that $\int_{\Omega}|f|^{p} \mathrm{~d} \mu<\infty$. Moreover,

$$
N:=\left\{f \in \mathscr{L}^{p}(\Omega, \mathscr{A}, \mu): \int_{\Omega}|f|^{p}=0\right\} .
$$

Note that $N$ is a linear subspace of $\mathscr{L}^{p}(\Omega, \mathscr{A}, \mu)$, and that $N$ is the space of all functions $f \in \mathscr{L}^{p}$ which vanish almost everywhere. Then

$$
L^{p}(\Omega, \mathscr{A}, \mu):=\mathscr{L}^{p}(\Omega, \mathscr{A}, \mu) / N
$$

Proposition 1.21. Let $X$ be a normed space and let $Y \subseteq X$ be a linear subspace. Then

$$
\|x+Y\|:=\inf \{\|x-y\|: y \in Y\}
$$

defines a norm on $X / Y$ if and only if $Y$ is closed in $X$. If $X$ is a Banach space and $Y \subseteq X$ closed, then $X / Y$ is also a Banach space.

Proof. We have to check that $\|\cdot\|$ satisfies all properties of a norm. Recall that $0_{X / Y}=Y$, and that for all $x \in X$

$$
\begin{aligned}
& \|x+Y\|=0 \\
& \Leftrightarrow \inf \{\|x-y\|: y \in Y\}=0 \\
& \Leftrightarrow \exists\left(y_{n}\right) \subseteq Y: \lim _{n \rightarrow \infty} y_{n}=x \\
& \Leftrightarrow(\Rightarrow \text { if } Y \text { closed }): x \in Y \\
& \Leftrightarrow x+Y=Y .
\end{aligned}
$$

Second, for every $x \in X$ and every $\lambda \in \mathbb{K} \backslash\{0\}$,

$$
\begin{aligned}
\|\lambda(x+Y)\| & =\|\lambda x+Y\| \\
& =\inf \{\|\lambda x-y\|: y \in Y\} \\
& =\inf \{\|\lambda(x-y)\|: y \in Y\} \\
& =|\lambda| \inf \{\|x-y\|: y \in Y\} \\
& =|\lambda|\|x+Y\| .
\end{aligned}
$$

Third, for every $x, z \in X$,

$$
\begin{aligned}
\|(x+Y)+(z+Y)\| & =\|(x+z)+Y\| \\
& =\inf \{\|x+z-y\|: y \in Y\} \\
& =\inf \left\{\left\|x+z-y_{1}-y_{2}\right\|: y_{1}, y_{2} \in Y\right\} \\
& \leq \inf \left\{\left\|x-y_{1}\right\|+\left\|z-y_{2}\right\|: y_{1}, y_{2} \in Y\right\} \\
& \leq \inf \{\|x-y\|: y \in Y\}+\inf \{\|z-y\|: y \in Y\} \\
& =\|x+Y\|+\|z+Y\| .
\end{aligned}
$$

Hence, $X / Y$ is a normed space if $Y$ is closed.
Assume next that $X$ is a Banach space. Let $\left(x_{n}\right) \subseteq X$ be such that the series $\sum_{n \geq 1} x_{n}+Y$ converges absolutely, that is, $\sum_{n \geq 1}\left\|x_{n}+Y\right\|<\infty$. By definition of the norm in $X / Y$, we find $\left(y_{n}\right) \subseteq Y$ such that $\left\|x_{n}-y_{n}\right\| \leq\left\|x_{n}+Y\right\|+2^{-n}$. Replacing $\left(x_{n}\right)$ by $\left(\hat{x}_{n}\right)=\left(x_{n}-y_{n}\right)$, we find that $x_{n}+Y=\hat{x}_{n}+Y$ and that the series $\sum_{n \geq 0} \hat{x}_{n}$ is absolutely convergent. Since $X$ is complete, by Lemma 1.13, the limit $\sum_{n \geq 1} \overline{\hat{x}}_{n}=$ $x \in X$ exists. As a consequence,

$$
\begin{aligned}
\left\|(x+Y)-\sum_{k=1}^{n}\left(\hat{x}_{k}+Y\right)\right\| & =\left\|\left(x-\sum_{k=1}^{n} \hat{x}_{k}\right)+Y\right\| \\
& \leq\left\|x-\sum_{k=1}^{n} \hat{x}_{k}\right\| \quad \rightarrow \quad 0
\end{aligned}
$$

that is, the series $\sum_{n \geq 1} x_{n}+Y$ converges. By Lemma 1.13, $X / Y$ is complete.

### 1.3 Bounded linear operators

In the following a linear mapping between two normed spaces $X$ and $Y$ will also be called a linear operator or just operator. If $Y=\mathbb{K}$, then we call linear operators also linear functionals. If $T: X \rightarrow Y$ is a linear operator between two normed spaces, then we denote by

$$
\operatorname{ker} T:=\{x \in X: T x=0\}
$$

its kernel or null space, and by

$$
\operatorname{ran} T:=\{T x: x \in X\}
$$

its range or image. Observe that we simply write $T x$ instead of $T(x)$, meaning that $T$ is applied to $x \in X$. The identity operator $X \rightarrow X, x \mapsto x$ is denoted by $I$.

Lemma 1.22. Let $T: X \rightarrow Y$ be a linear operator between two normed spaces $X$ and $Y$. Then the following are equivalent
(i) $T$ is continuous.
(ii) $T$ is continuous at 0 .
(iii) $\quad T B$ is bounded in $Y$, where $B=B(0,1)$ denotes the unit ball in $X$.
(iv) There exists a constant $C \geq 0$ such that for every $x \in X$

$$
\|T x\| \leq C\|x\| .
$$

Proof. The implication (i) $\Rightarrow$ (ii) is trivial.
(ii) $\Rightarrow$ (iii). If $T$ is continuous at 0 , then there exists some $\delta>0$ such that for every $x \in B(0, \delta)$ one has $T x \in B(0,1)$ (so the $\varepsilon$ from the $\varepsilon-\delta$ definition of continuity is chosen to be 1 here). By linearity, for every $x \in B=B(0,1)$

$$
\|T x\|=\frac{1}{\delta}\|T(\delta x)\| \leq \frac{1}{\delta}
$$

and this means that $T B$ is bounded.
(iii) $\Rightarrow($ iv ). The set $T B$ being bounded in $Y$ means that there exists some constant $C \geq 0$ such that for every $x \in B$ one has $\|T x\| \leq C$. By linearity, for every $x \in X \backslash\{0\}$,

$$
\|T x\|=\left\|T \frac{x}{\|x\|}\right\|\|x\| \leq C\|x\| .
$$

(iv) $\Rightarrow$ (i). Let $x \in X$, and assume that $\lim _{n \rightarrow \infty} x_{n}=x$. Then

$$
\left\|T x_{n}-T x\right\|=\left\|T\left(x_{n}-x\right)\right\| \leq C\left\|x_{n}-x\right\| \rightarrow 0 \quad \text { as } n \rightarrow \infty,
$$

so that $\lim _{n \rightarrow \infty} T x_{n}=T x$.

We call a continuous linear operator $T: X \rightarrow Y$ between two normed spaces $X$ and $Y$ also a bounded operator (since it maps the unit ball of $X$ to a bounded subset of $Y$ ). The set of all bounded linear operators is denoted by $\mathscr{L}(X, Y)$. Special cases: If $X=Y$, then we write $\mathscr{L}(X, X)=: \mathscr{L}(X)$. If $Y=\mathbb{K}$, then we write $\mathscr{L}(X, \mathbb{K})=: X^{\prime}$.

Lemma 1.23. The set $\mathscr{L}(X, Y)$ is a vector space and

$$
\begin{align*}
\|T\| & :=\inf \{C \geq 0:\|T x\| \leq C\|x\| \text { for all } x \in X\}  \tag{1.1}\\
& =\sup \{\|T x\|:\|x\| \leq 1\} \\
& =\sup \{\|T x\|:\|x\|=1\}
\end{align*}
$$

is a norm on $\mathscr{L}(X, Y)$.
Proof. We first show that the three quantities on the right-hand side of (1.1) are equal. In fact, the equality

$$
\sup \{\|T x\|:\|x\| \leq 1\}=\sup \{\|T x\|:\|x\|=1\}
$$

is easy to check so that it remains only to show that

$$
A:=\inf \{C \geq 0:\|T x\| \leq C\|x\| \text { for all } x \in X\}=\sup \{\|T x\|:\|x\|=1\}=: B
$$

If $C>A$, then for every $x \in X \backslash\{0\},\|T x\| \leq C\|x\|$ or $\left\|T \frac{x}{\|x\|}\right\| \leq C$. Hence, $C \geq B$ which implies that $A \geq B$. If $C>B$, then for every $x \in X \backslash\{0\},\left\|T \frac{x}{\|x\|}\right\| \leq C$, and therefore $\|T x\| \leq C\|x\|$. Hence, $C \geq A$ which implies that $A \leq B$.

Now we check that $\|\cdot\|$ is a norm on $\mathscr{L}(X, Y)$. First, for every $T \in \mathscr{L}(X, Y)$,

$$
\begin{aligned}
\|T\|=0 & \Leftrightarrow \sup \{\|T x\|:\|x\| \leq 1\}=0 \\
& \Leftrightarrow \forall x \in X,\|x\| \leq 1:\|T x\|=0 \\
& \Leftrightarrow(\|\cdot\| \text { is a norm on } Y) \forall x \in X,\|x\| \leq 1: T x=0 \\
& \Leftrightarrow(\Rightarrow \text { linearity of } T) \forall x \in X: T x=0 \\
& \Leftrightarrow T=0 .
\end{aligned}
$$

Second, for every $T \in \mathscr{L}(X, Y)$ and every $\lambda \in \mathbb{K}$

$$
\begin{aligned}
\|\lambda T\| & =\sup \{\|(\lambda T) x\|:\|x\| \leq 1\} \\
& =\sup \{|\lambda|\|T x\|:\|x\| \leq 1\} \\
& =|\lambda|\|T\| .
\end{aligned}
$$

Finally, for every $T, S \in \mathscr{L}(X, Y)$,

$$
\begin{aligned}
\|T+S\| & =\sup \{\|(T+S) x\|:\|x\| \leq 1\} \\
& \leq \sup \{\|T x\|+\|S x\|:\|x\| \leq 1\} \\
& \leq\|T\|+\|S\| .
\end{aligned}
$$

The proof is complete.
Remark 1.24. (a) Note that the infimum on the right-hand side of (1.1) in Lemma 1.23 is always attained. Thus, for every operator $T \in \mathscr{L}(X, Y)$ and every $x \in X$,

$$
\|T x\| \leq\|T\|\|x\| .
$$

This inequality shall be frequently used in the sequel! Note that on the other hand the suprema on the right-hand side of (1.1) are not always attained. (b) From Lemma 1.23 we can learn how to show that some operator $T: X \rightarrow Y$ is bounded and how to calculate the norm $\|T\|$. Usually (in most cases), one should prove in the first step some inequality of the form

$$
\|T x\| \leq C\|x\|, \quad x \in X
$$

because this inequality shows on the one hand that $T$ is bounded, and on the other hand it shows the estimate $\|T\| \leq C$. In the second step one should prove that the estimate $C$ was optimal by finding some $x \in X$ of norm $\|x\|=1$ such that $\|T x\|=C$, or by finding some sequence $\left(x_{n}\right) \subseteq X$ of norms $\left\|x_{n}\right\| \leq 1$ such that $\lim _{n \rightarrow \infty}\left\|T x_{n}\right\|=$ $C$, because this shows that $\|T\|=C$. Of course, the second step only works if one has not lost anything in the estimate of the first step. There are in fact many examples of bounded operators for which it is difficult to estimate their norm.

Example 1.25. 1. (Shift-operator). On $l^{p}(\mathbb{N})$ consider the left-shift operator

$$
L x=L\left(x_{n}\right)=\left(x_{n+1}\right) .
$$

Then

$$
\left\|L\left(x_{n}\right)\right\|_{p}=\left(\sum_{n}\left|x_{n+1}\right|^{p}\right)^{1 / p} \leq\left(\sum_{n}\left|x_{n}\right|^{p}\right)^{1 / p}
$$

so that $L$ is bounded and $\|L\| \leq 1$. On the other hand, for $x=(0,1,0,0, \ldots)$ one computes that $\|x\|_{p}=1$ and $\|L x\|_{p}=\|(1,0,0, \ldots)\|_{p}=1$, and one concludes that $\|L\|=1$.
2. (Shift-operator). Similarly, one shows that the right-shift operator $R$ on $l^{p}(\mathbb{N})$ defined by

$$
R x=R\left(x_{n}\right)=\left(0, x_{0}, x_{1}, \ldots\right)
$$

is bounded and $\|R\|=1$. Note that actually $\|R x\|_{p}=\|x\|_{p}$ for every $x \in l^{p}$.
3. (Multiplication operator). Let $m \in l^{\infty}$ and consider on $l^{p}$ the multiplication operator

$$
M x=M\left(x_{n}\right)=\left(m_{n} x_{n}\right)
$$

4. (Functionals on $C$ ). Consider the linear functional $\varphi: C([0,1]) \rightarrow \mathbb{K}$ defined by

$$
\varphi(f):=\int_{0}^{\frac{1}{2}} f(x) \mathrm{d} x
$$

Then

$$
|\varphi(f)| \leq \int_{0}^{\frac{1}{2}}|f(x)| \mathrm{d} x \leq \frac{1}{2}\|f\|_{\infty}
$$

so that $\varphi$ is bounded and $\|\varphi\| \leq \frac{1}{2}$. On the other hand, for the constant function $f=1$ one has $\|f\|_{\infty}=1$ and $|\varphi(f)|=\frac{1}{2}$, so that $\|\varphi\|=\frac{1}{2}$.
Lemma 1.26. Let $X, Y, Z$ be three Banach spaces, and let $T \in \mathscr{L}(X, Y)$ and $S \in$ $\mathscr{L}(Y, Z)$. Then $S T \in \mathscr{L}(X, Z)$ and

$$
\|S T\| \leq\|S\|\|T\|
$$

Proof. The boundedness of $S T$ is clear since compositions of continuous functions are again continuous. To obtain the bound on $S T$, we calculate

$$
\begin{aligned}
\|S T\| & =\sup _{\|x\| \leq 1}\|S T x\| \\
& \leq \sup _{\|x\| \leq 1}\|S\|\|T x\| \\
& =\|S\|\|T\| .
\end{aligned}
$$

Lemma 1.27. If $Y$ is a Banach space then $\mathscr{L}(X, Y)$ is a Banach space.
Proof. Assume that $Y$ is a Banach space and let $\left(T_{n}\right)$ be a Cauchy sequence in $\mathscr{L}(X, Y)$. By the estimate

$$
\left\|T_{n} x-T_{m} x\right\|=\left\|\left(T_{n}-T_{m}\right) x\right\| \leq\left\|T_{n}-T_{m}\right\|\|x\|,
$$

the sequence $\left(T_{n} x\right)$ is a Cauchy sequence in $Y$ for every $x \in X$. Since $Y$ is complete, the limit $\lim _{n \rightarrow \infty} T_{n} x$ exists for every $x \in X$. Define $T x:=\lim _{n \rightarrow \infty} T_{n} x$. Clearly, $T$ : $X \rightarrow Y$ is linear. Moreover, since any Cauchy sequence is bounded, we find that

$$
\|T x\| \leq \sup _{n}\left\|T_{n} x\right\| \leq C\|x\|
$$

for some constant $C \geq 0$, that is, $T$ is bounded. Moreover, for every $n \in \mathbb{N}$ we have the estimate

$$
\begin{aligned}
\left\|T-T_{n}\right\| & =\sup _{\|x\| \leq 1}\left\|T x-T_{n} x\right\| \\
& \leq \sup _{\|x\| \leq 1} \sup \left\|T_{m} x-T_{n} x\right\| \\
& \leq \sup _{m \geq n}\left\|T_{m}-T_{n}\right\| .
\end{aligned}
$$

Since that right-hand side of this inequality becomes arbitrarily small for large $n$, we see that $\lim _{n \rightarrow \infty} T_{n}=T$ exists, and so we have proved that $\mathscr{L}(X, Y)$ is a Banach space.
Remark 1.28. The converse of the statement in Lemma 1.27 is also true, that is, if $\mathscr{L}(X, Y)$ is a Banach space then necessarily $Y$ is a Banach space. For the proof,
however, one has to know that there are nontrivial operators in $\mathscr{L}(X, Y)$ as soon as $Y$ is nontrivial (that is, $Y \neq\{0\}$ ). For this, we need the Theorem of Hahn-Banach and its consequences discussed in Chapter 3.

Corollary 1.29. The space $X^{\prime}=\mathscr{L}(X, \mathbb{K})$ of all bounded linear functionals on $X$ is always a Banach space. The space $X^{\prime}$ is called the dual space of $X$.

Let $X, Y$ be two normed spaces. We call $T \in \mathscr{L}(X, Y)$ an isomorphism if $T$ is bijective and $T^{-1} \in \mathscr{L}(Y, X)$. We call $T \in \mathscr{L}(X, Y)$ an isometry if $\|T x\|=\|x\|$ for every $x \in X$. We say that space $X$ and $Y$ are isomorphic (and we write $X \cong Y$ ) if there exists an isomorphism $T \in \mathscr{L}(X, Y)$. We say that $X$ and $Y$ are isometrically isomorphic if there exists an isometric isomorphism $T \in \mathscr{L}(X, Y)$.

Remark 1.30. 1. Two norms $\|\cdot\|_{1},\|\cdot\|_{2}$ on a $\mathbb{K}$ vector space $X$ are equivalent if and only if the identity operator $I:\left(X,\|\cdot\|_{1}\right) \rightarrow\left(X,\|\cdot\|_{2}\right)$ is an isomorphism.
2. Saying that two normed spaces $X$ and $Y$ are isomorphic means that they are not only 'equal' as vector spaces (in the sense that we find a bijective linear operator) but also as normed spaces (that is, the bijection is continuous as well as its inverse).
3. If $T \in \mathscr{L}(X, Y)$ and $S \in \mathscr{L}(Y, Z)$ are isomorphisms, then $S T \in \mathscr{L}(X, Z)$ is an isomorphism and $(S T)^{-1}=T^{-1} S^{-1}$.
4. Every isometry $T \in \mathscr{L}(X, Y)$ is clearly injective. If it is also surjective, then $T$ is an isometric isomorphism, that is, the inverse $T^{-1}$ is also bounded (even isometric).
5. Clearly, if $T \in \mathscr{L}(X, Y)$ is isometric, then it is an isometric isomorphism from $X$ onto $\operatorname{ran} T$, and we may say that $X$ is isometrically embedded into $Y$ (via $T$ ).

Example 1.31. The right-shift operator from Example 1.25 (2) is isometric, but not surjective. In particular, $l^{p}$ is isometrically isomorphic to a proper subspace of $l^{p}$.

Exercise 1.32 Show that the spaces $\left(c,\|\cdot\|_{\infty}\right)$ of all convergent sequences and $\left(c_{0},\|\cdot\|_{\infty}\right)$ of all null sequences are isomorphic.

Exercise 1.33 Show that $\left(c_{0},\|\cdot\|_{\infty}\right)$ is (isometrically) isomorphic to a linear subspace of $\left(C([0,1]),\|\cdot\|_{\infty}\right)$, that is, find an isometry $T: c_{0} \rightarrow C([0,1])$.

Lemma 1.34 (Neumann series). Let $X$ be a Banach space and let $T \in \mathscr{L}(X)$ be such that $\|T\|<1$. Then $I-T$ is boundedly invertible, that is, it is an isomorphism. Moreover, $(I-T)^{-1}=\sum_{n \geq 0} T^{n}$.

Proof. Since $X$ is a Banach space, $\mathscr{L}(X)$ is also a Banach space by Lemma 1.27. By assumption on $\|T\|$, the series $\sum_{n \geq 0} T^{n}$ is absolutely convergent, and hence, by Lemma 1.13, it is convergent to some element $S \in \mathscr{L}(X)$. Moreover,

$$
(I-T) S=\lim _{n \rightarrow \infty}(I-T) \sum_{k=0}^{n} T^{k}=\lim _{n \rightarrow \infty}\left(I-T^{k+1}\right)=I
$$

and similarly, $S(I-T)=I$.

Corollary 1.35. Let $X$ and $Y$ be two Banach spaces. Then the set $\mathscr{I}(X, Y)$ of all isomorphisms in $\mathscr{L}(X, Y)$ is open, and the mapping $T \mapsto T^{-1}$ is continuous from $\mathscr{I}(X, Y)$ onto $\mathscr{I}(Y, X)$.

Proof. Let $\mathscr{I} \subseteq \mathscr{L}(X, Y)$ be the set of all isomorphisms, and assume that $\mathscr{I}$ is not empty (if it is empty, then it is also open). Let $T \in \mathscr{I}$. Then for every $S \in B\left(T, \frac{1}{\left\|T^{-1}\right\|}\right)$ we have

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

and since $\left\|T^{-1}(S-T)\right\| \leq\left\|T^{-1}\right\|\|S-T\|<1$, the operator $I+T^{-1}(S-T) \in \mathscr{L}(X)$ is an isomorphism by Lemma 1.34. As a composition of two isomorphisms, $S \in \mathscr{I}$, and hence $\mathscr{I}$ is open. The continuity is also a direct consequence of the above representation of $S$ (and thus of its inverse), using the Neumann series.

### 1.4 The Arzelà-Ascoli theorem

It is a consequence of Riesz' Lemma (Lemma 1.14) that the unit ball in an infinite dimensional Banach space is not compact; see also Theorem 1.16. But compact sets play an important role in many theorems from analysis, in particular when one wants to prove the existence of some fixed point, the existence of a solution to an algebraic equation, the existence of a solution of a differential equation, the existence of a solution of a partial differential equation etc. It is therefore important to identify the compact sets in Banach spaces, in particular in the classical Banach spaces. The Arzalà-Ascoli theorem characterizes the compact subsets of $C(K ; X)$, where $(K, d)$ is a compact metric space and $X$ is a Banach space.

We say that a subset $B \subseteq C(K ; X)$ is equicontinuous at some point $x \in K$ if for every $\varepsilon>0$ there exists $\delta>0$ such that for every $y \in K$ and every $f \in B$ the implication

$$
d(x, y)<\delta \quad \Rightarrow \quad\|f(x)-f(y)\|<\varepsilon
$$

holds.
Theorem 1.36 (Arzelà-Ascoli). Let $(K, d)$ be a compact metric space, $X$ be a Banach space and consider the Banach space $C(K ; X)$ of all continuous functions $K \rightarrow X$ equipped with the supremum norm $\|f\|_{\infty}=\sup _{x \in K}\|f(x)\|$. For a subset $B \subseteq C(K ; X)$, the following assertions are equivalent:
(i) The set $B$ is relatively compact.
(ii) The set $B$ is equicontinuous at every $x \in K$ and there exists a dense set $D \subseteq K$ such that for every $x \in D$ the set $B_{x}=\{f(x): f \in B\}$ is relatively compact.

We point out that, by the Heine-Borel theorem, the condition of pointwise relative compactness of $B$ can be replaced by mere pointwise or global boundedness as soon as the space $X$ is finite dimensional.

Corollary 1.37 (Arzelà-Ascoli). Let $(K, d)$ be a compact metric space, and consider the Banach space $C\left(K ; \mathbb{R}^{d}\right)$ of all continuous functions $K \rightarrow \mathbb{R}^{d}$ equipped with the supremum norm $\|f\|_{\infty}=\sup _{x \in K}\|f(x)\|$. For a subset $B \subseteq C\left(K ; \mathbb{R}^{d}\right)$, the following assertions are equivalent:
(i) The set $B$ is compact.
(ii) The set $B$ is closed, equicontinuous at every $x \in K$ and pointwise bounded in the sense that for every $x \in K$ the set $B_{x}=\{f(x): f \in B\}$ is bounded.

Proof (of Theorem 1.36). The proof of the Arzelà-Ascoli theorem is a nice application of Cantor's diagonal sequence argument which we see here for the first time, but which we will see again below when we prove that every bounded sequence in a reflexive Banach space admits a weakly convergent subsequence. Given a sequence, Cantor's argument allows us to construct a subsequence which satisfies a countable number of properties. It is instructive to learn the idea of Cantor's argument since it can be help in various situations.

We first assume that $B \subseteq C(K ; X)$ is relatively compact. Any relatively compact subset of a Banach space is bounded, and therefore $B$ is bounded, too. For every $x \in K$, the point evaluation $C(K ; X) \rightarrow X, f \mapsto f(x)$ is linear and continuous. Since continuous images of relatively compact sets are relatively compact, the image of $B$ under the point evaluation, that is the set $B_{x}=\{f(x): f \in B\}$, is relatively compact.

We show that $B$ is equicontinuous at every $x$. Assume that this was not the case. Then there exist $x \in K$ and $\varepsilon>0$ such that for every $n \geq 1$ there exist $y_{n} \in K$ and $f_{n} \in B$ such that $d\left(x, y_{n}\right)<\frac{1}{n}$ and $\left\|f_{n}(x)-f_{n}\left(y_{n}\right)\right\| \geq \varepsilon$. Since $B$ is relatively compact, there exists a subsequence of $\left(f_{n}\right)$ (which we denote for simplicity again by $\left(f_{n}\right)$ ) such that $\lim _{n \rightarrow \infty} f_{n}=f$ in $C(K ; X)$. Then, by the triangle inequality from below,

$$
\begin{aligned}
\liminf _{n \rightarrow \infty}\left\|f(x)-f\left(y_{n}\right)\right\| & =\liminf _{n \rightarrow \infty}\left\|f(x)-f_{n}(x)+f_{n}(x)-f_{n}\left(y_{n}\right)+f_{n}\left(y_{n}\right)-f\left(y_{n}\right)\right\| \\
& \geq \liminf _{n \rightarrow \infty}\left(\left\|f_{n}(x)-f_{n}\left(y_{n}\right)\right\|-2\left\|f-f_{n}\right\|_{\infty}\right) \\
& \geq \varepsilon .
\end{aligned}
$$

This inequality, however, contradicts to the continuity of $f$ (note that $\lim _{n \rightarrow \infty} y_{n}=x$ ), and therefore, $B$ is equicontinuous at every $x \in K$.

Assume now that $B$ satisfies the properties from assertion (ii). In order to show that $B$ is relatively compact, it suffices to show that every sequence $\left(f_{n}\right) \subseteq B$ admits a convergent subsequence, that is, $B$ is relatively sequentially compact. So let $\left(f_{n}\right) \subseteq B$ be an arbitrary sequence.

Recall that every compact metric space is separable. Moreover, every subset of a separable space is separable. Hence, there exists a sequence $\left(x_{m}\right)_{m \geq 1} \subseteq K$ which is dense in $K$.

Consider the sequence $\left(f_{n}\left(x_{1}\right)\right) \subseteq B_{x_{1}} \subseteq X$. Since $B_{x_{1}}$ is relatively compact by assumption, there exists a subsequence $\left(f_{\varphi_{1}(n)}\right)$ of $\left(f_{n}\right)$ such that $\lim _{n \rightarrow \infty} f_{\varphi_{1}(n)}\left(x_{1}\right)$ exists.

Consider next the sequence $\left(f_{\varphi_{1}(n)}\left(x_{2}\right)\right) \subseteq B_{x_{2}} \subseteq X$. Since $B_{x_{2}}$ is relatively compact by assumption, there exists a subsequence $\left(f_{\varphi_{2}(n)}\right)$ of $\left(f_{\varphi_{1}(n)}\right)$ such that $\lim _{n \rightarrow \infty} f_{\varphi_{2}(n)}\left(x_{2}\right)$ exists. Note that we have also the existence of the limit $\lim _{n \rightarrow \infty} f_{\varphi_{2}(n)}\left(x_{1}\right)$.

Iterating this argument, we obtain for every $m \geq 2$ a subsequence $\left(f_{\varphi_{m}(n)}\right)$ of $\left(f_{\varphi_{m-1}(n)}\right)$ such that $\lim _{n \rightarrow \infty} f_{\varphi_{m}(n)}\left(x_{i}\right)$ exists for every $1 \leq i \leq m$. These subsequences converge therefore pointwise at a finite number of elements of $K$.

We now consider the diagonal subsequence $\left(f_{\varphi(n)}\right)=\left(f_{\varphi_{n}(n)}\right)$. This diagonal subsequence has the property of being a subsequence of $\left(f_{\varphi_{m}(n)}\right)$ for every $m \geq 1$, up to a finite number of initial elements perhaps. It enjoys therefore the property that $\lim _{n \rightarrow \infty} f_{\varphi(n)}\left(x_{m}\right)$ exists for every $m \geq 1$, that is, it converges pointwise on a dense subset of $K$. We will show that $\left(f_{\varphi(n)}\right)$ converges everywhere and uniformly on $K$. Since $C(K ; X)$ is complete, it suffices to show that $\left(f_{\varphi(n)}\right)$ is a Cauchy sequence in $C(K ; X)$.

Let $\varepsilon>0$. Since $B$ is equicontinuous at every $x \in K$, for every $x \in K$ there exists $\delta_{x}>0$ such that for every $y \in K$ and every $f \in B$ the implication

$$
\begin{equation*}
d(x, y)<\delta \quad \Rightarrow \quad\|f(x)-f(y)\|<\varepsilon \tag{1.2}
\end{equation*}
$$

is true. We clearly have $K=\bigcup_{x \in K} B\left(x, \delta_{x}\right)$, and since $K$ is compact, we find finitely many points $x_{1}^{\prime}, \ldots, x_{k}^{\prime}$ such that $K=\bigcup_{i=1}^{k} B\left(x_{i}^{\prime}, \delta_{i}\right)$ (with $\delta_{i}=\delta_{x_{i}^{\prime}}$. Since the sequence $\left(x_{m}\right)$ is dense in $K$, for every $1 \leq i \leq k$ there exists $m_{i} \geq 1$ such that $x_{m_{i}} \in B\left(x_{i}^{\prime}, \delta_{i}\right)$. Since the sequence $\left(f_{\varphi(n)}\right)$ converges pointwise on $\left(x_{m}\right)$, there exists $n_{0} \geq 0$ such that

$$
\text { for every } n, n^{\prime} \geq n_{0} \text { and every } 1 \leq i \leq k \quad\left\|f_{\varphi(n)}\left(x_{m_{i}}\right)-f_{\varphi\left(n^{\prime}\right)}\left(x_{m_{i}}\right)\right\|<\varepsilon .
$$

Let now $x \in K$ be arbitrary. Then $x \in B\left(x_{i}^{\prime}, \delta_{i}\right)$ for some $1 \leq i \leq k$. Hence, for every $n, n^{\prime} \geq n_{0}$, by the preceding estimate and by the implication (1.2),

$$
\begin{aligned}
\left\|f_{\varphi(n)}(x)-f_{\varphi\left(n^{\prime}\right)}(x)\right\| \leq & \left\|f_{\varphi(n)}(x)-f_{\varphi(n)}\left(x_{i}^{\prime}\right)\right\|+ \\
& +\left\|f_{\varphi(n)}\left(x_{i}^{\prime}\right)-f_{\varphi(n)}\left(x_{m_{i}}\right)\right\|+ \\
& +\left\|f_{\varphi(n)}\left(x_{m_{i}}\right)-f_{\varphi\left(n^{\prime}\right)}\left(x_{m_{i}}\right)\right\|+ \\
& +\left\|f_{\varphi\left(n^{\prime}\right)}\left(x_{m_{i}}\right)-f_{\varphi\left(n^{\prime}\right)}\left(x_{i}^{\prime}\right)\right\|+ \\
& +\left\|f_{\varphi\left(n^{\prime}\right)}\left(x_{i}^{\prime}\right)-f_{\varphi\left(n^{\prime}\right)}(x)\right\|
\end{aligned}
$$

$$
\leq 5 \varepsilon
$$

Since $n_{0} \geq 0$ did not depend on $x \in K$, and since $\varepsilon>0$ was arbitrary, this proves that $\left(f_{\varphi(n)}\right)$ is a Cauchy sequence in $C(K ; X)$. We have therefore proved that every sequence in $B$ admits a convergent subsequence.

## Chapter 2

## Hilbert spaces

Let $H$ be a vector space over $\mathbb{K}$.

### 2.1 Inner product spaces

A function $\langle\cdot, \cdot\rangle: H \times H \rightarrow \mathbb{K}$ is called an inner product if for every $x, y, z \in H$ and every $\lambda \in \mathbb{K}$
(i) $\langle x, x\rangle \geq 0$ for every $x \in H$ and $\langle x, x\rangle=0$ if and only if $x=0$,
(ii) $\langle x, y\rangle=\overline{\langle y, x\rangle}$,
(iii) $\langle\lambda x+y, z\rangle=\lambda\langle x, z\rangle+\langle y, z\rangle$.

A pair $(H,\langle\cdot, \cdot\rangle)$ of a vector space over $\mathbb{K}$ and a scalar product is called an inner product space.
Example 2.1. 1. On the space $H=\mathbb{K}^{d}$,

$$
\langle x, y\rangle:=\sum_{i=1}^{d} x_{i} \bar{y}_{i}
$$

defines an inner product.
2. On the space $H=l^{2}:=\left\{\left(x_{n}\right) \subseteq \mathbb{K}: \sum\left|x_{n}\right|^{2}<\infty\right\}$,

$$
\langle x, y\rangle:=\sum_{n} x_{n} \overline{y_{n}}
$$

defines an inner product.
3. On the space $H=C([0,1])$, the Riemann integral

$$
\langle f, g\rangle:=\int_{0}^{1} f(x) \overline{g(x)} \mathrm{d} x
$$

defines an inner product.
4. On the space $H=L^{2}(\Omega)$, the integral

$$
\langle f, g\rangle:=\int_{\Omega} f \bar{g} \mathrm{~d} \mu
$$

defines an inner product.
Lemma 2.2. Let $\langle\cdot, \cdot\rangle$ be an inner product on a vector space H. Then, for every $x, y$, $z \in H$ and $\lambda \in \mathbb{K}$
(iv) $\langle x, \lambda y+z\rangle=\bar{\lambda}\langle x, y\rangle+\langle x, z\rangle$.

Proof.

$$
\langle x, \lambda y+z\rangle=\overline{\langle\lambda y+z, x\rangle}=\bar{\lambda} \overline{\langle y, x\rangle}+\overline{\langle z, x\rangle}=\bar{\lambda}\langle x, y\rangle+\langle x, z\rangle .
$$

In the following, if $H$ is an inner product space, then we put

$$
\|x\|:=\sqrt{\langle x, x\rangle}, \quad x \in H
$$

Lemma 2.3 (Cauchy-Schwarz inequality). Let $H$ be an inner product space. Then, for every $x, y \in H$,

$$
|\langle x, y\rangle| \leq\|x\|\|y\|
$$

and equality holds if and only if $x$ and $y$ are colinear.
Proof. Let $\lambda \in \mathbb{K}$. Then

$$
\begin{aligned}
0 & \leq\langle x+\lambda y, x+\lambda y\rangle \\
& =\langle x, x\rangle+\langle\lambda y, x\rangle+\langle x, \lambda y\rangle+|\lambda|^{2}\langle y, y\rangle \\
& =\langle x, x\rangle+\lambda \overline{\langle x, y\rangle}+\bar{\lambda}\langle x, y\rangle+|\lambda|^{2}\langle y, y\rangle,
\end{aligned}
$$

that is,

$$
\begin{equation*}
0 \leq\|x+\lambda y\|^{2}=\|x\|^{2}+2 \operatorname{Re} \bar{\lambda}\langle x, y\rangle+|\lambda|^{2}\|y\|^{2} \tag{2.1}
\end{equation*}
$$

Assuming that $y \neq 0$ (for $y=0$ the Cauchy-Schwarz inequality is trivial), we may put $\lambda:=-\langle x, y\rangle /\|y\|^{2}$. Then

$$
\begin{aligned}
0 & \leq\left\langle x-\frac{\langle x, y\rangle}{\|y\|^{2}} y, x-\frac{\langle x, y\rangle}{\|y\|^{2}} y\right\rangle \\
& =\|x\|^{2}-\frac{|\langle x, y\rangle|^{2}}{\|y\|^{2}},
\end{aligned}
$$

which is the Cauchy-Schwarz inequality. The calculation also shows that equality holds if and only if $x=\lambda y$, that is, if $x$ and $y$ are colinear.
Lemma 2.4. Every inner product space $H$ is a normed linear space for the norm

$$
\|x\|=\sqrt{\langle x, x\rangle}, \quad x \in H
$$

Proof. Properties (i) and (ii) in the definition of a norm follow from the properties (i) and (iii) (together with Lemma 2.2) in the definition of an inner product. The only difficulty is to show that $\|\cdot\|$ satisfies the triangle inequality. This, however, follows from putting $\lambda=1$ in (2.1) and estimating with the Cauchy-Schwarz inequality:

$$
\|x+y\|^{2} \leq(\|x\|+\|y\|)^{2} .
$$

A complete inner product space is called a Hilbert space.
Example 2.5. The spaces $\mathbb{K}^{d}$ (with Euclidean inner product), $l^{2}$ and $L^{2}(\Omega)$ are Hilbert spaces. More examples are given by the Sobolev spaces defined below.

Lemma 2.6 (Completion of an inner product space). Let $H$ be an inner product space. Then there exists a Hilbert space $K$ and a bounded linear operator j: $H \rightarrow K$ such that for every $x, y \in H$

$$
\langle x, y\rangle_{H}=\langle j(x), j(y)\rangle_{K},
$$

and such that $j(H)$ is dense in $K$. The Hilbert space $K$ is unique up to isometry. It is called the completion of $H$.

Lemma 2.7 (Parallelogram identity). Let $H$ be an inner product space. Then for every $x, y \in H$

$$
\|x+y\|^{2}+\|x-y\|^{2}=2\left(\|x\|^{2}+\|y\|^{2}\right) .
$$

Proof. The parallelogram identity follows immediately from (2.1) by putting $\lambda=$ $\pm 1$ and adding up.

Exercise 2.8 (von Neumann) Show that a norm satisfying the parallelogram identity comes from a scalar product. That means, the parallelogram identity characterises inner product spaces.

A subset $K$ of a real or complex vector space $X$ is convex if for every $x, y \in K$ and every $t \in[0,1]$ one has $t x+(1-t) y \in K$.

Theorem 2.9 (Projection onto closed, convex sets). Given a nonempty closed, convex subset $K$ of a Hilbert space $H$, and given a point $x \in H$, there exists a unique $y \in K$ such that

$$
\|x-y\|=\inf \{\|x-z\|: z \in K\} .
$$

Proof. Let $d:=\inf \{\|x-z\|: z \in K\}$, and choose $\left(y_{n}\right) \in K$ such that

$$
\begin{equation*}
\lim _{n \rightarrow \infty}\left\|x-y_{n}\right\|=d \tag{2.2}
\end{equation*}
$$

Applying the parallelogram identity to $\left(x-y_{n}\right) / 2$ and $\left(x-y_{m}\right) / 2$, we obtain

$$
\left\|x-\frac{y_{n}+y_{m}}{2}\right\|^{2}+\frac{1}{4}\left\|y_{n}-y_{m}\right\|^{2}=\frac{1}{2}\left(\left\|x-y_{n}\right\|^{2}+\left\|x-y_{m}\right\|^{2}\right) .
$$

Since $K$ is convex, $\frac{y_{n}+y_{m}}{2} \in K$ and hence $\left\|x-\frac{y_{n}+y_{m}}{2}\right\|^{2} \geq d^{2}$. Using this and (2.2), the last identity implies that $\left(y_{n}\right)$ is a Cauchy sequence. Since $H$ is complete, $y:=$ $\lim _{n \rightarrow \infty} y_{n}$ exists. Since $K$ is closed, $y \in K$. Moreover, $\|x-y\|=\lim _{n \rightarrow \infty}\left\|x-y_{n}\right\|=d$, so that $y$ is a minimizer for the distance to $x$. To see that there is only one such minimizer, suppose that $y^{\prime} \in K$ is a second one, and apply the parallelogram identity to $x-y$ and $x-y^{\prime}$.

Let $H$ be an inner product space. We say that two vectors $x, y \in H$ are orthogonal (and we write $x \perp y$ ), if $\langle x, y\rangle=0$. Given a subset $S \subseteq H$, we define the orthogonal space $S^{\perp}:=\{y \in H: x \perp y$ for all $x \in S\}$. If $S=K$ is a linear subspace of $H$, then we call $K^{\perp}$ also the orthogonal complement of $K$.

Theorem 2.10. Let $H$ be a Hilbert space, $S \subseteq H$ be a subset and $K$ a closed linear subspace. Then:
a) $S^{\perp}$ is a closed linear subspace of $H$,
b) $K$ and $K^{\perp}$ are complementary subspaces, i.e. every $x \in H$ can be decomposed uniquely as a sum of an $x_{0} \in K$ and an $x_{1} \in K^{\perp}$,
c) $\left(K^{\perp}\right)^{\perp}=K$ and $\left(S^{\perp}\right)^{\perp}=\overline{\operatorname{span}} S$.
d) $\operatorname{span} S$ is dense in $H$ if and only if $S^{\perp}=\{0\}$.

Proof. (a) It follows from the bilinearity of the inner product that $S^{\perp}$ is a linear subspace of $H$. Let $\left(y_{n}\right) \subseteq S^{\perp}$ be convergent to some $y \in H$. Then, for every $x \in S$, by the Cauchy-Schwarz inequality,

$$
\langle x, y\rangle=\lim _{n \rightarrow \infty}\left\langle x, y_{n}\right\rangle=0,
$$

that is, $y \in S^{\perp}$ and therefore $S^{\perp}$ is closed.
(b) For every $x \in H$ we let $x_{0} \in K$ be the unique element (Theorem 2.9) such that

$$
\left\|x-x_{0}\right\|=\inf \{\|x-y\|: y \in K\}
$$

Put $x_{1}=x-x_{0}$. For every $y \in K$ and every $\lambda \in \mathbb{K}$, by the minimum property of $x_{0}$,

$$
\begin{aligned}
\left\|x_{1}\right\|^{2} & \leq\left\|x_{1}-\lambda y\right\|^{2} \\
& =\left\|x_{1}\right\|^{2}-2 \operatorname{Re} \bar{\lambda}\left\langle x_{1}, y\right\rangle+|\lambda|^{2}\|y\|^{2} .
\end{aligned}
$$

This implies that $\left\langle x_{1}, y\right\rangle=0$, that is, $x_{1} \in K^{\perp}$. Every decomposition $x=x_{0}+x_{1}$ with $x_{0} \in K$ and $x_{1} \in K^{\perp}$ is unique since $x \in K \cap K^{\perp}$ implies $\langle x, x\rangle=0$, that is, $x=0$.
(c) and (d) follow immediately from (a) and (b).

Lemma 2.11 (Pythagoras). Let $H$ be an inner product space. Whenever $x, y \in H$ are orthogonal, then

$$
\|x+y\|^{2}=\|x\|^{2}+\|y\|^{2} .
$$

Proof. The claim follows from (2.1) and putting $\lambda=1$.

We call an operator $P: X \rightarrow X$ on a linear space $X$ a projection if $P^{2}=P$.
Lemma 2.12. Let $X$ be a normed space and let $P \in \mathscr{L}(X)$ be a bounded projection. Then the following are true:
a) $Q=I-P$ is a projection.
b) Either $P=0$ or $\|P\| \geq 1$.
c) The kernel $\operatorname{ker} P$ and the range $\operatorname{ran} P$ are closed in $X$.
d) Every $x \in X$ can be decomposed uniquely as a sum of an $x_{0} \in \operatorname{ker} P$ and an $x_{1} \in \operatorname{ran} P$, and $X \cong \operatorname{ker} P \oplus \operatorname{ran} P$.
Proof. (a) $Q^{2}=(I-P)^{2}=I-2 P+P^{2}=I-P=Q$.
(b) follows from $\|P\|=\left\|P^{2}\right\| \leq\|P\|^{2}$.
(c) Since $\{0\}$ is closed in $X$ and since $P$ is continuous, $\operatorname{ker} P=P^{-1}(\{0\})$ is closed. Similarly, $\operatorname{ran} P=\operatorname{ker}(I-P)$ is closed.
(d) For every $x \in X$ we can write $x=P x+(I-P) x=x_{1}+x_{2}$ with $x_{1} \in \operatorname{ran} P$ and $x_{2} \in \operatorname{ker} P$. The decomposition is unique since if $x \in \operatorname{ker} P \cap \operatorname{ran} P$, then $x=P x=0$. This proves that the vector spaces $X$ and $\operatorname{ker} P \oplus \operatorname{ran} P$ are isomorphic. That they are also isomorphic as normed spaces follows from the continuity of $P$.

Lemma 2.13. Let $H$ be a Hilbert space and $K \subseteq H$ be a closed linear subspace. For every $x \in H$ we let $x_{1}=P x$ be the unique element in $K$ which minimizes the distance to $x$ (Theorem 2.9). Then $P: H \rightarrow H$ is a bounded projection satisfying $\operatorname{ran} P=K$. Moreover, $\operatorname{ker} P=K^{\perp}$. We call $P$ the orthogonal projection onto $K$.

### 2.2 Orthogonal decomposition

We call a metric space separable if there exists a countable dense subset.
Example 2.14. The space $\mathbb{R}^{d}$ (or $\mathbb{C}^{d}$ ) is separable: one may take $\mathbb{Q}^{d}$ as an example of a dense countable subset. It is not too difficult to see that subsets of separable metric spaces are separable (note, however, that in general the dense subset has to be constructed carefully), and that finite products of separable metric spaces are separable.

Lemma 2.15. A normed space $X$ is separable if and only if there exists a sequence $\left(x_{n}\right) \subseteq X$ such that span $\left\{x_{n}: n \in \mathbb{N}\right\}$ is dense in $X$ (such a sequence is in general called a total sequence).
Proof. If $X$ is separable, then there exists a sequence $\left(x_{n}\right) \subseteq X$ such that $\left\{x_{n}: n \in \mathbb{N}\right\}$ is dense. In particular, the larger set span $\left\{x_{n}: n \in \mathbb{N}\right\}$ is dense.

If, on the other hand, there exists a total sequence $\left(x_{n}\right) \subseteq X$, and if we put $D=\mathbb{Q}$ in the case $\mathbb{K}=\mathbb{R}$ and $D=\mathbb{Q}+i \mathbb{Q}$ in the case $\mathbb{K}=\mathbb{C}$, then the set

$$
\left\{\sum_{i=1}^{m} \lambda_{i} x_{n_{i}}: m \in \mathbb{N}, \lambda_{i} \in D, n_{i} \in \mathbb{N}\right\}
$$

is dense in $X$ (in fact, the closure contains all finite linear combinations of the $x_{n}$, that is, it contains $\left.\operatorname{span}\left\{x_{n}: n \in \mathbb{N}\right\}\right)$. It is an exercise to show that this set is countable. The claim follows.

Corollary 2.16. The space $\left(C([0,1]),\|\cdot\|_{\infty}\right)$ is separable.
Proof. By Weierstrass' theorem, the subspace of all polynomials is dense in $C([0,1])$ (Weierstrass' theorem says that every continuous function $f:[0,1] \rightarrow \mathbb{R}$ can be uniformly approximated by polynomials). The polynomials, however, are the linear span of the monomials $f_{n}(t)=t^{n}$. The claim therefore follows from Lemma 2.15 .

Corollary 2.17. The space $l^{p}$ is separable if $1 \leq p<\infty$. The space $c_{0}$ is separable.
Proof. Let $e_{n}=\left(\delta_{n k}\right)_{k} \in l^{p}$ be the $n$-th unit vector in $l^{p}$ (here $\delta_{n k}$ denotes the Kronecker symbol: $\delta_{n k}=1$ if $n=k$ and $\delta_{n k}=0$ otherwise). Then $\operatorname{span}\left\{e_{n}: n \in \mathbb{N}\right\}=c_{00}$ (the space of all finite sequences) is dense in $l^{p}$ if $1 \leq p<\infty$. The claim for $l^{p}$ follows from Lemma 2.15. The argument for $c_{0}$ is similar.

Lemma 2.18. The space $l^{\infty}$ is not separable.
Proof. The set $\{0,1\}^{\mathbb{N}} \subseteq l^{\infty}$ of all sequences taking only values 0 or 1 is uncountable. Moreover, whenever $x, y \in\{0,1\}^{\mathbb{N}}, x \neq y$, then

$$
\|x-y\|_{\infty}=1 .
$$

Hence, the balls $B\left(x, \frac{1}{2}\right)$ with centers $x \in\{0,1\}^{\mathbb{N}}$ and radius $\frac{1}{2}$ are mutually disjoint. If $l^{\infty}$ was separable, that is, if there exists a dense countable set $D \subseteq l^{\infty}$, then in each $B\left(x, \frac{1}{2}\right)$ there exists at least one element $y \in D$, a contradiction.

Definition 2.19. Let $H$ be an inner product space. A family $\left(e_{l}\right)_{l \in I} \subseteq H$ is called
a) an orthogonal system if $\left(e_{l}, e_{k}\right)=0$ whenever $l \neq k$,
b) an orthonormal system if it is an orthogonal system and $\left\|e_{l}\right\|=1$ for every $l \in I$, and
c) an orthonormal basis if it is an orthonormal system and span $\left\{e_{l}: l \in I\right\}$ is dense in $H$.

Lemma 2.20 (Gram-Schmidt process). Let $\left(x_{n}\right)$ be a sequence in an inner product space $H$. Then there exists an orthonormal system $\left(e_{n}\right)$ such that $\operatorname{span}\left\{x_{n}\right\}=$ $\operatorname{span}\left\{e_{n}\right\}$.

Proof. Passing to a subsequence, if necessary, we may assume that the $\left(x_{n}\right)$ are linearly independent.

Let $e_{1}:=x_{1} /\left\|x_{1}\right\|$. Then $e_{1}$ and $x_{1}$ span the same linear subspace. Next, assume that we have constructed an orthonormal system $\left(e_{k}\right)_{1 \leq k \leq n}$ such that

$$
\operatorname{span}\left\{x_{k}: 1 \leq k \leq n\right\}=\operatorname{span}\left\{e_{k}: 1 \leq k \leq n\right\} .
$$

Let $e_{n+1}^{\prime}:=x_{n+1}-\sum_{k=1}^{n}\left\langle x_{n+1}, e_{k}\right\rangle e_{k}$. Since the $x_{n}$ are linearly independent, we find $e_{n+1}^{\prime} \neq 0$. Let $e_{n+1}:=e_{n+1}^{\prime} /\left\|e_{n+1}^{\prime}\right\|$. By construction, for every $1 \leq k \leq n$, $\left\langle e_{n+1}, e_{k}\right\rangle=0$, and

$$
\operatorname{span}\left\{x_{k}: 1 \leq k \leq n+1\right\}=\operatorname{span}\left\{e_{k}: 1 \leq k \leq n+1\right\} .
$$

Proceeding inductively, the claim follows.
Corollary 2.21. Every separable inner product space admits an orthonormal basis.
Example 2.22. Consider the inner product space $C([-1,1])$ equiped with the scalar product $\langle f, g\rangle=\int_{-1}^{1} f(t) \overline{g(t)} \mathrm{d} t$ and resulting norm $\|\cdot\|_{2}$. Let $f_{n}(t):=t^{n}(n \geq 0)$, so that span $\left\{f_{n}\right\}$ is the space of all polynomials on the interval $[-1,1]$. Applying the Gram-Schmidt process to the sequence $\left(f_{n}\right)$ yields a orthonormal sequence $\left(p_{n}\right)$ of polynomials. The $p_{n}$ are called Legendre polynomials.

Recall that the space of all polynomials is dense in $C([-1,1])$ by Weierstrass' theorem (even for the uniform norm; a fortiori also for the norm $\|\cdot\|_{2}$ ). Hence, the Legendre polynomials form an orthonormal basis in $C([-1,1])$.

Lemma 2.23 (Bessel's inequality). Let $H$ be an inner product space, $\left(e_{n}\right)_{n \in \mathbb{N}} \subseteq H$ an orthonormal system. Then, for every $x \in H$,

$$
\sum_{n \in \mathbb{N}}\left|\left\langle x, e_{n}\right\rangle\right|^{2} \leq\|x\|^{2} .
$$

Proof. Let $N \in \mathbb{N}$. Put $x_{N}=x-\sum_{n=1}^{N}\left\langle x, e_{n}\right\rangle e_{n}$ so that $x_{N} \perp e_{n}$ for every $1 \leq n \leq N$. By Pythagoras (Lemma 2.11),

$$
\begin{aligned}
\|x\|^{2} & =\left\|x_{N}\right\|^{2}+\left\|\sum_{n=1}^{N}\left\langle x, e_{n}\right\rangle e_{n}\right\|^{2} \\
& =\left\|x_{N}\right\|^{2}+\sum_{n=1}^{N}\left|\left\langle x, e_{n}\right\rangle\right|^{2} \\
& \geq \sum_{n=1}^{N}\left|\left\langle x, e_{n}\right\rangle\right|^{2} .
\end{aligned}
$$

Since $N$ was arbitrary, the claim follows.
Lemma 2.24. Let $H$ be a (separable) Hilbert space, $\left(e_{n}\right)_{n \in \mathbb{N}} \subseteq H$ an orthonormal system. Then:
a) For every $x \in H$, the series $\sum_{n \in \mathbb{N}}\left\langle x, e_{n}\right\rangle e_{n}$ converges.
b) $P: H \rightarrow H, x \mapsto \sum_{n \in \mathbb{N}}\left\langle x, e_{n}\right\rangle e_{n}$ is the orthogonal projection onto $\overline{\operatorname{span}}\left\{e_{n}: n \in\right.$ $\mathbb{N}\}$.

Proof. (a) Let $x \in H$. Since ( $e_{n}$ ) is an orthonormal system, by Pythagoras (Lemma 2.11), for every $l>k \geq 1$,

$$
\begin{aligned}
\left\|\sum_{n=1}^{l}\left\langle x, e_{n}\right\rangle e_{n}-\sum_{n=1}^{k}\left\langle x, e_{n}\right\rangle e_{n}\right\|^{2} & =\left\|\sum_{n=k+1}^{l}\left\langle x, e_{n}\right\rangle e_{n}\right\|^{2} \\
& =\sum_{n=k+1}^{l}\left|\left\langle x, e_{n}\right\rangle\right|^{2} .
\end{aligned}
$$

Hence, by Bessel's inequality, the sequence ( $\sum_{n=1}^{l}\left\langle x, e_{n}\right\rangle e_{n}$ ) of partial sums forms a Cauchy sequence. Since $H$ is complete, the series $\sum_{n \in \mathbb{N}}\left\langle x, e_{n}\right\rangle e_{n}$ converges.
(b) is an exercise.

Theorem 2.25. Let $H$ be a (separable) Hilbert space, $\left(e_{n}\right)_{n \in \mathbb{N}}$ an orthonormal system. Then the following are equivalent:
(i) $\left(e_{n}\right)_{n \in \mathbb{N}}$ is an orthonormal basis.
(ii) If $x \perp e_{n}$ for every $n \in \mathbb{N}$, then $x=0$.
(iii) $x=\sum_{n \in \mathbb{N}}\left\langle x, e_{n}\right\rangle e_{n}$ for every $x \in H$.
(iv) $\langle x, y\rangle=\sum_{n \in \mathbb{N}}\left\langle x, e_{n}\right\rangle\left\langle e_{n}, y\right\rangle$ for every $x, y \in H$.
(v) (Parseval's identity) For every $x \in H$,

$$
\|x\|^{2}=\sum_{n \in \mathbb{N}}\left|\left\langle x, e_{n}\right\rangle\right|^{2}
$$

Proof. (i) $\Rightarrow$ (ii) follows from Theorem 2.10.
(ii) $\Rightarrow$ (iii) follows from Lemma 2.24 (i). In fact, let $x_{0}=\sum_{n \in \mathbb{N}}\left\langle x, e_{n}\right\rangle e_{n}$ (which exists by Lemma 2.24 (i)). Then $\left\langle x-x_{0}, e_{n}\right\rangle=0$ for every $n \in \mathbb{N}$, and by assumption (ii), this implies $x=x_{0}$.
(iii) $\Rightarrow$ (iv) follows when multiplying $x$ scalarly with $y$, applying also the CauchySchwarz inequality for the sequences $\left(\left\langle x, e_{l}\right\rangle\right),\left(\left\langle e_{l}, y\right\rangle\right) \in l^{2}$.
(iv) $\Rightarrow(\mathrm{v})$ follows from putting $x=y$.
(v) $\Rightarrow$ (i). Let $x \in \operatorname{span}\left\{e_{n}: n \in \mathbb{N}\right\}^{\perp}$. Then Parseval's identity implies $\|x\|^{2}=0$, that is, $x=0$. By Theorem 2.10, $\operatorname{span}\left\{e_{n}: n \in \mathbb{N}\right\}$ is dense in $H$, that is, $\left(e_{n}\right)$ is an orthonormal basis.

A bounded linear operator $U \in \mathscr{L}(H, K)$ between two Hilbert spaces is called a unitary operator if it is invertible and for every $x, y \in H$,

$$
\langle x, y\rangle_{H}=\langle U x, U y\rangle_{K}
$$

Two Hilbert spaces $H$ and $K$ are unitarily equivalent if there exists a unitary operator $U \in \mathscr{L}(H, K)$.

Corollary 2.26. Every infinite dimensional separable Hilbert space $H$ is unitarily equivalent to $l^{2}$.

Proof. Choose an orthonormal basis $\left(e_{n}\right)_{n \in \mathbb{N}}$ of $H$ (which exists by Corollary 2.21), and define $U: H \rightarrow l^{2}$ by $U(x)=\left(\left\langle x, e_{n}\right\rangle\right)_{n \in \mathbb{N}}$. Then $\langle x, y\rangle_{H}=\langle U(x), U(y)\rangle_{l^{2}}$ by Theorem 2.25; in particular, $U$ is bounded, isometric and injective. The fact that $U$
is surjective, that is, that $\sum_{n} c_{n} e_{n}$ converges for every $c=\left(c_{n}\right) \in l^{2}$, follows as in the proof of Lemma 2.24 (i).

Clearly, if a sequence $\left(e_{n}\right)$ in a Hilbert space $H$ is an orthonormal basis, then necessarily $H$ is separable by Lemma 2.15 . Hence, the equivalent statements of Theorem 2.25 are only satisfied in separable Hilbert spaces. In most of the applications (if not all!), we will only deal with separable Hilbert spaces so that Theorem 2.25 is sufficient for our purposes.

However, what is true in general Hilbert spaces? The following sequence of results generalizes the preceeding results to arbitrary Hilbert spaces.

Let $X$ be a normed space, $\left(x_{i}\right)_{i \in I}$ be a family. We say that the series $\sum_{i \in I} x_{i}$ converges unconditionally if the set $I_{0}:=\left\{i \in I: x_{i} \neq 0\right\}$ is countable, and for every bijective $\varphi: \mathbb{N} \rightarrow I_{0}$ the series $\sum_{n=1}^{\infty} x_{\varphi(n)}$ converges.

Corollary 2.27 (Bessel's inequality, general case). Let $H$ be an inner product space, $\left(e_{l}\right)_{l \in I} \subseteq H$ an orthonormal system. Then, for every $x \in H$, the set $\{l \in I$ : $\left.\left\langle x, e_{l}\right\rangle \neq 0\right\}$ is countable and

$$
\begin{equation*}
\sum_{l \in I}\left|\left\langle x, e_{l}\right\rangle\right|^{2} \leq\|x\|^{2} \tag{2.3}
\end{equation*}
$$

Proof. By Bessel's inequality, the sets $\left\{l \in I:\left|\left\langle x, e_{l}\right\rangle\right| \geq 1 / n\right\}$ must be finite for every $n \in \mathbb{N}$. The countability of $\left\{l \in I:\left\langle x, e_{l}\right\rangle \neq 0\right\}$ follows. The inequality (2.3) is then a direct consequence of Bessel's inequality.

Lemma 2.28. Let $H$ be a Hilbert space, $\left(e_{l}\right)_{l \in I} \subseteq H$ an orthonormal system. Then:
a) For every $x \in H$, the series $\sum_{l \in I}\left\langle x, e_{l}\right\rangle e_{l}$ converges unconditionally.
b) $P: H \rightarrow H, x \mapsto \sum_{l \in I}\left\langle x, e_{l}\right\rangle e_{l}$ is the orthogonal projection onto $\overline{\operatorname{span}}\left\{e_{l}: l \in I\right\}$.

Corollary 2.29. Every Hilbert space admits an orthonormal basis.
Proof. If $H$ is separable, the claim follows directly from the Gram-Schmidt process and has already been stated in Corollary 2.21. In general, one may argue as follows:

The set of all orthonormal systems in $H$ forms a partially ordered set by inclusion. Given a totally ordered collection of orthonormal systems, the union of all vectors contained in all systems in this collection forms a supremum. By Zorn's lemma, there exists an orthonormal system $\left(e_{l}\right)_{l \in I}$ which is maximal. It follows from Bessel's inequality (2.3) that this system is actually an orthonormal basis.

Theorem 2.25 remains true for arbitrary Hilbert spaces when replacing the countable orthonormal system $\left(e_{n}\right)_{n \in \mathbb{N}}$ by an arbitrary orthonormal system $\left(e_{l}\right)_{l \in I}$.

## $2.3 *$ Fourier series

In the following we will identify the space $L^{1}(0,2 \pi)$ with

$$
L_{2 \pi}^{1}(\mathbb{R}):=\left\{f: \mathbb{R} \rightarrow \mathbb{C} \text { measurable, } 2 \pi \text {-periodic : } \int_{0}^{2 \pi}|f| \mathrm{d} \lambda<\infty\right\}
$$

Similarly, we identify $L^{2}(0,2 \pi)$ with $L_{2 \pi}^{2}(\mathbb{R})$, and we define

$$
C_{2 \pi}(\mathbb{R}):=\{f \in C(\mathbb{R}): f \text { is } 2 \pi \text {-periodic }\}
$$

For every $f \in L^{1}(0,2 \pi)=L_{2 \pi}^{1}(\mathbb{R})$ and every $n \in \mathbb{Z}$ we call

$$
\hat{f}(n):=\frac{1}{2 \pi} \int_{0}^{2 \pi} f(t) e^{-i n t} \mathrm{~d} t
$$

the $n$-th Fourier coefficient of $f$. The sequence $\hat{f}=(\hat{f}(n))$ is called the Fourier transform of $f$. The formal series $\frac{1}{\sqrt{2 \pi}} \sum_{n \in \mathbb{Z}} \hat{f}(n) e^{i n \cdot}$ is called the Fourier series of $f$.

Lemma 2.30. For every $f \in L^{1}(0,2 \pi)=L_{2 \pi}^{1}(\mathbb{R})$ we have $\hat{f} \in l^{\infty}(\mathbb{Z})$ and the Fourier transform ${ }^{\wedge}: L^{1}(0,2 \pi) \rightarrow l^{\infty}$ is a bounded, linear operator. More precisely,

$$
\|\hat{f}\|_{\infty} \leq \frac{1}{2 \pi}\|f\|_{1}, \quad f \in L^{1}(0,2 \pi)
$$

Proof. For every $f \in L^{1}(0,2 \pi)$ and every $n \in \mathbb{Z}$,

$$
|\hat{f}(n)|=\frac{1}{2 \pi}\left|\int_{0}^{2 \pi} f(t) e^{-i n t} \mathrm{~d} t\right| \leq \frac{1}{2 \pi} \int_{0}^{2 \pi}|f(t)| d t
$$

This proves that $\hat{f} \in l^{\infty}$ and the required bound on $\|\hat{f}\|_{\infty}$. Linearity of ${ }^{\wedge}$ is clear.
Lemma 2.31 (Riemann-Lebesgue). For every $f \in L^{1}(0,2 \pi)=L_{2 \pi}^{1}(\mathbb{R})$ we have $\hat{f} \in$ $c_{0}(\mathbb{Z})$, i.e.

$$
\lim _{|n| \rightarrow \infty}|\hat{f}(n)|=0
$$

Proof. Let $f \in L^{1}(0,2 \pi)=L_{2 \pi}^{1}(\mathbb{R})$ and $n \in \mathbb{Z}, n \neq 0$. Then

$$
\begin{aligned}
\hat{f}(n) & =\frac{1}{2 \pi} \int_{0}^{2 \pi} f(t) e^{-i n t} \mathrm{~d} t \\
& =\frac{1}{4 \pi} \int_{0}^{2 \pi} f(t) e^{-i n t}\left(1-e^{i \pi \frac{n}{n}}\right) \mathrm{d} t \\
& =\frac{1}{4 \pi} \int_{0}^{2 \pi} f(t)\left(e^{-i n t}-e^{-i n\left(t-\frac{\pi}{n}\right)}\right) \mathrm{d} t \\
& =\frac{1}{4 \pi} \int_{0}^{2 \pi}\left(f(t)-f\left(t+\frac{\pi}{n}\right)\right) e^{-i n t} \mathrm{~d} t
\end{aligned}
$$

so that

$$
|\hat{f}(n)| \leq \frac{1}{4 \pi} \int_{0}^{2 \pi}\left|f(t)-f\left(t+\frac{\pi}{n}\right)\right| \mathrm{d} t
$$

Hence, if $f=1_{O} \in L^{1}(0,2 \pi)$ for some open set $O \subseteq[0,2 \pi]$, then $\hat{f} \in c_{0}(\mathbb{Z})$ by Lebesgue dominated convergence theorem. On the other hand, since span $\left\{1_{O}: O \subseteq\right.$ $[0,2 \pi]$ open $\}$ is dense in $L^{1}(0,2 \pi)$, since the Fourier transform is bounded with values in $l^{\infty}(\mathbb{Z})$ (Lemma 2.30), and since $c_{0}(\mathbb{Z})$ is a closed subspace of $l^{\infty}(\mathbb{Z})$, we find that $\hat{f} \in c_{0}(\mathbb{Z})$ for every $f \in L^{1}(0,2 \pi)$.

Remark 2.32. At the end of the proof of the Lemma of Riemann-Lebesgue, we used the following general principle: if $T \in \mathscr{L}(X, Y)$ is a bounded linear operator between two normed linear spaces $X, Y$, and if $M \subseteq X$ is dense, then $\operatorname{ran} T \subseteq \overline{T(M)}$. We used in addition that $c_{0}(\mathbb{Z})$ is closed in $l^{\infty}(\mathbb{Z})$.

Theorem 2.33. Let $f \in C_{2 \pi}(\mathbb{R})$ be differentiable in some point $s \in \mathbb{R}$. Then

$$
f(s)=\sum_{n \in \mathbb{Z}} \hat{f}(n) e^{i n s}
$$

Proof. Note that for $f_{s}(t):=f(s+t)$,

$$
\hat{f}_{s}(n)=\frac{1}{2 \pi} \int_{0}^{2 \pi} f(s+t) e^{-i n t} \mathrm{~d} t=\frac{1}{2 \pi} \int_{0}^{2 \pi} f(t) e^{-i n(t-s)} \mathrm{d} t=e^{i n s} \hat{f}(n)
$$

Hence, replacing $f$ by $f_{s}$, if necessary, we may without loss of generality assume that $s=0$. Moreover, replacing $f$ by $f-f(0)$, if necessary, we may without loss of generality assume that $f(0)=0$. We hence have to show that if $f$ is differentiable in 0 and if $f(0)=0$, then $\sum_{n \in \mathbb{Z}} \hat{f}(n)=0$.

Let $g(t):=\frac{f(t)}{1-e^{i t}}$. Since $f$ is differentiable in $0, f(0)=0$, and since $f$ is $2 \pi$ periodic, the function $g$ belongs to $C_{2 \pi}(\mathbb{R})$. By the Lemma of Riemann-Lebesgue, $\hat{g} \in c_{0}(\mathbb{Z})$. Note that

$$
\hat{f}(n)=\frac{1}{2 \pi} \int_{0}^{2 \pi} g(t)\left(1-e^{i t}\right) e^{-i n t} \mathrm{~d} t=\hat{g}(n)-\hat{g}(n-1) .
$$

Hence,

$$
\begin{aligned}
\sum_{k=-n}^{n} \hat{f}(k) & =\sum_{k=-n}^{n} \hat{g}(k)-\hat{g}(k-1) \\
& =\hat{g}(n)-\hat{g}(-n-1) \rightarrow 0 \quad(n \rightarrow \infty)
\end{aligned}
$$

This is the claim.
Corollary 2.34. For every $f \in C_{2 \pi}^{1}(\mathbb{R}):=C_{2 \pi}(\mathbb{R}) \cap C^{1}(\mathbb{R})$ and every $t \in \mathbb{R}$

$$
f(t)=\sum_{n \in \mathbb{Z}} \hat{f}(n) e^{i n t}
$$

Remark 2.35. We will see that the convergence in the preceeding corollary is even uniform in $t \in \mathbb{R}$.

Throughout the following, we equip the space $L^{2}(0,2 \pi)=L_{2 \pi}^{2}(\mathbb{R})$ with the scalar product given by

$$
\langle f, g\rangle:=\frac{1}{2 \pi} \int_{0}^{2 \pi} f(t) \overline{g(t)} \mathrm{d} t
$$

which differs from the usual scalar product by the factor $\frac{1}{2 \pi}$.
Lemma 2.36. The space $C_{2 \pi}^{1}(\mathbb{R})$ is dense in $L_{2 \pi}^{2}(\mathbb{R})$.
Proof. We first prove that $C([0,2 \pi])$ is dense in $L^{2}(0,2 \pi)=L_{2 \pi}^{2}(\mathbb{R})$. For this, consider first a characteristic function $f=1_{(a, b)} \in L^{2}(0,2 \pi)$. Let $\left(g_{n}\right) \subseteq C([0,2 \pi])$ be defined by

$$
g_{n}(t):= \begin{cases}1, & t \in[a, b] \\ 1+n(t-a), & t \in[a-1 / n, a) \\ 1-n(t-b), & t \in(b, b+1 / n] \\ 0, & \text { else }\end{cases}
$$

It is then easy to see that $\lim _{n \rightarrow \infty}\left\|f-g_{n}\right\|_{L^{2}}=0$, so that $f=1_{(a, b)} \in \overline{C([0,2 \pi])}\|\cdot\|_{L^{2}}$.
In the second step, consider a characteristic function $f=1_{A}$ of an arbitrary Borel set $A \in \mathscr{B}([0,2 \pi])$, and let $\varepsilon>0$. By outer regularity of the Lebesgue measure, there exists an open set $O \supset A$ such that $\lambda(O \backslash A)<\varepsilon^{2}$. Recall that $O$ is the countable union of mutually disjoint intervals. Since $O$ has finite measure, there exist finitely many (mutually disjoint) intervals $\left(a_{n}, b_{n}\right) \subseteq O(1 \leq n \leq N)$ such that $\lambda\left(O \backslash \bigcup_{n=1}^{N}\left(a_{n}, b_{n}\right)\right) \leq \varepsilon^{2}$. By the preceeding step, for every $1 \leq n \leq N$ there exists $g_{n} \in C([0,2 \pi])$ such that $\left\|1_{\left(a_{n}, b_{n}\right)}-g_{n}\right\|_{2} \leq \frac{\varepsilon}{N}$. Let $g:=\sum_{n=1}^{N} g_{n} \in C([0,2 \pi])$. Then

$$
\begin{aligned}
\|f-g\|_{2} & \leq\left\|1_{A}-1_{O}\right\|_{2}+\left\|1_{O}-1_{\bigcup_{n=1}^{N}\left(a_{n}, b_{n}\right)}\right\|_{2}+\left\|1_{\cup_{n=1}^{N}\left(a_{n}, b_{n}\right)}-g\right\|_{2} \\
& \leq \varepsilon+\varepsilon+\left\|\sum_{n=1}^{N}\left(1_{\left(a_{n}, b_{n}\right)}-g_{n}\right)\right\|_{2} \\
& \leq 3 \varepsilon .
\end{aligned}
$$

 $A \in \mathscr{B}([0,2 \pi])\}=L^{2}(0,2 \pi)$, we find that $C([0,2 \pi])$ is dense in $L^{2}(0,2 \pi)$.

It remains to show that $C_{2 \pi}^{1}(\mathbb{R})$ is dense in $C([0,2 \pi])$ for the norm $\|\cdot\|_{2}$. So let $f \in C([0,2 \pi])$ and let $\varepsilon>0$. By Weierstrass' theorem, there exists a function $g_{0} \in C^{\infty}([0,2 \pi])$ (even a polynomial!) such that $\left\|f-g_{0}\right\|_{\infty} \leq \varepsilon$. Let $g_{1} \in C^{1}([0,2 \pi])$ be such that $g_{1}(2 \pi)=g_{1}^{\prime}(2 \pi)=0, g_{1}(0)=g_{0}(2 \pi)-g_{0}(0)$ and $g_{1}^{\prime}(0)=g_{0}^{\prime}(2 \pi)-$ $g_{0}^{\prime}(0)$ and $\left\|g_{1}\right\|_{2} \leq \varepsilon$. Such a function $g_{1}$ exists: it suffices for example to consider functions for which the derivative is of the form

$$
g_{1}^{\prime}(t)= \begin{cases}g_{0}(2 \pi)-g_{0}(0)+c t, & t \in\left[0, h_{1}\right] \\ g_{0}(2 \pi)-g_{0}(0)+c h_{1}+d\left(t-h_{1}\right), & t \in\left(h_{1}, h_{2}\right), \\ 0, & t \in\left[h_{2}, 2 \pi\right]\end{cases}
$$

with appropriate constants $0 \leq h_{1} \leq h_{2}$ and $c, d \in \mathbb{C}$. Having chosen $g_{1}$, we let $g=g_{0}+g_{1}$ and we calculate that

$$
\|f-g\|_{2} \leq\left\|f-g_{0}\right\|_{2}+\left\|g_{1}\right\|_{2} \leq 2 \varepsilon .
$$

Since $g$ extends to a function in $C_{2 \pi}^{1}(\mathbb{R})$, we have thus proved that $C_{2 \pi}^{1}(\mathbb{R})$ is dense in $L_{2 \pi}^{2}(\mathbb{R})$.
Remark 2.37. An adaptation of the above proof actually shows that for every $1 \leq$ $p<\infty$ and every compact interval $[a, b] \subseteq \mathbb{R}$, the space $C([a, b])$ is dense in $L^{p}(a, b)$. A further application of Weierstrass' theorem actually shows that the space of all polynomials is dense in $L^{p}(a, b)$. In particular, we may obtain the following result.
Corollary 2.38. The space $L^{p}(a, b)$ is separable if $1 \leq p<\infty$. The space $L^{\infty}(a, b)$ is not separable.
Corollary 2.39. Let $e_{n}(t):=e^{\text {int }}, n \in \mathbb{Z}, t \in \mathbb{R}$. Then $\left(e_{n}\right)_{n \in \mathbb{Z}}$ is an orthonormal basis in $L_{2 \pi}^{2}(\mathbb{R})$.
Proof. The fact that $\left(e_{n}\right)_{n \in \mathbb{Z}}$ is an orthonormal system in $L_{2 \pi}^{2}(\mathbb{R})$ is an easy calculation. We only have to prove that $\operatorname{span}\left\{e_{n}: n \in \mathbb{Z}\right\}$ is dense in $L_{2 \pi}^{2}(\mathbb{R})$. Note that $\hat{f}(n)=\left\langle f, e_{n}\right\rangle$ for every $f \in L_{2 \pi}^{2}(\mathbb{R})$ and every $n \in \mathbb{Z}$. By Lemma 2.24, we know that for every $f \in L_{2 \pi}^{2}(\mathbb{R})$

$$
g:=\sum_{n \in \mathbb{Z}} \hat{f}(n) e_{n} \text { exists in } L_{2 \pi}^{2}(\mathbb{R})
$$

In particular, a subsequence of $\left(\sum_{n=-k}^{k} \hat{f}(n) e_{n}\right)$ converges almost everywhere to $g$. But by Corollary 2.34 we know that $\left(\sum_{n=-k}^{k} \hat{f}(n) e_{n}\right)$ converges pointwise everywhere to $f$ if $f \in C_{2 \pi}^{1}(\mathbb{R})$. As a consequence, for every $f \in C_{2 \pi}^{1}(\mathbb{R})$,

$$
\lim _{k \rightarrow \infty} \sum_{n=-k}^{k} \hat{f}(n) e_{n}=f \text { in } L_{2 \pi}^{2}(\mathbb{R})
$$

so that span $\left\{e_{n}: n \in \mathbb{Z}\right\}$ is dense in $\left(C_{2 \pi}^{1}(\mathbb{R}),\|\cdot\|_{L_{2 \pi}^{2}}\right)$. Since $C_{2 \pi}^{1}(\mathbb{R})$ is dense in $L_{2 \pi}^{2}(\mathbb{R})$ by Lemma 2.36, we find that $\left(e_{n}\right)_{n \in \mathbb{Z}}$ is an orthonormal basis in $L_{2 \pi}^{2}(\mathbb{R})$.
Theorem 2.40 (Plancherel). For every $f \in L_{2 \pi}^{2}(\mathbb{R})$ we have $\hat{f} \in l^{2}(\mathbb{Z})$ and the Fourier transform ${ }^{\wedge}: L_{2 \pi}^{2}(\mathbb{R}) \rightarrow l^{2}(\mathbb{Z})$ is an isometric isomorphism. Moreover, for every $f \in L_{2 \pi}^{2}(\mathbb{R})$,

$$
\sum_{n \in \mathbb{Z}} \hat{f}(n) e_{n}=f \text { in } L_{2 \pi}^{2}(\mathbb{R})
$$

that is, the Fourier series of $f$ converges to $f$ in the $L^{2}$ sense.
Proof. By Corollary 2.39, the sequence $\left(e_{n}\right)_{n \in \mathbb{Z}}$ is an orthonormal basis in $L_{2 \pi}^{2}(\mathbb{R})$. Moreover, recall that for every $f \in L_{2 \pi}^{2}(\mathbb{R})$ and every $n \in \mathbb{Z}, \hat{f}(n)=\left\langle f, e_{n}\right\rangle$. Hence, by Theorem $2.25, \hat{f} \in l^{2}(\mathbb{Z}), f=\sum_{n \in \mathbb{Z}} \hat{f}(n) e_{n}$, and $\|f\|_{L_{2 \pi}^{2}}=\|\hat{f}\|_{l^{2}}$ (the last property being Parseval's identity).

Corollary 2.41. Let $f \in C_{2 \pi}(\mathbb{R})$ be such that $\hat{f} \in l^{1}(\mathbb{Z})$. Then

$$
\sum_{n \in \mathbb{Z}} \hat{f}(n) e_{n}=f \text { in } C_{2 \pi}(\mathbb{R})
$$

that is, the Fourier series of $f$ converges uniformly to $f$.
Proof. Note that for every $n \in \mathbb{Z},\left\|e_{n}\right\|_{\infty}=1$. The assumption $\hat{f} \in l^{1}(\mathbb{Z})$ therefore implies that the series $\sum_{n \in \mathbb{Z}} \hat{f}(n) e_{n}$ converges absolutely in $C_{2 \pi}(\mathbb{R})$, i.e. for the uniform norm $\|\cdot\|_{\infty}$. Since $\left(C_{2 \pi}(\mathbb{R}),\|\cdot\|_{\infty}\right)$ is complete, the series $\sum_{n \in \mathbb{Z}} \hat{f}(n) e_{n}$ converges uniformly to some element $g \in C_{2 \pi}(\mathbb{R})$. By Plancherel, $g=f$.

Remark 2.42. The assumption $\hat{f} \in l^{1}(\mathbb{Z})$ in Corollary 2.41 is essential. For general $f \in C_{2 \pi}(\mathbb{R})$, the Fourier series $\sum_{n \in \mathbb{Z}} \hat{f}(n) e_{n}$ need not not converge uniformly. Questions regarding the convergence of Fourier series (which type of convergence? for which function?) can go deeply into the theory of harmonic analysis and answers are sometimes quite involved. The $L^{2}$ theory gives in this context satisfactory answers with relatively easy proofs (see Plancherel's theorem). For continuous functions we state the following result without giving a proof.

Theorem 2.43 (Féjer). For every $f \in C_{2 \pi}(\mathbb{R})$ one has

$$
\lim _{K \rightarrow \infty} \frac{1}{K} \sum_{k=1}^{K} \sum_{n=-k}^{k} \hat{f}(n) e_{n}=f \text { in } C_{2 \pi}(\mathbb{R})
$$

that is, the Fourier series of $f$ converges in the Cesàro mean uniformly to $f$.

### 2.4 Linear functionals on Hilbert spaces

In this section, we discuss bounded functionals on Hilbert spaces. Compared to the case of bounded linear functionals on general Banach spaces, the case of bounded linear functionals on Hilbert spaces is considerably easy but it has far reaching consequences.

Theorem 2.44 (Riesz-Fréchet). Let $H$ be a Hilbert space. Then for every bounded linear functional $\varphi \in H^{\prime}$ there exists a unique $y \in H$ such that

$$
\varphi(x)=\langle x, y\rangle \quad \text { for every } x \in H
$$

Proof. Uniqueness. Let $y_{1}, y_{2} \in H$ be two elements such that

$$
\varphi(x)=\left\langle x, y_{1}\right\rangle=\left\langle x, y_{2}\right\rangle \quad \text { for every } x \in H
$$

Then $\left\langle x, y_{1}-y_{2}\right\rangle=0$ for every $x \in H$, in particular also for $x=y_{1}-y_{2}$. This implies $\left\|y_{1}-y_{2}\right\|^{2}=0$, that is, $y_{1}=y_{2}$.

Existence. We may assume that $\varphi \neq 0$ since the case $\varphi=0$ is trivial. Let $\tilde{y} \in$ $(\operatorname{ker} \varphi)^{\perp} \backslash\{0\}$. Since $H \neq \operatorname{ker} \varphi$ and since $\operatorname{ker} \varphi$ is closed, such a $\tilde{y}$ exists. Next, let

$$
y:=\overline{\varphi(\tilde{y})} /\|\tilde{y}\|^{2} \tilde{y} .
$$

Note that $\varphi(y)=\|y\|^{2}=\langle y, y\rangle$. Recall that every $x \in H$ can be uniquely written as $x=x_{0}+\lambda y$ with $x_{0} \in \operatorname{ker} \varphi$ and $\lambda \in \mathbb{K}$ so that $\lambda y \in(\operatorname{ker} \varphi)^{\perp}$. Note that $(\operatorname{ker} \varphi)^{\perp}$ is one-dimensional. Hence, for every $x \in H$,

$$
\begin{aligned}
\varphi(x) & =\varphi\left(x_{0}+\lambda y\right) \\
& =\varphi\left(x_{0}\right)+\lambda \varphi(y) \\
& =\lambda \varphi(y) \\
& =\lambda\langle y, y\rangle \\
& =\langle\lambda y, y\rangle \\
& =\left\langle x_{0}, y\right\rangle+\langle\lambda y, y\rangle \\
& =\langle x, y\rangle .
\end{aligned}
$$

The claim is proved.
Corollary 2.45. Let $J: H \rightarrow H^{\prime}$ be the mapping which maps to every $y \in H$ the functional $J y \in H^{\prime}$ given by $J y(x)=\langle x, y\rangle$. Then $J$ is antilinear if $\mathbb{K}=\mathbb{C}$ and linear if $\mathbb{K}=\mathbb{R}$. Moreover, $J$ is isometric and bijective.

Proof. The fact that $J$ is isometric follows from the Cauchy-Schwarz inequality. Antilinearity (or linearity in case $\mathbb{K}=\mathbb{R}$ ) follows from the sesquilinearity (resp. bilinearity) of the scalar product on $H$. Since $J$ is isometric, it is injective. The surjectivity of $J$ follows from Theorem 2.44.

Remark 2.46. The theorem of Riesz-Fréchet allows us to identify any (real) Hilbert space $H$ with its dual space $H^{\prime}$. Note, however, that there are situations in which one does not identify $H^{\prime}$ with $H$. This is for example the case when $V$ is a second Hilbert space which embeds continuously and densely into $H$, that is, for which there exists a bounded, injective $J: V \rightarrow H$ with dense range.

### 2.5 Weak convergence in Hilbert spaces

Let $H$ be a Hilbert space. We say that a sequence $\left(x_{n}\right) \subseteq H$ converges weakly to some element $x \in H$ if for every $y \in H$ one has $\lim _{n \rightarrow \infty}\left\langle x_{n}, y\right\rangle=\langle x, y\rangle$. We write $x_{n} \rightharpoonup x$ or $x_{n} \xrightarrow{\text { weak }} x$ if $\left(x_{n}\right)$ converges weakly to $x$.

Theorem 2.47. Every bounded sequence $\left(x_{n}\right)$ in a Hilbert space $H$ admits a weakly convergent subsequence, that is, there exists $x \in H$ and there exists a subsequence $\left(x_{n_{k}}\right)$ of $\left(x_{n}\right)$ such that $x_{n_{k}} \xrightarrow{\text { weak }} x$.

In the proof of this theorem, we will use the following general result.
Lemma 2.48. Let $X$ and $Y$ be two normed spaces, let $\left(T_{n}\right) \in \mathscr{L}(X, Y)$ be a bounded sequence of bounded operators. Assume that there exists a dense set $M \subseteq X$ such that $\lim _{n \rightarrow \infty} T_{n} x$ exists for every $x \in M$. Then $\lim _{n \rightarrow \infty} T_{n} x=:$ Tx exists for every $x \in X$ and $T \in \mathscr{L}(X, Y)$.
Proof. Define $T x:=\lim _{n \rightarrow \infty} T_{n} x$ for every $x \in \operatorname{span} M$. Then

$$
\|T x\|=\lim _{n \rightarrow \infty}\left\|T_{n} x\right\| \leq \sup _{n}\left\|T_{n}\right\|\|x\|,
$$

that is. $T: \operatorname{span} M \rightarrow Y$ is a bounded linear operator. Since $M$ is dense in $X, T$ admits a unique bounded extension $T: X \rightarrow Y$.

Let $x \in X$ and $\varepsilon>0$. Since $M$ is dense in $X$, there exists $y \in M$ such that $\|x-y\| \leq$ $\varepsilon$. By assumption, there exists $n_{0}$ such that for every $n \geq n_{0}$ we have $\left\|T_{n} y-T y\right\| \leq \varepsilon$. Hence, for every $n \geq n_{0}$,

$$
\begin{aligned}
\left\|T_{n} x-T x\right\| & \leq\left\|T_{n} x-T_{n} y\right\|+\left\|T_{n} y-T y\right\|+\|T y-T x\| \\
& \leq \sup _{n}\left\|T_{n}\right\|\|x-y\|+\varepsilon+\|T\|\|x-y\| \\
& \leq \varepsilon\left(\sup _{n}\left\|T_{n}\right\|+1+\|T\|\right),
\end{aligned}
$$

and therefore $\lim _{n \rightarrow \infty} T_{n} x=T x$.
Proof (of Theorem 2.47). As in the proof of the Arzela-Ascoli theorem (Theorem 1.36), we use Cantor's diagonal sequence argument. Let $\left(x_{n}\right)$ be a bounded sequence in $H$. We first assume that $H$ is separable, and we let $\left(y_{m}\right) \subseteq H$ be a dense sequence.

Since $\left(\left\langle x_{n}, y_{1}\right\rangle\right)$ is bounded by the boundedness of $\left(x_{n}\right)$, there exists a subsequence $\left(x_{\varphi_{1}(n)}\right)$ of $\left(x_{n}\right)\left(\varphi_{1}: \mathbb{N} \rightarrow \mathbb{N}\right.$ is increasing, unbounded) such that

$$
\lim _{n \rightarrow \infty}\left\langle x_{\varphi_{1}(n)}, y_{1}\right\rangle \text { exists. }
$$

Similarly, there exists a subsequence $\left(x_{\varphi_{2}(n)}\right)$ of $\left(x_{\varphi_{1}(n)}\right)$ such that

$$
\lim _{n \rightarrow \infty}\left\langle x_{\varphi_{2}(n)}, y_{2}\right\rangle \text { exists. }
$$

Note that for this subsequence, we also have that

$$
\lim _{n \rightarrow \infty}\left\langle x_{\varphi_{2}(n)}, y_{1}\right\rangle \text { exists. }
$$

Iterating this argument, we find a subsequence $\left(x_{\varphi_{3}(n)}\right)$ of $\left(x_{\varphi_{2}(n)}\right)$ and finally for every $m \in \mathbb{N}, m \geq 2$, a subsequence $\left(x_{\varphi_{m}(n)}\right)$ of $\left(x_{\varphi_{m-1}(n)}\right)$ such that

$$
\lim _{n \rightarrow \infty}\left\langle x_{\varphi_{m}(n)}, y_{j}\right\rangle \text { exists for every } 1 \leq j \leq m
$$

Let $\left(x_{n}^{\prime}\right):=\left(x_{\varphi_{n}(n)}\right)$ be the 'diagonal sequence'. Then $\left(x_{n}^{\prime}\right)$ is a subsequence of $\left(x_{n}\right)$ and

$$
\lim _{n \rightarrow \infty}\left\langle x_{n}^{\prime}, y_{m}\right\rangle \text { exists for every } m \in \mathbb{N}
$$

By Lemma 2.48 and the Riesz-Fréchet representation theorem (Theorem 2.44), there exists $x \in H$ such that

$$
\lim _{n \rightarrow \infty}\left\langle x_{n}^{\prime}, y\right\rangle=\langle x, y\rangle \text { for every } y \in H
$$

and the claim is proved in the case when $H$ is separable.
If $H$ is not separable as we first assumed, then one may replace $H$ by $\tilde{H}$ := $\overline{\operatorname{span}}\left\{x_{n}: n \in \mathbb{N}\right\}$ which is separable. By the above, there exists $x \in \tilde{H}$ and a subsequence of $\left(x_{n}\right)$ (which we denote again by $\left(x_{n}\right)$ ) such that for every $y \in \tilde{H}$,

$$
\lim _{n \rightarrow \infty}\left\langle x_{n}, y\right\rangle=\langle x, y\rangle
$$

that is, $\left(x_{n}\right)$ converges weakly in $\tilde{H}$. On the other hand, for every $y \in \tilde{H}^{\perp}$ and every $n$,

$$
\left\langle x_{n}, y\right\rangle=\langle x, y\rangle=0 .
$$

The decomposition $H=\tilde{H} \oplus \tilde{H}^{\perp}$ therefore yields that $\left(x_{n}\right)$ converges weakly in $H$.

## Chapter 3 <br> Dual spaces and weak convergence

### 3.1 The theorem of Hahn-Banach

Given a normed space $X$, we denote by $X^{\prime}:=\mathscr{L}(X, \mathbb{K})$ the space of all bounded linear functionals on $X$. It is called the dual space of $X$. Recall that $X^{\prime}$ is always a Banach space by Corollary 1.29 of Chapter 1.

However, a priori it is not clear whether there exists any bounded linear functional on a normed space $X$ (apart from the zero functional). This fundamental question and the analysis of dual spaces (analysis of functionals) shall be developed in this chapter.

The existence of nontrivial bounded functionals is guaranteed by the HahnBanach theorem which actually admits several versions. However, before stating the first version, we need the following definition.

Let $X$ be a real or complex vector space. A function $p: X \rightarrow \mathbb{R}$ is called sublinear if
(i) $p(\lambda x)=\lambda p(x)$ for every $\lambda>0, x \in X$, and
(ii) $p(x+y) \leq p(x)+p(y)$ for every $x, y \in X$.

Example 3.1. On a normed space $X$, the norm $\|\cdot\|$ is sublinear. Every linear $p$ : $X \rightarrow \mathbb{R}$ is sublinear.

Theorem 3.2 (Hahn-Banach; version of linear algebra, real case). Let $X$ be $a$ real vector space, $U \subseteq X$ a linear subspace, and $p: X \rightarrow \mathbb{R}$ sublinear. Let $\varphi: U \rightarrow \mathbb{R}$ be linear such that

$$
\varphi(x) \leq p(x) \text { for all } x \in U
$$

Then there exists a linear $\tilde{\varphi}: X \rightarrow \mathbb{R}$ such that $\tilde{\varphi}(x)=\varphi(x)$ for every $x \in U$ (that is, $\tilde{\varphi}$ is an extension of $\varphi$ ) and

$$
\begin{equation*}
\tilde{\varphi}(x) \leq p(x) \text { for all } x \in X \tag{3.1}
\end{equation*}
$$

The following lemma asserts that this version of Hahn-Banach is true in the special case when $X / U$ has dimension 1 . It is an essential step in the proof of Theorem 3.2.

Lemma 3.3. Take the assumptions of Theorem 3.2 and assume in addition that $\operatorname{dim} X / U=1$. Then the assertion of Theorem 3.2 is true.
Proof. If $\operatorname{dim} X / U=1$, then there exists $x_{0} \in X \backslash U$ such that every $x \in X$ can be uniquely written in the form $x=u+\lambda x_{0}$ with $u \in U$ and $\lambda \in \mathbb{R}$. So we define $\tilde{\varphi}: X \rightarrow \mathbb{R}$ by

$$
\tilde{\varphi}(x):=\tilde{\varphi}\left(u+\lambda x_{0}\right):=\varphi(u)+\lambda r,
$$

where $r \in \mathbb{R}$ is a parameter which has to be chosen such that (3.1) holds, that is, such that for every $u \in U, \lambda \in \mathbb{R}$,

$$
\begin{equation*}
\varphi(u)+\lambda r \leq p\left(u+\lambda x_{0}\right) . \tag{3.2}
\end{equation*}
$$

If $\lambda=0$, then this condition clearly holds for every $u \in U$ by the assumption on $\varphi$. If $\lambda>0$, then (3.2) holds for every $u \in U$ if and only if

$$
\begin{aligned}
& \lambda r \leq p\left(u+\lambda x_{0}\right)-\varphi(u) \text { for every } u \in U \\
& \Leftrightarrow r \leq p\left(\frac{u}{\lambda}+x_{0}\right)-\varphi\left(\frac{u}{\lambda}\right) \text { for every } u \in U \\
& \Leftrightarrow r \leq \inf _{v \in U} p\left(v+x_{0}\right)-\varphi(v)
\end{aligned}
$$

Similarly, if $\lambda<0$, then (3.2) holds for every $u \in U$ if and only if

$$
\begin{aligned}
& \lambda r \leq p\left(u+\lambda x_{0}\right)-\varphi(u) \text { for every } u \in U \\
& \Leftrightarrow-r \leq p\left(\frac{u}{-\lambda}-x_{0}\right)-\varphi\left(\frac{u}{-\lambda}\right) \text { for every } u \in U \\
& \Leftrightarrow r \geq \sup _{w \in U} \varphi(w)-p\left(w-x_{0}\right) .
\end{aligned}
$$

So it is possible to find an appropriate $r \in \mathbb{R}$ in the definition of $\tilde{\varphi}$ if and only if

$$
\varphi(w)-p\left(w-x_{0}\right) \leq p\left(v+x_{0}\right)-\varphi(v) \text { for all } v, w \in U,
$$

or, equivalently, if

$$
\varphi(w)+\varphi(v) \leq p\left(v+x_{0}\right)+p\left(w-x_{0}\right) \text { for all } v, w \in U .
$$

However, by the assumptions on $\varphi$ and $p$, for every $v, w \in U$,
$\varphi(w)+\varphi(v)=\varphi(w+v) \leq p(w+v)=p\left(v+x_{0}+w-x_{0}\right) \leq p\left(v+x_{0}\right)+p\left(w-x_{0}\right)$.
For the second step in the proof of Theorem 3.2, we need the Lemma of Zorn.
Lemma 3.4 (Zorn). Let $(M, \leq)$ be a ordered set. Assume that every totally ordered subset $T \subseteq M$ (i.e. for every $x, y \in T$ one either has $x \leq y$ or $y \leq x$ ) admits an
upper bound. Then for every $x \in M$ there exists a maximal element $m \geq x$ (that is, an element $m$ such that $m \leq \tilde{m}$ implies $m=\tilde{m}$ for every $\tilde{m} \in M$ ).

Proof (of Theorem 3.2). Define the following set

$$
\begin{gathered}
M:=\left\{\left(V, \varphi_{V}\right): V \subseteq X \text { linear subspace, } U \subseteq V, \varphi_{V}: V \rightarrow \mathbb{R}\right. \text { linear, s.t. } \\
\left.\varphi(x)=\varphi_{V}(x)(x \in U) \text { and } \varphi_{V}(x) \leq p(x)(x \in V)\right\},
\end{gathered}
$$

and equip it with the order relation $\leq$ defined by

$$
\left(V_{1}, \varphi_{V_{1}}\right) \leq\left(V_{2}, \varphi_{V_{2}}\right): \Leftrightarrow V_{1} \subseteq V_{2} \text { and } \varphi_{V_{1}}(x)=\varphi_{V_{2}}(x) \text { for all } x \in V_{1} .
$$

Then $(M, \leq)$ is an ordered set. Let $T=\left(\left(V_{i}, \varphi_{V_{i}}\right)\right)_{i \in I} \subseteq M$ be a totally ordered subset. Then the element $\left(V, \varphi_{V}\right) \in M$ defined by

$$
V:=\bigcup_{i \in I} V_{i} \text { and } \varphi_{V}(x)=\varphi_{V_{i}}(x) \text { for } x \in V_{i}
$$

is an upper bound of $T$. By the Lemma of Zorn, the set $M$ admits a maximal element $\left(X_{0}, \varphi_{X_{0}}\right)$. Assume that $X_{0} \neq X$. Then, by Lemma 3.3, we could construct an element which is strictly larger than $\left(X_{0}, \varphi_{X_{0}}\right)$, a contradiction to the maximality of $\left(X_{0}, \varphi_{X_{0}}\right)$. Hence, $X=X_{0}$, and $\tilde{\varphi}:=\varphi_{X_{0}}$ is an element we are looking for.

The complex version of the Hahn-Banach theorem reads as follows.
Theorem 3.5 (Hahn-Banach; version of linear algebra, complex case). Let $X$ be a complex vector space, $U \subseteq X$ a linear subspace, and $p: X \rightarrow \mathbb{R}$ sublinear. Let $\varphi: U \rightarrow \mathbb{C}$ be linear such that

$$
\operatorname{Re} \varphi(x) \leq p(x) \text { for all } x \in U .
$$

Then there exists a linear $\tilde{\varphi}: X \rightarrow \mathbb{C}$ such that $\tilde{\varphi}(x)=\varphi(x)$ for every $x \in U$ (that is $\tilde{\varphi}$ is an extension of $\varphi$ ) and

$$
\begin{equation*}
\operatorname{Re} \tilde{\varphi}(x) \leq p(x) \text { for all } x \in X \tag{3.3}
\end{equation*}
$$

Proof. We may consider $X$ also as a real vector space. Note that $\psi(x):=\operatorname{Re} \varphi(x)$ is an $\mathbb{R}$-linear functional on $X$. By Theorem 3.2, there exists an extension $\tilde{\psi}: X \rightarrow \mathbb{R}$ of $\psi$ satisfying

$$
\tilde{\psi}(x) \leq p(x) \text { for every } x \in X
$$

Let

$$
\tilde{\varphi}(x):=\tilde{\psi}(x)-i \tilde{\psi}(i x), \quad x \in X
$$

It is an exercise to show that $\tilde{\varphi}$ is $\mathbb{C}$-linear, that $\varphi(x)=\tilde{\varphi}(x)$ for every $x \in U$ and it is clear from the definition that $\operatorname{Re} \tilde{\varphi}(x)=\tilde{\psi}(x)$. Thus, $\tilde{\varphi}$ is a possible element we are looking for.
Theorem 3.6 (Hahn-Banach; extension of bounded linear functionals). Let X be a normed space and $U \subseteq X$ a linear subspace. Then for every bounded linear
$u^{\prime}: U \rightarrow \mathbb{K}$ there exists a bounded linear extension $x^{\prime}: X \rightarrow \mathbb{K}\left(\right.$ that is, $\left.\left.x^{\prime}\right|_{U}=u^{\prime}\right)$ such that $\left\|x^{\prime}\right\|=\left\|u^{\prime}\right\|$.

Proof. We first assume that $X$ is a real normed space. The function $p: X \rightarrow \mathbb{R}$ defined by $p(x):=\left\|u^{\prime}\right\|\|x\|$ is sublinear and

$$
u^{\prime}(x) \leq p(x) \text { for every } x \in U .
$$

By the first Hahn-Banach theorem (Theorem 3.2), there exists a linear $x^{\prime}: X \rightarrow \mathbb{R}$ extending $u^{\prime}$ such that

$$
x^{\prime}(x) \leq p(x)=\left\|u^{\prime}\right\|\|x\| \text { for every } x \in X
$$

Replacing $x$ by $-x$, this implies that

$$
\left|x^{\prime}(x)\right| \leq\left\|u^{\prime}\right\|\|x\| \text { for every } x \in X
$$

Hence, $x^{\prime}$ is bounded and $\left\|x^{\prime}\right\| \leq\left\|u^{\prime}\right\|$. On the other hand, one trivially has

$$
\left\|x^{\prime}\right\|=\sup _{\substack{x \in X \\\|x\| \leq 1}}\left|x^{\prime}(x)\right| \geq \sup _{\substack{x \in U \\\|x\| \| \leq 1}}\left|x^{\prime}(x)\right|=\sup _{\substack{x \in U \\\|x\| \leq 1}}\left|u^{\prime}(x)\right|=\left\|u^{\prime}\right\| .
$$

If $X$ is a complex normed space, then the second Hahn-Banach theorem (Theorem 3.5) implies that there exists a linear $x^{\prime}: X \rightarrow \mathbb{C}$ such that

$$
\operatorname{Re} x^{\prime}(x) \leq p(x)=\left\|u^{\prime}\right\|\|x\| \text { for every } x \in X
$$

In particular,

$$
\left|x^{\prime}(x)\right|=\sup _{\theta \in[0,2 \pi]} \operatorname{Re} x^{\prime}\left(e^{i \theta} x\right) \leq\left\|u^{\prime}\right\|\|x\| \text { for every } x \in X
$$

so that again $x^{\prime}$ is bounded and $\left\|x^{\prime}\right\| \leq\left\|u^{\prime}\right\|$. The inequality $\left\|x^{\prime}\right\| \geq\left\|u^{\prime}\right\|$ follows as above.

Corollary 3.7. If $X$ is a normed space, then for every $x \in X \backslash\{0\}$ there exists $x^{\prime} \in X^{\prime}$ such that

$$
\left\|x^{\prime}\right\|=1 \text { and } x^{\prime}(x)=\|x\| .
$$

In particular, $X^{\prime}$ separates the points of $X$, i.e. for every $x_{1}, x_{2} \in X, x_{1} \neq x_{2}$, there exists $x^{\prime} \in X^{\prime}$ such that $x^{\prime}\left(x_{1}\right) \neq x^{\prime}\left(x_{2}\right)$.

Proof. By the Hahn-Banach theorem (Theorem 3.6), there exists an extension $x^{\prime} \in$ $X^{\prime}$ of the functional $u^{\prime}: \operatorname{span}\{x\} \rightarrow \mathbb{K}$ defined by $u^{\prime}(\lambda x)=\lambda\|x\|$ such that $\left\|x^{\prime}\right\|=$ $\left\|u^{\prime}\right\|=1$.

For the proof of the second assertion, set $x:=x_{1}-x_{2}$.
Corollary 3.8. If $X$ is a normed space, then for every $x \in X$

$$
\begin{equation*}
\|x\|=\sup _{\substack{x^{\prime} \in X^{\prime} \\\left\|x^{\prime}\right\| \leq 1}}\left|x^{\prime}(x)\right| . \tag{3.4}
\end{equation*}
$$

Proof. For every $x^{\prime} \in X^{\prime}$ with $\left\|x^{\prime}\right\| \leq 1$ one has

$$
\left|x^{\prime}(x)\right| \leq\left\|x^{\prime}\right\|\|x\| \leq\|x\|,
$$

which proves one of the required inequalities. The other inequality follows from Corollary 3.7.

Remark 3.9. The equality (3.4) should be compared to the definition of the norm of an element $x^{\prime} \in X^{\prime}$ :

$$
\left\|x^{\prime}\right\|=\sup _{\substack{x \in X \\\|x\| \leq 1}}\left|x^{\prime}(x)\right| .
$$

From now on, it will be convenient to use the following notation. Given a normed space $X$ and elements $x \in X, x^{\prime} \in X^{\prime}$, we write

$$
\left\langle x^{\prime}, x\right\rangle:=\left\langle x^{\prime}, x\right\rangle_{X^{\prime} \times X}:=x^{\prime}(x) .
$$

For the bracket $\langle\cdot, \cdot\rangle$, we note the following properties. The function

$$
\begin{aligned}
\langle\cdot, \cdot\rangle: X^{\prime} \times X & \rightarrow \mathbb{K}, \\
\left(x^{\prime}, x\right) & \mapsto\left\langle x^{\prime}, x\right\rangle=x^{\prime}(x)
\end{aligned}
$$

is bilinear and for every $x^{\prime} \in X^{\prime}, x \in X$,

$$
\left|\left\langle x^{\prime}, x\right\rangle\right| \leq\left\|x^{\prime}\right\|\|x\| .
$$

The bracket $\langle\cdot, \cdot\rangle$ thus appeals to the notion of the scalar product on inner product spaces, and the last inequality appeals to the Cauchy-Schwarz inequality, but note, however, that the bracket is not a scalar product since it is defined on a pair of two different spaces. Moreover, even if $X=H$ is a complex Hilbert space, then the bracket differs from the scalar product in that it is bilinear instead of sesquilinear.

Corollary 3.10. Let $X$ be a normed space, $U \subseteq X$ a closed linear subspace and $x \in X \backslash U$. Then there exists $x^{\prime} \in X^{\prime}$ such that

$$
x^{\prime}(x) \neq 0 \text { and } x^{\prime}(u)=0 \text { for every } u \in U .
$$

Proof. Let $\pi: X \rightarrow X / U$ be the quotient map $(\pi(x)=x+U)$. Since $x \notin U$, we have $\pi(x) \neq 0$. By Corollary 3.7, there exists $\varphi \in(X / U)^{\prime}$ such that $\varphi(\pi(x)) \neq 0$. Then $x^{\prime}:=\varphi \circ \pi \in X^{\prime}$ is a functional we are looking for.

A linear operator $P: X \rightarrow X$ on a linear space $X$ is called a projection if $P^{2}=P$. A linear subspace $U$ of a normed space $X$ is called complemented if there exists a projection $P \in \mathscr{L}(X)$ such that $\operatorname{ran} P=U$.

Remark 3.11. If $P$ is a projection, then $Q=I-P$ is also a projection and $\operatorname{ran} P=$ $\operatorname{ker} Q$. Hence, if $P$ is a bounded projection on a normed space, then $\operatorname{ran} P$ is necessarily closed. Thus, a necessary condition for $U$ to be complemented is that $U$ is closed.

Corollary 3.12. Every finite dimensional subspace of a normed space is complemented.

Proof. Let $U$ be a finite dimensional subspace of a normed space $X$. Let $\left(b_{i}\right)_{1 \leq i \leq N}$ be a basis of $U$. By Corollary 3.10, there exist functionals $x_{i}^{\prime} \in X^{\prime}$ such that

$$
\left\langle x_{i}^{\prime}, b_{j}\right\rangle=\left\{\begin{array}{l}
1 \text { if } i=j \\
0 \text { otherwise }
\end{array}\right.
$$

Let $P: X \rightarrow X$ be defined by

$$
P x:=\sum_{i=1}^{N}\left\langle x_{i}^{\prime}, x\right\rangle b_{i}, \quad x \in X .
$$

Then $P b_{i}=b_{i}$ for every $1 \leq i \leq N$, and thus $P^{2}=P$, that is, $P$ is a projection. Moreover, $\operatorname{ran} P=U$ by construction. By the estimate

$$
\begin{aligned}
\|P x\| & \leq \sum_{i=1}^{N}\left|\left\langle x_{i}^{\prime}, x\right\rangle\right|\left\|b_{i}\right\| \\
& \leq\left(\sum_{i=1}^{N}\left\|x_{i}^{\prime}\right\|\left\|b_{i}\right\|\right)\|x\|,
\end{aligned}
$$

the projection $P$ is bounded.
The following lemma which does not depend on the Hahn-Banach theorem is stated for completeness.

Lemma 3.13. In a Hilbert space every closed linear subspace is complemented.
Proof. Take the orthogonal projection onto the closed subspace as a possible projection.

Corollary 3.14. If $X$ is a normed space such that $X^{\prime}$ is separable, then $X$ is separable, too.

Proof. Let $D^{\prime}=\left\{x_{n}^{\prime}: n \in \mathbb{N}\right\}$ be a dense subset of the unit sphere of $X^{\prime}$. For every $n \in \mathbb{N}$ we choose an element $x_{n} \in X$ such that $\left\|x_{n}\right\| \leq 1$ and $\left|\left\langle x_{n}^{\prime}, x_{n}\right\rangle\right| \geq \frac{1}{2}$. We claim that $D:=\operatorname{span}\left\{x_{n}: n \in \mathbb{N}\right\}$ is dense in $X$. If this was not true, i.e. if $\bar{D} \neq X$, then, by Corollary 3.10, we find an element $x^{\prime} \in X^{\prime} \backslash\{0\}$ such that $x^{\prime}\left(x_{n}\right)=0$ for every $n \in \mathbb{N}$. We may without loss of generality assume that $\left\|x^{\prime}\right\|=1$. Since $D^{\prime}$ is dense in the unit sphere of $X^{\prime}$, we find $n_{0} \in \mathbb{N}$ such that $\left\|x^{\prime}-x_{n_{0}}^{\prime}\right\| \leq \frac{1}{4}$. But then

$$
\frac{1}{2} \leq\left|\left\langle x_{n_{0}}^{\prime}, x_{n_{0}}\right\rangle\right|=\left|\left\langle x_{n_{0}}^{\prime}-x^{\prime}, x_{n_{0}}\right\rangle\right| \leq\left\|x_{n_{0}}^{\prime}-x^{\prime}\right\|\left\|x_{n_{0}}\right\| \leq \frac{1}{4}
$$

which is a contradiction. Hence, $\bar{D}=X$ and $X$ is separable by Lemma 2.15 of Chapter 2.

### 3.2 Weak* convergence and the theorem of Banach-Alaoglu

Let $X$ be a Banach space. We say that a sequence $\left(x_{n}^{\prime}\right) \subseteq X^{\prime}$ converges weak* to some element $x^{\prime} \in X^{\prime}$ if for every $x \in X$ one has $\lim _{n \rightarrow \infty}\left\langle x_{n}^{\prime}, x\right\rangle=\left\langle x^{\prime}, x\right\rangle$. We write $x_{n}^{\prime} \xrightarrow{\text { weak* }} x^{\prime}$ if $\left(x_{n}^{\prime}\right)$ converges weak ${ }^{*}$ to $x^{\prime}$.

Theorem 3.15 (Banach-Alaoglu). Let $X$ be a separable Banach space. Then every bounded sequence $\left(x_{n}^{\prime}\right) \subseteq X^{\prime}$ admits a weak ${ }^{*}$ convergent subsequence, that is, there exists $x^{\prime} \in X^{\prime}$ and there exists a subsequence $\left(x_{n_{k}}^{\prime}\right)$ of $\left(x_{n}^{\prime}\right)$ such that $x_{n_{k}}^{\prime} \xrightarrow{\text { weak* }} x^{\prime}$.

Proof. As in the proof of the Arzelá-Ascoli theorem (Theorem 1.36) and the theorem about weak sequential compactness of the unit ball in Hilbert spaces (Theorem 2.47), we use Cantor's diagonal sequence argument. Let $\left(x_{n}^{\prime}\right)$ be a bounded sequence in $X^{\prime}$.

Since $X$ is separable by assumption, we can choose a dense sequence $\left(x_{m}\right) \subseteq X$. Since $\left(\left\langle x_{n}^{\prime}, x_{1}\right\rangle\right)$ is bounded by the boundedness of $\left(x_{n}^{\prime}\right)$, there exists a subsequence $\left(x_{\varphi_{1}(n)}^{\prime}\right)$ of $\left(x_{n}^{\prime}\right)\left(\varphi_{1}: \mathbb{N} \rightarrow \mathbb{N}\right.$ is increasing, unbounded) such that

$$
\lim _{n \rightarrow \infty}\left\langle x_{\varphi_{1}(n)}^{\prime}, x_{1}\right\rangle \text { exists. }
$$

Similarly, there exists a subsequence $\left(x_{\varphi_{2}(n)}^{\prime}\right)$ of $\left(x_{\varphi_{1}(n)}^{\prime}\right)$ such that

$$
\lim _{n \rightarrow \infty}\left\langle x_{\varphi_{2}(n)}^{\prime}, x_{2}\right\rangle \text { exists. }
$$

Note that for this subsequence, we also have that

$$
\lim _{n \rightarrow \infty}\left\langle x_{\varphi_{2}(n)}^{\prime}, x_{1}\right\rangle \text { exists. }
$$

Iterating this argument, we find a subsequence $\left(x_{\varphi_{3}(n)}^{\prime}\right)$ of $\left(x_{\varphi_{2}(n)}^{\prime}\right)$ and finally for every $m \in \mathbb{N}, m \geq 2$, a subsequence $\left(x_{\varphi_{m}(n)}^{\prime}\right)$ of $\left(x_{\varphi_{m-1}(n)}^{\prime}\right)$ such that

$$
\lim _{n \rightarrow \infty}\left\langle x_{\varphi_{m}(n)}^{\prime}, x_{j}\right\rangle \text { exists for every } 1 \leq j \leq m
$$

Let $\left(y_{n}^{\prime}\right):=\left(x_{\varphi_{n}(n)}^{\prime}\right)$ be the 'diagonal sequence'. Then $\left(y_{n}^{\prime}\right)$ is a subsequence of $\left(x_{n}^{\prime}\right)$ and

$$
\lim _{n \rightarrow \infty}\left\langle y_{n}^{\prime}, x_{m}\right\rangle \text { exists for every } m \in \mathbb{N}
$$

By Lemma 2.48 of Chapter 2, there exists $x^{\prime} \in X^{\prime}$ such that

$$
\lim _{n \rightarrow \infty}\left\langle y_{n}^{\prime}, x\right\rangle=\left\langle x^{\prime}, x\right\rangle \text { for every } x \in X
$$

This is the claim.

### 3.3 Weak convergence and reflexivity

Given a normed space $X$, we call $X^{\prime \prime}:=\left(X^{\prime}\right)^{\prime}=\mathscr{L}\left(X^{\prime}, \mathbb{K}\right)$ the bidual of $X$.
Lemma 3.16. Let $X$ be a normed space. Then the mapping

$$
\begin{aligned}
J: X & \rightarrow X^{\prime \prime}, \\
x & \mapsto\left(x^{\prime} \mapsto\left\langle x^{\prime}, x\right\rangle\right),
\end{aligned}
$$

is well defined and isometric.
Proof. The linearity of $x^{\prime} \mapsto\left\langle x^{\prime}, x\right\rangle$ is clear, and from the inequality

$$
\left|J x\left(x^{\prime}\right)\right|=\left|\left\langle x^{\prime}, x\right\rangle\right| \leq\left\|x^{\prime}\right\|\|x\|
$$

follows that $J x \in X^{\prime \prime}$ (that is, $J$ is well defined) and $\|J x\| \leq\|x\|$. The fact that $J$ is isometric follows from Corollary 3.7.

A normed space $X$ is called reflexive if the isometry $J$ from Lemma 3.16 is surjective, i.e. if $J X=X^{\prime \prime}$. In other words: a normed space $X$ is reflexive if for every $x^{\prime \prime} \in X^{\prime \prime}$ there exists $x \in X$ such that

$$
\left\langle x^{\prime \prime}, x^{\prime}\right\rangle=\left\langle x^{\prime}, x\right\rangle \text { for all } x^{\prime} \in X^{\prime} .
$$

Remark 3.17. If a normed space is reflexive then $X$ and $X^{\prime \prime}$ are isometrically isomorphic (via the operator $J$ ). Since $X^{\prime \prime}$ is always complete, a reflexive space is necessarily a Banach space.

Note that it can happen that $X$ and $X^{\prime \prime}$ are isomorphic without $X$ being reflexive (the example of such a Banach space is however quite involved). We point out that reflexivity means that the special operator $J$ is an isomorphism.

Lemma 3.18. Every Hilbert space is reflexive.
Proof. By the Theorem of Riesz-Fréchet, we may identify $H$ with its dual $H^{\prime}$ and thus also $H$ with its bidual $H^{\prime \prime}$. The identification is done via the scalar product. It is an exercise to show that this identification of $H$ with $H^{\prime \prime}$ coincides with the mapping $J$ from Lemma 3.16.

Remark 3.19. It should be noted that for complex Hilbert spaces, the identification of $H$ with its dual $H^{\prime}$ is only antilinear, but after the second identification ( $H^{\prime}$ with $\left.H^{\prime \prime}\right)$ it turns out that the identification of $H$ with $H^{\prime \prime}$ is linear.

Lemma 3.20. Every finite dimensional Banach space is reflexive.
Proof. It suffices to remark that if $X$ is finite dimensional, then

$$
\operatorname{dim} X=\operatorname{dim} X^{\prime}=\operatorname{dim} X^{\prime \prime}<\infty .
$$

Surjectivity of the mapping $J$ (which is always injective) thus follows from linear algebra.

Theorem 3.21. The space $L^{p}(\Omega)$ is reflexive if $1<p<\infty((\Omega, \mathscr{A}, \mu)$ being an arbitrary measure space).

We will actually only prove the following special case.
Theorem 3.22. The spaces $l^{p}$ are reflexive if $1<p<\infty$.
The proof of Theorem 3.22 is based on the following lemma.
Lemma 3.23. Let $1 \leq p<\infty$ and let $q:=\frac{p}{p-1}$ be the conjugate exponent so that $\frac{1}{p}+\frac{1}{q}=1$. Then the operator

$$
\begin{aligned}
T: l^{q} & \rightarrow\left(l^{p}\right)^{\prime}, \\
\left(a_{n}\right) & \mapsto\left(\left(x_{n}\right) \mapsto \sum_{n} a_{n} x_{n}\right),
\end{aligned}
$$

is an isometric isomorphism, that is, $\left(l^{p}\right)^{\prime}=l^{q}$.
Proof. Linearity of $T$ is obvious. Assume first $p>1$, so that $q<\infty$. Note that for every $a:=\left(a_{n}\right) \in l^{q} \backslash\{0\}$ the sequence $\left(x_{n}\right):=\left(c \overline{a_{n}}\left|a_{n}\right|^{q-2}\right)\left(c=\|a\|_{q}^{-q / p}\right)$ belongs to $l^{p}$ and

$$
\|x\|_{p}^{p}=\|a\|_{q}^{-q} \sum_{n}\left|a_{n}\right|^{(q-1) p}=1 .
$$

This particular $x \in l^{p}$ shows that

$$
\|T a\|_{\left(l^{p}\right)^{\prime}} \geq \sum_{n} a_{n} x_{n}=\|a\|_{q}^{-q / p} \sum_{n}\left|a_{n}\right|^{q}=\|a\|_{q}^{q(p-1) / p}=\|a\|_{q} .
$$

On the other hand, by Hölder's inequality,

$$
\|T a\|_{\left(l^{p}\right)^{\prime}}=\sup _{\|x\|_{p} \leq 1}\left|\sum_{n} a_{n} x_{n}\right| \leq\|a\|_{q}
$$

so that $T$ is isometric in the case $p \in(1, \infty)$. The case $p=1$ is very similar and will be omitted.

In order to show that $T$ is surjective, let $\varphi \in\left(l^{p}\right)^{\prime}$. Denote by $e_{n}$ the $n$-th unit vector in $l^{p}$, and let $a_{n}:=\varphi\left(e_{n}\right)$. If $p=1$, then $\left(a_{n}\right) \in l^{\infty}=l^{q}$ by the trivial estimate

$$
\left|a_{n}\right|=\left|\varphi\left(e_{n}\right)\right| \leq\|\varphi\|\left\|e_{n}\right\|_{1}=\|\varphi\| .
$$

If $p>1$, then we may argue as follows. For every $N \in \mathbb{N}$,

$$
\begin{aligned}
\sum_{n=1}^{N}\left|a_{n}\right|^{q} & =\sum_{n=1}^{N} a_{n} \overline{a_{n}}\left|a_{n}\right|^{q-2} \\
& =\varphi\left(\sum_{n=1}^{N} \overline{a_{n}}\left|a_{n}\right|^{q-2} e_{n}\right) \\
& \leq\|\varphi\|\left(\sum_{n=1}^{N}\left|a_{n}\right|^{(q-1) p}\right)^{\frac{1}{p}} \\
& =\|\varphi\|\left(\sum_{n=1}^{N}\left|a_{n}\right|^{q}\right)^{\frac{1}{p}},
\end{aligned}
$$

which is equivalent to

$$
\left(\sum_{n=1}^{N}\left|a_{n}\right|^{q}\right)^{1-\frac{1}{p}}=\left(\sum_{n=1}^{N}\left|a_{n}\right|^{q}\right)^{\frac{1}{q}} \leq\|\varphi\| .
$$

Since the right-hand side of this inequality does not depend on $N \in \mathbb{N}$, we obtain that $a:=\left(a_{n}\right) \in l^{q}$ and $\|a\|_{q} \leq\|\varphi\|$.

Next, observe that for every $x \in l^{p}$ one has

$$
x=\sum_{n} x_{n} e_{n}=\lim _{N \rightarrow \infty} \sum_{n=1}^{N} x_{n} e_{n}
$$

the series converging in $l^{p}$ (here we need the restriction $p<\infty!$ ). Hence, for every $x \in l^{p}$, by the boundedness of $\varphi$,

$$
\begin{aligned}
\varphi(x) & =\lim _{N \rightarrow \infty} \varphi\left(\sum_{n=1}^{N} x_{n} e_{n}\right) \\
& =\lim _{N \rightarrow \infty} \sum_{n=1}^{N} x_{n} a_{n} \\
& =\sum_{n} x_{n} a_{n} \\
& =T a(x) .
\end{aligned}
$$

Hence, $T$ is surjective.
Proof (of Theorem 3.22). By Lemma 3.23, we may identify $\left(l^{p}\right)^{\prime}$ with $l^{q}$ and, if $1<p<\infty(!)$, also $\left(l^{p}\right)^{\prime \prime}=\left(l^{q}\right)^{\prime}$ with $l^{p}$. One just has to notice that this identification of $l^{p}$ with $\left(l^{p}\right)^{\prime \prime}=l^{p}$ (the identity map on $l^{p}$ ) coincides with the operator $J$ from Lemma 3.16, so that $l^{p}$ is reflexive if $1<p<\infty$.

Lemma 3.24. The spaces $l^{1}, L^{1}(\Omega)\left(\Omega \subseteq \mathbb{R}^{N}\right)$ and $C([0,1])$ are not reflexive.
Proof. For every $t \in[0,1]$, let $\delta_{t} \in C([0,1])^{\prime}$ be defined by

$$
\left\langle\delta_{t}, f\right\rangle:=f(t), \quad f \in C([0,1]) .
$$

Then $\left\|\delta_{t}\right\|=1$ and whenever $t \neq s$, then

$$
\left\|\delta_{t}-\delta_{s}\right\|=2
$$

In particular, the uncountably many balls $B\left(\delta_{t}, \frac{1}{2}\right)(t \in[0,1])$ are mutually disjoint so that $C([0,1])^{\prime}$ is not separable.

Now, if $C([0,1])$ were reflexive, then $C([0,1])^{\prime \prime}=C([0,1])$ would be separable (since $C([0,1])$ is separable), and then, by Corollary $3.14, C([0,1])^{\prime}$ would be separable; a contradiction to what has been said before. This proves that $C([0,1])$ is not reflexive.

The cases of $l^{1}$ and $L^{1}(\Omega)$ are proved similarly. They are separable Banach spaces with nonseparable dual.

Theorem 3.25. Every closed subspace of a reflexive Banach space is reflexive.
Proof. Let $X$ be a reflexive Banach space, and let $U \subseteq X$ be a closed subspace. Let $u^{\prime \prime} \in U^{\prime \prime}$. Then the mapping $x^{\prime \prime}: X^{\prime} \rightarrow \mathbb{K}$ defined by

$$
\left\langle x^{\prime \prime}, x^{\prime}\right\rangle=\left\langle u^{\prime \prime},\left.x^{\prime}\right|_{U}\right\rangle, \quad x^{\prime} \in X^{\prime}
$$

is linear and bounded, i.e. $x^{\prime \prime} \in X^{\prime \prime}$. By reflexivity of $X$, there exists $x \in X$ such that

$$
\begin{equation*}
\left\langle x^{\prime}, x\right\rangle=\left\langle u^{\prime \prime},\left.x^{\prime}\right|_{U}\right\rangle, \quad x^{\prime} \in X^{\prime} \tag{3.5}
\end{equation*}
$$

Assume that $x \notin U$. Then, by Corollary 3.10 , there exists $x^{\prime} \in X^{\prime}$ such that $\left.x^{\prime}\right|_{U}=0$ and $\left\langle x^{\prime}, x\right\rangle \neq 0$; a contradiction to the last equality. Hence, $x \in U$. We need to show that

$$
\begin{equation*}
\left\langle u^{\prime \prime}, u^{\prime}\right\rangle=\left\langle u^{\prime}, x\right\rangle, \forall u^{\prime} \in U^{\prime} . \tag{3.6}
\end{equation*}
$$

However, if $u^{\prime} \in U^{\prime}$, then, by Hahn-Banach we can choose an extension $x^{\prime} \in X^{\prime}$, i.e. $\left.x^{\prime}\right|_{U}=u^{\prime}$. The equation (3.6) thus follows from (3.5).

Corollary 3.26. The Sobolev spaces $W^{k, p}(\Omega)\left(\Omega \subseteq \mathbb{R}^{N}\right.$ open) are reflexive if $1<$ $p<\infty, k \in \mathbb{N}$.

Proof. For example, for $k=1$, the operator

$$
\begin{array}{rlr}
T: W^{1, p}(\Omega) & \rightarrow & L^{p}(\Omega)^{1+N}, \\
u & \mapsto & \left(u, \frac{\partial u}{\partial x_{1}}, \ldots, \frac{\partial u}{\partial x_{N}}\right),
\end{array}
$$

is isometric, so that we may consider $W^{1, p}(\Omega)$ as a closed subspace of $L^{p}(\Omega)^{1+N}$ which is reflexive by Theorem 3.21. The claim follows from Theorem 3.25.

Corollary 3.27. A Banach space is reflexive if and only if its dual is reflexive.

Proof. Assume that the Banach space $X$ is reflexive. Let $x^{\prime \prime \prime} \in X^{\prime \prime \prime}$ (the tridual!). Then the mapping $x^{\prime}: X \rightarrow \mathbb{K}$ defined by

$$
\left\langle x^{\prime}, x\right\rangle:=\left\langle x^{\prime \prime \prime}, J_{X}(x)\right\rangle, \quad x \in X
$$

is linear and bounded, i.e. $x^{\prime} \in X^{\prime}$ (here $J_{X}$ denotes the isometry $X \rightarrow X^{\prime \prime}$ ). Let $x^{\prime \prime} \in$ $X^{\prime \prime}$ be arbitrary. Since $X$ is reflexive, there exists $x \in X$ such that $J_{X} x=x^{\prime \prime}$. Hence,

$$
\left\langle x^{\prime \prime \prime}, x^{\prime \prime}\right\rangle=\left\langle x^{\prime \prime \prime}, J_{X} x\right\rangle=\left\langle x^{\prime}, x\right\rangle=\left\langle x^{\prime \prime}, x^{\prime}\right\rangle,
$$

which proves that $J_{X^{\prime}} x^{\prime}=x^{\prime \prime \prime}$, i.e. the isometry $J_{X^{\prime}}: X^{\prime} \rightarrow X^{\prime \prime \prime}$ is surjective. Hence, $X^{\prime}$ is reflexive.

On the other hand, assume that $X^{\prime}$ is reflexive. Then $X^{\prime \prime}$ is reflexive by the preceeding argument, and therefore $X$ (considered as a closed subspace of $X^{\prime \prime}$ via the isometry $J$ ) is reflexive by Theorem 3.25.

Let $X$ be a normed space. We say that a sequence $\left(x_{n}\right) \subseteq X$ converges weakly to some $x \in X$ if

$$
\lim _{n \rightarrow \infty}\left\langle x^{\prime}, x_{n}\right\rangle=\left\langle x^{\prime}, x\right\rangle \text { for every } x^{\prime} \in X^{\prime}
$$

Notations: if $\left(x_{n}\right)$ converges weakly to $x$, then we write $x_{n} \rightharpoonup x, w-\lim _{n \rightarrow \infty} x_{n}=x$, $x_{n} \rightarrow x$ in $\sigma\left(X, X^{\prime}\right)$, or $x_{n} \rightarrow x$ weakly.

Theorem 3.28. In a reflexive Banach space every bounded sequence admits a weakly convergent subsequence.

Proof. Let $\left(x_{n}\right)$ be a bounded sequence in a reflexive Banach space $X$. We first assume that $X$ is separable. Then $X^{\prime \prime}$ is separable by reflexivity, and $X^{\prime}$ is separable by Corollary 3.14. Let $\left(x_{m}^{\prime}\right) \subseteq X^{\prime}$ be a dense sequence.

Since $\left(\left\langle x_{1}^{\prime}, x_{n}\right\rangle\right)$ is bounded by the boundedness of $\left(x_{n}\right)$, there exists a subsequence $\left(x_{\varphi_{1}(n)}\right)$ of $\left(x_{n}\right)\left(\varphi_{1}: \mathbb{N} \rightarrow \mathbb{N}\right.$ is increasing, unbounded) such that

$$
\lim _{n \rightarrow \infty}\left\langle x_{1}^{\prime}, x_{\varphi_{1}(n)}\right\rangle \text { exists. }
$$

Similarly, there exists a subsequence $\left(x_{\varphi_{2}(n)}\right)$ of $\left(x_{\varphi_{1}(n)}\right)$ such that

$$
\lim _{n \rightarrow \infty}\left\langle x_{2}^{\prime}, x_{\varphi_{2}(n)}\right\rangle \text { exists. }
$$

Note that for this subsequence, we also have that

$$
\lim _{n \rightarrow \infty}\left\langle x_{1}^{\prime}, x_{\varphi_{2}(n)}\right\rangle \text { exists. }
$$

Iterating this argument, we find a subsequence $\left(x_{\varphi_{3}(n)}\right)$ of $\left(x_{\varphi_{2}(n)}\right)$ and finally for every $m \in \mathbb{N}, m \geq 2$, a subsequence $\left(x_{\varphi_{m}(n)}\right)$ of $\left(x_{\varphi_{m-1}(n)}\right)$ such that

$$
\lim _{n \rightarrow \infty}\left\langle x_{j}^{\prime}, x_{\varphi_{m}(n)}\right\rangle \text { exists for every } 1 \leq j \leq m
$$

Let $\left(y_{n}\right):=\left(x_{\varphi_{n}(n)}\right)$ be the 'diagonal sequence'. Then $\left(y_{n}\right)$ is a subsequence of $\left(x_{n}\right)$ and

$$
\lim _{n \rightarrow \infty}\left\langle x_{m}^{\prime}, y_{n}\right\rangle \text { exists for every } m \in \mathbb{N}
$$

By Lemma 2.48 of Chapter 2, there exists $x^{\prime \prime} \in X^{\prime \prime}$ such that

$$
\lim _{n \rightarrow \infty}\left\langle x^{\prime}, y_{n}\right\rangle=\left\langle x^{\prime}, x^{\prime \prime}\right\rangle \text { for every } x^{\prime} \in X^{\prime}
$$

Since $X$ is reflexive, there exists $x \in X$ such that $J x=x^{\prime \prime}$. For this $x$, we have by definition of $J$

$$
\lim _{n \rightarrow \infty}\left\langle x^{\prime}, y_{n}\right\rangle=\left\langle x^{\prime}, x\right\rangle \text { exists for every } x^{\prime} \in X^{\prime},
$$

that is, $\left(y_{n}\right)$ converges weakly to $x$.
If $X$ is not separable as we first assumed, then one may replace $X$ by $\tilde{X}:=$ $\overline{\operatorname{span}}\left\{x_{n}: n \in \mathbb{N}\right\}$ which is separable. By the above, there exists $x \in \tilde{X}$ and a subsequence of $\left(x_{n}\right)$ (which we denote again by $\left(x_{n}\right)$ ) such that for every $\tilde{x}^{\prime} \in \tilde{X}^{\prime}$,

$$
\lim _{n \rightarrow \infty}\left\langle\tilde{x}^{\prime}, x_{n}\right\rangle=\left\langle\tilde{x}^{\prime}, x\right\rangle,
$$

that is, $\left(x_{n}\right)$ converges weakly in $\tilde{X}$. If $x^{\prime} \in X^{\prime}$, then $\left.x^{\prime}\right|_{\tilde{X}} \in \tilde{X}^{\prime}$, and it follows easily that the sequence $\left(x_{n}\right)$ also converges weakly in $X$ to the element $x$.

## $3.4 *$ Minimization of convex functionals

Recall from page 31 that subset $K$ of a real or complex vector space is convex if for every $x, y \in K$ and every $t \in[0,1]$ one has $t x+(1-t) y \in K$.

Theorem 3.29 (Hahn-Banach; separation of convex sets). Let $X$ be a Banach space, $K \subseteq X$ a closed, nonempty, convex subset, and $x_{0} \in X \backslash K$. Then there exists $x^{\prime} \in X^{\prime}$ and $\varepsilon>0$ such that

$$
\operatorname{Re}\left\langle x^{\prime}, x\right\rangle+\varepsilon \leq \operatorname{Re}\left\langle x^{\prime}, x_{0}\right\rangle, \quad x \in K .
$$

Lemma 3.30. Let $K$ be an open, nonempty, convex subset of a Banach space $X$ such that $0 \in K$. Define the Minkowski functional $p: X \rightarrow \mathbb{R}$ by

$$
p(x)=\inf \left\{\lambda>0: \frac{x}{\lambda} \in K\right\} .
$$

Then $p$ is sublinear, there exists $M \geq 0$ such that

$$
p(x) \leq M\|x\|, \quad x \in X
$$

and $K=\{x \in X: p(x)<1\}$.
Proof. Since $B(0, r) \subseteq K$ for some $r>0$, we find that

$$
p(x) \leq \frac{1}{r}\|x\| \text { for every } x \in X
$$

The property $p(\alpha x)=\alpha p(x)$ for every $\alpha>0$ and every $x \in X$ is obvious.
Next, if $p(x)<1$, then there exists $\lambda \in(0,1)$ such that $x / \lambda \in K$. Hence, by convexity, $x=\lambda \frac{x}{\lambda}=\lambda \frac{x}{\lambda}+(1-\lambda) 0 \in K$. On the other hand, if $x \in K$, then $(1+\varepsilon) x \in$ $K$, since $K$ is open. Hence, $p(x) \leq(1+\varepsilon)^{-1}<1$, so that $K=\{x \in X: p(x)<1\}$.

Let finally $x, y \in X$. Then for every $\varepsilon>0, x /(p(x)+\varepsilon) \in K$ and $y /(p(y)+\varepsilon) \in K$. In particular, for every $t \in[0,1]$,

$$
\frac{t}{p(x)+\varepsilon} x+\frac{1-t}{p(y)+\varepsilon} y \in K
$$

Setting $t=(p(x)+\varepsilon) /(p(x)+p(y)+2 \varepsilon)$, one finds that

$$
\frac{x+y}{p(x)+p(y)+2 \varepsilon} \in K
$$

so that $p(x+y) \leq p(x)+p(y)+2 \varepsilon$. Since $\varepsilon>0$ was arbitrary, we find $p(x+y) \leq$ $p(x)+p(y)$. The claim is proved.

Proof (of Theorem 3.29). We prove the theorem for the case when $X$ is a real Banach space. The complex case is proved similarly.

We may without loss of generality assume that $0 \in K$; it suffices to translate $K$ and $x_{0}$ for this. Since $x_{0} \notin K$ and since $K$ is closed, we find that $d:=\operatorname{dist}\left(x_{0}, K\right)>0$. Put

$$
K_{d}:=\{x \in X: \operatorname{dist}(x, K)<d / 2\}
$$

so that $K_{d}$ is an open, convex subset such that $0 \in K_{d}$. Let $p$ be the corresponding Minkowski functional (see Lemma 3.30).

Define on the one-dimensional subspace $U:=\left\{\lambda x_{0}: \lambda \in \mathbb{R}\right\}$ the functional $u^{\prime}$ : $U \rightarrow \mathbb{R}$ by $\left\langle u^{\prime}, \lambda x_{0}\right\rangle=\lambda$. Then

$$
\left\langle u^{\prime}, u\right\rangle \leq p(u), \quad u \in U .
$$

By the Hahn-Banach theorem 3.2, there exists a linear extension $x^{\prime}: X \rightarrow \mathbb{R}$ such that

$$
\begin{equation*}
\left\langle x^{\prime}, x\right\rangle \leq p(x), \quad x \in X \tag{3.7}
\end{equation*}
$$

In particular, by Lemma 3.30,

$$
\left|\left\langle x^{\prime}, x\right\rangle\right| \leq M\|x\|,
$$

so that $x^{\prime} \in X^{\prime}$ and $\left\|x^{\prime}\right\| \leq M$. By construction, $\left\langle x^{\prime}, x_{0}\right\rangle=1$. Moreover, by (3.7) and Lemma 3.30, $\left\langle x^{\prime}, x\right\rangle<1$ for every $x \in K \subseteq K_{d}$, so that

$$
\left\langle x^{\prime}, x\right\rangle \leq\left\langle x^{\prime}, x_{0}\right\rangle(=1), \quad x \in K_{d} .
$$

Replacing the above argument with $\left(1-\varepsilon^{\prime}\right) x_{0}$ instead of $x_{0}$ (where $\varepsilon^{\prime}>0$ is chosen so small that $\left.\left(1-\varepsilon^{\prime}\right) x_{0} \notin K_{d}\right)$, we find that

$$
\left\langle x^{\prime}, x\right\rangle+\varepsilon^{\prime}\left\langle x^{\prime}, x_{0}\right\rangle \leq\left\langle x^{\prime}, x_{0}\right\rangle, \quad x \in K \subseteq K_{d},
$$

and putting $\varepsilon:=\varepsilon^{\prime}=\varepsilon^{\prime}\left\langle x^{\prime}, x_{0}\right\rangle>0$ yields the claim.
Corollary 3.31. Let $X$ be a Banach space and $K \subseteq X$ a closed, convex subset (closed for the norm topology). If $\left(x_{n}\right) \subseteq K$ converges weakly to some $x \in X$, then $x \in K$.

Proof. Assume the contrary, that is, $x \notin K$. By the Hahn-Banach theorem (Theorem 3.29), there exist $x^{\prime} \in X^{\prime}$ and $\varepsilon>0$ such that

$$
\operatorname{Re}\left\langle x^{\prime}, x_{n}\right\rangle+\varepsilon \leq \operatorname{Re}\left\langle x^{\prime}, x\right\rangle \text { for every } n \in \mathbb{N},
$$

a contradiction to the assumption that $x_{n} \rightharpoonup x$.
A function $f: K \rightarrow \mathbb{R}$ on a convex subset $K$ of a Banach space $X$ is called convex if for every $x, y \in K$, and every $t \in[0,1]$,

$$
\begin{equation*}
f(t x+(1-t) y) \leq t f(x)+(1-t) f(y) \tag{3.8}
\end{equation*}
$$

Corollary 3.32. Let $X$ be a Banach space, $K \subseteq X$ a closed, convex subset, and $f$ : $K \rightarrow \mathbb{R}$ a continuous, convex function. If $\left(x_{n}\right) \subseteq K$ converges weakly to $x \in K$, then

$$
f(x) \leq \liminf _{n \rightarrow \infty} f\left(x_{n}\right)
$$

Proof. For every $l \in \mathbb{R}$, the set $K_{l}:=\{x \in K: f(x) \leq l\}$ is closed (by continuity of $f$ ) and convex (by convexity of $f$ ). After extracting a subsequence, if necessary, we may assume that $l:=\liminf _{n \rightarrow \infty} f\left(x_{n}\right)=\lim _{n \rightarrow \infty} f\left(x_{n}\right)$. Then for every $\varepsilon>0$ the sequence $\left(x_{n}\right)$ is eventually in $K_{l+\varepsilon}$, i.e. except for finitely many $x_{n}$, the sequence $\left(x_{n}\right)$ lies in $K_{l+\varepsilon}$. Hence, by Corollary 3.31, $x \in K_{l+\varepsilon}$, which means that $f(x) \leq l+\varepsilon$. Since $\varepsilon>0$ was arbitrary, the claim follows.

Let $K \subseteq X$ be a convex subset of a real or complex vector space. A function $f: K \rightarrow \mathbb{R}$ is called convex if for every $x, y \in K$ and every $t \in[0,1]$ one has

$$
f(t x+(1-t) y) \leq t f(x)+(1-t) f(y)
$$

It is called strictly convex if for every $x, y \in K, x \neq y$ and every $t \in(0,1)$ the above inequality is strict.

Theorem 3.33. Let $X$ be a reflexive Banach space, $K \subseteq X$ a closed, convex, nonempty subset, and $f: K \rightarrow \mathbb{R}$ a continuous, convex function such that

$$
\lim _{\substack{\|x\| \rightarrow \infty \\ x \in K}} f(x)=+\infty \text { (coercivity). }
$$

Then there exists $x_{0} \in K$ such that

$$
f\left(x_{0}\right)=\inf \{f(x): x \in K\}>-\infty .
$$

Proof. Let $\left(x_{n}\right) \subseteq K$ be such that $\lim _{n \rightarrow \infty} f\left(x_{n}\right)=\inf \{f(x): x \in K\}$. By the coercivity assumption on $f$, the sequence $\left(x_{n}\right)$ is bounded. Since $X$ is reflexive, there exists a weakly convergent subsequence (Theorem 3.28); we denote by $x_{0}$ the limit. By Corollary 3.31, $x_{0} \in K$. By Corollary 3.32,

$$
f\left(x_{0}\right) \leq \lim _{n \rightarrow \infty} f\left(x_{n}\right)=\inf \{f(x): x \in K\} .
$$

The claim is proved.
Remark 3.34. Theorem 3.33 remains true if $f$ is only lower semicontinuous, i.e. if

$$
\liminf _{n \rightarrow \infty} f\left(x_{n}\right) \geq f(x)
$$

for every convergent $\left(x_{n}\right) \subseteq K$ with $x=\lim x_{n}$. In fact, already Corollary 3.32 remains true if $f$ is only lower semicontinuous (and then Corollary 3.32 says that lower semicontinuity of a convex function in the norm topology implies lower semicontinuity in the weak topology). It suffices for example to remark that the sets $K_{l}:=\{f \leq l\}(l \in \mathbb{R})$ are closed as soon as $f$ is lower semicontinuous.

## 3.5 * The von Neumann minimax theorem

In the following theorem, we call a function $f: K \rightarrow \mathbb{R}$ on a convex subset $K$ of a Banach space $X$ concave if $-f$ is convex, or, equivalently, if for every $x, y \in K$ and every $t \in[0,1]$,

$$
\begin{equation*}
f(t x+(1-t) y) \geq t f(x)+(1-t) f(y) \tag{3.9}
\end{equation*}
$$

A function $f: K \rightarrow \mathbb{R}$ is called strictly convex (resp. strictly concave) if for every $x$, $y \in K, x \neq y, f(x)=f(y)$ the inequality in (3.8) (resp. (3.9)) is strict for $t \in(0,1)$.

Theorem 3.35 (von Neumann minimax theorem). Let $K$ and $L$ be two closed, bounded, nonempty, convex subsets of reflexive Banach spaces $X$ and $Y$, respectively. Let $f: K \times L \rightarrow \mathbb{R}$ be a continuous function such that
$x \mapsto f(x, y)$ is strictly convex for every $y \in L$, and
$y \mapsto f(x, y)$ is concave for every $x \in K$.
Then there exists $(\bar{x}, \bar{y}) \in K \times L$ such that

$$
\begin{equation*}
f(\bar{x}, y) \leq f(\bar{x}, \bar{y}) \leq f(x, \bar{y}) \text { for every } x \in K, y \in L . \tag{3.10}
\end{equation*}
$$

Remark 3.36. A point $(\bar{x}, \bar{y}) \in K \times L$ satisfying (3.10) is called a saddle point of $f$.
A saddle point is a point of equilibrium in a two-person zero-sum game in the following sense: If the player controlling the strategy $x$ modifies his strategy when
the second player plays $\bar{y}$, he increases his loss; hence, it is his interest to play $\bar{x}$. Similarly, if the player controlling the strategy $y$ modifies his strategy when the first player plays $\bar{x}$, he diminishes his gain; thus it is in his interest to play $\bar{y}$. This property of equilibrium of saddle points justifies their use as a (reasonable) solution in a two-person zero-sum game ([Aubin (1979)]).

Proof. Define the function $F: L \rightarrow \mathbb{R}$ by $F(y):=\inf _{x \in K} f(x, y)(y \in L)$. By Theorem 3.33, for every $y \in L$ there exists $x \in K$ such that $F(y)=f(x, y)$. By strict convexity, this element $x$ is uniquely determined. We denote $x:=\Phi(y)$ and thus obtain

$$
\begin{equation*}
F(y)=\inf _{x \in K} f(x, y)=f(\Phi(y), y), \quad y \in L . \tag{3.11}
\end{equation*}
$$

By concavity of the function $y \mapsto f(x, y)$ and by the definition of $F$, for every $y_{1}$, $y_{2} \in L$ and every $t \in[0,1]$,

$$
\begin{aligned}
F\left(t y_{1}+(1-t) y_{2}\right) & =f\left(\Phi\left(t y_{1}+(1-t) y_{2}\right), t y_{1}+(1-t) y_{2}\right) \\
& \geq t f\left(\Phi\left(t y_{1}+(1-t) y_{2}\right), y_{1}\right)+(1-t) f\left(\Phi\left(t y_{1}+(1-t) y_{2}\right), y_{2}\right) \\
& \geq t F\left(y_{1}\right)+(1-t) F\left(y_{2}\right)
\end{aligned}
$$

so that $F$ is concave. Moreover, $F$ is upper semicontinuous: let $\left(y_{n}\right) \subseteq L$ be convergent to $y \in L$. For every $x \in K$ and every $n \in \mathbb{N}$ one has $F\left(y_{n}\right) \leq f\left(x, y_{n}\right)$, and taking the limes superior on both sides, we obtain, by continuity of $f$,

$$
\limsup _{n \rightarrow \infty} F\left(y_{n}\right) \leq \limsup _{n \rightarrow \infty} f\left(x, y_{n}\right)=f(x, y) .
$$

Since $x \in K$ was arbitrary, this inequality implies $\limsup _{n \rightarrow \infty} F\left(y_{n}\right) \leq F(y)$, i.e. $F$ is upper semicontinuous.

By Theorem 3.33 (applied to $-F$; use also Remark 3.34), there exists $\bar{y} \in L$ such that

$$
f(\Phi(\bar{y}), \bar{y})=F(\bar{y})=\sup _{y \in L} F(y) .
$$

We put $\bar{x}=\Phi(\bar{y})$ and show that $(\bar{x}, \bar{y})$ is a saddle point. Clearly, for every $x \in K$,

$$
\begin{equation*}
f(\bar{x}, \bar{y}) \leq f(x, \bar{y}) . \tag{3.12}
\end{equation*}
$$

Therefore it remains to show that for every $y \in L$,

$$
\begin{equation*}
f(\bar{x}, \bar{y}) \geq f(\bar{x}, y) . \tag{3.13}
\end{equation*}
$$

Let $y \in L$ be arbitrary and put $y_{n}:=\left(1-\frac{1}{n}\right) \bar{y}+\frac{1}{n} y$ and $x_{n}=\Phi\left(y_{n}\right)$. Then, by concavity,

$$
\begin{aligned}
F(\bar{y}) & \geq F\left(y_{n}\right)=f\left(x_{n}, y_{n}\right) \\
& \geq\left(1-\frac{1}{n}\right) f\left(x_{n}, \bar{y}\right)+\frac{1}{n} f\left(x_{n}, y\right) \\
& \geq\left(1-\frac{1}{n}\right) F(\bar{y})+\frac{1}{n} f\left(x_{n}, y\right),
\end{aligned}
$$

or

$$
F(\bar{y}) \geq f\left(x_{n}, y\right) \text { for every } n \in \mathbb{N}
$$

Since $K$ is bounded and closed, the sequence $\left(x_{n}\right) \subseteq K$ has a weakly convergent subsequence which converges to some element $x_{0} \in K$ (Theorem 3.28 and Corollary 3.31 ). By the preceeding inequality and Corollary 3.32,

$$
F(\bar{y}) \geq f\left(x_{0}, y\right)
$$

This is just the remaining inequality (3.13) if we can prove that $x_{0}=\bar{x}$. By concavity, for every $x \in K$ and every $n \in \mathbb{N}$,

$$
\begin{aligned}
f\left(x, y_{n}\right) & \geq f\left(x_{n}, y_{n}\right) \\
& \geq\left(1-\frac{1}{n}\right) f\left(x_{n}, \bar{y}\right)+\frac{1}{n} f\left(x_{n}, y\right) \\
& \geq\left(1-\frac{1}{n}\right) f\left(x_{n}, \bar{y}\right)+\frac{1}{n} F(y) .
\end{aligned}
$$

Letting $n \rightarrow \infty$ in this inequality and using Corollary 3.32 again, we obtain that for every $x \in K$,

$$
f(x, \bar{y}) \geq f\left(x_{0}, \bar{y}\right) .
$$

Hence, $x_{0}=\Phi(\bar{y})=\bar{x}$ and the theorem is proved.

## Chapter 4 <br> Uniform boundedness, bounded inverse and closed graph

This chapter is devoted to the other fundamental theorems in functional analysis; other than the Hahn-Banach theorem which has been discussed in the previous chapter. These fundamental results are

- the uniform boundedness principle or the Banach-Steinhaus theorem,
- the bounded inverse theorem (and the related open mapping theorem), and
- the closed graph theorem.

All these fundamental results rely on an abstract lemma for metric spaces.

### 4.1 The lemma of Baire

Lemma 4.1 (Baire). Let $(M, d)$ be a complete metric space, and let $\left(O_{n}\right)$ be a sequence of open and dense subsets of $M$. Then $\bigcap_{n} O_{n}$ is dense in $M$.

Proof. We can assume that $M$ is not empty since the statement is trivial otherwise. Let $x_{0} \in M$ and $\varepsilon>0$ be arbitrary. We have to prove that $\bigcap_{n} O_{n} \cap B\left(x_{0}, \varepsilon\right)$ is not empty.

Since $O_{1}$ is dense and open in $M$, the intersection $B\left(x_{0}, \varepsilon\right) \cap O_{1}$ is open and nonempty. Hence, there exists $\varepsilon_{1}>0$ (w.l.o.g. $\varepsilon_{1} \leq \varepsilon / 2$ ) and $x_{1} \in B\left(x_{0}, \varepsilon\right) \cap O_{1}$ such that

$$
B\left(x_{1}, \varepsilon_{1}\right) \subseteq B\left(x_{0}, \varepsilon\right) \cap O_{1}
$$

Choosing $\varepsilon_{1}$ a little bit smaller, if necessary, we can even assume that

$$
\overline{B\left(x_{1}, \varepsilon_{1}\right)} \subseteq B\left(x_{0}, \varepsilon\right) \cap O_{1}
$$

Since $O_{2}$ is dense and open in $M$, the intersection $B\left(x_{1}, \varepsilon_{1}\right) \cap O_{2}$ is open and nonempty. Hence, there exists $\varepsilon_{2}>0$ (w.l.o.g. $\varepsilon_{2} \leq \varepsilon_{1} / 2$ ) and $x_{2} \in B\left(x_{1}, \varepsilon_{1}\right) \cap O_{2}$ such that

$$
\overline{B\left(x_{2}, \varepsilon_{2}\right)} \subseteq B\left(x_{1}, \varepsilon_{1}\right) \cap O_{2} \subseteq B\left(x_{0}, \varepsilon\right) \cap O_{1} \cap O_{2}
$$

Proceeding inductively, we can construct sequences $\left(\varepsilon_{n}\right) \subseteq(0, \infty)$ and $\left(x_{n}\right) \subseteq M$ such that
(i) $\varepsilon_{n} \leq \varepsilon_{n-1} / 2$ and
(ii) for every $n \in \mathbb{N}$

$$
\overline{B\left(x_{n}, \varepsilon_{n}\right)} \subseteq B\left(x_{n-1}, \varepsilon_{n-1}\right) \cap O_{n} \subseteq B\left(x_{0}, \varepsilon\right) \cap \bigcap_{j=1}^{n} O_{j} .
$$

In particular, $x_{m} \in B\left(x_{n}, \varepsilon_{n}\right)$ for every $m \geq n$ and $\lim _{n \rightarrow \infty} \varepsilon_{n}=0$. Hence, the sequence $\left(x_{n}\right)$ is a Cauchy sequence in $M$. Since $M$ is complete, there exists $x:=\lim _{n \rightarrow \infty} x_{n} \in$ $M$. By the above,

$$
x \in \overline{B\left(x_{n}, \varepsilon_{n}\right)} \text { for every } n \in \mathbb{N}
$$

or

$$
x \in \bigcap_{n} \overline{B\left(x_{n}, \varepsilon_{n}\right)} \subseteq B\left(x_{0}, \varepsilon\right) \cap \bigcap_{n} O_{n}
$$

The claim is proved.
Lemma 4.2 (Baire). Let $(M, d)$ be a complete, nonempty, metric space, and let $\left(A_{n}\right)$ be a sequence of closed subsets in $M$ such that $M=\bigcup_{n} A_{n}$. Then there exists $n_{0} \in \mathbb{N}$ such that $A_{n_{0}}$ has nonempty interior.

Proof. Assume the contrary, i.e. that every $A_{n}$ has empty interior. In this case, the sets $O_{n}:=M \backslash A_{n}$ are open and dense. By assumption,

$$
\emptyset=M \backslash \bigcup_{n} A_{n}=\bigcap_{n} O_{n}
$$

a contradiction to Lemma 4.1 and the assumption that $M$ is nonempty.
Remark 4.3. The assumption in Lemma 4.1 or Lemma 4.2 that $M$ is complete is necessary in general. For example,

$$
\mathbb{Q}=\bigcup_{x \in \mathbb{Q}}\{x\}
$$

and this union is countable. Each one point set $\{x\}$ is closed but in this example, none of these sets has nonempty interior.

Remark 4.4. As a corollary to the lemma of Baire one obtains for example that there exists a continuous function $f \in C([0,1])$ which is nowhere differentiable. In fact, the set of such functions is dense in $C([0,1])$; see [Werner (1997)].

### 4.2 The uniform boundedness principle

Theorem 4.5 (Uniform boundedness principle; Banach-Steinhaus). Let $X, Y$ be Banach spaces and let $\left(T_{i}\right)_{i \in I} \subseteq \mathscr{L}(X, Y)$ be a family of bounded linear operators such that

$$
\sup _{i \in I}\left\|T_{i} x\right\|<\infty \text { for every } x \in X
$$

Then

$$
\sup _{i \in I}\left\|T_{i}\right\|<\infty .
$$

Remark 4.6. Theorem 4.5 is in general not true if $X$ is only a normed space. For example, let $X=c_{00}(=Y)$ be the space of all finite sequences equipped with the supremum norm (or any other reasonable norm). Let

$$
T_{n} x=T_{n}\left(x_{m}\right)=\left(a_{n m} x_{m}\right)
$$

with

$$
a_{n m}=\left\{\begin{array}{l}
m \text { if } m \leq n \\
0 \text { if } m>n
\end{array}\right.
$$

Then $\sup _{n}\left\|T_{n} x\right\|$ is finite for every $x \in X$, but $\left\|T_{n}\right\|=n$ is unbounded.
Remark 4.7. The fact that in Theorem 4.5 we suppose also $Y$ to be a Banach space is not important. In fact, if $Y$ is not complete, then we may embed $Y$ into its completion $\tilde{Y}$ and consider every operator $T_{i} \in \mathscr{L}(X, Y)$ also as an operator in $\mathscr{L}(X, \tilde{Y})$.

Proof (Proof of Theorem 4.5). Let $A_{n}:=\left\{x \in X: \sup _{i \in I}\left\|T_{i} x\right\| \leq n\right\}$. Since arbitrary intersections of closed sets are closed, and by the boundedness of the $T_{i}$, the sets $A_{n}$ are closed for every $n \in \mathbb{N}$. By assumption, $X=\bigcup_{n} A_{n}$.

Hence, by the lemma of Baire (Lemma 4.2), there exists $n_{0} \in \mathbb{N}$ such that $A_{n_{0}}$ has nonempty interior, i.e. there exist $n_{0} \in \mathbb{N}, x_{0} \in X$ and $\varepsilon>0$ such that

$$
\sup _{i \in I}\left\|T_{i} x\right\| \leq n_{0} \text { for every } x \in B\left(x_{0}, \varepsilon\right),
$$

or, in other words, there exists $n_{0} \in \mathbb{N}, x_{0} \in X$ and $\varepsilon>0$ such that

$$
\left\|T_{i}\left(x_{0}+\varepsilon x\right)\right\| \leq n_{0} \text { for every } x \in B(0,1), i \in I .
$$

This implies, by the triangle inequality,

$$
\varepsilon\left\|T_{i} x\right\| \leq n_{0}+\left\|T_{i} x_{0}\right\| \leq 2 n_{0} \text { for every } x \in B(0,1), i \in I
$$

The claim is proved.
Corollary 4.8. Let $X, Y$ be Banach spaces and let $\left(T_{n}\right) \subseteq \mathscr{L}(X, Y)$ be a strongly convergent sequence of bounded linear operators, i.e.

$$
T x:=\lim _{n \rightarrow \infty} T_{n} x \text { exists for every } x \in X
$$

Then $\sup _{n \in \mathbb{N}}\left\|T_{n}\right\|=: M<\infty$ and $T \in \mathscr{L}(X, Y)$.
Proof. Linearity of $T$ is clear. Since $\left(T_{n}\right)$ is strongly convergent, the sequence $\left(T_{n} x\right)$ is bounded for every $x \in X$. By the uniform bounded principle (Theorem 4.5), $\sup _{n \in \mathbb{N}}\left\|T_{n}\right\|=: M<\infty$. As a consequence, for every $x \in X$,

$$
\|T x\|=\lim _{n \rightarrow \infty}\left\|T_{n} x\right\| \leq M\|x\|,
$$

so that $T$ is bounded.
Corollary 4.9. Every weakly convergent sequence in a Banach space is bounded.
Proof. Let $X$ be a Banach space and $\left(x_{n}\right) \subseteq X$ be weakly convergent. Considering the $x_{n}$ as elements in $X^{\prime \prime}=\mathscr{L}\left(X^{\prime}, \mathbb{K}\right)$ by the embedding $J: X \rightarrow X^{\prime \prime}$, the claim follows from Corollary 4.8.

### 4.3 Open mapping theorem, bounded inverse theorem

Theorem 4.10 (Open mapping theorem). Let $X, Y$ be two Banach spaces and let $T \in \mathscr{L}(X, Y)$ be surjective. Then there exists $r>0$ such that

$$
\begin{equation*}
T B_{X}(0,1) \supseteq B_{Y}(0, r) . \tag{4.1}
\end{equation*}
$$

Proof. First step: We show that there exists $r>0$ such that

$$
\begin{equation*}
B(0,2 r) \subseteq \overline{T B(0,1)} \tag{4.2}
\end{equation*}
$$

For this, we remark first that by surjectivity,

$$
Y=T X=\bigcup_{n} T B(0, n)=\bigcup_{n} \overline{T B(0, n)} .
$$

By the Lemma of Baire, there exists $n_{0}$ such that $\overline{T B\left(0, n_{0}\right)}$ has nonempty interior, i.e. there exist $x \in \overline{T B\left(0, n_{0}\right)}$ and $\varepsilon>0$ such that

$$
B(x, \varepsilon) \subseteq \overline{T B\left(0, n_{0}\right)}
$$

By symmetry,

$$
B(-x, \varepsilon) \subseteq \overline{T B\left(0, n_{0}\right)}
$$

and adding both 'inequalities' together, we obtain

$$
B(0, \varepsilon) \subseteq \overline{T B\left(0, n_{0}\right)},
$$

which implies the required inclusion (4.2) if we put $r=\frac{\varepsilon}{2 n_{0}}$.

Second step: We prove (4.1). Let $y \in B(0, r)$, where $r>0$ is as in (4.2) from the first step. Then, by (4.2), for every $\varepsilon>0$ there exists $x \in B\left(0, \frac{1}{2}\right)$ such that $\| y-$ $T x \|<\varepsilon$. In particular, if we choose $\varepsilon=\frac{r}{2}$, then there exists $x_{1} \in B\left(0, \frac{1}{2}\right)$ such that $\left\|y-T x_{1}\right\|<\frac{r}{2}$.

Similarly, since $y-T x_{1} \in B\left(0, \frac{r}{2}\right)$, there exists $x_{2} \in B\left(0, \frac{1}{4}\right)$ such that $\|(y-$ $\left.T x_{1}\right)-T x_{2} \| \leq \frac{r}{4}$. Iterating this construction, we find a sequence $\left(x_{n}\right)$ such that $x_{n} \in B\left(0,2^{-n}\right)$ and such that $\left\|y-\sum_{j=1}^{n} T x_{n}\right\| \leq 2^{-n} r$. Since $X$ is complete and since $\sum_{n} x_{n}$ is absolutely convergent with $\sum_{n}\left\|x_{n}\right\|<1$, the limit $x=\sum_{n} x_{n}$ exists and $x \in B(0,1)$. By the preceeding estimates, $\|y-T x\|=0$ or $T x=y$. Thus we have proved (4.1).

Remark 4.11. It is not difficult to prove that if an operator $T \in \mathscr{L}(X, Y)$ satisfies (4.1), then $T O$ is open for every open $O \subseteq X$. A function which maps open sets into open sets is called open; whence the name of the open mapping theorem.

Corollary 4.12 (Bounded inverse theorem). Let $X, Y$ be two Banach spaces and let $T \in \mathscr{L}(X, Y)$ be bijective. Then $T^{-1} \in \mathscr{L}(Y, X)$.

Proof. Linearity of $T^{-1}$ is clear. By the open mapping theorem (Theorem 4.10), we have

$$
T^{-1} B_{Y}(0,1) \subseteq B_{X}\left(0, \frac{1}{r}\right)
$$

for some $r>0$. Hence, $T^{-1}$ is bounded.
Corollary 4.13. Let $\|\cdot\|_{1}$ and $\|\cdot\|_{2}$ be two norms on a vector space $X$ such that $\left(X,\|\cdot\|_{1}\right)$ and $\left(X,\|\cdot\|_{2}\right)$ are complete. If there exists a constant $C>0$ such that

$$
\|x\|_{2} \leq C\|x\|_{1} \text { for every } x \in X
$$

then the two norms are equivalent.
Proof. It suffices to consider the identity $I:\left(X,\|\cdot\|_{1}\right) \rightarrow\left(X,\|\cdot\|_{2}\right)$. It is bounded by assumption, and clearly it is bijective. By the bounded inverse theorem (Corollary 4.12), the inverse $I^{-1}:\left(X,\|\cdot\|_{2}\right) \rightarrow\left(X,\|\cdot\|_{1}\right)$ is bounded, i.e. there exists $c>0$ such that

$$
\|x\|_{1} \leq c\|x\|_{2} \text { for every } x \in X
$$

### 4.4 Closed graph theorem

Let $X, Y$ be two Banach spaces, and let $\operatorname{dom} T \subseteq X$ be a linear subspace. A linear operator $T: \operatorname{dom} T \rightarrow Y$ is called a closed operator if the graph

$$
\text { Graph } T:=\{(x, T x): x \in \operatorname{dom} T\}
$$

is closed in $X \times Y$.

Lemma 4.14. A linear operator $T: X \supseteq \operatorname{dom} T \rightarrow Y$ is closed if and only if

$$
\left.\begin{array}{l}
\operatorname{dom} T \ni x_{n} \rightarrow x \text { in } X \text { and }  \tag{4.3}\\
T x_{n} \rightarrow y \text { in } Y
\end{array}\right\} \Rightarrow x \in \operatorname{dom} T \text { and } T x=y
$$

Proof. Exercise.
Lemma 4.15. Every bounded linear operator $T \in \mathscr{L}(X, Y)(X, Y$ Banach spaces $)$ is closed.

Proof. This is an immediate consequence of Lemma 4.14.
Lemma 4.16. A linear operator $T: X \supseteq \operatorname{dom} T \rightarrow Y$ is closed if and only if the space dom $T$ equipped with the graph norm

$$
\|x\|_{\operatorname{dom} T}:=\|x\|_{X}+\|T x\|_{Y}, \quad x \in X
$$

is complete.
Proof. $\Rightarrow$ Assume that $T$ is closed. Let $\left(x_{n}\right)$ be a Cauchy sequence in $(\operatorname{dom} T, \| \cdot$ $\left.\|_{\operatorname{dom} T}\right)$. Then $\left(x_{n}\right)$ is a Cauchy sequence in $X$ and $\left(T x_{n}\right)$ is a Cauchy sequence in $Y$. Since $X$ and $Y$ are complete, there exist $x \in X$ and $y \in Y$ such that $x_{n} \rightarrow x$ and $T x_{n} \rightarrow y$. Since $T$ is closed, and by Lemma 4.14, this implies $x \in \operatorname{dom} T$ and $T x=y$. Moreover,

$$
\left\|x_{n}-x\right\|_{\operatorname{dom} T}=\left\|x_{n}-x\right\|_{X}+\left\|T x_{n}-T x\right\|_{Y} \rightarrow 0
$$

so that $\left(x_{n}\right)$ converges in $\left(\operatorname{dom} T,\|\cdot\|_{\operatorname{dom} T}\right)$. Hence, $\operatorname{dom} T$ equipped with the graph norm is complete.
$\Leftarrow$ Assume that $\left(\operatorname{dom} T,\|\cdot\|_{\operatorname{dom} T}\right)$ is complete. Assume that $\operatorname{dom} T \ni x_{n} \rightarrow x \in X$ and $T x_{n} \rightarrow y \in Y$. Then $\left(x_{n}\right)$ and $\left(T x_{n}\right)$ are Cauchy sequences in $X$ and $Y$, respectively. By the definition of $\|\cdot\|_{\operatorname{dom} T}$, this implies that $\left(x_{n}\right)$ is a Cauchy sequence in $\left(\operatorname{dom} T,\|\cdot\|_{\operatorname{dom} T}\right)$. By completeness, there exists $\bar{x} \in \operatorname{dom} T$ such that $x_{n} \rightarrow \bar{x}$ in dom $T$ (with respect to the graph norm). Since convergence of $\left(x_{n}\right)$ in dom $T$ implies the convergence of $\left(x_{n}\right)$ in $X$, and since $\left(x_{n}\right)$ converges to $x$ in $X$, we find $x=\bar{x} \in \operatorname{dom} T$ by the uniqueness of the limit. Moreover, since $T$ is always bounded from dom $T$ (when equipped with the graph norm) into $Y$, we have $T x=\lim _{n \rightarrow \infty} T x_{n}=y$. Hence, by Lemma 4.14, $T$ is closed.

Example 4.17. Let $X=Y=C([0,1])$ be equipped with the supremum norm, and let $\operatorname{dom} T:=C^{1}([0,1]) \subseteq X$. Let $T f:=f^{\prime}$ for $f \in \operatorname{dom} T$. Then $T$ is a closed operator. In fact, the graph norm $\|\cdot\|_{\operatorname{dom} T}$ coincides with the canonical norm on $C^{1}([0,1])$, i.e.

$$
\|f\|_{C^{1}}:=\|f\|_{\infty}+\left\|f^{\prime}\right\|_{\infty},
$$

and $\left(C^{1}([0,1]),\|\cdot\|_{C^{1}}\right)$ is complete.
Theorem 4.18 (Closed graph theorem). Let $X, Y$ be two Banach spaces and let $T: X \rightarrow Y$ be a closed operator. Then $T$ is bounded.

Remark 4.19. The important assumption in Theorem 4.18, besides the assumption that $T$ is closed, is the assumption that dom $T=X$ ! The Example 4.17 shows that closed operators need not be bounded in general; when considered from (dom $T,\|\cdot\|_{X}$ ) with values in $Y$. Note that in Example 4.17, dom $T$ is not complete when equipped with the norm coming from $X$.

Proof (Proof of Theorem 4.18). By assumption $\left(X,\|\cdot\|_{X}\right)$ is a Banach space, and by closedness of $T$ and Lemma 4.16, also $\left(X,\|\cdot\|_{\operatorname{dom} T}\right)$ is a Banach space, where $\|\cdot\|_{\text {dom } T}$ denotes the graph norm. Moreover, trivially,

$$
\|x\|_{X} \leq\|x\|_{\operatorname{dom} T} \text { for every } x \in X
$$

By Corollary 4.13, the two norms $\|\cdot\|_{X}$ and $\|\cdot\|_{\text {dom } T}$ are equivalent, that is, there exists a constant $C \geq 0$ such that

$$
\|x\|_{X}+\|T x\|_{Y} \leq C\|x\|_{X} \text { for every } x \in X
$$

As a consequence, $T$ is bounded.
Example 4.20 (Sobolev embedding). Let $-\infty<a<b<\infty$. Then the embedding

$$
\begin{aligned}
J: W^{1, p}(a, b) & \rightarrow C([a, b]), \\
u & \mapsto u
\end{aligned}
$$

is well defined and bounded, that is, there exists a constant $C \geq 0$ such that

$$
\|u\|_{\infty} \leq C\|u\|_{W^{1, p}} \text { for every } u \in W^{1, p}(a, b) .
$$

Recall that this embedding is well defined since every function $u \in W^{1, p}(a, b)$ is continuous on $[a, b]$ by Theorem 9.8 of Chapter 9 .

In order to see that $J$ is also bounded, we apply the closed graph theorem together with the characterization in Lemma 4.14: let $\left(u_{n}\right) \subseteq W^{1, p}(a, b)$ be such that $u=$ $\lim _{n \rightarrow \infty} u_{n}$ exists in $W^{1, p}(a, b)$ and such that $v=\lim _{n \rightarrow \infty} u_{n}$ exists in $C([a, b])$. The convergence in $W^{1, p} \subseteq L^{p}$ implies that $u_{n} \rightarrow u$ almost everywhere if we extract a subsequence. The convergence in $C$ implies that $u_{n} \rightarrow v$ everywhere. Hence $u=v$ almost everyhwere, and since both functions are continuous, we obtain $u=v$. Hence, the embedding is closed. By the closed graph theorem, the embedding $W^{1, p} \rightarrow C$ is bounded.

Exercise 4.21 Let $T: X \supseteq \operatorname{dom} T \rightarrow Y$ be a closed, injective operator. Define

$$
\begin{aligned}
\operatorname{dom} T^{-1} & :=\operatorname{ran} T=\{T x: x \in \operatorname{dom} T\} \subseteq Y, \\
T^{-1} y & :=x \text { where } x \in \operatorname{dom} T \text { is the unique element such that } T x=y .
\end{aligned}
$$

Then $T^{-1}$ is a closed operator.
If in addition $T$ is surjective, then $T^{-1}: Y \rightarrow X$ is bounded.

## 4.5*Vector-valued analytic functions

Let $X$ be a complex Banach space and let $\Omega \subseteq \mathbb{C}$ be an open subset. We say that a function $f: \Omega \rightarrow X$ is analytic (or: holomorphic) if

$$
f^{\prime}\left(z_{0}\right):=\lim _{z \rightarrow z_{0}} \frac{f(z)-f\left(z_{0}\right)}{z-z_{0}} \text { exists for every } z_{0} \in \Omega
$$

We say that $f: \Omega \rightarrow X$ is weakly analytic (or: weakly holomorphic) if $x^{\prime} \circ f$ : $\Omega \rightarrow \mathbb{C}$ is analytic for every $x^{\prime} \in X^{\prime}$.

Theorem 4.22. A function $f: \Omega \rightarrow X$ is analytic if and only if it is weakly analytic.
Proof. Cleary, if $f$ is analytic, then $f$ is weakly analytic. So we only have to prove the other direction.

By considering $X$ as a closed subspace of $X^{\prime \prime}$ (via the embedding $J$ ), and by replacing then $X$ by $X^{\prime \prime}$ (so that the function $f$ becomes $X^{\prime \prime}$-valued), we can assume that $X$ is a dual space. But doing this, we no longer assume that $f$ is weakly analytic. The assertion which we have to prove is then the following:

Let $X$ be a complex Banach space, and let $X^{\prime}$ be its dual. Let $f: \Omega \rightarrow X^{\prime}$ be such that $\langle f, x\rangle: \Omega \rightarrow \mathbb{C}$ is analytic for every $x \in X$. Then $f$ is analytic.

In fact, it suffices to prove that for fixed $z_{0} \in \Omega$ there exists $M \geq 0$ such that for every $y, z \in \Omega \backslash\left\{z_{0}\right\}$ 'close' to $z_{0}$,

$$
\begin{equation*}
\left\|\frac{f(z)-f\left(z_{0}\right)}{z-z_{0}}-\frac{f(y)-f\left(z_{0}\right)}{y-z_{0}}\right\| \leq M|z-y| . \tag{4.4}
\end{equation*}
$$

Let $K:=\overline{B\left(z_{0}, r\right)} \backslash\left\{z_{0}\right\}$, where $r>0$ is chosen so small that $K \subseteq \Omega$. Let

$$
\tilde{K}=(K \times K) \backslash\{(z, z): z \in K\}
$$

be the cartesian product of $K$ and $K$ from which we take out the 'diagonal'.
By assumption, for every $x \in X$, the function $\langle f, x\rangle$ is analytic. Hence, for every $x \in X$ we have

$$
\sup _{(y, z) \in \tilde{K}}\left|\left\langle\frac{\frac{f(z)-f\left(z_{0}\right)}{z-z_{0}}-\frac{f(y)-f\left(z_{0}\right)}{y-z_{0}}}{y-z}, x\right\rangle\right|<\infty .
$$

By the uniform boundedness principle, this implies

$$
\sup _{(y, z) \in \tilde{K}}\left\|\frac{\frac{f(z)-f\left(z_{0}\right)}{z-z_{0}}-\frac{f(y)-f\left(z_{0}\right)}{y-z_{0}}}{y-z}\right\|=: M<\infty,
$$

which actually implies (4.4) for every $y, z \in K$.
By Theorem 4.22, many important properties of 'classical' analytic functions $\Omega \rightarrow \mathbb{C}$ carry over to vector-valued analytic functions $\Omega \rightarrow X$. For example:

- Every analytic function $f: \Omega \rightarrow X$ is infinitely many times differentiable.
- Every analytic function $f: \Omega \rightarrow X$ can be locally developed into a power series of the form $\sum_{n=0}^{\infty} a_{n}\left(z-z_{0}\right)^{n}$ with $a_{n} \in X$. In fact: $a_{n}=\frac{1}{n!} f^{(n)}\left(z_{0}\right)$.
- Cauchy's integral formula $f(z)=\frac{1}{2 \pi i} \int_{\gamma} \frac{f(y)}{z-y} d y$ holds true for appropriate paths $\gamma$. Note, however, that we have not yet defined integrals of vector-valued functions.

An important example of a vector-valued analytic function will be the resolvent of an operator $T \in \mathscr{L}(X)$; see the Chapter 5 .

## Chapter 5 <br> Spectral theory of operators on Banach spaces, compact operators, nuclear operators

### 5.1 Spectrum of closed operators

Let $X$ be a Banach space. A linear operator between two Banach spaces $X$ and $Y$ is a pair $(A, \operatorname{dom} A)$ where $\operatorname{dom} A \subseteq X$ is a linear subspace and $A: \operatorname{dom} A \rightarrow Y$ is a linear mapping. We call $\operatorname{dom} A$ the domain of $A$. Furthermore, we define the kernel, the range, and the graph of $A$ respectively by

$$
\begin{aligned}
\operatorname{ker} A & :=\{x \in X: A x=0\}, \\
\operatorname{ran} A & :=\{y \in Y: \exists x \in \operatorname{dom} A \text { s.t. } A x=y\} \text { and } \\
\operatorname{graph} A & :=\{(x, y) \in X \times Y: x \in \operatorname{dom} A \text { and } A x=y\} .
\end{aligned}
$$

We say that a linear operator from $X$ into $Y$ is densely defined if its domain is dense in $X$. If the domain is clear from the context, then we simply speak of a linear operator $A$ on $X$. For a bounded, linear operator $A$ we always assume, unless otherwise stated, that $\operatorname{dom} A=X$. Recall that an operator $A$ on $X$ is closed if its graph graph $A$ is closed in $X \times X$. We recall that an operator $A$ on $X$ is closed if and only if its domain, equipped with the graph norm, is complete. We also recall the Closed Graph Theorem (Theorem 4.18) which says that every closed operator $A$ with domain $\operatorname{dom} A=X$ is automatically bounded.

For every $\lambda \in \mathbb{K}$ we write $\lambda-A:=\lambda I-A$, where $I$ is the identity operator on $X$ and $\operatorname{dom}(\lambda-A):=\operatorname{dom} A$. We define the resolvent set of $A$ by

$$
\begin{gathered}
\rho(A):=\{\lambda \in \mathbb{K}: \lambda-A: \operatorname{dom} A \rightarrow X \text { is bijective and } \\
\left.(\lambda-A)^{-1} \text { is bounded on } X\right\} .
\end{gathered}
$$

We emphasize that the inverse $(\lambda-A)^{-1}$ is considered as an operator from $X$ into $X$, and not as an operator from $X$ into $\operatorname{dom} A$, although it effectively maps into $\operatorname{dom} A$. For every $\lambda \in \rho(A)$ we write

$$
R(\lambda, A):=(\lambda-A)^{-1}
$$

and we call $R(\lambda, A)$ the resolvent of $A$ at $\lambda$. The mapping $\rho(A) \rightarrow \mathscr{L}(X), \lambda \mapsto$ $R(\lambda, A)$ is called the resolvent of $A$.

The set

$$
\sigma(A):=\mathbb{K} \backslash \rho(A)
$$

is called the spectrum of $A$. Moreover, we define the point spectrum, the approximative point spectrum, the continuous spectrum and the residual spectrum, respectively, by

$$
\begin{aligned}
& \sigma_{p}(A):=\{\lambda \in \mathbb{K}: \lambda-A \text { is not injective }\} \\
& =\{\lambda \in \mathbb{K}: \exists x \in X \backslash\{0\} \text { s.t. } A x=\lambda x\} \\
& \sigma_{a p}(A):=\left\{\lambda \in \mathbb{K}: \exists\left(x_{n}\right) \subseteq \operatorname{dom} A \text { s.t. }\left\|x_{n}\right\|=1 \text { and }(\lambda-A) x_{n} \rightarrow 0\right\}, \\
& \sigma_{c}(A):=\{\lambda \in \mathbb{K}: \lambda-A \text { is injective, has dense range, but } \\
& \left.(\lambda-A)^{-1}: \operatorname{ran} A \rightarrow X \text { is not bounded }\right\} \text {, and } \\
& \sigma_{r}(A):=\{\lambda \in \mathbb{K}: \operatorname{ran}(\lambda-A) \text { is not dense in } X\} .
\end{aligned}
$$

Our first lemma shows that if we look for operators with non-empty resolvent set, then we necessarily have to search in the class of closed operators.

Lemma 5.1. If the resolvent set of a linear operator $A$ on a Banach space $X$ is nonempty, then A is closed.

Proof. Let $A$ be a linear operator on a Banach space $X$. Assume that the resolvent set is non-empty, and let $\lambda \in \rho(A)$. Then $\lambda-A$ is bijective and $(\lambda-A)^{-1}$ is a bounded, linear operator on $X$. In particular, $(\lambda-A)^{-1}$ is closed. This means that

$$
\operatorname{graph}(\lambda-A)^{-1}=\left\{(y, x) \in X \times X:(\lambda-A)^{-1} y=x\right\}
$$

is closed in $X \times X$. Hence,

$$
\operatorname{graph}(\lambda-A)=\{(x, y) \in X \times X: x \in \operatorname{dom} A \text { and }(\lambda-A) x=y\}
$$

is closed in $X \times X$. This easily implies that $A$ has closed graph.
Lemma 5.2 (Resolvent identity). For every $\lambda, \mu \in \rho(A)$ one has

$$
R(\lambda, A)-R(\mu, A)=(\mu-\lambda) R(\mu, A) R(\lambda, A)
$$

Proof. For every $\lambda, \mu \in \rho(A)$

$$
\mu-\lambda=(\mu-A)-(\lambda-A) .
$$

Multiplying both sides by $R(\mu, A)$ and $R(\lambda, A)$, one obtains the claim.
Lemma 5.3 (The resolvent is analytic). The resolvent set $\rho(A)$ is open in $\mathbb{K}$ and the resolvent $\rho(A) \rightarrow \mathscr{L}(X), \lambda \mapsto R(\lambda, A)$ is analytic, which means that it can locally near every point $\lambda \in \rho(A)$ be developped into a power series which converges to the resolvent itself.

Proof. Let $\lambda \in \rho(A)$ and $\mu \in \mathbb{K}$. Then

$$
\mu-A=\mu-\lambda+\lambda-A=((\mu-\lambda) R(\lambda, A)+I)(\lambda-A),
$$

and the right-hand side is invertible if $|\mu-\lambda|<1 /\|R(\lambda, A)\|$ by the Neumann series. Hence, $\rho(A)$ is open in $\mathbb{K}$. The Neumann series precisely yields

$$
R(\mu, A)=\sum_{n=0}^{\infty}(-1)^{n} R(\lambda, A)^{n+1}(\mu-\lambda)^{n}
$$

so that the function $\lambda \mapsto R(\lambda, A)$ can be locally developped into a power series. As a consequence, this function is analytic.

Remark 5.4. One may also employ the resolvent identity in order to prove that the function $\lambda \mapsto R(\lambda, A)$ is analytic; but in this case one should at least prove continuity of the resolvent $R(\cdot, A)$.

Lemma 5.5 (Growth of the resolvent near the spectrum). For every $\lambda \in \rho(A)$ one has

$$
\|R(\lambda, A)\| \geq \operatorname{dist}(\lambda, \sigma(A))^{-1}
$$

Proof. As we have seen in the proof of the preceding Lemma 5.3, for $\lambda \in \rho(A)$ the condition

$$
|\mu-\lambda|\|R(\lambda, A)\|<1
$$

implies $\mu \in \rho(A)$. The claim follows.
Lemma 5.6 (The topological boundary of the spectrum belongs to the approximative point spectrum). For every linear operator $A$ one has

$$
\partial \sigma(A) \subseteq \sigma_{a p}(A)
$$

Proof. If $\lambda \in \partial \sigma(A)$, then there exists $\left(\lambda_{n}\right) \subseteq \rho(A)$ such that $\lim _{n \rightarrow \infty} \lambda_{n}=\lambda$. By Lemma 5.5, $\lim _{n \rightarrow \infty}\left\|R\left(\lambda_{n}, A\right)\right\|=\infty$. By the definition of the operator norm, there exists a sequence $\left(y_{n}\right) \subseteq X,\left\|y_{n}\right\|=1$, such that

$$
\lim _{n \rightarrow \infty}\left\|R\left(\lambda_{n}, A\right) y_{n}\right\|=\infty
$$

Put $x_{n}:=\frac{R\left(\lambda_{n}, A\right) y_{n}}{\left\|R\left(\lambda_{n}, A\right) y_{n}\right\|}$, so that $x_{n} \in \operatorname{dom} A$ and $\left\|x_{n}\right\|=1$. Then

$$
\lambda x_{n}-A x_{n}=\left(\lambda-\lambda_{n}\right) x_{n}+\frac{y_{n}}{\left\|R\left(\lambda_{n}, A\right) y_{n}\right\|} \rightarrow 0 \quad(n \rightarrow \infty) .
$$

As a consequence, $\lambda \in \sigma_{a p}(A)$.
Lemma 5.7. For a bounded, linear operator $T \in \mathscr{L}(X)$ one has

$$
\{\lambda \in \mathbb{C}:|\lambda|>\|T\|\} \subseteq \rho(T)
$$

and

$$
R(\lambda, T)=\sum_{n=0}^{\infty} \frac{T^{n}}{\lambda^{n+1}}, \quad|\lambda|>\|T\| .
$$

Proof. Use the identity

$$
\lambda-T=\lambda\left(I-\frac{T}{\lambda}\right)
$$

and the Neumann series.
Remark 5.8. In fact, $\lambda \in \rho(T)$ as soon as

$$
|\lambda|>\liminf _{n \rightarrow \infty}\left\|T^{n}\right\|^{\frac{1}{n}}=: r(T) .
$$

The number $r(T) \geq 0$ is called the spectral radius of $T$.
Lemma 5.9. For every bounded, linear operator $T \in \mathscr{L}(X)$ with $X \neq\{0\}$ a complex Banach space, the spectrum $\sigma(T)$ is nonempty and compact.

Proof. The compactness of $\sigma(T)$ follows Lemma 5.3 and 5.7. If $\sigma(T)$ was empty, then, by Lemma 5.3, the resolvent $\lambda \mapsto R(\lambda, T)$ is an entire function. On the other hand, by Lemma 5.7,

$$
\lim _{|\lambda| \rightarrow \infty}\|R(\lambda, T)\|=0
$$

By Liouville's theorem, this implies $R(\lambda, T) \equiv 0$, which is only possible if $X=\{0\}$ is the trivial space.

Let $(A, \operatorname{dom} A)$ be a densely defined, linear operator between two Banach spaces $X$ and $Y$. We defined the adjoint operator or dual operator $\left(A^{\prime}, \operatorname{dom} A^{\prime}\right)$ between $Y^{\prime}$ and $X^{\prime}$ by

$$
\begin{aligned}
\operatorname{dom} A^{\prime} & :=\left\{y^{\prime} \in Y^{\prime}: \exists x^{\prime} \in X^{\prime} \forall x \in \operatorname{dom} A:\left\langle x^{\prime}, x\right\rangle_{X^{\prime}, X}=\left\langle y^{\prime}, A x\right\rangle_{Y^{\prime}, Y}\right\} \text { and } \\
A^{\prime} y^{\prime} & :=x^{\prime}
\end{aligned}
$$

Lemma 5.10. For every linear operator $(A, \operatorname{dom} A)$ between $X$ and $Y$, the adjoint operator $\left(A^{\prime}, \operatorname{dom} A^{\prime}\right)$ between $Y^{\prime}$ and $X^{\prime}$ is closed.

Proof. Let $\left(y_{n}^{\prime}\right)$ be any sequence in $\operatorname{dom} A^{\prime}$ such that $y_{n}^{\prime} \rightarrow y^{\prime}$ in $Y^{\prime}$ and $A^{\prime} y_{n}^{\prime} \rightarrow x^{\prime}$ in $X^{\prime}$. Then, for every $x \in \operatorname{dom} A$,

$$
\begin{aligned}
\left\langle x^{\prime}, x\right\rangle_{X^{\prime}, X} & =\lim _{n}\left\langle A^{\prime} y_{n}^{\prime}, x\right\rangle_{X^{\prime}, X} \\
& =\lim _{n}\left\langle y_{n}^{\prime}, A x\right\rangle_{Y^{\prime}, Y} \\
& =\left\langle y^{\prime}, A x\right\rangle_{Y^{\prime}, Y} .
\end{aligned}
$$

By definition of the adjoint operator, this equality implies $y^{\prime} \in \operatorname{dom} A^{\prime}$ and $A^{\prime} y^{\prime}=x^{\prime}$. As a consequence, $\left(A^{\prime}, \operatorname{dom} A^{\prime}\right)$ is closed.

Lemma 5.11. For every linear operator $(A, \operatorname{dom} A)$ on a Banach space $X$, one has $\operatorname{dom}(\lambda-A)^{\prime}=\operatorname{dom} A^{\prime}$ and $(\lambda-A)^{\prime}=\lambda-A^{\prime}$

Proof. Exercise.
If $T \in \mathscr{L}(X, Y)$ is a bounded, linear operator between two Banach spaces $X$ and $Y$, then, for every $y^{\prime} \in Y^{\prime}$, the linear mapping $X \rightarrow \mathbb{K}, x \mapsto\left\langle y^{\prime}, T x\right\rangle$ is bounded on $X$. We denote this linear mapping by $T^{\prime} y^{\prime} \in X^{\prime}$. The resulting operator $T^{\prime}: Y^{\prime} \rightarrow X^{\prime}$ is just the adjoint operator as defined above; its domain dom $T^{\prime}$ is equal to $Y^{\prime}$. For every $x \in X$ and every $y^{\prime} \in Y^{\prime}$,

$$
\left\langle y^{\prime}, T x\right\rangle_{Y^{\prime}, Y}=\left\langle T^{\prime} y^{\prime}, x\right\rangle_{X^{\prime}, X} .
$$

Lemma 5.12. For every bounded, linear operator $T \in \mathscr{L}(X, Y)$, the adjoint $T^{\prime}$ : $Y^{\prime} \rightarrow X^{\prime}$ is bounded and $\|T\|=\left\|T^{\prime}\right\|$.
Proof. For every $y^{\prime} \in Y^{\prime}$,

$$
\left\|T^{\prime} y^{\prime}\right\|=\sup _{\|x\| \leq 1}\left|\left\langle T^{\prime} y^{\prime}, x\right\rangle\right|=\sup _{\|x\| \leq 1}\left|\left\langle y^{\prime}, T x\right\rangle\right| \leq\|T\|\left\|y^{\prime}\right\|,
$$

which proves that $T^{\prime}$ is bounded and that $\left\|T^{\prime}\right\| \leq\|T\|$. On the other hand, by HahnBanach (Corollary 3.8 of Chapter 3),

$$
\begin{aligned}
\left\|T^{\prime}\right\| & =\sup _{\left\|y^{\prime}\right\| \leq 1}\left\|T^{\prime} y^{\prime}\right\| \\
& =\sup _{\left\|y^{\prime}\right\| \leq 1} \sup _{\|x\| \leq 1}\left|\left\langle T^{\prime} y^{\prime}, x\right\rangle\right| \\
& =\sup _{\|x\| \leq 1} \sup _{\left\|y^{\prime}\right\| \leq 1}\left|\left\langle y^{\prime}, T x\right\rangle\right| \\
& =\sup _{\|x\| \leq 1}\|T x\| \\
& =\|T\|,
\end{aligned}
$$

and the claim is proved.
Lemma 5.13. For every closed, densely defined, linear operator $(A, \operatorname{dom} A)$ one has $\sigma(A)=\sigma\left(A^{\prime}\right)$. For every $\lambda \in \rho(A)$ one has

$$
R(\lambda, A)^{\prime}=R\left(\lambda, A^{\prime}\right)
$$

Proof. Let $\lambda \in \rho(A)$. For every $x^{\prime} \in \operatorname{dom} A^{\prime}$ and every $x \in X$ we have

$$
\begin{aligned}
\left\langle R(\lambda, A)^{\prime}\left(\lambda-A^{\prime}\right) x^{\prime}, x\right\rangle & =\left\langle\left(\lambda-A^{\prime}\right) x^{\prime}, R(\lambda, A) x\right\rangle \\
& =\left\langle x^{\prime},(\lambda-A) R(\lambda, A) x\right\rangle \\
& =\left\langle x^{\prime}, x\right\rangle,
\end{aligned}
$$

so that $R(\lambda, A)^{\prime}$ is a right-inverse of $\lambda-A^{\prime}$. Moreover, for every $x^{\prime} \in X^{\prime}$ and every $x \in \operatorname{dom} A$ we have

$$
\begin{aligned}
\left\langle\left(\lambda-A^{\prime}\right) R(\lambda, A)^{\prime} x^{\prime}, x\right\rangle & =\left\langle R(\lambda, A)^{\prime} x^{\prime},(\lambda-A) x\right\rangle \\
& =\left\langle x^{\prime}, R(\lambda, A)(\lambda-A) x\right\rangle \\
& =\left\langle x^{\prime}, x\right\rangle .
\end{aligned}
$$

Since $\operatorname{dom} A$ is dense in $X$, this equality implies that $R(\lambda, A)^{\prime}$ is also a left-inverse of $\lambda-A^{\prime}$. Hence, $\lambda \in \rho\left(A^{\prime}\right)$ and $R\left(\lambda, A^{\prime}\right)=R(\lambda, A)^{\prime}$.

Let $X$ be a Banach space and $X^{\prime}$ its dual. For every subset $M \subseteq X$ we define the annihilator

$$
M^{\perp}:=\left\{x^{\prime} \in X^{\prime}:\left\langle x^{\prime}, x\right\rangle=0 \forall x \in M\right\} .
$$

For every subset $M^{\prime} \subseteq X^{\prime}$, we define the preannihilator

$$
M_{\perp}^{\prime}:=\left\{x \in X:\left\langle x^{\prime}, x\right\rangle=0 \forall x^{\prime} \in M^{\prime}\right\} .
$$

It is easy to show that $M^{\perp}$ and $M_{\perp}^{\prime}$ are closed linear subspaces of $X^{\prime}$ and $X$, respectively.

Lemma 5.14. Let $X$ be a Banach space and let $(A, \operatorname{dom} A)$ be a closed, linear operator on $X$. Then:
a) $(\operatorname{ran} A)^{\perp}=\operatorname{ker} A^{\prime}$.
b) $\overline{\operatorname{ran} A}=\left(\operatorname{ker} A^{\prime}\right)_{\perp}$.
c) $(\operatorname{ker} A)^{\perp} \supseteq \overline{\operatorname{ran} A^{\prime}}$
d) $\operatorname{ker} A=\left(\operatorname{ran} A^{\prime}\right)_{\perp}$.

Proof. In order to prove (a), we observe

$$
\begin{aligned}
x^{\prime} \in(\operatorname{ran} A)^{\perp} & \Leftrightarrow \forall x \in X:\left\langle x^{\prime}, A x\right\rangle=0 \\
& \Leftrightarrow x^{\prime} \in \operatorname{dom} A^{\prime} \text { and } \forall x \in X:\left\langle A^{\prime} x^{\prime}, x\right\rangle=0 \\
& \Leftrightarrow x^{\prime} \in \operatorname{dom} A^{\prime} \text { and } A^{\prime} x^{\prime}=0 \\
& \Leftrightarrow x^{\prime} \in \operatorname{ker} A^{\prime} .
\end{aligned}
$$

(b) If $x \in \operatorname{ran} A, x=A y$ for some $y \in \operatorname{dom} A$, and if $x^{\prime} \in \operatorname{ker} A^{\prime}$, then

$$
\left\langle x^{\prime}, x\right\rangle=\left\langle x^{\prime}, A y\right\rangle=\left\langle A^{\prime} x^{\prime}, y\right\rangle=0 .
$$

Hence, $\operatorname{ran} A \subseteq\left(\operatorname{ker} A^{\prime}\right)_{\perp}$, and since the latter space is closed, we obtain $\overline{\operatorname{ran} T} \subseteq$ $\left(\operatorname{ker} T^{\prime}\right)_{\perp}$. Assume that the inclusion is strict. Then there exists $x_{0} \in\left(\operatorname{ker} A^{\prime}\right)_{\perp}$ which does not belong to $\overline{\operatorname{ran} A}$. By Hahn-Banach (Theorem 3.29 of Chapter 3), there exist $x^{\prime} \in X^{\prime}$ and $\varepsilon>0$ such that

$$
\begin{equation*}
\operatorname{Re}\left\langle x^{\prime}, x\right\rangle+\varepsilon \leq \operatorname{Re}\left\langle x^{\prime}, x_{0}\right\rangle, \quad x \in \overline{\operatorname{ran} A} . \tag{5.1}
\end{equation*}
$$

Since $\overline{\operatorname{ran} A}$ is a subspace of $X$, in particular $x \in \overline{\operatorname{ran} A}$ implies $\lambda x \in \overline{\operatorname{ran} A}$ for every $\lambda \in \mathbb{K}$, we deduce from this inequality that $\left\langle x^{\prime}, x\right\rangle=0$ for every $x \in \overline{\operatorname{ran} A}$. Hence, by
(a), $x^{\prime} \in \operatorname{ker} A^{\prime}$. But then $\left\langle x^{\prime}, x_{0}\right\rangle=0$, too, and this is a contradiction to (5.1). Hence, we have proved (b).
(c) If $x^{\prime} \in \operatorname{ran} A^{\prime}, x^{\prime}=A^{\prime} y^{\prime}$ for some $y^{\prime} \in \operatorname{dom} A^{\prime}$, and if $x \in \operatorname{ker} A$, then

$$
\left\langle x^{\prime}, x\right\rangle=\left\langle A^{\prime} y^{\prime}, x\right\rangle=\left\langle y^{\prime}, A x\right\rangle=0 .
$$

This implies $\operatorname{ran} A^{\prime} \subseteq(\operatorname{ker} A)^{\perp}$, and since the latter space is closed, we obtain (c).
(d) Similarly as in (a), we observe

$$
\begin{aligned}
x \in \operatorname{ker} A & \Leftrightarrow x \in \operatorname{dom} A \text { and } A x=0 \\
& \Leftrightarrow x \in \operatorname{dom} A \text { and } \forall x^{\prime} \in X^{\prime}:\left\langle x^{\prime}, A x\right\rangle=0 \\
& \Leftrightarrow x \in \operatorname{dom} A \text { and } \forall x^{\prime} \in \operatorname{dom} A^{\prime}:\left\langle A^{\prime} x^{\prime}, x\right\rangle=0 \\
& \Leftrightarrow x \in\left(\operatorname{ran} A^{\prime}\right)_{\perp} .
\end{aligned}
$$

Lemma 5.15. For every linear operator $(A, \operatorname{dom} A)$ on $X$ one has

$$
\sigma_{r}(A)=\sigma_{p}\left(A^{\prime}\right)
$$

Proof. Let $\lambda \in \sigma_{r}(A)$. Then, by definition of the residual spectrum, $\operatorname{ran}(\lambda-A)$ is not dense in $X$. By the Hahn-Banach theorem (see in particular Corollary 3.10), there exists a bounded, linear functional $x^{\prime} \in X^{\prime} \backslash\{0\}$ which vanishes on $\operatorname{ran}(\lambda-A)$, that is,

$$
\left\langle x^{\prime}, \lambda x-A x\right\rangle=0 \text { for every } x \in \operatorname{dom} A .
$$

In other words, $(\operatorname{ran}(\lambda-A))^{\perp} \neq\{0\}$. By Lemma 5.14 (a), this means $\operatorname{ker}\left(\lambda-A^{\prime}\right) \neq$ $\{0\}$, or, by definition of the point spectrum, $\lambda \in \sigma_{p}\left(A^{\prime}\right)$.

Conversely, if $\lambda \in \sigma_{p}\left(A^{\prime}\right)$, then $\operatorname{ker}\left(\lambda-A^{\prime}\right) \neq\{0\}$. This implies $\left(\operatorname{ker}\left(\lambda-A^{\prime}\right)\right)_{\perp} \neq$ $X$. By Lemma 5.14 (b), this means that $\operatorname{ran}(\lambda-A)$ is not dense in $X$. Hence, $\lambda \in$ $\sigma_{r}(A)$.

### 5.2 Compact operators

A linear operator $T: X \rightarrow Y$ between two Banach spaces $X$ and $Y$ is called a compact operator if $T B(0,1)$ is relatively compact in $Y$. The set of all compact linear operators from $X$ into $Y$ is denoted by $\mathscr{K}(X, Y)$. We denote $\mathscr{K}(X):=\mathscr{K}(X, X)$.

Remark 5.16. A linear operator $T: X \rightarrow Y$ is compact if and only if for every sequence $\left(x_{n}\right) \subseteq B(0,1)$ there exists a subsequence (again denoted by $\left(x_{n}\right)$ ) such that ( $T x_{n}$ ) is convergent (or Cauchy).

Since relatively compact subsets of normed spaces are necessarily bounded, every compact operator is bounded.

Lemma 5.17. Let $X, Y, Z$ be Banach spaces. Then:
a) The set $\mathscr{K}(X, Y)$ is a closed linear subspace of $\mathscr{L}(X, Y)$.
b) If $T \in \mathscr{K}(X, Y)$ and $S \in \mathscr{L}(Y, Z)$, then $S T \in \mathscr{K}(X, Z)$.
c) If $T \in \mathscr{L}(X, Y)$ and $S \in \mathscr{K}(Y, Z)$, then $S T \in \mathscr{K}(X, Z)$.
d) The set $\mathscr{K}(X)$ is a two-sided ideal in $\mathscr{L}(X)$.

Proof. (a) If $T, S \in \mathscr{K}(X, Y), \lambda \in \mathbb{K}$, then clearly $\lambda T \in \mathscr{K}(X, Y)$. Moreoever, if $\left(x_{n}\right) \subseteq B(0,1)$ is any sequence, then we can choose a subsequence (again denoted by $\left(x_{n}\right)$ ) such that $\left(T x_{n}\right)$ converges. From this subsequence, we extract another subsequence (again denoted by $\left(x_{n}\right)$ ) such that ( $S x_{n}$ ) converges. Then $\left(T x_{n}+S x_{n}\right)$ converges, and therefore $T+S \in \mathscr{K}(X, Y)$. Hence, $\mathscr{K}(X, Y)$ is a linear subspace of $\mathscr{L}(X, Y)$.

In order to see that $\mathscr{K}(X, Y)$ is closed in $\mathscr{L}(X, Y)$, let $\left(T_{n}\right) \subseteq \mathscr{K}(X, Y)$ be convergent to some element in $T \in \mathscr{L}(X, Y)$. Let $\left(x_{j}\right) \subseteq B(0,1)$ be any sequence. A diagonal sequence argument implies that we can choose a subsequence (again denoted by $\left.\left(x_{j}\right)\right)$ such that

$$
\lim _{j \rightarrow \infty} T_{n} x_{j} \text { exists for every } n \in \mathbb{N}
$$

Let $\varepsilon>0$ be arbitrary, and choose $n \in \mathbb{N}$ so large such that $\left\|T-T_{n}\right\|<\varepsilon$. Choose $j_{0} \in \mathbb{N}$ so large that $\left\|T_{n} x_{j}-T_{n} x_{k}\right\|<\varepsilon$ for every $j, k \geq j_{0}$. Then, for every $j, k \geq j_{0}$,

$$
\left\|T x_{j}-T x_{k}\right\| \leq\left\|T x_{j}-T_{n} x_{j}\right\|+\left\|T_{n} x_{j}-T_{n} x_{k}\right\|+\left\|T_{n} x_{k}-T x_{k}\right\|<3 \varepsilon .
$$

Hence, $\left(T x_{j}\right)$ is a Cauchy sequence. Since $Y$ is complete, $\left(T x_{j}\right)$ is convergent. As a consequence, for every sequence $\left(x_{j}\right) \subseteq B(0,1)$ we have extracted a subsequence (again denoted by $\left.\left(x_{j}\right)\right)$ such that $\left(T x_{j}\right)$ converges. This means that $T \in \mathscr{K}(X, Y)$. Hence, $\mathscr{K}(X, Y)$ is closed in $\mathscr{L}(X, Y)$.
(b), (c) Let $T \in \mathscr{L}(X, Y)$ and $S \in \mathscr{L}(Y, Z)$. If $T$ is compact, then $T B(0,1)$ is relatively compact, and since $S$ is continuous, $\operatorname{STB}(0,1)$ is relatively compact in $Z$ by Lemma 0.19 of chapter 0 . Hence, $S T \in \mathscr{K}(X, Z)$. If on the other hand $T$ is only bounded and $S$ is compact, then $T B(0,1)$ is bounded in $Y$, and therefore $\operatorname{STB}(0,1)$ is relatively compact in $Z$, i.e. $S T \in \mathscr{K}(X, Z)$.
(d) This is an immediate consequence of (b) and (c).

Lemma 5.18. Let $X, Y$ be Banach spaces. Then:
a) If $T \in \mathscr{L}(X, Y)$ has finite rank, that is, if $\operatorname{dim} \operatorname{ran} T<\infty$, then $T \in \mathscr{K}(X, Y)$.
b) If $\left(T_{n}\right) \subseteq \mathscr{K}(X, Y)$ is a uniformly convergent sequence of finite rank operators, then $T:=\lim _{n \rightarrow \infty} T_{n} \in \mathscr{K}(X, Y)$.

Proof. Assertion (a) follows from the Theorem of Heine-Borel, while (b) is a consequence of Lemma 5.17.

Example 5.19 (Rank-1-operator). For every $x^{\prime} \in X^{\prime}$ and $y \in Y$ we may define the operator $T: X \rightarrow Y$ by

$$
T x:=\left\langle x^{\prime}, x\right\rangle y \quad(x \in X) .
$$

Then $T$ has rank 1 (unless $x^{\prime}=0$ or $y=0$ in which case $T=0$ ), and it is therefore a compact operator. Operators of the form above are also denoted by $x^{\prime} \otimes y$. Every rank-1-operator is of this form.

Lemma 5.20. A Banach space $X$ is finite dimensional if and only if the identity operator $I \in \mathscr{L}(X)$ is compact.
Proof. This is an immediate consequence of Theorem 1.15 of Chapter 1 which itself was a consequence of the Lemma of Riesz (Lemma 1.14).

A difficult problem is in general to decide which operators are compact. By the very definition of compact operators, it is thus important to know which subsets of (infinite dimensional) Banach spaces are relatively compact. Boundedness of the subset alone does not suffice as the Lemma of Riesz shows (see also the preceeding lemma). In the case when the underlying Banach space is $C(K)$ ( $K$ a compact metric space) we have already seen a satisfactory characterization of relatively compact subsets; see the Theorem of Arzela-Ascoli (Theorem 1.36).
Example 5.21 (Sobolev embedding). Consider the embedding $J: W^{1, p}(a, b) \rightarrow$ $C([a, b])$ from Example 4.20 of Chapter 4. The closed graph theorem showed that $J$ is bounded, i.e. there exists $C \geq 0$ such that

$$
\|u\|_{\infty} \leq C\|u\|_{W^{1, p}}, \quad u \in W^{1, p}(a, b) .
$$

We can show in addition that the embedding is compact if $p>1$. Let

$$
M:=\left\{u \in W^{1, p}(a, b):\|u\|_{W^{1, p}}<1\right\}=J B(0,1) \subseteq C([a, b])
$$

be the image of the unit ball under $J$. By boundedness of $J, M$ is bounded in $C([a, b])$. Moreover, by Hölder's inequality (we assume $p>1$ ), for every $t, s \in[a, b](t \geq s)$ and every $u \in M$,

$$
|u(t)-u(s)|=\left|\int_{s}^{t} u^{\prime}(r) d r\right| \leq \int_{s}^{t}\left|u^{\prime}(r)\right| d r \leq\left\|u^{\prime}\right\|_{p}(t-s)^{\frac{p-1}{p}} \leq(t-s)^{\frac{p-1}{p}}
$$

This implies that $M$ is equicontinuous if $p>1$ (choose for every $\varepsilon>0$ the $\delta$ equal to $\varepsilon^{\frac{p}{p-1}}$ in order to check equicontinuity).

By the Arzela-Ascoli Theorem (Theorem 1.36), $M$ is relatively compact in $C([a, b])$, and therefore the embedding $W^{1, p}(a, b) \hookrightarrow C([a, b])$ is compact if $p>1$.
Exercise 5.22 (Sobolev embedding) Show that the embedding $W^{1,1}(a, b) \hookrightarrow$ $C([a, b])$ is not compact.

Exercise 5.23 (Multiplication operators in sequence spaces) Let $X=l^{p} \quad(1 \leq$ $p<\infty)$ or let $X=c_{0}$. Let $m \in l^{\infty}$ and define the associated multiplication operator $M \in \mathscr{L}(X)$ by

$$
M x=M\left(x_{n}\right)=\left(m_{n} x_{n}\right), \quad x \in X .
$$

Show that $M$ is compact if and only if $m \in c_{0}$.
Hint: Use Lemma 5.18.

Exercise 5.24 (Kernel operators) Let $\Omega \subseteq \mathbb{R}^{n}$ be a compact (!) set. Let $k \in C(\Omega \times$ $\Omega$ ), and define the associated kernel operator $K \in \mathscr{L}(C(\Omega))$ by

$$
K f(t)=\int_{\Omega} k(t, s) f(s) d s, \quad t \in \Omega, f \in C(\Omega)
$$

Then $K$ is compact.
Theorem 5.25 (Schauder). An operator $T \in \mathscr{L}(X, Y)$ is compact if and only if $T^{\prime} \in \mathscr{L}\left(Y^{\prime}, X^{\prime}\right)$ is compact.

Proof. Assume that $T \in \mathscr{K}(X, Y)$, and let $K:=\overline{T B_{X}(0,1)} \subseteq Y$. Then $K$ is compact. Let $M:=B_{Y^{\prime}}(0,1)$ be considered as a subset of $C(K)$. Then clearly $M$ is bounded, and it is not difficult to see that $M$ is also equicontinuous. By the theorem of ArzelaAscoli, $M$ is relatively compact in $C(K)$. This means that for every sequence $\left(y_{n}^{\prime}\right) \in$ $B_{Y^{\prime}}(0,1)$ there exists a convergent subsequence (convergent in $C(K)$ !). If we denote this subsequence again by $\left(y_{n}^{\prime}\right)$, then we obtain

$$
0=\lim _{n, m \rightarrow \infty}\left\|y_{n}^{\prime}-y_{m}^{\prime}\right\|_{C(K)} \geq \lim _{n, m \rightarrow \infty} \sup _{\|x\| \leq 1}\left|\left\langle y_{n}^{\prime}-y_{m}^{\prime}, T x\right\rangle\right|=\lim _{n, m \rightarrow \infty}\left\|T^{\prime} y_{n}^{\prime}-T^{\prime} y_{m}^{\prime}\right\|_{X^{\prime}}
$$

which just means that $T^{\prime}$ is compact.
Assume on the other hand that $T^{\prime} \in \mathscr{K}\left(Y^{\prime}, X^{\prime}\right)$. By what we have just proved, this implies $T^{\prime \prime} \in \mathscr{K}\left(X^{\prime \prime}, Y^{\prime \prime}\right)$. Hence, if $\left(x_{n}\right) \in B_{X}(0,1)$ is any sequence, then there exists a subsequence (again denoted by $\left(x_{n}\right)$ ) such that $\left(T^{\prime \prime} x_{n}\right)$ is convergent in $Y^{\prime \prime}$ (note that we have considered $\left(x_{n}\right)$ also as a sequence in $X^{\prime \prime}$ via the embedding $J$ ). However, $T^{\prime \prime} x_{n}=T x_{n}$, and the claim is proved.

Theorem 5.26 (Riesz-Schauder). Let $X$ be a Banach space, and $T \in \mathscr{K}(X)$. Then:
a) $\operatorname{ker}(I-T)$ is finite dimensional.
b) $\operatorname{ran}(I-T)$ is closed and $\operatorname{ran}(I-T)=\operatorname{ker}\left(I-T^{\prime}\right)_{\perp}$.
c) $\operatorname{ker}(I-T)=\{0\}$ if and only if $\operatorname{ran}(I-T)=X$.
d) $\operatorname{dim} \operatorname{ker}(I-T)=\operatorname{dim} \operatorname{ker}\left(I-T^{\prime}\right)=\operatorname{dim}(X / \operatorname{ran}(I-T))$.

An immediate consequence of the Riesz-Schauder Theorem is Fredholm's alternative.

Corollary 5.27 (Fredholm alternative). Let $X$ be a Banach space, and $T \in \mathscr{K}(X)$. Then, either for every $y \in Y$ the equation

$$
\begin{equation*}
x-T x=y, \tag{5.2}
\end{equation*}
$$

there exists a solution $x \in X$, and in this case the solution $x$ is unique, or the homogeneous equation

$$
x-T x=0
$$

has a a finite number of linearly independent solutions $\left(x_{i}\right)_{1 \leq i \leq n}$ and the equation (5.2) has a solution if and only if y satisfies $n$ equations of orthogonality $\left\langle x_{i}^{\prime}, y\right\rangle=0$, where the $x_{i}^{\prime} \in \operatorname{ker}\left(I-T^{\prime}\right)$ are linearly independent.

Remark 5.28. If $T \in \mathscr{K}(X)$, then, by property (c) of Theorem 5.26, $I-T$ is injective if and only if $I-T$ is surjective. In finite dimensions, this property of linear mappings is well-known. This property of operators of the form $I-T$ with $T$ compact is however not shared by arbitrary bounded operators on infinite-dimensional Banach spaces. For example, the left-shift $L$ on $l^{p}(\mathbb{N})$ defined by $L x=L\left(x_{n}\right):=\left(x_{n+1}\right)$ is surjective but not injective.

Remark 5.29. An operator $S \in \mathscr{L}(X, Y)$ such that $\operatorname{ker} S$ is finite dimensional and such that $\operatorname{ran} S$ is closed and has finite codimension (that is, $\operatorname{dim}(X / \operatorname{ran} S)<\infty$ ) is called a Fredholm operator, and

$$
\operatorname{ind} S:=\operatorname{dim} \operatorname{ker} S-\operatorname{dim}(X / \operatorname{ran} S)
$$

is called the Fredholm index of $S$. By Theorem 5.26, $S=I-T \in \mathscr{L}(X)$ is a Fredholm operator of Fredholm index 0 if $T \in \mathscr{K}(X)$.

Proof (of Theorem 5.26). (a) $\operatorname{On} \operatorname{ker}(I-T)$ we have $T=I$, and since $T$ is compact, $\operatorname{ker}(I-T)$ must be finite dimensional.
(b) Let $\left(x_{n}\right) \subseteq X$ be such that $u_{n}:=x_{n}-T x_{n} \rightarrow u \in X$. We have to show that $u \in \operatorname{ran}(I-T)$. Since $\operatorname{ker}(I-T)$ is finite dimensional, for every $n \in \mathbb{N}$ there exists $y_{n} \in \operatorname{ker}(I-T)$ such that

$$
\operatorname{dist}\left(x_{n}, \operatorname{ker}(I-T)\right)=\left\|x_{n}-y_{n}\right\| .
$$

We show that the sequence $\left(x_{n}-y_{n}\right)$ is bounded. Otherwise, after extracting a subsequence, we may assume that $\lim _{n \rightarrow \infty}\left\|x_{n}-y_{n}\right\|=\infty$. Putting $w_{n}:=\frac{x_{n}-y_{n}}{\left\|x_{n}-y_{n}\right\|}$, we find that $w_{n}-T w_{n}=u_{n} /\left\|x_{n}-y_{n}\right\| \rightarrow 0$. After extracting a subsequence, we may assume that $T w_{n} \rightarrow z$ ( $T$ is compact). But then $w_{n} \rightarrow z$, too, and therefore $z \in \operatorname{ker}(I-T)$. On the other hand,

$$
\operatorname{dist}\left(w_{n}, \operatorname{ker}(I-T)\right)=\frac{\operatorname{dist}\left(x_{n}, \operatorname{ker}(I-T)\right)}{\left\|x_{n}-y_{n}\right\|}=1,
$$

a contradiction. Hence, the sequence $\left(x_{n}-y_{n}\right)$ is bounded.
But then, by compactness of $T$, we can extract a subsequence (again denoted by $\left(x_{n}-y_{n}\right)$ ) such that $T\left(x_{n}-y_{n}\right) \rightarrow v$. Hence,

$$
x_{n}-y_{n}=u_{n}+T\left(x_{n}-y_{n}\right) \rightarrow u+v .
$$

We deduce that $T(u+v)=v$, or $u=(u+v)-T(u+v)$, so that $u \in \operatorname{ran}(I-T)$. Hence, $\operatorname{ran}(I-T)$ is closed.

Since the equality $\overline{\operatorname{ran}(I-T)}=\operatorname{ker}\left(I-T^{\prime}\right)_{\perp}$ always holds true (Lemma 5.14), we have thus proved (b).
(c) Assume first that $I-T$ is injective, i.e. $\operatorname{ker}(I-T)=\{0\}$. Assume that $X_{1}:=$ $\operatorname{ran}(I-T) \neq X$, that is, $X_{1}$ is a closed (by (b)) proper subspace of $X$. Then $\left.T\right|_{X_{1}} \in$ $\mathscr{K}\left(X_{1}\right)$, so that, by (b) again, $X_{2}=(I-T) X_{1}$ is a closed subspace of $X_{1}$. Since $I-T$ is injective, $X_{2} \neq X_{1}$. Iterating this argument and putting $X_{n}=(I-T)^{n} X$, we obtain a decreasing sequence $\left(X_{n}\right)$ of closed subspaces of $X$ such that $X_{n+1} \neq X_{n}$. By the Lemma of Riesz, for every $n \geq 1$ there exists $x_{n} \in X_{n}$ such that $\left\|x_{n}\right\|=1$ and $\operatorname{dist}\left(x_{n}, X_{n+1}\right) \geq \frac{1}{2}$. For every $n>m$ we have

$$
T x_{n}-T x_{m}=-\left(x_{n}-T x_{n}\right)+\left(x_{m}-T x_{m}\right)+x_{n}-x_{m}
$$

and

$$
-\left(x_{n}-T x_{n}\right)+\left(x_{m}-T x_{m}\right)+x_{n} \in X_{m+1} .
$$

Hence, $\left\|T x_{n}-T x_{m}\right\| \geq \frac{1}{2}$ whenever $n \neq m$, a contradiction to the assumption that $T$ is compact. Hence, $\operatorname{ran}(I-T)=X$.

Assume now on the other hand that $\operatorname{ran}(I-T)=X$. Then, by Lemma 5.14, $\operatorname{ker}\left(I-T^{\prime}\right)=\{0\}$. Since $T^{\prime}$ is compact by Schauder's theorem, this implies ran $(I-$ $\left.T^{\prime}\right)=X^{\prime}$ by the preceeding step. By Lemma $5.14, \operatorname{ker}(I-T)=\{0\}$.
(d) For every closed subspace $U$ of $X$ the dual $(X / U)^{\prime}$ is isomorphic to $U^{\perp}$. In particular, for $U=\operatorname{ran}(I-T)$ one obtains (using Lemma 5.14)

$$
\operatorname{ker}\left(I-T^{\prime}\right)=(\operatorname{ran}(I-T))^{\perp} \cong(X / \operatorname{ran}(I-T))^{\prime} \cong X / \operatorname{ran}(I-T)
$$

The last isomorphy holds since we know by the first isomorphy that $(X / \operatorname{ran}(I-T))^{\prime}$ is finite dimensional. In particular,

$$
\operatorname{dim} \operatorname{ker}\left(I-T^{\prime}\right)=\operatorname{dim} X / \operatorname{ran}(I-T)
$$

so that we have proved the second inequality.
It remains to prove that

$$
\operatorname{dim} X / \operatorname{ran}(I-T)=\operatorname{dim} \operatorname{ker}(I-T)
$$

Since $T x=x$ for every $x \in \operatorname{ker}(I-T)$, we see that $T$ leaves $\operatorname{ker}(I-T)$ invariant. In particular, the operator

$$
\begin{aligned}
\tilde{T}: X / \operatorname{ker}(I-T) & \rightarrow X / \operatorname{ker}(I-T), \\
x+\operatorname{ker}(I-T) & \mapsto T x+\operatorname{ker}(I-T),
\end{aligned}
$$

is well-defined and one easily checks that $\tilde{T}$ is compact since $T$ is compact. By construction, $\operatorname{ker}(I-\tilde{T})=\{0\}$ so that, by $(\mathrm{c}), \operatorname{ran}(I-\tilde{T})=X / \operatorname{ker}(I-T)$. This means that for every $y \in X$ there exists $x \in X$ and $x_{0} \in \operatorname{ker}(I-T)$ such that

$$
(I-T) x=y-x_{0}
$$

or

$$
y=(I-T) x+x_{0}=: x_{1}+x_{0} .
$$

In particular, every $y \in X$ can be written as a sum $x_{1}+x_{0}$ of an element $x_{1} \in \operatorname{ran}(I-$ $T)$ and an element $x_{0} \in \operatorname{ker}(I-T)$. Hence,

$$
\operatorname{dim} \operatorname{ker}\left(I-T^{\prime}\right)=\operatorname{dim} X / \operatorname{ran}(I-T) \leq \operatorname{dim} \operatorname{ker}(I-T)
$$

Replacing $T$ by $T^{\prime}$ (which is compact by Schauder's theorem), we obtain

$$
\operatorname{dimker}\left(I-T^{\prime \prime}\right) \leq \operatorname{dimker}\left(I-T^{\prime}\right) \leq \operatorname{dim} \operatorname{ker}(I-T)
$$

On the other hand, since $I-T^{\prime \prime}$ extends $I-T$, one trivially has

$$
\operatorname{dimker}(I-T) \leq \operatorname{dimker}\left(I-T^{\prime \prime}\right)
$$

The claim is proved
Theorem 5.30 (Spectrum of a compact operator). Let X be a Banach space and let $T \in \mathscr{K}(X)$. Then:
a) If $X$ is infinite-dimensional, then $0 \in \sigma(T)$.
b) $\sigma(T) \backslash\{0\}=\sigma_{p}(T) \backslash\{0\}$.
c) Either $\sigma(T)$ is finite or $\sigma(T) \backslash\{0\}=\left\{\lambda_{n}: n \in \mathbb{N}\right\}$ for some sequence $\left(\lambda_{n}\right) \subseteq \mathbb{C}$ such that $\lim _{n \rightarrow \infty} \lambda_{n}=0$.

Proof. (a) If $0 \in \rho(T)$, then $T^{-1}$ exists and is bounded. Hence, $I=T T^{-1}$ is compact; a contradiction to the assumption that $X$ is infinite dimensional.
(b) Let $\lambda \in \sigma(T) \backslash\{0\}$. If $\lambda \notin \sigma_{p}(T)$, then $\operatorname{ker}(\lambda-T)=\{0\}$. By the RieszSchauder Theorem (Theorem 5.26), this implies $\operatorname{ran}(\lambda-T)=X$ so that $\lambda-T$ is bijective; a contradiction to the assumption $\lambda \in \sigma(T)$.
(c) It suffices to prove that $\sigma(T) \cap\{\lambda \in \mathbb{C}:|\lambda| \geq R\}$ is finite for every $R>0$. If this was not the case, then we find a sequence $\left(\lambda_{n}\right) \subseteq \sigma(T) \backslash\{0\}$ such that $\lambda_{n} \neq \lambda_{m}$ for $n \neq m$ and $\left|\lambda_{n}\right| \geq R>0$. By (b), for every $n \in \mathbb{N}$ there exists $x_{n} \in X \backslash\{0\}$ such that $\lambda_{n} x_{n}-T x_{n}=0$. Note that the family $\left(x_{n}\right)$ are linearly independent. Otherwise, we find a smallest $n \in \mathbb{N}$ such that the family $\left(x_{i}\right)_{1 \leq i \leq n}$ is linearly independent, but $x_{n+1}=\sum_{i=1}^{n} \alpha_{i} x_{i}$ for some scalars $\alpha_{i}$. Then

$$
\sum_{i=1}^{n} \alpha_{i} \lambda_{n+1} x_{i}=\lambda_{n+1} x_{n+1}=T x_{n+1}=\sum_{i=1}^{n} \alpha_{i} \lambda_{i} x_{i}
$$

and this implies $\alpha_{i}\left(\lambda_{n+1}-\lambda_{i}\right)=0$ for every $1 \leq i \leq n$. Since $\lambda_{n+1} \neq \lambda_{i}$ for $1 \leq i \leq n$, we obtain $\alpha_{i}=0$; a contradiction to $x_{n+1} \neq 0$. Let $X_{n}:=\operatorname{span}\left\{x_{i}: 1 \leq i \leq n\right\}$. Then $\left(X_{n}\right)$ is an increasing sequence of closed subspaces of $X$ such that $X_{n} \neq X_{n+1}$ (the latter by linear independence of the vectors $x_{n}$ ). By the Lemma of Riesz, for every $n \geq 2$ there exists $y_{n} \in X_{n}$ such that $\left\|y_{n}\right\|=1$ and dist $\left(y_{n}, X_{n-1}\right) \geq \frac{1}{2}$. Then, for every $n>m \geq 2$,

$$
\begin{aligned}
\left\|T y_{n}-T y_{m}\right\| & =\left\|-\left(\lambda_{n} y_{n}-T y_{n}\right)+\left(\lambda_{m} y_{m}-T y_{m}\right)+\lambda_{n} y_{n}-\lambda_{m} y_{m}\right\| \\
& \geq \operatorname{dist}\left(\lambda_{n} y_{n}, X_{n-1}\right) \\
& \geq \frac{\lambda_{n}}{2} \geq \frac{R}{2} .
\end{aligned}
$$

This is a contradiction to the compactness of $T$, and hence (c) is proved.

### 5.3 Nuclear operators

Let $X$ and $Y$ be two Banach spaces. An operator $T: X \rightarrow Y$ is called nuclear operator, if there exist sequences $\left(x_{k}^{\prime}\right)$ in $X^{\prime}$ and $\left(y_{k}\right)$ in $Y$ such that
(i) $\sum_{k}\left\|x_{k}^{\prime}\right\|\left\|y_{k}\right\|<\infty$, and
(ii) $T x=\sum_{k}\left\langle x_{k}^{\prime}, x\right\rangle y_{k}$ for every $x \in X$.

Taking up the notation from Example 5.19 (Rank-1-operators), the condition (ii) is equivalent to

$$
T=\sum_{k} x_{k}^{\prime} \otimes y_{k},
$$

the series converging absolutely in $\mathscr{L}(X, Y)$, thanks to condition (i). Note that the representation of $T$ in the above form is not unique in the sense that the sequences $\left(x_{k}^{\prime}\right)$ and $\left(y_{k}\right)$ are not uniquely determined by $T$. We denote by $\mathscr{N}(X, Y)$ the space of all nuclear operators from $X$ into $Y ; \mathscr{N}(X):=\mathscr{N}(X, X)$. When being equipped with the norm

$$
\|T\|_{\mathscr{N}}:=\inf \left\{\sum_{k}\left\|x_{k}^{\prime}\right\|\left\|y_{k}\right\|: x_{k}^{\prime} \in X^{\prime}, y \in Y, T=\sum_{k} x_{k}^{\prime} \otimes y_{k}\right\},
$$

the space $\mathscr{N}(X, Y)$ becomes a Banach space (sic!).
Lemma 5.31. Every nuclear operator is compact, that is, in other words, $\mathscr{N}(X, Y) \subseteq \mathscr{K}(X, Y)$. Moreover, the embedding $\mathscr{N}(X, Y) \rightarrow \mathscr{K}(X, Y), T \mapsto T$, is continuous.
Proof. Let $T \in \mathscr{N}(X, Y)$. By definition, there exist sequences $\left(x_{k}^{\prime}\right)_{k}$ in $X^{\prime}$ and $\left(y_{k}\right)$ in $Y$ such that $\sum_{k}\left\|x_{k}^{\prime}\right\|\left\|y_{k}\right\|<\infty$, and

$$
T=\sum_{k} x_{k}^{\prime} \otimes y_{k}=\lim _{K \rightarrow \infty} \sum_{k \leq K} x_{k}^{\prime} \otimes y_{k},
$$

the limit being taken with respect to the norm in $\mathscr{L}(X, Y)$. In particular, $T$ is limit in $\mathscr{L}(X, Y)$ of the finite rank operators $\sum_{k \leq K} x_{k}^{\prime} \otimes y_{k}$, and hence $T$ is compact by Lemma 5.18.

Moreover,

$$
\|T\|=\left\|\sum_{k} x_{k}^{\prime} \otimes y_{k}\right\| \leq \sum_{k}\left\|x_{k}^{\prime} \otimes y_{k}\right\|=\sum_{k}\left\|x_{k}^{\prime}\right\|\left\|y_{k}\right\|
$$

for every representation of $T$. Taking the infimum over all representations of $T$, we obtain $\|T\| \leq\|T\|_{\mathscr{N}}$, so that the embedding $\mathscr{N}(X, Y) \subseteq \mathscr{K}(X, Y)$ is continuous.

Exercise 5.32 (Multiplication operators in sequence spaces) Let $X=l^{p} \quad(1 \leq$ $p<\infty)$ or let $X=c_{0}$. Let $m \in l^{\infty}$ and define the associated multiplication operator $M \in \mathscr{L}(X)$ as in Exercise 5.23:

$$
M x=M\left(x_{n}\right)=\left(m_{n} x_{n}\right), \quad x \in X
$$

Show that $M$ is nuclear if and only if $m \in \ell^{1}$.

## $5.4 *$ The mean ergodic theorem

A bounded, linear operator $T$ on a Banach space $X$ is called powerbounded if $\sup _{n \geq 0}\left\|T^{n}\right\|<\infty$. Clearly, the spectral radius of a powerbounded linear operator is less than or equal to 1 , which implies that its spectrum is contained in the closed unit disk $\overline{\mathbb{D}}:=\{\lambda \in \mathbb{C}:|\lambda| \leq 1\}$. Here, we are particularly interested in the asymptotic behaviour of orbits of powers of $T$, or, in other words, in the asymptotic behaviour of the discrete, linear dynamical system $\left(T^{n}\right)$.

Lemma 5.33. Let $T \in \mathscr{L}(X)$ be a powerbounded operator. Then:
a) For every $x \in \operatorname{ker}(I-T)$ and every $n \in \mathbb{N}$ one has $T^{n} x=x$.
b) For every $x \in \overline{\operatorname{ran}(I-T)}$ one has

$$
\lim _{N \rightarrow \infty} \frac{1}{N} \sum_{n=0}^{N-1} T^{n} x=0
$$

that is, the orbit $\left(T^{n} x\right)$ converges in the Cesaro mean to 0 .
c) $\operatorname{ker}(I-T) \cap \overline{\operatorname{ran}(I-T)}=\{0\}$.

Proof. (a) If $x \in \operatorname{ker}(I-T)$, then $T x=x$. An iteration gives $T^{n} x=x$ for every $n \in \mathbb{N}$. (b) First let $x \in \operatorname{ran}(I-T)$. Then $x=y-T y$ for some $y \in X$. Hence,

$$
\begin{aligned}
\frac{1}{N} \sum_{n=0}^{N-1} T^{n} x & =\frac{1}{N} \sum_{n=0}^{N-1} T^{n}(y-T y) \\
& =\frac{1}{N} \sum_{n=0}^{N-1}\left(T^{n} y-T^{n+1} y\right) \\
& =\frac{1}{N}\left(y-T^{N} y\right) \\
& \rightarrow 0 \text { as } N \rightarrow \infty
\end{aligned}
$$

due to the assumption that $T$ is powerbounded. The assumption that $T$ is powerbounded also implies that the Cesaro means $\frac{1}{N} \sum_{n=0}^{N-1} T^{n}$ are uniformly bounded. A
simple $3 \varepsilon$-argument implies that

$$
\lim _{N \rightarrow \infty} \frac{1}{N} \sum_{n=0}^{N-1} T^{n} x=0
$$

for every $x \in \overline{\operatorname{ran}(I-T)}$.
(c) If $x \in \operatorname{ker}(I-T) \cap \operatorname{ran}(I-T)$, then, by part (a),

$$
x=\frac{1}{N} \sum_{n=0}^{N-1} T^{n} x \text { for every } N \in \mathbb{N} .
$$

By part (b), the right-hand side of this equality converges to 0 as $N \rightarrow \infty$. Hence $x=0$.

Theorem 5.34 (Mean ergodic theorem). Let $T \in \mathscr{L}(X)$ be a powerbounded operator. Then, for every $x \in X$, the following assertions are equivalent:
(i) $x \in \operatorname{ker}(I-T) \oplus \overline{\operatorname{ran}(I-T)}$, that is, $x=x_{0}+x_{1}$ for some $x_{0} \in \operatorname{ker}(I-T)$ and some $x_{1} \in \overline{\operatorname{ran}(I-T)}$.
(ii) The limit $\lim _{N \rightarrow \infty} \frac{1}{N} \sum_{n=0}^{N-1} T^{n} x$ exists in $X$.
(iii) The limit $\lim _{N \rightarrow \infty} \frac{1}{N} \sum_{n=0}^{N-1} T^{n} x$ exists weakly in $X$.
(iv) The sequence $\left(\frac{1}{N} \sum_{n=0}^{N-1} T^{n} x\right)$ of Cesaro means has a weakly convergent subsequence.
If one of the equivalent conditions (i)-(iv) holds true, then

$$
\lim _{N \rightarrow \infty} \frac{1}{N} \sum_{n=0}^{N-1} T^{n} x=x_{0}
$$

We say that a sequence $\left(x_{n}\right)$ in a Banach space $X$ converges in Cesaro mean to some element $x \in X$ if

$$
\lim _{N \rightarrow \infty} \frac{1}{N} \sum_{n=0}^{N-1} x_{n}=x
$$

One can prove (exercise!) that if a sequence ( $x_{n}$ ) converges in the usual sense to some element $x \in X$, then it also converges in the Cesaro mean to the same element. However, the converse is not true: the sequence $\left((-1)^{n}\right)$ does obviously not converge in $\mathbb{R}$, but

$$
\lim _{N \rightarrow \infty} \frac{1}{N} \sum_{n=0}^{N-1}(-1)^{n}=\lim _{N \rightarrow \infty} \frac{1}{N} \frac{1}{2}\left(1+(-1)^{N+1}\right)=0
$$

that is, this sequence converges in the Cesaro mean to 0 . We also say that the Cesaro average of this sequence is 0 .

If one of the equivalent conditions (i)-(iv) in the Mean Ergodic Theorem above holds true, then the final conclusion is that the sequence $\left(T^{n} x\right)$ of iterates of $T$ ap-
plied to $x$ converges in Cesaro mean to $x_{0}$. Note that the sequence $\left(T^{n} x\right)$ need not converge in the usual sense.

Proof (of Theorem 5.34). The implication (i) $\Rightarrow$ (ii) follows from Lemma 5.33, while the implications (ii) $\Rightarrow$ (iii) and (iii) $\Rightarrow$ (iv) are trivial. So let us prove the remaining implication (iv) $\Rightarrow$ (i). Assume that the sequence $\left(\frac{1}{N} \sum_{n=0}^{N-1} T^{n} x\right)$ admits a weak accumulation point. Then there exists $x_{0} \in X$ and an increasing sequence $\left(N_{k}\right)$ in $\mathbb{N}$ such that

$$
w-\lim _{k \rightarrow \infty} \frac{1}{N_{k}} \sum_{n=0}^{N_{k}-1} T^{n} x=x_{0} .
$$

Since every bounded, linear operator is also weak-weak continuous, this implies

$$
\begin{aligned}
(I-T) x_{0} & =w-\lim _{k \rightarrow \infty} \frac{1}{N_{k}} \sum_{n=0}^{N_{k}-1} T^{n}(I-T) x \\
& =w-\lim _{k \rightarrow \infty} \frac{1}{N_{k}} \sum_{n=0}^{N_{k}-1}\left(T^{n} x-T^{n+1} x\right) \\
& =w-\lim _{k \rightarrow \infty} \frac{1}{N_{k}}\left(x-T^{N_{k}} x\right) \\
& =0
\end{aligned}
$$

so that $x_{0} \in \operatorname{ker}(I-T)$. On the other hand, for every $k$ one has

$$
\begin{aligned}
x-\frac{1}{N_{k}} \sum_{n=0}^{N_{k}-1} T^{n} x & =\frac{1}{N_{k}} \sum_{n=0}^{N_{k}-1}\left(x-T^{n} x\right) \\
& =\frac{1}{N_{k}} \sum_{0}^{N_{k}-1} \sum_{j=0}^{n-1} T^{j}(I-T) x \\
& =(I-T)\left[\frac{1}{N_{k}} \sum_{0}^{N_{k}-1} \sum_{j=0}^{n-1} T^{j} x\right] \in \operatorname{ran}(I-T) .
\end{aligned}
$$

Hence,

$$
\begin{aligned}
x-x_{0} & =x-\text { weak }-\lim _{k \rightarrow \infty} \frac{1}{N_{k}} \sum_{n=0}^{N_{k}-1} T^{n} x \\
& =\text { weak }-\lim _{k \rightarrow \infty}\left[x-\frac{1}{N_{k}} \sum_{n=0}^{N_{k}-1} T^{n} x\right] \\
& =: x_{1} \in \overline{\operatorname{ran}(I-T)}
\end{aligned}
$$

and we have proved that (i) holds.
Corollary 5.35 (Mean ergodic theorem in reflexive spaces). Let $T \in \mathscr{L}(X)$ be a powerbounded operator on a reflexive Banach space X. Then

$$
X=\operatorname{ker}(I-T) \oplus \overline{\operatorname{ran}(I-T)}
$$

and if $P \in \mathscr{L}(X)$ denotes the projection onto $\operatorname{ker}(I-T)$ along $\overline{\operatorname{ran}(I-T)}$, then, for every $x \in X$

$$
\lim _{N \rightarrow \infty} \frac{1}{N} \sum_{n=0}^{N-1} T^{n} x=P x
$$

that is, the iterates of $T$ converge strongly, and in the Cesaro mean, to the projection $P$. If 1 is not an eigenvalue of $T$, then, for every $x \in X$,

$$
\lim _{N \rightarrow \infty} \frac{1}{N} \sum_{n=0}^{N-1} T^{n} x=0
$$

Proof. If suffices to note that for every $x \in X$ the sequence $\left(\frac{1}{N} \sum_{n=0}^{N-1} T^{n} x\right)$ of Cesaro means is bounded in $X$. Since $X$ is assumed to be reflexive, this sequence thus admits a weakly convergent subsequence by Theorem 3.28. The claims thus follow from the Mean Ergodic Theorem (Theorem 5.34).

Since Hilbert spaces are in particular reflexive spaces, we immediately obtain the following corollary, due to von Neumann.

Corollary 5.36 (von Neumann mean ergodic theorem). Let $T$ be a contraction on a Hilbert space H. Then, for every $f \in H$, the Cesaro limit

$$
\lim _{N \rightarrow \infty} \frac{1}{N} \sum_{n=0}^{N-1} T^{n} f=: P f
$$

exists in $H, P$ being the projection onto $\operatorname{ker}(I-T)$ along $\overline{\operatorname{ran}(I-T)}$. If 1 is not an eigenvalue of $T$, then, for every $f \in H$,

$$
\lim _{N \rightarrow \infty} \frac{1}{N} \sum_{n=0}^{N-1} T^{n} f=0
$$

## Convergence in the Abel mean of powerbounded operators

Let $T \in \mathscr{L}(X)$ be a powerbounded operator, and let $M \geq 0$ be a constant such that $\left\|T^{n}\right\| \leq M$ for every $n \geq 0$. From the Neumann series (see also the short proof of Lemma 5.7 and the Remark 5.8), we obtain for every $\lambda \in \mathbb{K}$ with $|\lambda|>1$ the estimate

$$
\begin{aligned}
\|R(\lambda, T)\| & =\left\|\sum_{n \geq 0} \frac{T^{n}}{\lambda^{n+1}}\right\| \\
& \leq M \sum_{n \geq 0} \frac{1}{|\lambda|^{n+1}} \\
& =M \frac{1}{|\lambda|-1} .
\end{aligned}
$$

In particular,

$$
\begin{equation*}
\|(\lambda-1) R(\lambda, T)\| \leq M \text { for every real } \lambda>1 \tag{5.3}
\end{equation*}
$$

Lemma 5.37. Let $T \in \mathscr{L}(X)$ be a powerbounded operator. Then:
a) For every $x \in \operatorname{ker}(I-T)$ and every real $\lambda>1$ one has $(\lambda-1) R(\lambda, T) x=x$.
b) For every $x \in \overline{\operatorname{ran}(I-T)}$ one has $\lim _{\lambda \rightarrow 1+}(\lambda-1) R(\lambda, A) x=0$.
c) $\operatorname{ker}(I-T) \cap \overline{\operatorname{ran}(I-T)}=\{0\}$.

Proof. (a) Let $x \in \operatorname{ker}(I-T)$. Then

$$
0=x-T x=-(\lambda-1) x+(\lambda-T) x \text { for every real } \lambda>1
$$

Multiplying this equality with $R(\lambda, T)$ yields the claim.
(b) Assume first that $x \in \operatorname{ran}(I-T)$, that is, $x=y-T y$ for some $y \in X$. Then

$$
\begin{aligned}
\lim _{\lambda \rightarrow 1+}(\lambda-1) R(\lambda, T) x & =\lim _{\lambda \rightarrow 1+}(\lambda-1) R(\lambda, T)((1-\lambda) y+\lambda y-T y) \\
& =\lim _{\lambda \rightarrow 1+}\left[(\lambda-1)^{2} R(\lambda, T) y+(\lambda-1) y\right] \\
& =0 .
\end{aligned}
$$

The full claim follows from this equality, from the estimate (5.3), and from a simple density argument (compare with Lemma 2.48).
(c) Let $x \in \operatorname{ker}(I-T) \cap \overline{\operatorname{ran}(I-T)}$. Then the previous two points yield

$$
x=(\lambda-1) R(\lambda, T) x \text { for every real } \lambda>1
$$

and

$$
\lim _{\lambda \rightarrow 1+}(\lambda-1) R(\lambda, T) x=0
$$

which is only possible if $x=0$.
Theorem 5.38 (Mean ergodic theorem). Let $T \in \mathscr{L}(X)$ be a powerbounded operator. Then, for every $x \in X$, the following assertions are equivalent:
(i) $\frac{x \in \operatorname{ker}(I-T) \oplus \overline{\operatorname{ran}(I-T)}}{\operatorname{ran}(I-T)}$, that is, $x=x_{0}+x_{1}$ for some $x_{0} \in \operatorname{ker}(I-T)$ and some $x_{1} \in$ $\overline{\operatorname{ran}(I-T)}$.
(ii) The limit $\lim _{\lambda \rightarrow 1+}(\lambda-1) R(\lambda, A) x$ exists strongly (in $X$ ).
(iii) The limit $\lim _{\lambda \rightarrow 1+}(\lambda-1) R(\lambda, A) x$ exists weakly.
(iv) The net $((\lambda-1) R(\lambda, A) x)_{\lambda \searrow 1}$ admits a weakly convergent subsequence in the sense that there exists a sequence $\left(\lambda_{n}\right)$ in $\mathbb{R}, \lambda_{n} \rightarrow 1+$, such that $\left(\left(\lambda_{n}-1\right) R\left(\lambda_{n}, A\right) x\right)_{n}$ converges weakly.
(v) The limit $\lim _{N \rightarrow \infty} \frac{1}{N} \sum_{n=0}^{N-1} T^{n} x$ exists strongly.

If one of the equivalent conditions (i)-(v) holds true, then

$$
\lim _{\lambda \rightarrow 1+}(\lambda-1) R(\lambda, A) x=\lim _{n \rightarrow \infty} \frac{1}{n} \sum_{0}^{n-1} T^{k} x=x_{0} .
$$

We say that a sequence $\left(x_{n}\right)$ in a Banach space $X$ converges in Abel mean to some element $x \in X$ if the power series $\sum_{n=0}^{\infty} x_{n} \lambda^{n}$ converges (absolutely) for every $\lambda \in \mathbb{D}$, and if

$$
\lim _{\lambda \rightarrow 1-}(1-\lambda) \sum_{n=0}^{\infty} x_{n} \lambda^{n}=x
$$

One can prove that if a sequence $\left(x_{n}\right)$ converges in Cesaro mean to some element $x \in X$, then it also converges in the Abel mean to the same element. The converse, however, is not true. In general, we have thus the implications

$$
\begin{aligned}
& \left(x_{n}\right) \text { converges in the usual sense to } x \in X \\
& \Downarrow \\
& \left(x_{n}\right) \text { converges in the Cesaro mean to } x \in X \\
& \Downarrow \\
& \left(x_{n}\right) \text { converges in the Abel mean to } x \in X .
\end{aligned}
$$

The second Mean Ergodic Theorem (Theorem 5.38) says that the algebraic condition (i) is equivalent to convergence in the Abel mean of the sequence $\left(T^{n} x\right)$ of iterates of $T$ applied to $x$ (condition (iv)), which in turn is equivalent to convergence in the Cesaro mean (condition (v)). Hence, in this special situation, convergence in the Abel mean and in the Cesaro mean are equivalent.

Proof (of Theorem 5.38). The implication (i) $\Rightarrow$ (ii) follows from the preceding Lemma 5.33, assertions (a) and (b). The lemma also yields the equality $\lim _{\lambda \rightarrow 1+}(\lambda-1) R(\lambda, A) x=x_{0}$.
The implications (ii) $\Rightarrow$ (iii) and (iii) $\Rightarrow$ (iv) are trivial.
(iv) $\Rightarrow$ (i) We assume that there exists $x_{0} \in X$ and a sequence $\left(\lambda_{n}\right)$ in $\mathbb{R}, \lambda_{n} \rightarrow 1+$, such that weak- $\lim _{n}\left(\lambda_{n}-1\right) R\left(\lambda_{n}, A\right) x=x_{0}$. Then, for every $x^{\prime} \in X^{\prime}$,

$$
\begin{aligned}
\left\langle x^{\prime}, x_{0}\right\rangle & =\lim _{n}\left\langle x^{\prime},\left(\lambda_{n}-1\right) R\left(\lambda_{n}, T\right) x\right\rangle \\
& =\lim _{n}\left\langle x^{\prime},\left(1-\lambda_{n}+\lambda_{n}-T+T\right)\left(\lambda_{n}-1\right) R\left(\lambda_{n}, T\right) x\right\rangle \\
& =\lim _{n}\left\langle x^{\prime},-\left(\lambda_{n}-1\right)^{2} R\left(\lambda_{n}, T\right) x+\left(\lambda_{n}-1\right) x+T\left(\lambda_{n}-1\right) R\left(\lambda_{n}, T\right) x\right\rangle \\
& =\lim _{n}\left\langle x^{\prime}, T\left(\lambda_{n}-1\right) R\left(\lambda_{n}, T\right) x\right\rangle \\
& =\left\langle x^{\prime}, T x_{0}\right\rangle .
\end{aligned}
$$

Hence $x_{0}=T x_{0}$, or, in other words, $x_{0} \in \operatorname{ker}(I-T)$. It remains to show that $x_{1}:=x-x_{0} \in$ $\overline{\operatorname{ran}(I-T)}$. Note that for every $n$ one has

$$
\begin{aligned}
x-\left(\lambda_{n}-1\right) R\left(\lambda_{n}, T\right) x & =x-\left(\lambda_{n}-T+T-1\right) R\left(\lambda_{n}, T\right) x \\
& =(I-T) R\left(\lambda_{n}, T\right) x \in \operatorname{ran}(I-T) .
\end{aligned}
$$

Hence,

$$
\begin{aligned}
x_{1} & =x-x_{0} \\
& =x-\text { weak }-\lim _{n \rightarrow \infty}\left(\lambda_{n}-1\right) R\left(\lambda_{n}, T\right) x \\
& =\text { weak }-\lim _{n \rightarrow \infty}\left[x-\left(\lambda_{n}-1\right) R\left(\lambda_{n}, T\right) x\right] \in \overline{\operatorname{ran}(I-T)},
\end{aligned}
$$

which proves that (i) holds.
The equivalence (i) $\Leftrightarrow(\mathrm{v})$ follows from the Mean Ergodic Theorem 5.34.

## The mean ergodic theorem for general resolvents

The preceding situation can still be generalized. We now consider a general closed, linear operator $(A, \operatorname{dom} A)$ on a Banach space $X$, and we study the relation between the behaviour of the resolvent of $A$ near the boundary of the spectrum and some algebraic properties of $A$.

Lemma 5.39. Let $(A, \operatorname{dom} A)$ be a closed, linear operator on a Banach space $X$. Let $\lambda_{0} \in \mathbb{K}$ be such that there exists a sequence $\left(\lambda_{n}\right)$ in $\rho(A)$ satisfying $\lim _{n} \lambda_{n}=\lambda_{0}$ and $\left\|\left(\lambda_{n}-\lambda_{0}\right) R\left(\lambda_{n}, A\right)\right\| \leq M$ for every $n$ and some constant $M \geq 0$. Then:
a) For every $x \in \operatorname{ker}\left(\lambda_{0}-A\right)$ one has $\left(\lambda_{n}-\lambda_{0}\right) R\left(\lambda_{n}, A\right) x=x$ for every $n$.
b) For every $x \in \overline{\operatorname{ran}\left(\lambda_{0}-A\right)}$ one has $\lim _{n}\left(\lambda_{n}-\lambda_{0}\right) R\left(\lambda_{n}, A\right) x=0$.
c) $\operatorname{ker}\left(\lambda_{0}-A\right) \cap \overline{\operatorname{ran}\left(\lambda_{0}-A\right)}=\{0\}$.

Proof. (a) Let $x \in \operatorname{ker}\left(\lambda_{0}-A\right)$. Then $x \in \operatorname{dom} A$ and

$$
0=\left(\lambda_{0}-A\right) x=\left(\lambda_{0}-\lambda_{n}\right) x+\left(\lambda_{n}-A\right) x \text { for every } n
$$

Multiplying this equality with $R\left(\lambda_{n}, A\right)$ yields the claim.
(b) Assume first that $x \in \operatorname{ran}\left(\lambda_{0}-A\right)$, that is, $x=\left(\lambda_{0}-A\right) y$ for some $y \in \operatorname{dom} A$. Then

$$
\begin{aligned}
\lim _{n}\left(\lambda_{n}-\lambda_{0}\right) R\left(\lambda_{n}, A\right) x & =\lim _{n}\left(\lambda_{n}-\lambda_{0}\right) R\left(\lambda_{n}, A\right)\left(\lambda_{0}-\lambda_{n}+\lambda_{n}-A\right) y \\
& =\lim _{n}\left[\left(\lambda_{n}-\lambda_{0}\right)^{2} R\left(\lambda_{n}, A\right) y+\left(\lambda_{n}-\lambda_{0}\right) y\right] \\
& =0
\end{aligned}
$$

The full claim follows from this equality, from the assumption that the sequence $\left(\left(\lambda_{n}-\right.\right.$ $\left.\left.\lambda_{0}\right) R\left(\lambda_{n}, A\right)\right)_{n}$ is bounded in $\mathscr{L}(X)$, and from a simple density argument (compare with Lemma 2.48).
(c) Let $x \in \operatorname{ker}\left(\lambda_{0}-A\right) \cap \operatorname{ran}\left(\lambda_{0}-A\right)$. Then the previous two points give

$$
x=\left(\lambda_{n}-\lambda_{0}\right) R\left(\lambda_{n}, A\right) x \rightarrow 0 \text { as } n \rightarrow \infty,
$$

that is, $x=0$.
Theorem 5.40 (Mean ergodic theorem for resolvents). Let $(A, \operatorname{dom} A)$ be a closed, linear operator on a Banach space $X$. Let $\lambda_{0} \in \mathbb{K}$ be such that there exists a sequence $\left(\lambda_{n}\right)$ in $\rho(A)$ satisfying $\lim _{n} \lambda_{n}=\lambda_{0}$ and $\left\|\left(\lambda_{n}-\lambda_{0}\right) R\left(\lambda_{n}, A\right)\right\| \leq M$ for every $n$ and some constant $M \geq 0$. Then, for every $x \in X$, the following assertions are equivalent:
(i) $x \in \operatorname{ker}\left(\lambda_{n}-A\right) \oplus \overline{\operatorname{ran}\left(\lambda_{0}-A\right)}$, that is, $x=x_{0}+x_{1}$ for some $x_{0} \in \operatorname{ker}\left(\lambda_{0}-A\right)$ and some $x_{1} \in \overline{\operatorname{ran}\left(\lambda_{0}-A\right)}$.
(ii) The sequence $\left(\left(\lambda_{n}-\lambda_{0}\right) R\left(\lambda_{n}, A\right) x\right)_{n}$ converges strongly (in $X$ ).
(iii) The sequence $\left(\left(\lambda_{n}-\lambda_{0}\right) R\left(\lambda_{n}, A\right) x\right)_{n}$ converges weakly.
(iv) The sequence $\left(\left(\lambda_{n}-\lambda_{0}\right) R\left(\lambda_{n}, A\right) x\right)_{n}$ admits a weakly convergent subsequence.

If one of the equivalent conditions (i)-(iv) holds true, then

$$
\lim _{n}\left(\lambda_{n}-\lambda_{0}\right) R\left(\lambda_{n}, A\right) x=x_{0}
$$

Proof. The implication (i) $\Rightarrow$ (ii) follows from the preceding Lemma 5.39, assertions (a) and (b). It also yields the equality $\lim _{n}\left(\lambda_{n}-\lambda_{0}\right) R\left(\lambda_{n}, A\right) x=x_{0}$.
The implications (ii) $\Rightarrow$ (iii) and (iii) $\Rightarrow$ (iv) are trivial.
So let us prove the implication (iv) $\Rightarrow$ (i). We assume that $\left(\left(\lambda_{n}-\lambda_{0}\right) R\left(\lambda_{n}, A\right) x\right)_{n}$ admits a weakly convergent subsequence. After passing to a subsequence, if necessary, we may in fact assume that the sequence $\left(\left(\lambda_{n}-\lambda_{0}\right) R\left(\lambda_{n}, A\right) x\right)_{n}$ itself converges weakly, say, to some element $x_{0} \in X$. Then, for every $x^{\prime} \in X^{\prime}$,

$$
\begin{aligned}
\left\langle x^{\prime}, \lambda_{0} x_{0}\right\rangle & =\lim _{n}\left\langle x^{\prime}, \lambda_{0}\left(\lambda_{n}-\lambda_{0}\right) R\left(\lambda_{n}, A\right) x\right\rangle \\
& =\lim _{n}\left\langle x^{\prime},\left(\lambda_{0}-\lambda_{n}+\lambda_{n}-A+A\right)\left(\lambda_{n}-\lambda_{0}\right) R\left(\lambda_{n}, A\right) x\right\rangle \\
& =\lim _{n}\left\langle x^{\prime},\left(\lambda_{0}-\lambda_{n}\right)^{2} R\left(\lambda_{n}, A\right) x+\left(\lambda_{0}-\lambda_{n}\right) x+A\left(\lambda_{n}-\lambda_{0}\right) R\left(\lambda_{n}, A\right) x\right\rangle \\
& =\lim _{n}\left\langle x^{\prime}, A\left(\lambda_{n}-\lambda_{0}\right) R\left(\lambda_{n}, A\right) x\right\rangle \\
& =\left\langle x^{\prime}, A x_{0}\right\rangle
\end{aligned}
$$

Since $A$ is closed, this equality implies $x_{0} \in \operatorname{dom} A$ and $\lambda_{0} x=A x$. In other word, $x_{0} \in \operatorname{ker}\left(\lambda_{0}-A\right)$. It remains to show that $x_{1}:=x-x_{0} \in \overline{\operatorname{ran}\left(\lambda_{0}-A\right)}$. Note that for every $n$ one has $R\left(\lambda_{n}, A\right) x \in \operatorname{dom} A$ and

$$
\begin{aligned}
x-\left(\lambda_{n}-1\right) R\left(\lambda_{n}, A\right) x & =x-\left(\lambda_{n}-A+A-1\right) R\left(\lambda_{n}, A\right) x \\
& =(I-A) R\left(\lambda_{n}, A\right) x \in \operatorname{ran}(I-A) .
\end{aligned}
$$

Hence,

$$
\begin{aligned}
x_{1} & =x-x_{0} \\
& =x-\text { weak }-\lim _{n \rightarrow \infty}\left(\lambda_{n}-1\right) R\left(\lambda_{n}, A\right) x \\
& =\text { weak }-\lim _{n \rightarrow \infty}\left[x-\left(\lambda_{n}-1\right) R\left(\lambda_{n}, A\right) x\right] \in \overline{\operatorname{ran}(I-A)},
\end{aligned}
$$

which proves that (i) holds.
Corollary 5.41 (Mean ergodic theorem for resolvents in reflexive spaces). In addition to the assumption of the preceding Theorem 5.40, assume that the underlying Banach space $X$ is reflexive. Then $X=\operatorname{ker}\left(\lambda_{0}-A\right) \oplus \overline{\operatorname{ran}\left(\lambda_{0}-A\right)}$ and the for every $x \in X$ the limit

$$
\lim _{n}\left(\lambda_{n}-\lambda_{0}\right) R\left(\lambda_{n}, A\right) x=: x_{0}
$$

exists, and the limit $x_{0}$ coincides with the projection of $x$ onto $\operatorname{ker}\left(\lambda_{0}-A\right)$ along $\overline{\operatorname{ran}\left(\lambda_{0}-A\right)}$.
Proof. By assumption, for every $x \in X$, the sequence $\left(\left(\lambda_{n}-\lambda_{0}\right) R\left(\lambda_{n}, A\right) x\right)_{n}$ is bounded. Since $X$ is reflexive and by Theorem 3.28 , for every $x \in X$ the sequence $\left(\left(\lambda_{n}-\lambda_{0}\right) R\left(\lambda_{n}, A\right) x\right)_{n}$ admits a weakly convergent subsequence. The claim follows from Theorem 5.40.

## Chapter 6 <br> Banach algebras

### 6.1 Banach algebras and the theorem of Gelfand

A normed space $A$ is called a normed algebra if it is an algebra, and if

$$
\|a b\| \leq\|a\|\|b\| \text { for every } a, b \in A
$$

A complete, normed algebra is also called Banach algebra.
Examples 6.1. 1. Let $X$ be a normed space. Then the space $A=\mathscr{L}(X)$ of all bounded, linear operators on $X$ is a normed algebra for the usual multiplication which is the composition of operators (Lemma 1.26). It is a Banach algebra as soon as $X$ is a Banach space (Lemma 1.27).
2. Let $X$ be a Banach space. Then the space $A=\mathscr{K}(X)$ of all compact, linear operators on $X$ is a Banach algebra. Actually, $\mathscr{K}(X)$ is a closed, two-sided ideal in $\mathscr{L}(X)$.
3. Let $K$ be a compact space. Then $A=C(K)$ is a Banach algebra for the usual (pointwise) multiplication of functions. Similarly, if $\Omega$ is a locally compact space, then the space of continuous functions $\Omega \rightarrow \mathbb{K}$ vanishing at infinity, $C_{0}(\Omega)$, is a Banach algebra. Finally, if $M$ is an arbitrary topological space, then the space of continuous, bounded functions $M \rightarrow \mathbb{K}, C_{b}(M)$, is a Banach algebra. All spaces of continuous functions in this example are equipped with the supremum norm.
4. Let $\Omega$ be a measure space. Then $A=L^{\infty}(\Omega)$ is a Banach algebra for the usual (pointwise) multiplication.
5. Let $A=L^{1}\left(\mathbb{R}^{N}\right)$ be equipped with the convolution product

$$
f * g(x):=\int_{\mathbb{R}^{N}} f(x-y) g(y) d y \quad\left(f, g \in L^{1}\left(\mathbb{R}^{N}\right), x \in \mathbb{R}^{N}\right)
$$

Then $A$ is a Banach algebra.
Proof. Let $f, g \in L^{1}\left(\mathbb{R}^{N}\right)$. By Tonelli's theorem,

$$
\begin{aligned}
\int_{\mathbb{R}^{N}} \int_{\mathbb{R}^{N}}|f(x-y) g(y)| d y d x & =\int_{\mathbb{R}^{N}} \int_{\mathbb{R}^{N}}|f(x-y) g(y)| d x d y \\
& =\int_{\mathbb{R}^{N}}|f(x)| d x \int_{\mathbb{R}^{N}}|g(y)| d y \\
& =\|f\|_{L^{1}}\|g\|_{L^{1}}<\infty .
\end{aligned}
$$

This inequality first implies that $f * g(x)$ exists for almost every $x \in \mathbb{R}^{N}$, and second that

$$
\int_{\mathbb{R}^{N}}|f * g(x)| d x \leq\|f\|_{L^{1}}\|g\|_{L^{1}}<\infty
$$

that is, $f * g \in L^{1}\left(\mathbb{R}^{N}\right)$. In particular, the convolution product is well-defined. However, the above inequality also implies a particular case of Young's inequality

$$
\|f * g\|_{L^{1}} \leq\|f\|_{L^{1}}\|g\|_{L^{1}}
$$

which implies that $L^{1}\left(\mathbb{R}^{N}\right)$ equipped with the convolution product is a Banach algebra.
6. Let $A=L^{1}\left(\mathbb{R}_{+}\right)$be equipped with the convolution product

$$
f * g(t):=\int_{0}^{t} f(t-s) g(s) d s \quad\left(f, g \in L^{1}\left(\mathbb{R}_{+}\right), t \in \mathbb{R}_{+}\right)
$$

Then $A$ is a Banach algebra.
7. Let $A$ be a Banach algebra, and let $I \subseteq A$ be a closed, two-sided ideal. Then the factor space $A / I$ is a Banach algebra for the multiplication

$$
(a+I) \cdot(b+I)=a b+I \quad(a, b \in A)
$$

note that this product is well-defined since $I$ is a two-sided ideal.
A Banach algebra $A$ is unital if it admits a neutral element for the multiplication, usually denoted by 1 or by $e$.

Remark 6.2 (Adjunction of a unit). Let $A$ be a Banach algebra without unit. Consider the product space

$$
\bar{A}:=A \times \mathbb{C}
$$

equipped with the sum norm. Then $\bar{A}$ is a unital Banach algebra for the multiplication given by

$$
(a, \lambda)(b, \mu):=(a b+\mu a+\lambda b, \lambda \mu) \quad((a, \lambda),(b, \mu) \in \bar{A})
$$

The unit element is the element $(0,1)$.
Given a unital Banach algebra $A$, we say that an element $a \in A$ is invertible (respectively, left-invertible, right-invertible), if there exists an element $b \in A$ such that

$$
a b=b a=1 \quad(\text { respectively, } b a=1 \text { or } a b=1) .
$$

If $a$ is invertible, then the element $b \in A$ satisfying $a b=b a=1$ is uniquely determined. We write $b=: a^{-1}$, and we call $a^{-1}$ the inverse of $a$. We define the resolvent set of an element $a \in A$ by

$$
\rho(a):=\{\lambda \in \mathbb{K}: \lambda-a \text { is invertible }\},
$$

and the spectrum by

$$
\sigma(a):=\mathbb{K} \backslash \rho(a) .
$$

For every $\lambda \in \rho(a)$ we write $R(\lambda, a):=(\lambda-a)^{-1}$, and we call $R(\lambda, a)$ the resolvent of $a$ at $\lambda$. The function $R(\cdot, a)$ is simply called the resolvent of $a$.

Several of the lemmas on the structure of the resolvent set and the spectrum of a bounded, linear operator on a Banach space, which are stated in the preceding chapter, remain true in the general context of Banach algebras and elements in Banach algebras. We start with the resolvent identity.

Lemma 6.3 (Resolvent identity). Let A be a unital Banach algebra, and $a \in A$. Then, for every $\lambda, \mu \in \rho(a)$ one has

$$
R(\lambda, a)-R(\mu, a)=(\mu-\lambda) R(\mu, a) R(\lambda, a)
$$

Proof. For every $\lambda, \mu \in \rho(a)$

$$
\mu-\lambda=(\mu-a)-(\lambda-a) .
$$

Multiplying both sides by $R(\mu, a)$ and $R(\lambda, a)$, one obtains the claim.
Lemma 6.4 (Neumann series). Let A be a unital Banach algebra, and let $a \in A$ be such that $\|a\|<1$. Then $1-a$ is invertible, and

$$
(1-a)^{-1}=\sum_{n=0}^{\infty} a^{n},
$$

the series being absolutely convergent in $A$.
Lemma 6.5 (The resolvent is analytic). Let A be a unital Banach algebra. For every $a \in A$ the resolvent set $\rho(a)$ is open in $\mathbb{K}$ and the resolvent $\rho(a) \rightarrow A, \lambda \mapsto$ $R(\lambda, a)$ is analytic.

Proof. Let $\lambda \in \rho(a)$ and $\mu \in \mathbb{K}$. Then

$$
\mu-a=\mu-\lambda+\lambda-a=((\mu-\lambda) R(\lambda, a)+I)(\lambda-a),
$$

and the right-hand side is invertible if $|\mu-\lambda|<1 /\|R(\lambda, a)\|$ by the Neumann series. Hence, $\rho(a)$ is open in $\mathbb{K}$. The Neumann series (Lemma 6.4) precisely yields

$$
R(\mu, a)=\sum_{n=0}^{\infty}(-1)^{n} R(\lambda, a)^{n+1}(\mu-\lambda)^{n}
$$

that is, the resolvent $\lambda \mapsto R(\lambda, a)$ can be locally developped into a power series. In other words, the resolvent is analytic.

Lemma 6.6 (Growth of the resolvent near the spectrum). For every $\lambda \in \rho(a)$ one has

$$
\|R(\lambda, a)\| \geq \operatorname{dist}(\lambda, \sigma(a))^{-1}
$$

Proof. As we have seen in the proof of the preceding Lemma 6.5 , for $\lambda \in \rho(a)$ the condition

$$
|\mu-\lambda|\|R(\lambda, a)\|<1
$$

implies $\mu \in \rho(a)$. The claim follows.
Lemma 6.7. For every $a \in A$ one has

$$
\{\lambda \in \mathbb{K}:|\lambda|>\|a\|\} \subseteq \rho(a),
$$

and

$$
R(\lambda, a)=\sum_{n=0}^{\infty} \frac{a^{n}}{\lambda^{n+1}} \quad(|\lambda|>\|a\|) .
$$

Proof. Use the identity

$$
\lambda-a=\lambda\left(I-\frac{a}{\lambda}\right)
$$

and the Neumann series.
Remark 6.8. Similarly as in Remark 5.8, we can remark here that $\lambda \in \rho(a)$ as soon as

$$
|\lambda|>\liminf _{n \rightarrow \infty}\left\|a^{n}\right\|^{\frac{1}{n}}=: r(a) .
$$

As in the case of bounded, linear operators, the number $r(a) \geq 0$ is called the spectral radius of $a$.

Lemma 6.9. Let $A \neq\{0\}$ be a complex, unital Banach algebra. Then for every $a \in A$ the spectrum $\sigma(a)$ is nonempty and compact, and

$$
r(a)=\sup \{|\lambda|: \lambda \in \sigma(a)\} .
$$

Proof. The compactness of $\sigma(a)$ follows Lemma 6.5 and 6.7. If $\sigma(a)$ was empty, then, by Lemma 6.5, the resolvent $\lambda \mapsto R(\lambda, a)$ is an entire function. On the other hand, by Lemma 6.7,

$$
\lim _{|\lambda| \rightarrow \infty}\|R(\lambda, a)\|=0
$$

By Liouville's theorem, this implies $R(\lambda, a) \equiv 0$, which is only possible if $A=\{0\}$ is the trivial algebra.

Theorem 6.10 (Gelfand-Mazur). Let $A \neq\{0\}$ be a complex, unital Banach algebra such that every element $a \neq 0$ is invertible (that is, $A$ is a division algebra). Then $A=\mathbb{C}$.

Proof. Let $a \in A$. Then, by Lemma 6.9 , there exists $\lambda \in \mathbb{C}$ such that $\lambda-a$ is not invertible. By assumption, this implies $\lambda-a=0$, or, in other words, $a=\lambda$ is a scalar multiple of the unit element.

A (two-sided) ideal $I$ in a Banach algebra is called maximal ideal if $I \neq A$ and if there does not exist an other (two-sided) ideal $J$ in $A$ such that $I \subsetneq J \subsetneq A$.

Lemma 6.11. Every ideal in a unital Banach algebra is contained in a maximal ideal.

Proof. Let $I$ be an ideal in a unital Banach algebra $A$ with unit denoted by 1 . Define the set $\mathscr{M}:=\{J: J$ is an ideal in $A$ and $I \subseteq J \subsetneq A\}$, and equip it with the order relation $\leq$ given by inclusion: $J_{1} \leq J_{2} \Leftrightarrow J_{1} \subseteq J_{2}$. Let $\mathscr{J} \subseteq \mathscr{M}$ be a totally ordered subset and define $\bar{J}:=\bigcup_{J \in \mathscr{J}} J$. Then clearly $\bar{J}$ is an ideal in $A$ which contains $I$. On the other hand, $\bar{J} \neq A$, since all the ideals $J$ are strictly contained in $A$, and since therefore $1 \notin J$ for every $J \in \mathscr{J}$. Hence, $\bar{J} \in \mathscr{M}$. Clearly, $\bar{J}$ is a supremum for $\mathscr{J}$, and we have proved that every totally ordered set admits a supremum. By the Lemma of Zorn, $\mathscr{M}$ admits a maximal element which, by definition, must be a maximal ideal of $A$.

Lemma 6.12. Every maximal ideal in a Banach algebra is closed.
Proof. Let $A$ be a Banach algebra, and let $I$ be a maximal ideal. Assume first that $A$ is unital. By the Neumann series, the set $G(A)$ of all invertible elements in $A$ is open, and since $1 \in G(A)$, this set is also nonempty. Clearly, $I \cap G(A)=\emptyset$, for if $I$ contained an invertible element, then $1 \in I$, which is only possible if $I=A$. By the preceding two arguments, $I \subseteq \bar{I} \subseteq A \backslash G(A) \neq A$, and clearly, the closure of $I$ is also an ideal. Since $I$ is a maximal ideal, we obtain $I=\bar{I}$, that is, $I$ is closed.

Now if $A$ is not unital, then we consider the unital algebra $\bar{A}$ from Remark 6.2, which results from $A$ by adjunction of a unit element. Then $I$ is also an ideal in $\bar{A}$, which is contained, by Lemma 6.11, in a maximal ideal $J$. By the first part of this proof, $J$ is closed. As a consequence, $I=J \cap A$ is closed.

Let $A$ be a Banach algebra. A character is a nonzero algebra homomorphism $A \rightarrow \mathbb{K}$.

Lemma 6.13. Every character on a Banach algebra is automatically continuous.
Proof. Let $A$ be a Banach algebra, and let $\chi: A \rightarrow \mathbb{K}$ be a character. Assume first that $A$ is unital. Since $\chi$ is an algebra homomorphism, then $\operatorname{ker} \chi$ is an ideal. Consider the associated, commutative diagram

where $q_{\chi}, b_{\chi}$ and $i_{\chi}$ are the canonical surjection (quotient map) onto $A / \operatorname{ker} \chi$, the canonical bijection onto ran $\chi$ (here, $\mathbb{K}$ ), and the canonical injection from ran $\chi$ into $\mathbb{K}$ (here, the identity map). The kernel ker $\chi$ having codimension 1 ( $b_{\chi}$ being bijective), it must be a maximal ideal. By Lemma 6.12 , $\operatorname{ker} \chi$ is closed, and hence the canonical surjection $q_{\chi}$ is continuous onto the normed quotient space $A / \operatorname{ker} \chi$. Since the other two homomorphisms $b_{\chi}$ and $i_{\chi}$ are homomorphisms between finitedimensional (in fact: one-dimensional) normed spaces, they are continuous, too. Hence, $\chi$ is continuous.

If $A$ is not a unital Banach algebra, then we consider the unital algebra $\bar{A}$ from Remark 6.2 , which results from $A$ by adjunction of a unit element. We then define the linear functional

$$
\begin{aligned}
\bar{\chi}: \bar{A} & \rightarrow \mathbb{K}, \\
(a, \lambda) & \mapsto \chi(a)+\lambda
\end{aligned}
$$

For every $(a, \lambda),(b, \mu) \in \bar{A}$ we have, since $\chi$ is an algebra homomorphism,

$$
\begin{aligned}
\bar{\chi}((a, \lambda)(b, \mu)) & =\bar{\chi}(a b+\lambda b+\mu a, \lambda \mu) \\
& =\chi(a b+\lambda b+\mu a)+\lambda \mu \\
& =\chi(a) \chi(b)+\lambda \chi(b)+\mu \chi(a)+\lambda \mu \\
& =(\chi(a)+\lambda)(\chi(b)+\mu) \\
& =\bar{\chi}(a, \lambda) \bar{\chi}(b, \mu),
\end{aligned}
$$

so that $\bar{\chi}$ is a character (= algebra homomorphism). By the first part of the proof, $\bar{\chi}$ is continuous, which implies that $\chi$ is continuous, too.

Let $A$ be a Banach algebra, and let $A^{\prime}$ be its dual space. The set of all characters is denoted by $\sigma(A)$, and it is called the spectrum of the algebra $A$, or the Gelfand space of the algebra $A$. By the preceding lemma, the Gelfand space is a subset of $A^{\prime}$. The following lemma says that the Gelfand space is in fact a subset of the unit ball of $A^{\prime}$.

Lemma 6.14. Let A be a Banach algebra. Then, for every character $\chi \in \sigma(A)$ one has $\|\chi\|_{A^{\prime}} \leq 1$, with equality if $A$ is a unital Banach algebra and if $\|1\|=1$.

Proof. Let $\chi \in \sigma(A)$, and let $a \in A$ be such that $\|a\| \leq 1$. Then, for every $n \in \mathbb{N}$,

$$
\begin{aligned}
|\langle\chi, a\rangle|^{n} & =\left|\langle\chi, a\rangle^{n}\right| \\
& =\left|\left\langle\chi, a^{n}\right\rangle\right| \\
& \leq\|\chi\|\left\|a^{n}\right\| \\
& \leq\|\chi\|\|a\|^{n} \\
& \leq\|\chi\| .
\end{aligned}
$$

Since the right-hand side is finite, we necessarily obtain $|\langle\chi, a\rangle| \leq 1$, and hence $\|\chi\| \leq 1$.

If $A$ is unital, and if $\|1\|=1$, then $\|\chi\| \geq|\langle\chi, 1\rangle|=1$, which together with the preceding inequality implies $\|\chi\|=1$.

Remark 6.15. If $A$ is a unital Banach algebra, then one does not necessarily have $\|1\|=1$. However, there always exists an equivalent Banach algebra norm $\|\cdot\|_{1}$ for which one has $\|1\|_{1}=1$. Such a norm is for example given by

$$
\|a\|_{1}:=\sup _{\|b\| \leq 1}\|a b\| .
$$

By the preceding lemma, the Gelfand space of any Banach algebra $A$ is a subset of the closed unit ball in $A^{\prime}$. The closed unit ball in $A^{\prime}$, however, when being equipped with the topology which is induced by the weak* topology on $A^{\prime}$, is, by the Theorem of Banach-Alaoglu, a compact space. In the following, we shall always consider the Gelfand space as a topological space, equipped with the topology which is induced by the weak* topology on $A^{\prime}$, too. The following theorem, in combination with Lemma 6.11, shows in particular that the Gelfand space of a nontrivial, complex, commutative, unital Banach algebra is nonempty.

Theorem 6.16 (Gelfand-Mazur). In a complex, commutative Banach algebra, there is a one-to-one correspondence between the set of maximal ideals and the Gelfand space. In fact, every maximal ideal is the kernel of some unique character, and, conversely, the kernel of every character is a closed, maximal ideal.

In particular, the Gelfand space of a nontrivial, complex, commutative, unital Banach algebra is nonempty.

Proof. Let $I$ be a maximal ideal of a complex, commutative Banach algebra A. Assume first that $A$ is unital. Since $A$ is commutative, the left, right and two-sided ideals all coincide, and hence there is no other left or right ideal strictly included between $I$ and $A$. As a consequence, the quotient algebra $A / I$ is a complex, commutative, unital Banach algebra without any left or right ideals, except the trivial ones. However, if $B$ is a complex, commutative, unital Banach algebra with any left or right ideal, except the trivial ones, then for every $a \in B$, the ideal $a B$ generated by $a$ must be either equal to $\{0\}$ or $B$. Since $B$ is unital, $a B$ contains $a$, and thus $a B=B$ for every nonzero $B$. But then every nonzero element is invertible which implies, by the first theorem of Gelfand-Mazur (Theorem 6.10), that $B$ is isomorphic to $\mathbb{C}$. As a consequence, $A / I$ is isomorphic to $\mathbb{C}$. Now, the quotient map $\chi: A \rightarrow A / I=\mathbb{C}$ is a character, and $I=\operatorname{ker} \chi$.

If $A$ is not unital, then we consider the unital Banach algebra $\bar{A}$ from Remark 6.2, which we obtain from $A$ by adjunction of a unit. By Lemma 6.11, there exists a maximal ideal $J \subseteq \bar{A}$ such that $J \supseteq I \times\{0\}$. By the preceding argument, there exists a character $\bar{\chi}$ on $\bar{A}$ such that $\operatorname{ker} \bar{\chi}=J$. The restriction of $\bar{\chi}$ to $A \times\{0\}=A$ is a character on $A$ such that $\operatorname{ker} \chi=I$.

Conversely, if $\chi \in \sigma(A)$ is a character, then $\operatorname{ker} \chi$ is an ideal of codimension 1 , hence a maximal ideal. Moreover, since $\chi$ is continuous by Lemma 6.13 , $\operatorname{ker} \chi$ is closed.

The existence of a character in a complex, commutative, unital Banach algebra now follows from this first part of the proof and the fact that there exists a maximal ideal (Lemma 6.11) and that every maximal ideal is closed (Lemma 6.12). Hence, $\sigma(A)$ is nonempty.

Lemma 6.17. Let A be a Banach algebra. The set $\sigma(A) \cup\{0\}$ is a closed subset of the closed unit ball $\bar{B}_{A^{\prime}}(0,1)$. If $A$ is a unital Banach algebra, then the Gelfand space $\sigma(A)$ itself is a closed subset of $\bar{B}_{A^{\prime}}(0,1)$. In particular, if $A$ is a unital Banach algebra, then the Gelfand space $\sigma(A)$ is a compact, nonempty space. In general, the Gelfand space is a locally compact space (which may, however, be empty).

Proof. Let $\left(\chi_{\alpha}\right)_{\alpha}$ be a net in $\sigma(A) \cup\{0\}$, which converges in $\bar{B}_{A^{\prime}}(0,1)$ to some element $a^{\prime}$. Then, for every $a, b \in A$

$$
\begin{aligned}
\left\langle a^{\prime}, a b\right\rangle & =\lim _{\alpha}\left\langle\chi_{\alpha}, a b\right\rangle \\
& =\lim _{\alpha}\left\langle\chi_{\alpha}, a\right\rangle\left\langle\chi_{\alpha}, b\right\rangle \\
& =\lim _{\alpha}\left\langle a^{\prime}, a\right\rangle\left\langle a^{\prime}, b\right\rangle .
\end{aligned}
$$

In other words, $a^{\prime}$ is a multiplicative functional, which means either $a^{\prime} \in \sigma(A)$, or $a^{\prime}=0$. As a consequence, $\sigma(A) \cup\{0\}$ is closed in $\bar{B}_{A^{\prime}}(0,1)$.

If, in addition, $A$ is a unital Banach algebra, and if $\left(\chi_{\alpha}\right)_{\alpha}$ is a net in $\sigma(A)$ which converges to some $a^{\prime} \in \bar{B}_{A^{\prime}}(0,1)$, then, by the preceding argument, $a^{\prime} \in \sigma(A)$, or $a^{\prime}=0$. However,

$$
\left\langle a^{\prime}, 1\right\rangle=\lim _{\alpha}\left\langle\chi_{\alpha}, 1\right\rangle=1,
$$

which actually excludes the possibility $a^{\prime}=0$. Hence, $\sigma(A)$ is closed in $\bar{B}_{A^{\prime}}(0,1)$.
Example 6.18 (Gelfand space of $C(K)$ ). We consider the Banach algebra $C(K)$, where $K$ is a compact Hausdorff space. We claim that

$$
\sigma(C(K)) \text { is homeomorphic to } K
$$

or, with a slight abuse of language, the Gelfand space of $C(K)$ is equal to $K$. In fact, for every $x \in K$ the Dirac functional $\delta_{x}: C(K) \rightarrow \mathbb{K}, f \mapsto f(x)$ is a character, so that $K$ can be naturally identified with a subset of $\sigma(C(K))$. On the other hand, every character in $\sigma(C(K))$ must be a Dirac functional. In fact, let us argue from the point of view of maximal ideals. If $I$ is a maximal ideal, then there must be some $x \in K$ such that $f(x)=0$ for every $f \in I$. In fact, if this was not true, then there exists $f \in I$ which never vanishes on $K$ (sic!). By continuity of $f$ and compactness of $K$, $|f|$ is uniformly bounded away from 0 , and $f^{-1}$ exists in $C(K)$. Since $I$ is an ideal, we obtain $1=f f^{-1} \in I$, and therefore $I=C(K)$, a contradiction to the assumption that $I$ is a maximal ideal. On the other hand, again since $I$ is a maximal ideal, there exists exactly one $x \in K$ such that $f(x)=0$ for every $f \in I$. Hence, $I=\operatorname{ker} \delta_{x}$ for the corresponding Dirac functional. Thus, the mapping $K \mapsto \sigma(C(K)), x \mapsto \delta_{x}$ is a bijection, which is clearly also continuous thanks to continuity of elements in $C(K)$
and the definition of the weak* topology. By compactness of $K$ and $\sigma(C(K))$, this bijection is a homeomorphism.

Examples 6.19 (Gelfand space of $L^{1}\left(\mathbb{R}^{N}\right)$ or $L^{1}\left(\mathbb{R}_{+}\right)$).

1. We consider the Banach algebra $L^{1}\left(\mathbb{R}^{N}\right)$, equipped with the convolution product $*$, as in Example 6.1.5. The dual space of $L^{1}\left(\mathbb{R}^{N}\right)$ can be identified with $L^{\infty}\left(\mathbb{R}^{N}\right)$, the duality being given by

$$
\langle f, g\rangle_{L^{\infty}, L^{1}}:=\int_{\mathbb{R}^{N}} f g .
$$

Let $\chi \in L^{\infty}\left(\mathbb{R}^{N}\right)$ be a character. Then, by Lemma $6.14,\|\chi\|_{\infty} \leq 1$, and by definition of character, for every $f, g \in L^{1}\left(\mathbb{R}^{N}\right)$,

$$
\begin{aligned}
\int_{\mathbb{R}^{N}} \chi(x) f(x) d x \int_{\mathbb{R}^{N}} \chi(y) g(y) d y & =\langle\chi, f\rangle_{L^{\infty}, L^{1}}\langle\chi, g\rangle_{L^{\infty}, L^{1}} \\
& =\langle\chi, f * g\rangle_{L^{\infty}, L^{1}} \\
& =\int_{\mathbb{R}^{N}} \chi(x) \int_{\mathbb{R}^{N}} f(x-y) g(y) d y d x \\
& =\int_{\mathbb{R}^{N}} \int_{\mathbb{R}^{N}} \chi(x+y) f(x) g(y) d y d x .
\end{aligned}
$$

It is not difficult to deduce from this equality, that every character $\chi$ satisfies the functional equation

$$
\chi(x+y)=\chi(x) \chi(y) \text { for almost every } x, y \in \mathbb{R}^{N} .
$$

Since $\chi$ is measurable, bounded and nonzero, this functional equation implies that there exists $\xi \in \mathbb{R}^{N}$ such that

$$
\chi(x)=e^{i \xi x} \text { for every } x \in \mathbb{R}^{N}
$$

Thus, the Gelfand space of $L^{1}\left(\mathbb{R}^{N}\right)$ is given by

$$
\boldsymbol{\sigma}\left(L^{1}\left(\mathbb{R}^{N}\right)\right)=\left\{e^{i \xi \cdot}: \xi \in \mathbb{R}^{N}\right\}
$$

One can show that this space, equipped with the weak* topology, is homeomorphic to the space $\mathbb{R}^{N}$, equipped with the usual Euclidean topology.
2. Now we consider the Banach algebra $L^{1}\left(\mathbb{R}_{+}\right)$, equipped with the convolution product, as in Example 6.1.6. As in the previous example, one shows that every character $\chi \in L^{\infty}\left(\mathbb{R}_{+}\right)$satisfies the functional equation

$$
\chi(t+s)=\chi(t) \chi(s) \text { for almost every } t, s \in \mathbb{R}_{+} .
$$

This implies that there exists $\lambda \in \mathbb{C}$ with $\operatorname{Re} \lambda \geq 0$ such that

$$
\chi(t)=e^{-\lambda t} \text { for every } t \in \mathbb{R}_{+} .
$$

Hence,

$$
\sigma\left(L^{1}\left(\mathbb{R}_{+}\right)\right)=\left\{e^{-\lambda \cdot}: \lambda \in \mathbb{C}, \operatorname{Re} \lambda \geq 0\right\} .
$$

One can show that this space is homeomorphic to the closed right half-plane $\{\lambda \in \mathbb{C}: \operatorname{Re} \lambda \geq 0\}$.

Let $A$ be a Banach algebra with Gelfand space $\sigma(A)$, and let $a \in A$. Then we define the function

$$
\begin{aligned}
\hat{a}: \sigma(A) \cup\{0\} & \rightarrow \mathbb{K}, \\
\chi & \mapsto \hat{a}(\chi):=\langle\chi, a\rangle,
\end{aligned}
$$

and we note that this function is continuous and vanishing at infinity. In fact, if $\left(\chi_{\alpha}\right)_{\alpha}$ is a convergent net in $\sigma(A) \cup\{0\}, \lim _{\alpha} \chi_{\alpha}=: \chi$, then, by definition of the weak* topology,

$$
\lim _{\alpha} \hat{a}\left(\chi_{\alpha}\right)=\lim _{\alpha}\left\langle\chi_{\alpha}, a\right\rangle=\langle\chi, a\rangle=\hat{a}(\chi) .
$$

As a consequence, $\hat{a} \in C(\sigma(A) \cup\{0\})$. In the following, we consider the function $\hat{a}$ only to be defined on the Gelfand space itself. If $A$ is a unital Banach algebra, then $\sigma(A)$ is already compact by the preceding lemma, and $\hat{a} \in C(\sigma(A))$. If $A$ is a non-unital Banach algebra, then the Gelfand space $\sigma(A)$ is only locally compact, and $\hat{a} \in C_{0}(\sigma(A))$, the space of continuous functions vanishing at infinity. Since $C(K)=C_{0}(K)$ for every compact space $K$, we may always write $\hat{a} \in C_{0}(\sigma(A))$.

Theorem 6.20 (Gelfand). Let A be a complex, commutative Banach algebra, and let $\sigma(A)$ be its Gelfand space (considered as a locally compact space for the weak* topology). Then the Gelfand transform

$$
\begin{aligned}
\wedge & : A \rightarrow C_{0}(\sigma(A)), \\
& a \mapsto \hat{a},
\end{aligned}
$$

where $\hat{a}(\chi):=\langle\chi, a\rangle(\chi \in \sigma(A))$, is a bounded Banach algebra homomorphism.
Proof. We have already shown above that the Gelfand transform is well-defined. By Lemma 6.14,

$$
\begin{aligned}
\|\hat{a}\|_{C_{0}(\sigma(A))} & =\sup _{\chi \in \sigma(A)}|\hat{a}(\chi)| \\
& =\sup _{\chi \in \sigma(A)}|\langle\chi, a\rangle| \\
& \leq \sup _{\chi \in \sigma(A)}\|\chi\|\|a\| \\
& \leq\|a\|,
\end{aligned}
$$

so that ${ }^{\wedge}$ is actually a contraction. It is clear that ${ }^{\wedge}$ is linear. Moreover, for every $a$, $b \in A$ and every $\chi \in \sigma(A)$ one has

$$
\begin{aligned}
\widehat{a b}(\chi) & =\langle\chi, a b\rangle \\
& =\langle\chi, a\rangle\langle\chi, b\rangle \\
& =\hat{a}(\chi) \hat{b}(\chi)
\end{aligned}
$$

that is, $\widehat{a b}=\hat{a} \hat{b}$. We have proved that ${ }^{\wedge}$ is an algebra homomorphism.
Theorem 6.21. Let A be a complex, commutative, unital Banach algebra. Then, for every $a \in A$,

$$
\sigma(a)=\{\langle\chi, a\rangle: \chi \in \sigma(A)\} .
$$

Proof. " $\subseteq$ " Let $\lambda \in \sigma(a)$. Then $\lambda-a$ is not invertible, which means that $\lambda-a$ is contained in some maximal ideal. Hence, there exists a character $\chi \in \sigma(A)$ such that $\langle\chi, \lambda-a\rangle=0$. However, $\langle\chi, \lambda\rangle=\lambda\langle\chi, 1\rangle=\lambda$, and hence $\lambda \in\{\langle\chi, a\rangle: \chi \in \sigma(A)\}$. " $\supseteq$ " Now assume that $\lambda \in\{\langle\chi, a\rangle: \chi \in \sigma(A)\}$. Then there exists $\chi \in \sigma(A)$ such that $0=\lambda-\langle\chi, a\rangle=\langle\chi, \lambda-a\rangle$. In other words, $\lambda-a$ is contained in the kernel of some character $\chi$, or, equivalently, in some maximal ideal. As a consequence, $\lambda-a$ is not invertible, that is, $\lambda \in \sigma(a)$.

## 6.2 $C^{*}$-algebras and the theorem of Gelfand-Naimark

An involution on a (complex) Banach algebra $A$ is a mapping $*: A \rightarrow A, a \mapsto a^{*}$ such that
a) $(a+b)^{*}=a^{*}+b^{*}$ and $(\lambda a)^{*}=\bar{\lambda} a^{*}$ for every $a, b \in A, \lambda \in \mathbb{C}$,
b) $(a b)^{*}=b^{*} a^{*}$ for every $a, b \in A$,
c) $a^{* *}=a$ for every $a \in A$.

If $A$ is unital, and if $e$ is the unit element, then automatically

$$
e^{*}=e .
$$

Indeed, for every $a \in A$, by properties (2) and (3),

$$
\begin{aligned}
e^{*} a & =e^{*} a^{* *} \\
& =\left(a^{*} e\right)^{*} \\
& =a^{* *} \\
& =a,
\end{aligned}
$$

and similarly, $a e^{*}=a$. Hence, $e^{*}$ is a unit element, too. By uniqueness of the unit element, $e=e^{*}$.

A Banach algebra with involution is called a Banach $*$-algebra. A Banach $*-$ algebra such that

$$
\begin{equation*}
\left\|a a^{*}\right\|=\|a\|^{2} \text { for every } a \in A \tag{6.1}
\end{equation*}
$$

is called a $C^{*}$-algebra. An example of a (commutative) $C^{*}$-algebra is the algebra $C(K)$ ( $K$ a compact space) with the natural involution $f \mapsto \bar{f}$ (pointwise complex conjugation). The following theorem says that - up to isomorphism - this is the only example of a commutative, unital $C^{*}$-algebra.

Theorem 6.22 (Gelfand-Naimark). Let $A$ be a unital, commutative $C^{*}$-algebra. Then the Gelfand transform ${ }^{\wedge}: A \rightarrow C(\sigma(A))$ is an isometric $*$-isomorphism, that is, it is an isometric isomorphism and

$$
\overline{\hat{a}}=\hat{a^{*}} \text { for every } a \in A .
$$

Proof. Let $\chi \in \sigma(A)$ be a character. Let $a \in A$ be a selfadjoint element, that is, $a=a^{*}$, and write $\chi(a)=\alpha+i \beta$ with $\alpha, \beta \in \mathbb{R}$. Let $e \in A$ be the unit element and define, for every $t \in \mathbb{R}$

$$
b:=a+i t e .
$$

Then

$$
\chi(b)=\chi(a)+i t \chi(e)=\alpha+i(\beta+t)
$$

and

$$
b b^{*}=(a+\text { ite } e)(a+\text { ite })^{*}=(a+\text { ite } e)(a-\text { ite })=a^{2}+t^{2} e .
$$

As a consequence,

$$
\begin{aligned}
\alpha^{2}+(\beta+t)^{2} & =|\chi(b)|^{2} \\
& \leq\|b\|^{2} \\
& =\left\|b b^{*}\right\| \\
& \leq\|a\|^{2}+t^{2},
\end{aligned}
$$

and therefore

$$
\alpha^{2}+\beta^{2}+2 \beta t \leq\|a\|^{2}
$$

This is only possible if $\beta=0$. Hence, we have proved $\chi(a) \in \mathbb{R}$.
If $a \in A$ is an arbitrary element, then $a=u+i v$ with $u, v \in A$ such that $u=u^{*}$ and $v=v^{*}$ (in fact, take $u=\frac{a+a^{*}}{2}$ and $v=\frac{a-a^{*}}{2 i}$ ). Then $a^{*}=u-i v$ and therefore, using the first step,

$$
\begin{aligned}
\chi\left(a^{*}\right) & =\chi(u)-i \chi(v) \\
& =\overline{\chi(u)+i \chi(v)} \\
& =\overline{\chi(a)} .
\end{aligned}
$$

By the definition of the Gelfand transform, this is equivalent to saying that $\overline{\hat{a}}=\hat{a^{*}}$, that is, ${ }^{\wedge}$ is a $*$-homomorphism.

Next, we prove that the Gelfand transform is isometric. For every $a \in A$ we have

$$
\begin{array}{rlr}
\|\hat{a}\|_{\infty} & =\sup _{\chi \in \sigma(A)}|\chi(a)| & \\
& =\sup _{\lambda \in \sigma(a)}|\lambda| & \quad \text { (by Theorem 6.21) } \\
& =r(a) & \\
& =\lim _{n \rightarrow \infty}\left\|a^{n}\right\|^{\frac{1}{n}} . &
\end{array}
$$

Note that, by properties (2) and (3), $\left(a a^{*}\right)^{*}=a^{* *} a^{*}=a a^{*}$, and therefore, by property (6.1),

$$
\left\|a a^{*}\right\|^{2}=\left\|\left(a a^{*}\right)\left(a a^{*}\right)^{*}\right\|=\left\|\left(a a^{*}\right)^{2}\right\|
$$

By induction, this inequality implies

$$
\left\|a a^{*}\right\|^{2^{m}}=\left\|\left(a a^{*}\right)^{2^{m}}\right\| \text { for every } m \in \mathbb{N}
$$

This equality, again property (6.1) and the commutativity imply

$$
\left\|a^{2^{m}}\right\|^{2}=\left\|a^{2^{m}}\left(a^{*}\right)^{2^{m}}\right\|=\left\|\left(a a^{*}\right)^{2^{m}}\right\|=\left\|a a^{*}\right\|^{2^{m}}=\|a\|^{2 \cdot 2^{m}}
$$

and thus

$$
\|\hat{a}\|_{\infty}=\|a\|,
$$

which yields that the Gelfand transform is isometric.
Like any isometric, linear operator, the Gelfand transform is injective and the range $\hat{A}$ is a closed subalgebra of $C(\sigma(A))$. Since, by the first step, the Gelfand transform is a $*$-homomorphism, the algebra $\hat{A}$ is closed under taking complex conjugation. Clearly, $1=\hat{e} \in \hat{A}$. Also, $\hat{A}$ separates the points of $\sigma(A)$. Thus, by the Stone-Weierstraß theorem, $\hat{A}=C(\sigma(A))$, that is, the Gelfand transform is surjective.

Corollary 6.23. Let A be a unital, commutative $C^{*}$-algebra generated by a single element $a \in A$, that is, the linear span of elements of the form $a^{n}\left(a^{*}\right)^{m}\left(n, m \in \mathbb{N}_{0}\right)$ is dense in $A$. Then

$$
\begin{aligned}
\hat{a}: \sigma(A) & \rightarrow \sigma(a), \\
\chi & \mapsto \chi(a)=\hat{a}(\chi),
\end{aligned}
$$

is a homeomorphism. Moreover, if we denote by the inverse of the Gelfand transform (see Theorem 6.22), then

$$
\begin{aligned}
\Phi: C(\sigma(a)) & \rightarrow A, \\
f & \mapsto(f \circ \hat{a})
\end{aligned}
$$

is an isometric $*$-isomorphism such that

$$
\Phi(1)=e \text { and } \Phi(i d)=a
$$

Proof. Note that by Theorem 6.21, $\hat{a}$ maps the Gelfand space continuously onto the spectrum $\sigma(a) \subseteq \mathbb{C}$. We show that the mapping $\hat{a}$ is injective. In fact, if $\chi_{1}, \chi_{2} \in$ $\sigma(A)$ are such that $\chi_{1}(a)=\chi_{2}(a)$, then, by the fact that the Gelfand transform is a *-homomorphism (Theorem 6.22), $\chi_{1}\left(a^{*}\right)=\chi_{2}\left(a^{*}\right)$. Hence, by the multiplicativity of $\chi_{1}$ and $\chi_{2}, \chi_{1}\left(a^{n}\left(a^{*}\right)^{m}\right)=\chi_{2}\left(a^{n}\left(a^{*}\right)^{m}\right)$ for every $n, m \in \mathbb{N}_{0}$. By assumption, $\chi_{1}$ and $\chi_{2}$ therefore coincide on a dense subspace of $A$. By continuity of the characters, $\chi_{1}=\chi_{2}$, and injectivity of $\hat{a}$ is proven. The function $\hat{a}$ being a continuous bijection between the compact spaces $\sigma(A)$ and $\sigma(a)$, it is necessarily a homeomorphism.

The mapping $C(K) \rightarrow C(\sigma(A)), f \mapsto f \circ \hat{a}$ is then an isometric $*$-isomorphism, and the same is true for the inverse ${ }^{2}: C(\sigma(A)) \rightarrow A$ of the Gelfand transform. Thus, $\Phi$ is an isometric $*$-isomorphism. The properties $\Phi(1)=e$ and $\Phi(\mathrm{id})=a$ follow easily from the definition of $\Phi$.

## Chapter 7 <br> Operators on Hilbert spaces

### 7.1 Spectral theorem for compact selfadjoint operators

Let $H, K$ be two Hilbert spaces, $T \in \mathscr{L}(H, K)$. For every $y \in K$ the mapping $H \rightarrow \mathbb{K}$, $x \mapsto\langle T x, y\rangle_{K}$ is a bounded linear functional on $H$ which admits a unique representation by $T^{*} y \in H$ such that

$$
\langle T x, y\rangle_{K}=\left\langle x, T^{*} y\right\rangle_{H} \quad(x \in H)
$$

The resulting linear operator $T^{*}: K \rightarrow H$ is called the (Hilbert space) adjoint of $T$.

Lemma 7.1. Let $H_{1}, H_{2}$, and $H_{3}$ be three Hilbert spaces, $T, S \in \mathscr{L}\left(H_{1}, H_{2}\right), R \in$ $\mathscr{L}\left(H_{2}, H_{3}\right)$ and $\lambda \in \mathbb{K}$. Then:
a) $(T+S)^{*}=T^{*}+S^{*}$.
b) $(\lambda T)^{*}=\bar{\lambda} T^{*}$.
c) $(R T)^{*}=T^{*} R^{*}$.
d) $T^{*} \in \mathscr{L}\left(H_{2}, H_{1}\right)$ and $\left\|T^{*}\right\|=\|T\|$.
e) $T^{* *}=T$.
f) $\left\|T^{*} T\right\|=\left\|T T^{*}\right\|=\|T\|^{2}$.
g) $\operatorname{ker} T=\left(\operatorname{ran} T^{*}\right)^{\perp}$ and $\operatorname{ker} T^{*}=(\operatorname{ran} T)^{\perp}$ (orthogonal spaces).

Proof. The properties (a)-(c) are simple exercises. Concerning (d), note that

$$
\begin{aligned}
\left\|T^{*}\right\| & :=\sup _{\|y\|_{H^{2}} \leq 1}\left\|T^{*} y\right\|_{H_{1}} \\
& =\sup _{\|y\|_{H^{2}} \leq 1} \sup _{\|x\|_{H^{1}} \leq 1}\left|\left\langle T^{*} y, x\right\rangle_{H_{1}}\right| \\
& =\sup _{\|x\|_{H^{1}} \leq 1} \sup _{\|y\|_{H^{2}} \leq 1}\left|\langle y, T x\rangle_{H_{2}}\right| \\
& =\sup _{\|x\|_{H^{1}} \leq 1}\|T x\|_{H_{2}} \\
& =\|T\| .
\end{aligned}
$$

Next, for every $x \in H_{1}, y \in H_{2}$,

$$
\begin{aligned}
\left\langle T^{* *} x, y\right\rangle_{H_{2}} & =\left\langle x, T^{*} y\right\rangle_{H_{1}} \\
& =\overline{\left\langle T^{*} y, x\right\rangle_{H_{1}}} \\
& =\overline{\langle y, T x\rangle_{H_{2}}} \\
& =\langle T x, y\rangle_{H_{2}},
\end{aligned}
$$

which implies (e). Finally, note that

$$
\begin{aligned}
\left\|T^{*} T\right\| & =\sup _{\|x\| \leq 1}\left\|T^{*} T x\right\| \\
& =\sup _{\|x\| \leq 1} \sup _{\|y\| \leq 1}\left|\left\langle T^{*} T x, y\right\rangle\right| \\
& =\sup _{\|x\| \leq 1} \sup _{\|y\| \leq 1}|\langle T x, T y\rangle| \\
& \geq \sup _{\|x\| \leq 1}|\langle T x, T x\rangle| \\
& =\sup _{\|x\| \leq 1}\|T x\|^{2} \\
& =\|T\|^{2},
\end{aligned}
$$

while the inequality $\left\|T^{*} T\right\| \leq\left\|T^{*}\right\|\|T\|=\|T\|^{2}$ (using also (d)) is trivial. Hence, we have proved ( f ). The property ( g ) is also left as an exercise.

Remark 7.2. Let $\mathscr{A}$ be a complex Banach algebra. A mapping $*: \mathscr{A} \rightarrow \mathscr{A}$ is called an involution if for every $a, b \in \mathscr{A}, \lambda \in \mathbb{C}$,

$$
(a+b)^{*}=a^{*}+b^{*},(a b)^{*}=b^{*} a^{*},(\lambda a)^{*}=\bar{\lambda} a^{*},\left(a^{*}\right)^{*}=a .
$$

If a complex Banach algebra $\mathscr{A}$ admits an involution $*$ such that for every $a \in \mathscr{A}$,

$$
\left\|a^{*} a\right\|=\|a\|^{2},
$$

then $\mathscr{A}$ is called a $C^{*}$-algebra.

If $H$ is a Hilbert space, then $\mathscr{L}(H)$ is a $C^{*}$-algebra for the involution $T \mapsto T^{*}$, where $T^{*}$ is the (Hilbert space) adjoint of $T$. This follows from Lemma 7.1.

The simplest $C^{*}$-algebra is $\mathbb{C}$ (the involution being the complex conjugation). In the space of matrices $\mathbb{C}^{N \times N}=\mathscr{L}\left(\mathbb{C}^{N}\right)$, the involution as defined above, that is, the Hilbert space adjoint with respect to the Euclidean inner product, is given by $A^{*}=\bar{A}^{t}$ (complex conjugation and transposition).

Given a compact space $K$, the space $C(K)$ is also a $C^{*}$-algebra for the usual algebra structure and the involution $f \mapsto f^{*}$ given by $f^{*}(x):=f \overline{(x)}(x \in K)$.

Let $H$ be a complex Hilbert space. An operator $T \in \mathscr{L}(H)$ is called selfadjoint if $T=T^{*}$, or, equivalently, if for every $x, y \in H$,

$$
\langle T x, y\rangle=\langle x, T y\rangle .
$$

We say that the operator $T$ is positive semidefinite and we write $T \geq 0$, if it is selfadjoint, and

$$
\langle T x, x\rangle \geq 0 \quad \text { for every } x \in H
$$

An operator $T \in \mathscr{L}(H)$ is called normal if $T T^{*}=T^{*} T$. An operator $U \in \mathscr{L}(H, K)$ between two Hilbert spaces is called unitary if $U$ is an isomorphism and $U^{*} U=I_{H}$ and $U U^{*}=I_{K}$.

Remark 7.3. In every $C^{*}$-algebra $\mathscr{A}$ one can define that an element $a$ is selfadjoint if $a=a^{*}$. The selfadjoint elements of $\mathscr{A}=\mathbb{C}$ are the real numbers. The selfadjoint elements of $\mathbb{C}^{N \times N}$ are the hermitian matrices, that is, the matrices $A$ for which $A=$ $\bar{A}^{t}$.

Theorem 7.4 (Hellinger-Toeplitz). Let $T: H \rightarrow H$ be linear and symmetric, that is,

$$
\langle T x, y\rangle=\langle x, T y\rangle \text { for every } x, y \in H
$$

Then $T$ is bounded.
Proof. Let $\left(x_{n}\right) \subseteq H$ be convergent to $x \in H$ and such that $\left(T x_{n}\right)$ converges to $y \in H$. Then, for every $z \in H$,

$$
\langle T x, z\rangle=\langle x, T z\rangle=\lim _{n \rightarrow \infty}\left\langle x_{n}, T z\right\rangle=\lim _{n \rightarrow \infty}\left\langle T x_{n}, z\right\rangle=\langle y, z\rangle .
$$

Hence, $T x=y$. This means that $T$ is closed, and by the closed graph theorem, $T$ is bounded.

Lemma 7.5. Let $T \in \mathscr{L}(H)$ be a selfadjoint operator on a Hilbert space $H$. Then

$$
\begin{equation*}
\sigma(T) \subseteq \overline{W(T)} \subseteq \mathbb{R} \tag{7.1}
\end{equation*}
$$

where

$$
\begin{equation*}
W(T):=\{\langle T x, x\rangle:\|x\|=1\} . \tag{7.2}
\end{equation*}
$$

is the numerical range of $T$.

Proof. Since $\langle T x, x\rangle=\langle x, T x\rangle=\overline{\langle T x, x\rangle}$ by symmetry, we obtain $W(T) \subseteq \mathbb{R}$.
Let $\lambda \in \mathbb{K}$ be such that $d:=\operatorname{dist}(\lambda, W(T))>0$. Then, for every $x \in H$ such that $\|x\|=1$,

$$
d=d\|x\| \leq|\lambda-\langle T x, x\rangle|=|\langle(\lambda-T) x, x\rangle| \leq\|(\lambda-T) x\| .
$$

By linearity, this estimates remains true for every $x \in H$. This estimate then implies that $\lambda-T$ is injective and that $\operatorname{ran}(\lambda-T)$ is closed. If $\operatorname{ran}(\lambda-T) \neq H$, then there exists $x_{0} \in(\operatorname{ran}(\lambda-T))^{\perp}$ such that $\left\|x_{0}\right\|=1$. For this $x_{0}$ we have

$$
0=\left\langle(\lambda-T) x_{0}, x_{0}\right\rangle=\lambda-\left\langle T x_{0}, x_{0}\right\rangle \geq d>0
$$

a contradiction. Hence, $\lambda-T$ is invertible, or $\lambda \in \rho(T)$. Thus we have proved also the first inclusion in (7.1).

Lemma 7.6. Let $T \in \mathscr{L}(H)$ be a selfadjoint operator on a Hilbert space H. Then

$$
\sup W(T) \in \sigma(T) \text { and } \inf W(T) \in \sigma(T)
$$

where $W(T)$ is the numerical range defined in (7.2).
Proof. Let $\lambda:=\sup W(T)$. By definition of $W(T)$, the form $a(x, y):=\lambda\langle x, y\rangle-$ $\langle T x, y\rangle$ is sesquilinear in the case of a complex Hilbert space, or bilinear and symmetric in the case of a real Hilbert space. Moreover, this form is positive semidefinite, that is, $a(x, x) \geq 0$ for every $x \in H$.

By the Cauchy-Schwarz inequality applied to the form $a(x, y)$, for every $x, y \in H$,

$$
|\langle\lambda x-T x, y\rangle| \leq\langle\lambda x-T x, x\rangle^{\frac{1}{2}}\langle\lambda y-T y, y\rangle^{\frac{1}{2}}
$$

This inequality implies that there exists a constant $C \geq 0$ such that for every $x \in H$,

$$
\|\lambda x-T x\| \leq C\langle\lambda x-T x, x\rangle^{\frac{1}{2}}
$$

Let $\left(x_{n}\right) \subseteq H,\left\|x_{n}\right\|=1$ be such that $\left\langle T x_{n}, x_{n}\right\rangle \rightarrow \lambda$. Then the preceeding inequality implies that $\lim _{n \rightarrow \infty}\left\|\lambda x_{n}-T x_{n}\right\|=0$. Hence, $\lambda \in \sigma_{a p}(T) \subseteq \sigma(T)$.

The proof that $\inf W(T) \in \sigma(T)$ is similar.
Lemma 7.7. Let $T \in \mathscr{L}(H)$ be a selfadjoint operator on a Hilbert space $H$. Then

$$
\|T\|=\sup _{\|x\|=1}|\langle T x, x\rangle|=\sup _{\lambda \in \sigma(T)}|\lambda| .
$$

Proof. The second equality follows from Lemma 7.6 combined with Lemma 7.5. Moreover, the inequality

$$
\sup _{\|x\|=1}|\langle T x, x\rangle| \leq\|T\|
$$

is obvious, by the definition of $\|T\|$ and the Cauchy-Schwarz inequality. Using the fact that $T=T^{*}$, one easily calculates for every $x, y \in H$,

$$
4 \operatorname{Re}\langle T x, y\rangle=\langle T(x+y), x+y\rangle-\langle T(x-y), x-y\rangle .
$$

Hence,

$$
\begin{aligned}
\|T\| & =\sup _{\|x\|=1}\|T x\| \\
& =\sup _{\|x\|=1} \sup _{\|y\|=1} \operatorname{Re}\langle T x, y\rangle \\
& =\frac{1}{4} \sup _{\|x\|=1} \sup _{\|y\|=1}[\langle T(x+y), x+y\rangle-\langle T(x-y), x-y\rangle] \\
& \leq \frac{1}{4} \sup _{\|x\|=1} \sup _{\|y\|=1}[|\langle T(x+y), x+y\rangle|+|\langle T(x-y), x-y\rangle|] \\
& \leq \sup _{\|z\|=1}|\langle T z, z\rangle| \frac{1}{4} \sup _{\|x\|=1} \sup _{\|y\|=1}\left[\|x+y\|^{2}+\|x-y\|^{2}\right] \\
& \leq \sup _{\|z\|=1}|\langle T z, z\rangle| \frac{1}{2} \sup _{\|x\|=1} \sup _{\|y\|=1}\left[\|x\|^{2}+\|y\|^{2}\right] \\
& \leq \sup _{\|z\|=1}|\langle T z, z\rangle|,
\end{aligned}
$$

which is just the remaining inequality.
Lemma 7.8. Let $T \in \mathscr{L}(H)$ be a selfadjoint operator on a complex Hilbert space, and let $x, y \in H$ be two eigenvectors corresponding to two distinct eigenvalues $\lambda$, $\mu \in \sigma_{p}(T)$. Then $\langle x, y\rangle=0$.
Proof. Since $T$ is selfadjoint and $\lambda, \mu \in \mathbb{R}$ (Lemma 7.5),

$$
\lambda\langle x, y\rangle=\langle\lambda x, y\rangle=\langle T x, y\rangle=\langle x, T y\rangle=\langle x, \mu y\rangle=\mu\langle x, y\rangle .
$$

Since $\lambda \neq \mu$, this equality can only hold if $\langle x, y\rangle=0$.
Theorem 7.9 (Spectral theorem for compact, selfadjoint operators). Let $H$ be a separable Hilbert space, and let $T \in \mathscr{K}(H)$ be a compact, selfadjoint operator. Then there exists an orthonormal basis $\left(e_{n}\right)_{n}$ of $H$, and a family $\left(\lambda_{n}\right)_{n}$ of real numbers such that $\lim _{n \rightarrow \infty} \lambda_{n}=0$ and

$$
T e_{n}=\lambda_{n} e_{n} \quad \text { for every } n
$$

that is, there is an orthonormal basis $\left(e_{n}\right)_{n}$ consisting only of eigenvectors of $T$. In other words, $T$ is unitarily equivalent to the multiplication operator $M: \ell^{2} \rightarrow \ell^{2}$, $M\left(x_{n}\right)_{n}:=\left(\lambda_{n} x_{n}\right)_{n}$, that is, there exists a unitary operator $U: H \rightarrow \ell^{2}$ such that the diagram

$$
\begin{array}{lll}
H \xrightarrow{T} & H \\
\downarrow_{U} & & \\
\ell^{2} \xrightarrow{M} & { }_{U}^{*}=U^{-1} \\
\ell^{2}
\end{array}
$$

## commutes

Proof. By the spectral theory of compact operators, $\sigma(T)$ is at most countable, every $\mu \in \sigma(T) \backslash\{0\}$ is an eigenvalue, and its eigenspace $H_{\mu}:=\operatorname{ker}(\mu-T)$ is finitedimensional.

Let $\left(\mu_{n}\right)$ be the (finite or countable) family of all nonzero eigenvalues of $T$ ( $\mu_{n} \neq \mu_{m}$ if $n \neq m$ ), and let $d_{n}:=\operatorname{dim} \operatorname{ker}\left(\mu_{n}-T\right)$ be their multiplicities. Let $\left(f_{k}^{n}\right)_{1 \leq k \leq d_{n}}$ be an orthonormal basis of $H_{\mu_{n}}=\operatorname{ker}\left(\mu_{n}-T\right)$. If the kernel $H_{0}:=\operatorname{ker} T$ is nontrivial, then choose also an orthonormal basis $\left(f_{k}\right)_{0 \leq k<\operatorname{dim} H_{0}}$ of $H_{0}$. Next, let $\left(e_{n}\right)$ be the family which is obtained by taking successively the union over all eigenvectors $f_{k}^{n}$ and $f_{k}$, and let $\left(\lambda_{n}\right)$ be the family which is obtained by taking the eigenvalues corresponding to $f_{k}^{n}$ or $f_{k}$. For simplicity, assume that the kernel $H_{0}=\operatorname{ker} T$ is trivial. Then $e_{1}=f_{1}^{1}, \ldots, e_{d_{1}}=f_{d_{1}}^{1}, e_{d_{1}+1}=f_{1}^{2}, \ldots, e_{d_{1}+d_{2}}=f_{d_{2}}^{2}$, etc., and $\lambda_{1}=\mu_{1}$, $\ldots, \lambda_{d_{1}}=\mu_{1}, \lambda_{d_{1}+1}=\mu_{2}, \ldots, \lambda_{d_{1}+d_{2}}=\mu_{2}$, etc.

The family $\left(e_{n}\right)$ thus obtained is an orthonormal system by construction and by Lemma 7.8. Moreover, by construction, $T e_{n}=\lambda_{n} e_{n}$ for every $n$. It remains only to show that $\operatorname{span}\left\{e_{n}: n\right\}=: H^{0}$ is dense in $H$.

Let $H^{1}:=\left(H^{0}\right)^{\perp}$ be the orthogonal complement. For every $x \in H^{1}$ and every $n$, since $T$ is selfadjoint,

$$
\left\langle T x, e_{n}\right\rangle=\left\langle x, T e_{n}\right\rangle=\left\langle x, \lambda_{n} e_{n}\right\rangle=\bar{\lambda}_{n}\left\langle x, e_{n}\right\rangle=0
$$

Hence, $T H^{1} \subset H^{1}$, that is, $T$ leaves the space $H^{1}$ invariant. We may thus consider the restriction $T^{1}:=\left.T\right|_{H_{1}} \in \mathscr{L}\left(H^{1}\right)$ which inherits the property from $T$ to be compact and selfadjoint. Since all eigenvectors of $T$ are contained in $H^{0}, T^{1}$ does not have any eigenvalue. In other words, $\sigma\left(T^{1}\right) \subseteq\{0\}$. By Lemma 7.7, this implies $T^{1}=0$. However, as we just remarked, $T^{1}$ does also not admit any eigenvector for the only possible eigenvalue 0 . Hence, $H^{1}=\operatorname{ker} T^{1}=\{0\}$, which implies that $H^{0}$ is dense in $H$.

To complete the proof, consider the operator $U: H \rightarrow \ell^{2}$ given by $U x:=$ $\left(\left\langle x, e_{n}\right\rangle\right)_{n}$. This operator does the work, that is, $U$ is unitary and $T=U^{*} M U$, as one easily shows.

Remark 7.10. Let $T \in \mathscr{K}(H)$ be a compact, selfadjoint operator on a general (not necessarily separable) Hilbert space. Then $H=\operatorname{ker} T \oplus(\operatorname{ker} T)^{\perp}$, where $(\operatorname{ker} T)^{\perp}=$ $\overline{\operatorname{ran} T}$ is separable (any compact, metric space is separable, and $\operatorname{ran} T$ is spanned by the relatively compact set $T B_{H}(0,1)$. Applying the above spectral theorem (which holds only on separable Hilbert spaces) to the restriction of $T$ to $\overline{\operatorname{ran} T}$, we obtain an orthonormal basis of $\operatorname{ran} T$ which consists only of eigenvectors of $T$. This (at most countable) orthonormal basis can be completed by an orthonormal basis of $\operatorname{ker} T$, which consists necessarily of eigenvectors to the eigenvalue 0 . As a conclusion, we obtain an orthonormal eigenbasis of $H$ which consists only of eigenvectors of $T$. We thus see that the assumption of separability of $H$ can be dropped in the spectral theorem.

We may immediately generalize the spectral theorem to the larger class of normal operators. For this, we also need the following variant of Schauder's theorem.

Lemma 7.11. Let $H, K$ be two Hilbert spaces and $T \in \mathscr{L}(H, K)$. Then $T$ is compact if and only if $T^{*}$ is compact.

Proof. It is intructive to represent the Hilbert space adjoint $T^{*}$ by using the Banach space adjoint $T^{\prime} \in \mathscr{L}\left(K^{\prime}, H^{\prime}\right)$ and the (antilinear) isomorphisms $\Phi_{H}: H^{\prime} \rightarrow H$ and $\Phi_{K}: K^{\prime} \rightarrow K$ which one obtains from the Theorem of Riesz-Fréchet (Theorem 2.44). In fact,

$$
T^{*}=\Phi_{H} T^{\prime} \Phi_{K}^{-1}
$$

If $T$ is compact, then $T^{\prime}$ is compact by Schauder's theorem (Theorem 5.25), and hence $T^{*}$ is compact due to the above representation. Conversely, if $T^{*}$ is compact, then, by what has just been said, $T^{* *}$ is compact. However, $T^{* *}=T$ (Lemma 7.1 (e)), and the claim is proved.

Theorem 7.12 (Spectral theorem for compact, normal operators). Let $H$ be a complex, separable Hilbert space, and let $T \in \mathscr{K}(H)$ be a compact, normal operator. Then there exists an orthonormal basis $\left(e_{n}\right)_{n \in I} \subseteq H(I \subseteq \mathbb{N})$ of $H$, and a sequence $\left(\lambda_{n}\right)_{n \in I} \subseteq \mathbb{C}$ such that $\lim _{n \rightarrow \infty} \lambda_{n}=0$ and

$$
T e_{n}=\lambda_{n} e_{n} \quad \text { for every } n \in I
$$

that is, $\left(e_{n}\right)$ is an orthonormal basis consisting only of eigenvectors of $T$.
Proof. We define

$$
\operatorname{Re} T:=\frac{T+T^{*}}{2} \text { and } \operatorname{Im} T:=\frac{T-T^{*}}{2 i}
$$

Since $T$ is normal, the operators $\operatorname{Re} T$ and $\operatorname{Im} T$ commute. Moreover, they are easily seen to be selfadjoint and compact (for compactness, we use Lemma 7.11). We show that $\operatorname{Re} T$ and $\operatorname{Im} T$ can be diagonalized simultaneously.

By the spectral theory of compact operators, $\sigma(\operatorname{Re} T)$ is at most countable, every $\alpha \in \sigma(\operatorname{Re} T) \backslash\{0\}$ is an eigenvalue, and its eigenspace $H_{\alpha}:=\operatorname{ker}(\alpha-T)$ is finitedimensional.

Let $\left(\alpha_{n}\right)$ be the (finite or countable) family of all nonzero eigenvalues of $\operatorname{Re} T$ ( $\alpha_{n} \neq \alpha_{m}$ if $n \neq m$ ), and let $d_{n}:=\operatorname{dim} \operatorname{ker}\left(\alpha_{n}-T\right)$ be their multiplicities. For every $e \in H_{\alpha_{n}}$ one has

$$
\operatorname{Re} T e=\alpha_{n} e
$$

We apply $\operatorname{Im} T$ on both sides of this equality, and use the fact that $\operatorname{Re} T$ and $\operatorname{Im} T$ commute, and we find that the vector $\operatorname{Im} T e$ is also an eigenvector of $\operatorname{Re} T$ for the eigenvalue $\alpha_{n}$. In other words, the eigenspaces $H_{\alpha_{n}}$ are left invariant under $\operatorname{Im} T$. By applying the spectral theorem for compact, selfadjoint operators to the restrictions of $\operatorname{Im} T$ to $H_{\alpha_{n}}$, we find for every $n$ an orthonormal basis $\left(f_{k}^{n}\right)_{1 \leq k \leq d_{n}}$ of $H_{\alpha_{n}}$, and we find a family $\left(\beta_{k}^{n}\right)_{1 \leq k \leq d_{n}}$ of real numbers such that

$$
\operatorname{Im} T f_{k}^{n}=\beta_{k}^{n} f_{k}^{n} \text { for every } 1 \leq k \leq d_{n}
$$

Of course, we still have

$$
\operatorname{Re} T f_{k}^{n}=\alpha_{n} f_{k}^{n} \text { for every } 1 \leq k \leq d_{n}
$$

If $H_{0}:=\operatorname{ker} \operatorname{Re} T$ is nontrivial, then we may repeat the arguments from above in order to see that $\operatorname{Im} T$ leaves $H_{0}$ invariant. We may also apply the spectral theorem for compact, selfadjoint operators to the restriction of $\operatorname{Im} T$ to $H_{0}$, and we find an orthonormal basis $\left(f_{k}\right)_{0 \leq k<\operatorname{dim} H_{0}}$ and a sequence $\left(\beta_{k}\right)_{0 \leq k<\operatorname{dim} H_{0}}$ of real numbers such that

$$
\operatorname{Im} T f_{k}=\beta_{k} f_{k} \text { for every } 0 \leq k<\operatorname{dim} H_{0} .
$$

Of course, we have

$$
\operatorname{Re} T f_{k}=0 \text { for every } 0 \leq k<\operatorname{dim} H_{0} .
$$

From the above relations and from the equality $T=\operatorname{Re} T+i \operatorname{Im} T$ we obtain

$$
T f_{k}^{n}=\left(\alpha_{n}+i \beta_{k}^{n}\right) f_{k}^{n}=: \mu_{k}^{n} f_{k}^{n} \text { for every } 1 \leq k \leq d_{n}
$$

and

$$
T f_{k}=i \beta_{k} f_{k}=: \mu_{k} f_{k} \text { for every } 0 \leq k<\operatorname{dim} H_{0},
$$

that is, the $f_{k}^{n}$ and $f_{k}$ are eigenvectors of $T$ for the complex eigenvalues $\mu_{k}^{n}$ and $\mu_{k}$, respectively.

Next, let $\left(e_{n}\right)$ be the family which is obtained by taking successively the union over all eigenvectors $f_{k}^{n}$ and $f_{k}$, and let $\left(\lambda_{n}\right)$ be the family which is obtained by taking the eigenvalues corresponding to $f_{k}^{n}$ or $f_{k}$. For simplicity, assume that the kernel $H_{0}=\operatorname{ker} \operatorname{Re} T$ is trivial. Then $e_{1}=f_{1}^{1}, \ldots, e_{d_{1}}=f_{d_{1}}^{1}, e_{d_{1}+1}=f_{1}^{2}, \ldots, e_{d_{1}+d_{2}}=$ $f_{d_{2}}^{2}$, etc., and $\lambda_{1}=\mu_{1}^{1}, \ldots, \lambda_{d_{1}}=\mu_{d_{1}}^{1}, \lambda_{d_{1}+1}=\mu_{1}^{2}, \ldots, \lambda_{d_{1}+d_{2}}=\mu_{d_{2}}^{2}$, etc.

The family $\left(e_{n}\right)$ thus obtained is orthonormal by construction and by Lemma 7.8 (applied to $\operatorname{Re} T$ ). Moreover, by construction, $T e_{n}=\lambda_{n} e_{n}$ for every $n$. It remains only to show that $\operatorname{span}\left\{e_{n}: n \in \mathbb{N}\right\}=: H^{0}$ is dense in $H$. For this, one proceeds similarly as in the proof of the spectral theorem for compact, selfadjoint operators. One shows that $\operatorname{Re} T$ and $\operatorname{Im} T$ leave $H^{1}=\left(H^{0}\right)^{\perp}$ invariant but admit no eigenvectors in $H^{1}$. This implies for example $\operatorname{Re} T=0$ in $H^{1}$, and thus $H^{1}=\{0\}$. As a consequence, $H^{0}$ is dense, and $\left(e_{n}\right)$ an orthonormal basis.

### 7.2 Spectral theorem for bounded, normal operators

## The continuous functional calculus

Theorem 7.13 (Spectral theorem for bounded, normal operators - the continuous functional calculus). Let $T \in \mathscr{L}(H)$ be a normal operator, and let $K:=\sigma(T)$ be its spectrum. Then there exists an $C^{*}$-algebra homomorphism

$$
\Phi: C(K) \rightarrow \mathscr{L}(H)
$$

with the following properties:
(i) $\Phi(\mathrm{id})=T$ and in particular $\Phi(p)=p(T)$ for every polynomial $p$.
(ii) $\Phi$ is isometric, that is, $\|\Phi(f)\|_{\mathscr{L}(H)}=\|f\|_{\infty}$ for every $f \in C(K)$.
(iii) $\Phi$ is positive in the sense that if $f \geq 0$, then $\Phi(f) \geq 0$.
(iv) (Spectral mapping theorem) For every $f \in C(K)$ one has $\sigma(\Phi(f))=$ $f(\sigma(T))=f(K)$.
(v) (Spectral mapping theorem for the point spectrum) For every $f \in C(K)$ and every $\lambda \in \sigma_{p}(T)$ and every corresponding eigenvector $x \in H$ (that is, $T x=\lambda x$ ) one has $\Phi(f) x=f(\lambda) x$.

Proof. This theorem is a direct consequence of Corollary to the Gelfand-Naimark theorem (Theorem 6.22, applied to the commutative $C^{*}$-subalgebra of $\mathscr{L}(H)$ generated by $T$ (commutativity of this subalgebra follows from the assumption that $T$ is normal).

## The Riesz-Markov representation theorem

Let $K$ be a compact space. We denote by $\mathscr{B}(K)$ the Borel- $\sigma$-algebra on $K$, that is, the smallest $\sigma$-algebra on $K$ which contains the open sets. A Borel measure on $K$ is a measure on the Borel- $\sigma$-algebra $\mathscr{B}(K)$, that is, a $\sigma$-additive function $\mu: \mathscr{B}(K) \rightarrow$ $[0,+\infty]$ (we consider here only nonnegative measures). A Borel measure $\mu$ on $K$ is regular if for every Borel measurable set $A \subseteq K$
(i) $\mu(A)=\inf \{\mu(O): O \supseteq A, O$ open $\}$, and
(ii) $\mu(A)=\sup \{\mu(C): C \subseteq A, C$ compact $\}$.

We say that $\mu$ is finite if $\mu(K)<\infty$. The following Riest-Markov representation theorem characterizes positive, linear functionals on $C(K)$. We say that a functional $\varphi \in C(K)^{\prime}$ is positive if $\varphi(f) \geq 0$ for every function $f \in C(K)$ taking its values in $\mathbb{R}_{+}$(the notion of positivity makes also sense on the complex space $C(K)$ ). Finally, we define $B(K)$ to be the space of all bounded, Borel measurable functions $K \rightarrow \mathbb{C}$. Equipped with the sup-norm, $B(K)$ is a Banach space which contains $C(K)$ as a closed, linear subspace.

Theorem 7.14 (Riesz-Markov representation theorem). Let $K$ be a compact space. Then, for every positive functional $\varphi \in C(K)^{\prime}$ there exists a finite, regular Borel measure $\mu$ on $K$ such that

$$
\varphi(f)=\int_{K} f d \mu \text { for every } f \in C(K)
$$

Proof. Let $\varphi \in C(K)^{\prime}$ be a positive functional. If necessary, we restrict $\varphi$ to the (real) subspace of real-valued continuous functions. By Hahn-Banach, we may extend the
functional $\varphi$ to a functional $\tilde{\varphi}$ on the space $B(K)$ of bounded Borel measurable functions, such that $\|\tilde{\varphi}\|_{B(K)^{\prime}}=\|\varphi\|_{C(K)^{\prime}}$.

Since $\varphi$ is positive, then $\varphi(1)=\|\varphi\|_{C(K)^{\prime}}$. Hence, for every Borel function $f \in$ $B(K)$ satisfying $\|f\|_{\infty} \leq 1$ one has

$$
|\tilde{\varphi}(f)| \leq\|\tilde{\varphi}\|_{B(K)^{\prime}}=\|\varphi\|_{C(K)^{\prime}}=\varphi(1) .
$$

In particular, if $f \in B(K)$ is a positive Borel function such that $\|f\|_{\infty} \leq 1$, then $1-f$ is also a positive Borel function, $\|1-f\|_{\infty} \leq 1$, and thus $|\tilde{\varphi}(f)| \leq \varphi(1)$ and $|\tilde{\varphi}(1-f)| \leq \varphi(1)$. On the other hand,

$$
0 \leq \varphi(1)=\tilde{\varphi}(1)=\tilde{\varphi}(f)+\tilde{\varphi}(1-f)
$$

which, together with the preceding estimates, is only possible if $\tilde{\varphi}(f) \geq 0$ (and $\tilde{\varphi}(1-f) \geq 0)$. We have thus proved that the extension $\tilde{\varphi}$ is a positive linear functional on $B(K)$.

For every Borel measurable set $A \subseteq K$, we now define

$$
\mu(A):=\tilde{\varphi}\left(\chi_{A}\right) \geq 0
$$

where $\chi_{A} \in B(K)$ is the characteristic function of the set $A$. We claim that $\mu$ is a bounded, regular, Borel measure which represents $\varphi$ as stated in the theorem.

First, $\mu$ is finitely additive by additivity of $\tilde{\varphi}$, and $\mu$ is monotone $(\mu(A) \leq \mu(B)$ whenever $A \subseteq B$ ) by positivity of $\tilde{\varphi}$.

## The spectral theorem for bounded, normal operators

Lemma 7.15. Let $T \in \mathscr{L}(H)$ be a normal operator on a separable Hilbert space, for which there exists $x \in X$ such that $\operatorname{span}\left\{T^{n}\left(T^{*}\right)^{m} x: n, m \in \mathbb{N}_{0}\right\}$ is dense in $H$. Let $K=\sigma(T)$ be the spectrum of $T$. Then there exists a regular, finite Borel measure $\mu$ on $K$ and a unitary operator $U: H \rightarrow L^{2}(K ; d \mu)$ such that the diagram

commutes. Here, $M: L^{2}(K ; d \mu) \rightarrow L^{2}(K ; d \mu)$ is the multiplication operator given by

$$
M f(\omega)=\omega f(\omega) \quad\left(f \in L^{2}(K ; d \mu), \omega \in K\right)
$$

In other words, $T$ is unitarily equivalent to a multiplication operator.
Proof. Let $x \in H$ be any vector, and let $\Phi: C(K) \rightarrow \mathscr{L}(H)$ be the continuous functional calculus associated with $T$ (Theorem 7.13). Then the linear mapping

$$
\begin{aligned}
\varphi_{x}: C(K) & \rightarrow \mathbb{C}, \\
f & \mapsto\langle\Phi(f) x, x\rangle,
\end{aligned}
$$

is bounded and positive (Theorem 7.13 (ii), (iii)). By the Riesz-Markov representation theorem (Theorem 7.14), there exists a finite, regular Borel measure $\mu_{x}$ on $K$ such that

$$
\varphi_{x}(f)=\langle\Phi(f) x, x\rangle=\int_{K} f d \mu \text { for every } f \in C(K)
$$

As a consequence of this equality and by using the properties of $\Phi$, for every $f \in$ $C(K)$,

$$
\begin{aligned}
\|\Phi(f) x\|_{H}^{2} & =\left\langle\Phi(f)^{*} \Phi(f) x, x\right\rangle_{H} \\
& =\langle\Phi(\bar{f} f) x, x\rangle_{H} \\
& =\int_{K}|f|^{2} d \mu
\end{aligned}
$$

This equality shows first that if $f_{1}, f_{2} \in C(K)$ coincide $\mu$-almost everywhere, then $\Phi\left(f_{1}\right) x=\Phi\left(f_{2}\right) x$. Hence, the operator

$$
\begin{aligned}
U^{*}: L^{2}(K ; d \mu) & \rightarrow H, \\
f & \mapsto U^{*} f=\Phi(f) x
\end{aligned}
$$

is well defined first for equivalence classes of continuous functions, but then, by the above equality and by continuous extension, everywhere on $L^{2}(K ; d \mu)$. Moreover, the operator thus defined is isometric.

Now we suppose that the vector $x \in H$, which was arbitrary in the beginning, is as in the statement. Then the operator $U^{*}$ is isometric and invertible, and thus a unitary operator. In fact, $U^{*}$ being isometric, it is injective and has closed range. Moreover, the range of $U^{*}$ contains the set $\left\{T^{n}\left(T^{*}\right)^{m} x: n, m \in \mathbb{N}_{0}\right\}=\left\{\Phi\left(\mathrm{Id}^{n} i d^{m}\right) x: n, m \in\right.$ $\left.\mathbb{N}_{0}\right\}$ which is dense in $H$ by the assumption and the choice of $x$. Hence, $U^{*}$ is surjective.

Finally, for every $f \in C(K)$,

$$
\begin{aligned}
T U^{*} f & =T \Phi(f) x \\
& =\Phi(i d) \Phi(f) x \\
& =\Phi(i d \cdot f) x \\
& =U^{*}(i d \cdot f),
\end{aligned}
$$

and thus $U T U^{*}\left(U=\left(U^{*}\right)^{-1}\right)$ is the multiplication operator given in the statement.
Lemma 7.16. Let $T \in \mathscr{L}(H)$ be a normal operator on a Hilbert space $H$. Then there exists a family $\left(H_{i}\right)_{i \in I}$ of closed subspaces such that
a) the $H_{i}$ are mutually orthogonal,
b) $H=\bigoplus_{i \in I} H_{i}$,
c) each $H_{i}$ is invariant under $T$ and $T^{*}$, and
d) for every $i \in I$ there exists $x \in H_{i}$ such that $\left\{T^{n}\left(T^{*}\right)^{m} x: n, m \in \mathbb{N}_{0}\right\}$ is dense in $H_{i}$.
Proof.
Theorem 7.17 (Spectral theorem for bounded, normal operators). Let $T \in$ $\mathscr{L}(H)$ be a normal operator on a separable Hilbert space $H$. Then there exists a measure space $(\Omega, \mathscr{A}, \mu)$, a function $m \in L^{\infty}(\Omega ; d \mu)$, and a unitary operator $U: H \rightarrow L^{2}(\Omega ; d \mu)$ such that the diagram

commutes. Here, $M: L^{2}(\Omega ; d \mu) \rightarrow L^{2}(\Omega ; d \mu)$ is the multiplication operator given by

$$
M f(\omega)=m(\omega) f(\omega) \quad\left(f \in L^{2}(\Omega ; d \mu), \omega \in \Omega\right)
$$

In other words, $T$ is unitarily equivalent to a multiplication operator.
Proof. Choose a family $\left(H_{i}\right)_{i \in I}$ (with $I \subseteq \mathbb{N}$ ) as in Lemma 7.16. By Lemma 7.15, for every $i \in I$ there exists a finite, regular Borel measure $\mu_{i}$ on $\sigma\left(\left.T\right|_{H_{i}}\right) \subseteq \sigma(T)=: K$ and a unitary operator $U_{i}^{*}: L^{2}\left(\sigma(T) ; d \mu_{i}\right) \rightarrow H_{i}$ such that $\left.U_{i} T\right|_{H_{i}} U_{i}^{*}=M_{i}$, where $M_{i}$ : $L^{2}\left(K ; d \mu_{i}\right) \rightarrow L^{2}\left(K ; d \mu_{i}\right)$ is the multiplication operator given by $M_{i} f(\omega)=\omega f(\omega)$.

Set $\Omega:=K \times I=\bigcup_{i \in I} \sigma(T) \times\{i\}$, and let $\mu$ be the Borel measure on $\Omega$ whose restriction to $\sigma(T) \times\{i\} \cong \sigma(T)$ coincides with $\mu_{i}$, that is,

$$
\mu\left(\bigcup_{i \in I} B_{i} \times\{i\}\right):=\sum_{i \in I} \mu_{i}\left(B_{i}\right) \quad\left(B_{i} \in \mathscr{B}(K)\right) .
$$

Note that

$$
L^{2}(\Omega ; d \mu) \cong \bigoplus_{i \in I} L^{2}\left(K ; d \mu_{i}\right)
$$

in a canonical way, and that, via this identification, $U^{*}=\bigoplus_{i \in I} U_{i}^{*}$ defines a unitary operator from $L^{2}(\Omega ; d \mu)$ onto $H=\bigoplus_{i \in I} H_{i}$. It is now a short exercise to show that $U T U^{*}=M$, where $M L^{2}(\Omega ; d \mu) \rightarrow L^{2}(\Omega ; d \mu)$ is the multiplication operator given by

$$
M f(\omega, i)=\omega f(\omega, i)
$$

## The measurable functional calculus

In the following, given a Borel measurable $K \subseteq \mathbb{R}$, we define the space

$$
B(K):=\{f: K \rightarrow \mathbb{C}: f \text { is bounded and Borel measurable }\} .
$$

Equipped with the supremum norm $\|\cdot\|_{\infty}$, this space is a $C^{*}$-algebra for the natural (pointwise) scalar multiplication, addition and multiplication. Clearly, if $K$ is compact, $B(K)$ contains $C(K)$ as a closed subspace.

Theorem 7.18 (Spectral theorem - the measurable functional calculus). Let $T \in$ $\mathscr{L}(H)$ be a normal operator on a separable Hilbert space $H$. Let the measure space $(\Omega, \mathscr{A}, \mu)$, the unitary operator $U: H \rightarrow L^{2}(\Omega ; d \mu)$, the function $m \in L^{\infty}(\Omega ; d \mu)$ and the multiplication operator $M \in \mathscr{L}\left(L^{2}(\Omega ; d \mu)\right)$ be as in the Spectral Theorem (Theorem 7.17). Let $K:=\sigma(T)$. Then the operator

$$
\begin{aligned}
\tilde{\Phi}: B(K) & \rightarrow \mathscr{L}(H), \\
f & \mapsto \tilde{\Phi}(f):=U^{*} f(M) U,
\end{aligned}
$$

where $f(M) \in \mathscr{L}\left(L^{2}(\Omega ; d \mu)\right)$ is the multiplication operator given by

$$
f(M) g(\omega):=f(m(\omega)) g(\omega) \quad\left(g \in L^{2}(\Omega ; d \mu), \omega \in \Omega\right),
$$

is a $C^{*}$-algebra homomorphism which extends the continuous functional calculus $\Phi$ from Theorem 7.13, and which has the properties:
(i) $\|\tilde{\Phi}\|=1$,
(ii) $\tilde{\Phi}(f) \geq 0$ whenever $f \geq 0$, and
(iii) if $\left(f_{n}\right)$ is a bounded sequence in $B(K)$ which converges $\mu$-almost everywhere to a function $f \in B(K)$, then, for every $x \in H$,

$$
\lim _{n} \tilde{\Phi}\left(f_{n}\right) x=\tilde{\Phi}(f) x
$$

Remark 7.19. Note that we can choose the multiplication operator $M$ such that the range of $m$ is a subset of $K$, so that the expression $f(m(\omega))$ is well defined.

Proof. In the special case $T=M$, that is, when $T$ already is a multiplication operator (and $U=U^{*}=I$ ), the properties of $\tilde{\Phi}$ are easy to verify, even property (iii), which relies only on Lebesgue's dominated convergence theorem. The case of general $T$ follows then from this special case and the Spectral Theorem (Theorem 7.17).

## Spectral measures and spectral decomposition

### 7.3 Spectral theorem for unbounded selfadjoint operators

In the preceding two sections, we have actually proved more than just solvability of an elliptic and a hyperbolic partial differential equation. We have proved that the Dirichlet-Laplace operator is selfadjoint, that it has a compact resolvent, and that
therefore it is diagonalisable. In this last section, we discuss the spectral theorem for unbounded selfadjoint operators with compact resolvent.

Let $H$ be a complex Hilbert space, and let $A: H \supseteq \operatorname{dom} A \rightarrow H$ be a densely defined (that is, the domain $\operatorname{dom} A$ is dense in $H$ ) and linear operator. We define

$$
\begin{aligned}
\operatorname{dom} A^{*} & :=\left\{x \in H: \exists y \in H \forall z \in \operatorname{dom} A:\langle A z, x\rangle_{H}=\langle z, y\rangle_{H}\right\}, \\
A^{*} x & :=y .
\end{aligned}
$$

The operator $\left(A^{*}, \operatorname{dom} A^{*}\right)$ is called the (Hilbert space) adjoint of $A$. For every $x \in \operatorname{dom} A, y \in \operatorname{dom} A^{*}$ one has

$$
\langle A x, y\rangle=\left\langle x, A^{*} y\right\rangle .
$$

Remark 7.20. The adjoint $A^{*}$ is well-defined in the sense that the element $y \in H$ is uniquely determined (use that $\operatorname{dom} A$ is dense in $H$ ).

Lemma 7.21. Let $A: \operatorname{dom} A \rightarrow H$ be a densely defined, linear operator. Then $A^{*}:$ $\operatorname{dom} A^{*} \rightarrow H$ is closed.

Proof. Let $\left(x_{n}\right) \subseteq \operatorname{dom} A^{*}$ be convergent to some $x \in H$ and such that $\left(A^{*} x_{n}\right)$ converges to $y \in H$. Then, for every $z \in \operatorname{dom} A$,

$$
\begin{aligned}
\langle z, y\rangle & =\lim _{n \rightarrow \infty}\left\langle z, A^{*} x_{n}\right\rangle \\
& =\lim _{n \rightarrow \infty}\left\langle A z, x_{n}\right\rangle \\
& =\langle A z, x\rangle .
\end{aligned}
$$

By definition of $A^{*}$ this implies $x \in \operatorname{dom} A^{*}$ and $A^{*} x=y$. Hence, $A^{*}$ is closed.
Let $H$ be a complex Hilbert space, and let $A: H \supseteq \operatorname{dom} A \rightarrow H$ be a densely defined, linear operator. We say that $A$ is symmetric if for every $x, y \in \operatorname{dom} A$,

$$
\langle A x, y\rangle=\langle x, A y\rangle .
$$

We say that $A$ is selfadjoint if $A=A^{*}$.
Remark 7.22. Saying that $A$ is selfadjoint, that is, that $A=A^{*}$, means that $\operatorname{dom} A=$ $\operatorname{dom} A^{*}$ and $A=A^{*}$. By Lemma 7.21, every selfadjoint operator is necessarily closed. Note, however, that a symmetric closed linear operator $A$ need in general not be selfadjoint! However, if $\operatorname{dom} A=H$, then symmetric implies selfadjoint by the Theorem of Hellinger-Toeplitz (Theorem 7.4).

Remark 7.23. If a bounded operator $A: H \rightarrow H$ ( $\operatorname{dom} A=H!$ ) is selfadjoint in the sense of the definition for unbounded operators (see page 124), then $A$ is selfadjoint in the sense of the definition for bounded operators (see page 113), and vice versa.
Lemma 7.24. Let $A: \operatorname{dom} A \rightarrow H$ be densely defined and symmetric. Then the following are equivalent:
(i) A is selfadjoint.
(ii) $A$ is closed and $\operatorname{ker}\left(A^{*} \pm i\right)=\{0\}$.
(iii) $\operatorname{ran}(A \pm i)=H$.

Proof. We first remark that if $(A, \operatorname{dom} A)$ is symmetric, then $\operatorname{ker}(A \pm i)=\{0\}$. In fact, let $x \in H$ be such that $(A-i) x=0$. Since $A$ is symmetric,

$$
i\|x\|^{2}=\langle i x, x\rangle=\langle A x, x\rangle=\langle x, A x\rangle=-i\|x\|^{2} .
$$

Hence, $x=0$. Similarly, one proves $\operatorname{ker}(A+i)=\{0\}$.
(i) $\Rightarrow$ (ii). Now assume that $A$ is selfadjoint. By Lemma $7.21, A^{*}$ is closed, and therefore $A\left(=A^{*}\right)$ is closed. Since $A$ is symmetric, and since $A^{*}=A$, we find $\operatorname{ker}\left(A^{*} \pm i\right)=\{0\}$ by the above argument.
(ii) $\Rightarrow$ (iii). Similarly as in Lemma 5.14 one proves that

$$
\operatorname{ker}\left(A^{*}-i\right)=(\operatorname{ran}(A+i))^{\perp}
$$

where $\perp$ now means the Hilbert space orthogonal. Hence, if $\operatorname{ker}\left(A^{*}-i\right)=\{0\}$, then $\operatorname{ran}(A+i)$ is dense in $H$. We prove that $\operatorname{ran}(A+i)$ is also closed. Since $A$ is symmetric, we have $\langle A x, x\rangle \in \mathbb{R}$ for every $x \in \operatorname{dom} A$. Hence, for every $x \in \operatorname{dom} A$,

$$
\begin{aligned}
\|(A+i) x\| & =\|A x\|^{2}+\|x\|^{2}+2 \operatorname{Re}\langle A x, i x\rangle \\
& =\|A x\|^{2}+\|x\|^{2} \geq\|x\|^{2} .
\end{aligned}
$$

Let $\left(x_{n}\right) \subseteq \operatorname{dom} A$ be such that $\lim _{n \rightarrow \infty}(A+i) x_{n}=y \in H$ exists. By the preceding inequality, this implies that $\left(x_{n}\right)$ is a Cauchy sequence in $H$. Hence, $x:=\lim _{n \rightarrow \infty} x_{n} \in$ $H$ exists. Since $A+i$ is closed, we obtain $x \in \operatorname{dom} A$ and $(A+i) x=y$. We have shown that $\operatorname{ran}(A+i)$ is closed. Similarly, one shows that $\operatorname{ran}(A-i)$ is closed.
(iii) $\Rightarrow$ (i). Since $A$ is symmetric, $\operatorname{dom} A \subseteq \operatorname{dom} A^{*}$ and $A x=A^{*} x$ for every $x \in \operatorname{dom} A$. It remains to show that $\operatorname{dom} A^{*} \subseteq \operatorname{dom} A$. Let $y \in \operatorname{dom} A^{*}$. Since $\operatorname{ran}(A+i)=H$, there exists $x \in \operatorname{dom} A$ such that $\left(A^{*}+i\right) y=(A+i) x$. By the inclu$\operatorname{sion}(A, \operatorname{dom} A) \subseteq\left(A^{*}, \operatorname{dom} A^{*}\right),\left(A^{*}+i\right) y=\left(A^{*}+i\right) x$. Since $\operatorname{ran}(A-i)=H$ implies $\operatorname{ker}\left(A^{*}+i\right)=\{0\}$ (compare again with Lemma 5.14), this implies $x=y \in \operatorname{dom} A$.

Exercise 7.25 The Dirichlet-Laplace operator A defined in (7.4) is selfadjoint.
Lemma 7.26. Let $A: \operatorname{dom} A \rightarrow H$ be densely defined and closed. Then, for every $\lambda \in \rho(A)$ one has $\bar{\lambda} \in \rho\left(A^{*}\right)$ and

$$
R(\lambda, A)^{*}=R\left(\bar{\lambda}, A^{*}\right) .
$$

Proof. For every $x \in \operatorname{dom} A$ and every $y \in \operatorname{dom} A^{*}$ one has

$$
\begin{aligned}
\left\langle x, R(\lambda, A)^{*}\left(\bar{\lambda}-A^{*}\right) y\right\rangle & =\left\langle R(\lambda, A) x,\left(\bar{\lambda}-A^{*}\right) y\right\rangle \\
& =\langle(\lambda-A) R(\lambda, A) x, y\rangle \\
& =\langle x, y\rangle
\end{aligned}
$$

and

$$
\begin{aligned}
\left\langle x,\left(\bar{\lambda}-A^{*}\right) R(\lambda, A)^{*} y\right\rangle & =\left\langle(\lambda-A) x, R(\lambda, A)^{*} y\right\rangle \\
& =\langle R(\lambda, A)(\lambda-A) x, y\rangle \\
& =\langle x, y\rangle,
\end{aligned}
$$

so that $\bar{\lambda}-A^{*}$ is invertible and $R\left(\bar{\lambda}, A^{*}\right)=R(\lambda, A)^{*}$.
Theorem 7.27 (Spectral mapping theorem). Let $A: \operatorname{dom} A \rightarrow H$ be densely defined, closed. Assume that $\rho(A)$ is not empty. Then, for every $\lambda \in \rho(A)$,

$$
(\lambda-\sigma(A))^{-1}=\sigma\left((\lambda-A)^{-1}\right) \backslash\{0\} .
$$

Proof. The proof is an exercise.
We say that a closed, linear operator $(A, \operatorname{dom} A)$ on a Banach space $X$ has compact resolvent if $\rho(A)$ is nonempty, and if there exists $\lambda \in \rho(A)$ such that $R(\lambda, A)$ is compact.

Lemma 7.28. Let $(A, \operatorname{dom} A)$ be a closed, linear operator on a Banach space $X$ such that $\rho(A) \neq \emptyset$. Then the following are equivalent:
(i) A has compact resolvent.
(ii) For every $\lambda \in \rho(A)$, the resolvent $R(\lambda, A)$ is compact.
(iii) The embedding $j:\left(\operatorname{dom} A,\|\cdot\|_{\operatorname{dom} A}\right) \rightarrow\left(X,\|\cdot\|_{X}\right), x \mapsto x$ is compact.

Proof. The implication (ii) $\Rightarrow$ (i) is trivial, while the converse (i) $\Rightarrow$ (ii) is a consequence of the resolvent identity

$$
R(\mu, A)=R(\lambda, A)+(\lambda-\mu) R(\mu, A) R(\lambda, A)
$$

(i) $\Rightarrow$ (iii) Assume that $\lambda \in \rho(A)$ is such that $R(\lambda, A)$ is compact. Let $\left(x_{n}\right)$ be a bounded sequence in $\left(\operatorname{dom} A,\|\cdot\|_{\operatorname{dom} A}\right)$, that is, there exists $C \geq 0$ such that

$$
\left\|x_{n}\right\|_{X}+\left\|A x_{n}\right\|_{X} \leq C \text { for every } n
$$

Since $R(\lambda, A)$ is invertible from $X$ onto $\operatorname{dom} A$, there exists a sequence $\left(y_{n}\right)$ in $X$ such that $R(\lambda, A) y_{n}=x_{n}$. Using the equality $A R(\lambda, A)=\lambda R(\lambda, A)-I$, the above estimate for the $x_{n}$ yields

$$
\left\|R(\lambda, A) y_{n}\right\|_{X}+\left\|\lambda R(\lambda, A) y_{n}-y_{n}\right\|_{X} \leq C \text { for every } n .
$$

This estimate yields that $\left(y_{n}\right)$ is necessarily bounded in $X$. Since $R(\lambda, A)$ is compact, there exists a subsequence of $\left(y_{n}\right)$ (which we denote for simplicity again by $\left(y_{n}\right)$ ) such that $\left(R(\lambda, A) y_{n}\right)=\left(x_{n}\right)$ converges in $X$. In other words, for every bounded sequence $\left(x_{n}\right)$ in $\left(\operatorname{dom} A,\|\cdot\|_{\operatorname{domA} A}\right)$ we can extract a subsequence which converges in $X$. Hence, the embedding $j:\left(\operatorname{dom} A,\|\cdot\|_{\operatorname{dom} A}\right) \rightarrow\left(X,\|\cdot\|_{X}\right), x \mapsto x$ is compact.
(iii) $\Rightarrow$ (i) Choose any $\lambda \in \rho(A)$. Then the operator $j:\left(\operatorname{dom} A,\|\cdot\|_{\operatorname{dom} A}\right) \rightarrow(X, \| \cdot$ $\|_{X}$ ), $x \mapsto \lambda x-A x$ is continuous (by definition of the graph norm) and invertible (by the choice of $\lambda$ ). By the bounded inverse theorem, $R(\lambda, A)$ is a bounded linear operator from $\left(X,\|\cdot\|_{X}\right)$ onto $\left(\operatorname{dom} A,\|\cdot\|_{\operatorname{dom} A}\right)$. Composing this operator with $j$, we obtain that $R(\lambda, A)$ is a compact operator on $X$.

Lemma 7.29. Consider the meromorphic functions $f$ and $g$ on $\mathbb{C}$ given by

$$
\begin{aligned}
& f(z):=\frac{i-z}{i+z} \quad \text { and } \\
& g(z):=i \frac{1-z}{1+z} \quad(z \in \mathbb{C}) .
\end{aligned}
$$

Then, for every $z \in \mathbb{C}$ :
a) If $z \in \mathbb{R}$, then $|f(z)|=1$. If $|z|=1$, then $f(z) \in i \mathbb{R}$.
b) If $z \in i \mathbb{R}$, then $|g(z)|=1$. If $|z|=1$, then $g(z) \in \mathbb{R}$.
c) $f(g(z))=g(f(z))=z$.

The functions $f$ and $g$ in the preceding lemma are two special Möbius transforms. A general Möbius transform has the form $f(z)=\frac{a z+b}{c z+d}$ and it always has the property that it maps straight lines to straight lines or circles, and circles to straight lines or circles. Möbius transforms are the affine mappings on the Riemann sphere. We use their properties in the following lemma in order to transform selfadjoint operators (which have spectrum in the real line) to unitary operators (which have spectrum in the unit circle) and back.

Lemma 7.30 (Cayley transform). Let $H$ be a Hilbert space.
a) Let $A: \operatorname{dom} A \rightarrow H$ be a densely defined, selfadjoint operator. Then its Cayley transform

$$
f(A):=U:=(i-A)(i+A)^{-1}
$$

is a unitary operator such that $\operatorname{rg}(I+U)=\operatorname{dom} A$. In particular, $I+U$ has dense range.
b) If $U \in \mathscr{L}(H)$ is a unitary operator such that $I+U$ has dense range, then

$$
g(U):=A:=i(I-U)(I+U)^{-1}
$$

with maximal domain $\operatorname{dom} A:=\operatorname{rg}(I+U) \rightarrow H$ is selfadjoint.
c) Let $A: \operatorname{dom} A \rightarrow H$ be a densely defined, selfadjoint operator. Then $g(f(A))=$ A.

Theorem 7.31 (Spectral theorem for unbounded selfadjoint operators with compact resolvent). Let $A: \operatorname{dom} A \rightarrow H$ be densely defined, selfadjoint, having compact resolvent. Then there exists an orthonormal basis $\left(e_{n}\right) \subseteq H$ and a sequence $\left(\lambda_{n}\right) \subseteq \mathbb{R}$ such that $\lim _{n \rightarrow \infty}\left|\lambda_{n}\right|=\infty$,

$$
e_{n} \in \operatorname{dom} A \text { and } A e_{n}=\lambda_{n} e_{n} \text { for every } n
$$

Moreover, $\sigma(A)=\sigma_{p}(A)=\left\{\lambda_{n}: n\right\}$.
Proof. Let $\lambda \in \rho(A)$ be such that $R(\lambda, A) \in \mathscr{K}(H)$. By Theorem 5.30, $\sigma(R(\lambda, A))$ is countable. Hence, by Theorem 7.27, $\sigma(A)$ is countable. In particular, there exists $\mu \in$ $\rho(A) \cap \mathbb{R}$. By Lemma 7.28 (that is, by the resolvent identity), $R(\mu, A)$ is compact, too. Moreover, since $\mu \in \mathbb{R}$, for every $x, y \in H$,

$$
\begin{aligned}
\langle R(\mu, A) x, y\rangle & =\langle R(\mu, A) x,(\mu-A) R(\mu, A) y\rangle \\
& =\langle(\mu-A) R(\mu, A) x, R(\mu, A) y\rangle \\
& =\langle x, R(\mu, A) y\rangle,
\end{aligned}
$$

so that $R(\mu, A)$ is selfadjoint. By the spectral theorem for selfadjoint compact operators, there exists an orthonormal basis $\left(e_{n}\right)$ of $H$ and a sequence $\left(\mu_{n}\right) \subseteq \mathbb{R} \backslash\{0\}$ such that $\lim _{n \rightarrow \infty} \mu_{n}=0$ and such that

$$
\mu_{n} e_{n}=R(\mu, A) e_{n} \text { for every } n
$$

This equation implies on the one hand that $e_{n} \in \operatorname{dom} A$ and on the other hand, when we multiply by $\mu-A$,

$$
\lambda_{n} e_{n}=A e_{n} \text { for every } n
$$

with $\lambda_{n}=\mu-\frac{1}{\mu_{n}}$. Clearly, $\lim _{n \rightarrow \infty}\left|\lambda_{n}\right|=\infty$, and by the spectral mapping theorem (Theorem 7.27), $\sigma(A)=\sigma_{p}(A)=\left\{\lambda_{n}: n\right\}$. The claim is proved.

### 7.4 Hilbert-Schmidt operators and trace class operators

## 7.5 * Elliptic partial differential equations

Let $\Omega \subseteq \mathbb{R}^{N}$ be open and bounded, $\lambda \in \mathbb{C}$, and consider the elliptic partial differential equation

$$
\begin{cases}\lambda u-\Delta u=f & \text { in } \Omega  \tag{7.3}\\ u=0 & \operatorname{in} \partial \Omega\end{cases}
$$

where $\Delta$ stands for the Laplace operator and $f \in L^{2}(\Omega)$.
Recall from Chapter 2 that a function $u \in H_{0}^{1}(\Omega)$ is a weak solution of (7.3) if for every $\varphi \in H_{0}^{1}(\Omega)$ one has

$$
\lambda \int_{\Omega} u \bar{\varphi}+\int_{\Omega} \nabla u \overline{\nabla \varphi}=\int_{\Omega} f \bar{\varphi} .
$$

Let $H:=L^{2}(\Omega)$ and define

$$
\begin{align*}
\operatorname{dom} A & :=\left\{u \in H_{0}^{1}(\Omega): \exists f \in L^{2}(\Omega) \forall \varphi \in H_{0}^{1}(\Omega):\right.  \tag{7.4}\\
& \left.\int_{\Omega} \nabla u \overline{\nabla \varphi}=-\int_{\Omega} f \bar{\varphi}\right\} \\
A u & :=f,
\end{align*}
$$

so that $A: \operatorname{dom} A \rightarrow L^{2}(\Omega)$ is a linear operator on $L^{2}(\Omega)$. By definition, $u \in \operatorname{dom} A$ and $-A u=f$ if and only if $u$ is a weak solution of (7.3) for $\lambda=0$. Moreover, a function $u \in H_{0}^{1}(\Omega)$ is a weak solution of (7.3) if and only if

$$
\begin{equation*}
u \in \operatorname{dom} A \text { and } \lambda u-A u=f \tag{7.5}
\end{equation*}
$$

In this sense, we may say that $A$ is the realization of the Laplace operator with Dirichlet boundary conditions. The problem (7.5) is a functional analytic reformulation of (7.3). Instead of solving a partial differential equation we now have to solve an algebraic equation. Clearly, the operator $A$ is linear.
Theorem 7.32. There exists an orthonormal basis $\left(e_{n}\right)$ of $L^{2}(\Omega)$ and a sequence $\left(\lambda_{n}\right) \subset \mathbb{R}_{-}$such that $\lim _{n \rightarrow \infty} \lambda_{n}=-\infty$ and for every $n \in \mathbb{N}$

$$
e_{n} \in \operatorname{dom} A \text { and } \lambda_{n} e_{n}-A e_{n}=0
$$

Moreover, $\sigma(A)=\sigma_{p}(A)=\left\{\lambda_{n}: n \in \mathbb{N}\right\}$.
Remark 7.33. Theorem 7.32 gives also a description of the spectrum of the Dirichlet-Laplace operator $A$. Every spectral value is an eigenvalue. Every eigenspace is finite dimensional and there exists an orthonormal basis consisting only of eigenvectors. For every $\lambda \notin \sigma(A)$ and every $f \in L^{2}(\Omega)$ there exists a unique weak solution $u \in H_{0}^{1}(\Omega)$ of (7.3).

Theorem 7.32 also implies that the Dirichlet-Laplace operator is unitarily equivalent to a multiplication operator on an $l^{2}$ space, that is, the Dirichlet-Laplace operator is diagonalizable.

In order to prove Theorem 7.32, we need the following theorem which will not be proved here. We only remark that in the case when $\Omega \subset \mathbb{R}$ is a bounded interval we have given a proof in Example 5.21. For a proof for general $\Omega$, see [Brézis (1992)].
Theorem 7.34 (Rellich-Kondrachov). Let $\Omega \subset \mathbb{R}^{N}$ be open and bounded. Then the embedding

$$
H_{0}^{1}(\Omega) \rightarrow L^{2}(\Omega), \quad u \mapsto u
$$

is compact.
Proof (of Theorem 7.32). Let $u, v \in \operatorname{dom} A$. Then,

$$
\begin{aligned}
\langle A u, v\rangle_{L^{2}} & =\int_{\Omega} A u \bar{v}=-\int_{\Omega} \nabla u \overline{\nabla v} \\
& =-\overline{\int_{\Omega} \nabla v \overline{\nabla u}}=\overline{\int_{\Omega} A v \bar{u}} \\
& =\overline{\langle A v, u\rangle_{L^{2}}}=\langle u, A v\rangle
\end{aligned}
$$

This equality means that $A$ is symmetric.
By Theorem 9.23 of Chapter 2, for every $f \in L^{2}(\Omega)$ there exists a unique weak solution $u \in H_{0}^{1}(\Omega)$ of (7.3) with $\lambda=1$. This means that $I-A: \operatorname{dom} A \rightarrow H$ is bijective. Let $J:=(I-A)^{-1}: H \rightarrow \operatorname{dom} A \subseteq H$ be the inverse. For every $u, v \in H$, $u=u_{1}-A u_{1}, v=v_{1}-A v_{1}$, by the symmetry of $A$,

$$
\langle J u, v\rangle=\left\langle u_{1}, v_{1}-A v_{1}\right\rangle=\left\langle u_{1}-A u_{1}, v_{1}\right\rangle=\left\langle u, v_{1}\right\rangle=\langle u, J v\rangle .
$$

Hence, $J$ is symmetric. By the Theorem of Hellinger-Toeplitz (Theorem 7.4), J : $H \rightarrow H$ is bounded, and thus also selfadjoint. Since $J$ is also a linear operator from $H$ into $H_{0}^{1}(\Omega)$, and since $J$ is closed when considered as such an operator, we obtain in fact that $J: H \rightarrow H_{0}^{1}(\Omega)$ is bounded by the closed graph theorem. Since the embedding $H_{0}^{1}(\Omega) \rightarrow L^{2}(\Omega)$ is compact by the Rellich-Kondrachov theorem, we obtain that $J \in \mathscr{K}(H)$.

By the spectral theorem for selfadjoint compact operators, there exists an orthonormal basis $\left(e_{n}\right)$ of $H=L^{2}(\Omega)$ and a sequence $\left(\mu_{n}\right) \subset \mathbb{R}$ such that $\lim _{n \rightarrow \infty} \mu_{n}=$ 0 and

$$
\mu_{n} e_{n}=J e_{n} \text { for every } n \in \mathbb{N} .
$$

Since $\operatorname{ran} J=\operatorname{dom} A$, we obtain also that $e_{n} \in \operatorname{dom} A$ for every $n \in \mathbb{N}$. Multiplying the above equation by $I-A$, we obtain

$$
\lambda_{n} e_{n}-A e_{n}=0 \text { for every } n \in \mathbb{N},
$$

with $\lambda_{n}:=\frac{\mu_{n}-1}{\mu_{n}} \in \mathbb{R}$. Since, by Theorem 9.23 of Chapter $2, \lambda-A$ is invertible for every $\lambda>0$, we obtain $\lambda_{n} \in \mathbb{R}_{-}$. Clearly, the sequence $\left(\lambda_{n}\right)$ is unbounded since $\mu_{n} \rightarrow 0$.

Now let $\lambda \notin\left\{\lambda_{n}: n \in \mathbb{N}\right\}$, and let $f \in L^{2}(\Omega)$. If $\lambda=1$ (or even $\lambda>0$ ), then we have seen above that the operator $\lambda-A: \operatorname{dom} A \rightarrow H$ is bijective. So we can assume that $\lambda \neq 1$. Then $\frac{1}{1-\lambda} \in \rho(J)$ and we can define $u:=R(1, A) R\left(\frac{1}{1-\lambda}, J\right) \frac{f}{\lambda-1}$. Clearly, $u \in \operatorname{dom} A$, and an easy calculation shows that $\lambda u-A u=f$. Moreover, every solution of $\lambda u-A u=f$ is of the form above, and thus $\lambda-A$ is bijective.

The claim is proved.
Corollary 7.35. The operator $A$ is closed and

$$
\operatorname{dom} A=\left\{u \in L^{2}(\Omega):\left(\lambda_{n}\left\langle u, e_{n}\right\rangle\right) \in \ell^{2}\right\} .
$$

Proof. If an operator $A: X \supseteq \operatorname{dom} A \rightarrow X$ on a Banach space $X$ has nonempty resolvent set, then $A$ is necessarily closed. In fact, $(\lambda-A)^{-1}$ is bounded for some $\lambda \in \rho(A) \neq \emptyset$; in particular, $(\lambda-A)^{-1}$ is closed, and thus $\lambda-A$ is closed.

Note that the Dirichlet-Laplace operator $A$ defined above has nonempty resolvent set by Theorem 7.32, and thus $A$ is closed.

The remaining claim follows easily from the fact that, by Theorem 7.32, $A$ is unitarily equivalent to the (unbounded) multiplication operator

$$
\begin{aligned}
\operatorname{dom} M & :=\left\{\left(x_{n}\right) \in l^{2}:\left(\lambda_{n} x_{n}\right) \in \ell^{2}\right\} \\
M\left(x_{n}\right) & :=\left(\lambda_{n} x_{n}\right)
\end{aligned}
$$

where the unitary operator is given by

$$
\begin{aligned}
U: L^{2}(\Omega) & \rightarrow \ell^{2}, \\
u & \mapsto\left(\left\langle u, e_{n}\right\rangle\right),
\end{aligned}
$$

that is, $A=U^{-1} M U$.

## 7.6 * The heat equation

Let $\Omega \subset \mathbb{R}^{N}$ be open and bounded, and consider the heat equation

$$
\begin{cases}u_{t}-\Delta u=0 & \text { in } \mathbb{R}_{+} \times \Omega  \tag{7.6}\\ u=0 & \text { in } \mathbb{R}_{+} \times \partial \Omega \\ u(0, x)=u_{0}(x) & \text { in } \Omega\end{cases}
$$

where $\Delta$ denotes the Laplace operator, and $u_{0} \in L^{2}(\Omega)$.
We call a function $u \in C\left(\mathbb{R}_{+} ; L^{2}(\Omega)\right)$ a mild solution of (7.6) if $u(0)=u_{0}$ and if for every $\varphi \in \operatorname{dom} A$ the function $t \mapsto\langle u(t), \varphi\rangle_{L^{2}}$ is continuously differentiable and if

$$
\frac{d}{d t}\langle u, \varphi\rangle_{L^{2}}=\langle u, A \varphi\rangle_{L^{2}} .
$$

Here, $A$ is the realization of the Dirichlet-Laplace operator on $L^{2}(\Omega)$ defined in (7.4).

Theorem 7.36. For every $u_{0} \in L^{2}(\Omega)$ there exists a unique mild solution $u$ of (7.6).
Proof. Let $A$ be the realization of the Dirichlet-Laplace operator as defined in the previous section. By Theorem 7.32, there exists an orthonormal basis $\left(e_{n}\right)$ and an unbounded sequence $\left(\lambda_{n}\right) \subset \mathbb{R}_{-}$such that for every $n \in \mathbb{N}$ one has $\lambda_{n} e_{n}=A e_{n}$.

Assume that $u$ is a mild solution of the heat equation (7.6). Then, for every $n \in \mathbb{N}$,

$$
\frac{d}{d t}\left\langle u(t), e_{n}\right\rangle_{L^{2}}=\left\langle u(t), A e_{n}\right\rangle_{L^{2}}=\lambda_{n}\left\langle u(t), e_{n}\right\rangle_{L^{2}}
$$

This implies

$$
\left\langle u(t), e_{n}\right\rangle_{L^{2}}=e^{\lambda_{n} t}\left\langle u_{0}, e_{n}\right\rangle_{L^{2}}, \quad t \geq 0 .
$$

Hence, since $\left(e_{n}\right)$ is an orthonormal basis,

$$
\begin{equation*}
u(t)=\sum_{n \in \mathbb{N}} e^{\lambda_{n} t}\left\langle u_{0}, e_{n}\right\rangle_{L^{2}} e_{n}, \quad t \geq 0 . \tag{7.7}
\end{equation*}
$$

This proves uniqueness of mild solutions.
On the other hand, let $u_{0} \in L^{2}(\Omega)$ and define $u(t)$ as in (7.7). Since $\left|e^{\lambda_{n} t}\right| \leq 1$ for every $t \geq 0$ and since $t \mapsto e^{\lambda_{n} t}$ is continuous, $u(t) \in L^{2}(\Omega)$ for every $t \geq 0$, and the function $t \mapsto u(t), \mathbb{R}_{+} \rightarrow L^{2}(\Omega)$ is continuous. Moreover, $u(0)=u_{0}$.

Let $\varphi \in \operatorname{dom} A$. By Corollary 7.35, $\left(\lambda_{n}\left\langle\varphi, e_{n}\right\rangle\right) \in \ell^{2}$. As a consequence, $t \mapsto$ $\langle u(t), \varphi\rangle_{L^{2}}$ is continuously differentiable and, by the symmetry of $A$,

$$
\begin{aligned}
\frac{d}{d t}\langle u, \varphi\rangle_{L^{2}} & =\sum_{n \in \mathbb{N}} \lambda_{n} e^{\lambda_{n} t}\left\langle u_{0}, e_{n}\right\rangle_{L^{2}}\left\langle e_{n}, \varphi\right\rangle_{L^{2}} \\
& =\sum_{n \in \mathbb{N}} e^{\lambda_{n} t}\left\langle u_{0}, e_{n}\right\rangle_{L^{2}}\left\langle A e_{n}, \varphi\right\rangle_{L^{2}} \\
& =\sum_{n \in \mathbb{N}} e^{\lambda_{n} t}\left\langle u_{0}, e_{n}\right\rangle_{L^{2}}\left\langle e_{n}, A \varphi\right\rangle_{L^{2}} \\
& =\langle u, A \varphi\rangle_{L^{2}}, \quad t \geq 0 .
\end{aligned}
$$

This proves existence of mild solutions.
Remark 7.37. The concrete form (7.7) of the solution $u$ of the heat equation (7.6) allows us to prove that in fact

$$
u \in C^{\infty}\left((0, \infty) ; L^{2}(\Omega)\right)
$$

or even

$$
u \in C^{\infty}\left((0, \infty) ; \operatorname{dom} A^{k}\right) \text { for every } k \geq 1
$$

where $\operatorname{dom} A^{k}$ is the domain of $A^{k}$ equipped with the graph norm. The heat equation thus has a regularizing effect in space and time; even if $u_{0}$ belongs 'only' to $L^{2}(\Omega)$, then $u(t)$ belongs already to $\operatorname{dom} A^{k}$ for every $k \geq 1$. Moreover, the solution is $C^{\infty}$ with values in $\operatorname{dom} A^{k}$ for every $k \geq 1$.

## 7.7 * The wave equation

Let $\Omega \subset \mathbb{R}^{N}$ be open and bounded, and consider the wave equation

$$
\begin{cases}u_{t t}-\Delta u=0 & \text { in } \mathbb{R}_{+} \times \Omega  \tag{7.8}\\ u=0 & \text { in } \mathbb{R}_{+} \times \partial \Omega \\ u(0, x)=u_{0}(x) & \text { in } \Omega \\ u_{t}(0, x)=u_{1}(x) & \text { in } \Omega\end{cases}
$$

where $\Delta$ denotes the Laplace operator, $u_{0} \in H_{0}^{1}(\Omega)$, and $u_{1} \in L^{2}(\Omega)$.
We call a function $u \in C\left(\mathbb{R}_{+} ; H_{0}^{1}(\Omega)\right) \cap C^{1}\left(\mathbb{R}_{+} ; L^{2}(\Omega)\right)$ a mild solution of (7.8) if $u(0)=u_{0}, u_{t}(0)=u_{1}$, if for every $\varphi \in H_{0}^{1}(\Omega)$ the function $t \mapsto\langle u, \varphi\rangle_{L^{2}}$ is twice continuously differentiable and if

$$
\frac{d^{2}}{d t^{2}}\langle u(t), \varphi\rangle_{L^{2}}+\int_{\Omega} \nabla u(t) \overline{\nabla \varphi}=0 .
$$

Theorem 7.38. For every $u_{0} \in H_{0}^{1}(\Omega)$ and every $u_{1} \in L^{2}(\Omega)$ there exists a unique mild solution of (7.8).

For the proof of Theorem 7.38, we need the following result which we shall not prove here; compare with Corollary 7.35.

Lemma 7.39. Let A be the Dirichlet-Laplace operator as defined in (7.4), and let $\left(e_{n}\right)$ and $\left(\lambda_{n}\right)$ be as in Theorem 7.32. Then

$$
H_{0}^{1}(\Omega)=\left\{u \in L^{2}(\Omega):\left(\sqrt{-\lambda_{n}}\left\langle u, e_{n}\right\rangle\right) \in \ell^{2}\right\} .
$$

Proof (of Theorem 7.38). Let $A$ be the realization of the Dirichlet-Laplace operator as defined in Section 7.5. By Theorem 7.32, there exists an orthonormal basis $\left(e_{n}\right)$ and an unbounded sequence $\left(\lambda_{n}\right) \subset \mathbb{R}_{-}$such that for every $n \in \mathbb{N}$ one has $\lambda_{n} e_{n}=$ $A e_{n}$.

Assume that $u$ is a mild solution of the wave equation (7.8). Then, for every $n \in \mathbb{N}$,

$$
\frac{d^{2}}{d t^{2}}\left\langle u(t), e_{n}\right\rangle_{L^{2}}=\left\langle u(t), A e_{n}\right\rangle_{L^{2}}=\lambda_{n}\left\langle u(t), e_{n}\right\rangle_{L^{2}}
$$

Setting $\alpha_{n}:=\sqrt{-\lambda_{n}}$, this implies

$$
\left\langle u(t), e_{n}\right\rangle_{L^{2}}=\cos \left(\alpha_{n} t\right)\left\langle u_{0}, e_{n}\right\rangle_{L^{2}}+\frac{1}{\alpha_{n}} \sin \left(\alpha_{n} t\right)\left\langle u_{1}, e_{n}\right\rangle_{L^{2}}, \quad t \geq 0
$$

Hence, since $\left(e_{n}\right)$ is an orthonormal basis,

$$
\begin{equation*}
u(t)=\sum_{n \in \mathbb{N}} \cos \left(\alpha_{n} t\right)\left\langle u_{0}, e_{n}\right\rangle_{L^{2}} e_{n}+\sum_{n \in \mathbb{N}} \frac{1}{\alpha_{n}} \sin \left(\alpha_{n} t\right)\left\langle u_{1}, e_{n}\right\rangle_{L^{2}} e_{n}, \quad t \geq 0 \tag{7.9}
\end{equation*}
$$

This proves uniqueness of mild solutions.
On the other hand, let $u_{0} \in H_{0}^{1}(\Omega)$ and $u_{1} \in L^{2}(\Omega)$, and define $u(t)$ as in (7.9). Since $\left|\cos \left(\alpha_{n} t\right)\right| \leq 1$ and $\left|\sin \left(\alpha_{n} t\right)\right| \leq 1$ for every $t \geq 0$ and since cos and sin are continuous, by Lemma 7.39, $u(t) \in H_{0}^{1}(\Omega)$ for every $t \geq 0$, and the function $t \mapsto u(t)$, $\mathbb{R}_{+} \rightarrow H_{0}^{1}(\Omega)$ is continuous. Moreover, $u(0)=u_{0}$. By the same reasons, $t \mapsto u(t)$, $\mathbb{R}_{+} \rightarrow L^{2}(\Omega)$ is continuously differentiable and $u_{t}(0)=u_{1}$.

Let $\varphi \in H_{0}^{1}(\Omega)$. By Lemma 7.39, $\left(\alpha_{n}\left\langle\varphi, e_{n}\right\rangle\right) \in \ell^{2}$. As a consequence, $t \mapsto$ $\langle u(t), \varphi\rangle$ is twice continuously differentiable and, by the symmetry of $A$,

$$
\begin{aligned}
\frac{d^{2}}{d t^{2}}\langle u, \varphi\rangle= & -\sum_{n \in \mathbb{N}} \lambda_{n} \cos \left(\alpha_{n} t\right)\left\langle u_{0}, e_{n}\right\rangle_{L^{2}}\left\langle e_{n}, \varphi\right\rangle_{L^{2}}- \\
& -\sum_{n \in \mathbb{N}} \alpha_{n} \sin \left(\alpha_{n} t\right)\left\langle u_{1}, e_{n}\right\rangle_{L^{2}}\left\langle e_{n}, \varphi\right\rangle_{L^{2}} \\
= & -\sum_{n \in \mathbb{N}} \cos \left(\alpha_{n} t\right)\left\langle u_{0}, e_{n}\right\rangle_{L^{2}}\left\langle A e_{n}, \varphi\right\rangle_{L^{2}}- \\
& -\sum_{n \in \mathbb{N}} \frac{1}{\alpha_{n}} \sin \left(\alpha_{n} t\right)\left\langle u_{1}, e_{n}\right\rangle_{L^{2}}\left\langle A e_{n}, \varphi\right\rangle_{L^{2}} \\
=- & -\sum_{n \in \mathbb{N}} \cos \left(\alpha_{n} t\right)\left\langle u_{0}, e_{n}\right\rangle_{L^{2}} \int_{\Omega} \nabla e_{n} \nabla \varphi- \\
& -\sum_{n \in \mathbb{N}} \frac{1}{\alpha_{n}} \sin \left(\alpha_{n} t\right)\left\langle u_{1}, e_{n}\right\rangle_{L^{2}} \int_{\Omega} \nabla e_{n} \nabla \varphi \\
= & -\int_{\Omega} \nabla u \nabla \varphi, \quad t \geq 0 .
\end{aligned}
$$

This proves existence of mild solutions.
Remark 7.40. The concrete form (7.9) of the solution $u$ of the wave equation (7.8) shows that it can be uniquely extended to a solution $u$ defined on $\mathbb{R}$. However, for the wave equation (7.8) there is no regularizing effect as for the heat equation (7.6).

## 7.8 * The Schrödinger equation

Let $\Omega \subset \mathbb{R}^{N}$ be open and bounded, and consider the Schrödinger equation

$$
\begin{cases}u_{t}-i \Delta u=0 & \text { in } \mathbb{R}_{+} \times \Omega  \tag{7.10}\\ u=0 & \text { in } \mathbb{R}_{+} \times \partial \Omega \\ u(0, x)=u_{0}(x) & \text { in } \Omega\end{cases}
$$

where $\Delta$ denotes the Laplace operator, $i=\sqrt{-1}$ is the complex unity, and $u_{0} \in$ $L^{2}(\Omega)$.

We call a function $u \in C\left(\mathbb{R}_{+} ; L^{2}(\Omega)\right)$ a mild solution of (7.10) if $u(0)=u_{0}$ and if for every $\varphi \in \operatorname{dom} A$ the function $t \mapsto\langle u, \varphi\rangle_{L^{2}}$ is continuously differentiable and if

$$
\frac{d}{d t}\langle u, \varphi\rangle_{L^{2}}=i\langle u, A \varphi\rangle_{L^{2}}, \quad t \geq 0
$$

Here, $A$ is the realization of the Dirichlet-Laplace operator on $L^{2}(\Omega)$ defined in (7.4).

Theorem 7.41. For every $u_{0} \in L^{2}(\Omega)$ there exists a unique mild solution $u$ of (7.10).

Proof. Let $A$ be the realization of the Dirichlet-Laplace operator as defined in (7.4). By Theorem 7.32, there exists an orthonormal basis $\left(e_{n}\right)$ and an unbounded sequence $\left(\lambda_{n}\right) \subset \mathbb{R}_{-}$such that for every $n \in \mathbb{N}$ one has $\lambda_{n} e_{n}=A e_{n}$.

Assume that $u$ is a mild solution of the Schrödinger equation (7.10). Then, for every $n \in \mathbb{N}$,

$$
\frac{d}{d t}\left\langle u(t), e_{n}\right\rangle_{L^{2}}=i\left\langle u(t), A e_{n}\right\rangle_{L^{2}}=i \lambda_{n}\left\langle u(t), e_{n}\right\rangle_{L^{2}}
$$

This implies

$$
\left\langle u(t), e_{n}\right\rangle_{L^{2}}=e^{i \lambda_{n} t}\left\langle u_{0}, e_{n}\right\rangle_{L^{2}}, \quad t \geq 0 .
$$

Hence, since $\left(e_{n}\right)$ is an orthonormal basis,

$$
\begin{equation*}
u(t)=\sum_{n \in \mathbb{N}} e^{i \lambda_{n} t}\left\langle u_{0}, e_{n}\right\rangle_{L^{2}} e_{n}, \quad t \geq 0 \tag{7.11}
\end{equation*}
$$

This proves uniqueness of mild solutions.
On the other hand, let $u_{0} \in L^{2}(\Omega)$ and define $u(t)$ as in (7.11). Since $\left|e^{i \lambda_{n} t}\right| \leq 1$ for every $t \geq 0$ and since $t \mapsto e^{i \lambda_{n} t}$ is continuous, $u(t) \in L^{2}(\Omega)$ for every $t \geq 0$, and the function $t \mapsto u(t), \mathbb{R}_{+} \rightarrow L^{2}(\Omega)$ is continuous. Moreover, $u(0)=u_{0}$.

Let $\varphi \in \operatorname{dom} A$. By Corollary 7.35, $\left(\lambda_{n}\left\langle\varphi, e_{n}\right\rangle\right) \in \ell^{2}$. As a consequence, $t \mapsto$ $\langle u(t), \varphi\rangle_{L^{2}}$ is continuously differentiable and, by the symmetry of $A$,

$$
\begin{aligned}
\frac{d}{d t}\langle u, \varphi\rangle_{L^{2}} & =\sum_{n \in \mathbb{N}} i \lambda_{n} e^{i \lambda_{n} t}\left\langle u_{0}, e_{n}\right\rangle_{L^{2}}\left\langle e_{n}, \varphi\right\rangle_{L^{2}} \\
& =i \sum_{n \in \mathbb{N}} e^{i \lambda_{n} t}\left\langle u_{0}, e_{n}\right\rangle_{L^{2}}\left\langle A e_{n}, \varphi\right\rangle_{L^{2}} \\
& =i \sum_{n \in \mathbb{N}} e^{i \lambda_{n} t}\left\langle u_{0}, e_{n}\right\rangle_{L^{2}}\left\langle e_{n}, A \varphi\right\rangle_{L^{2}} \\
& =i\langle u, A \varphi\rangle_{L^{2}}, \quad t \geq 0 .
\end{aligned}
$$

This proves existence of mild solutions.
Remark 7.42. The concrete form (7.11) of the solution $u$ of the Schrödinger equation (7.10) shows that it can be uniquely extended to a solution $u$ defined on $\mathbb{R}$. However, similarly as for the wave equation (7.8), there is no regularizing effect for the Schrödinger equation (7.10).

## Chapter 8 <br> Calculus on Banach spaces

### 8.1 Differentiable functions between Banach spaces

Let $X, Y$ be two Banach spaces, and let $U \subseteq X$ be open. A function $f: U \rightarrow Y$ is differentiable at $x \in U$ if there exists a bounded linear operator $T \in \mathscr{L}(X, Y)$ such that

$$
\begin{equation*}
\lim _{\|h\| \rightarrow 0} \frac{f(x+h)-f(x)-T h}{\|h\|}=0 \tag{8.1}
\end{equation*}
$$

We say that $f$ is differentiable if it is differentiable at every point $x \in U$. If $f$ is differentiable at a point $x \in U$, then $T \in \mathscr{L}(X, Y)$ is uniquely determined. We write $D f(x):=f^{\prime}(x):=T$ and call $D f(x)=f^{\prime}(x)$ the derivative of $f$ at $x$.

Lemma 8.1. If a function $f: U \rightarrow Y$ is differentiable at $x \in U$, then it is continuous at $x$. In particular, every differentiable function is continuous.

Proof. Let $\left(x_{n}\right) \subseteq U$ be convergent to $x$. By definition (equation (8.1)) and continuity of $f^{\prime}(x)$,

$$
\begin{aligned}
\left\|f\left(x_{n}\right)-f(x)\right\| & \leq\left\|f\left(x_{n}\right)-f(x)-f^{\prime}(x)\left(x-x_{n}\right)\right\|+\left\|f^{\prime}(x)\left(x-x_{n}\right)\right\| \\
& \rightarrow 0,
\end{aligned}
$$

as $n \rightarrow \infty$.
Let $X, Y$ be two Banach spaces, and let $U \subseteq X$ be open. A function $f: U \rightarrow Y$ is called continuously differentiable if it is differentiable and if $f^{\prime}: U \rightarrow \mathscr{L}(X, Y)$ is continuous. We denote by

$$
C^{1}(U ; Y):=\left\{f: U \rightarrow Y: f \text { differentiable and } f^{\prime} \in C(U ; \mathscr{L}(X, Y))\right\}
$$

the space of all continuously differentiable functions. Moreover, for $k \geq 2$, we denote by

$$
C^{k}(U ; Y):=\left\{f: U \rightarrow Y: f \text { differentiable and } f^{\prime} \in C^{k-1}(U ; \mathscr{L}(X, Y))\right\}
$$

the space of all $k$ times continuously differentiable functions.
Let $X_{i}(1 \leq i \leq n)$ and $Y$ be Banach spaces. Let $U \subseteq \bigotimes_{i=1}^{n} X_{i}$ be open. We say that a function $f: U \rightarrow Y$ is at $a=\left(a_{i}\right)_{1 \leq i \leq n} \in U$ partially differentiable with respect to the $i$-th coordinate if the function

$$
f_{i}: U_{i} \subseteq X_{i} \rightarrow Y, \quad x_{i} \mapsto f\left(a_{1}, \ldots, x_{i}, \ldots, a_{n}\right)
$$

is differentiable in $a_{i}$. We write $\frac{\partial f}{\partial x_{i}}(a):=f_{i}^{\prime}\left(a_{i}\right) \in \mathscr{L}\left(X_{i}, Y\right)$.

### 8.2 Local inverse function theorem and implicit function theorem

Let $X$ and $Y$ be two Banach spaces and let $U \subseteq X$ be an open subset. The following are two classical theorems in differential calculus.

Theorem 8.2 (Local inverse function theorem). Let $f: U \rightarrow Y$ be continuously differentiable and $\bar{x} \in U$ such that $f^{\prime}(\bar{x}): X \rightarrow Y$ is an isomorphism, that is, bounded, bijective and the inverse is also bounded. Then there exist neighbourhoods $V \subseteq U$ of $\bar{x}$ and $W \subseteq Y$ of $f(\bar{x})$ such that $f: V \rightarrow W$ is a $C^{1}$ diffeomorphism, that is $f$ is continuously differentiable, bijective and the inverse $f^{-1}: W \rightarrow V$ is continuously differentiable, too.

Theorem 8.3 (Implicit function theorem). Assume that $X=X_{1} \times X_{2}$ for two Banach spaces $X_{1}, X_{2}$, and let $f: X \supset U \rightarrow Y$ be continuously differentiable. Let $\bar{x}=\left(\bar{x}_{1}, \bar{x}_{2}\right) \in U$ be such that $\frac{\partial f}{\partial x_{2}}(\bar{x}): X_{2} \rightarrow Y$ is an isomorphism. Then there exist neighbourhoods $U_{1} \subseteq X_{1}$ of $\bar{x}_{1}$ and $U_{2} \subseteq X_{2}$ of $\bar{x}_{2}, U_{1} \times U_{2} \subseteq U$, and a continuously differentiable function $g: U_{1} \rightarrow U_{2}$ such that

$$
\left\{x \in U_{1} \times U_{2}: f(x)=f(\bar{x})\right\}=\left\{\left(x_{1}, g\left(x_{1}\right)\right): x_{1} \in U_{1}\right\} .
$$

For the proof of the local inverse theorem, we need the following lemma.
Lemma 8.4. Let $f: U \rightarrow Y$ be continuously differentiable such that $f: U \rightarrow f(U)$ is a homeomorphism, that is, continuous, bijective and with continuous inverse. Then $f$ is a $C^{1}$ diffeomorphism if and only if for every $x \in U$ the derivative $f^{\prime}(x): X \rightarrow Y$ is an isomorphism.

Proof. Assume first that $f$ is a $C^{1}$ diffeomorphism. When we differentiate the identities $x=f^{-1}(f(x))$ and $y=f\left(f^{-1}(y)\right)$, which are true for every $x \in U$ and every $y \in f(U)$, then we find

$$
\begin{aligned}
I_{X} & =\left(f^{-1}\right)^{\prime}(f(x)) f^{\prime}(x) \quad \text { for every } x \in U \text { and } \\
I_{Y} & =f^{\prime}\left(f^{-1}(y)\right)\left(f^{-1}\right)^{\prime}(y) \\
& =f^{\prime}(x)\left(f^{-1}\right)^{\prime}(f(x)) \quad \text { for every } x=f^{-1}(y) \in U .
\end{aligned}
$$

As a consequence, $f^{\prime}(x)$ is an isomorphism for every $x \in U$.
For the converse, assume that $f^{\prime}(x)$ is an isomorphism for every $x \in U$. For every $x_{1}, x_{2} \in U$ one has, by differentiability,

$$
f\left(x_{2}\right)=f\left(x_{1}\right)+f^{\prime}\left(x_{1}\right)\left(x_{2}-x_{1}\right)+o\left(x_{2}-x_{1}\right),
$$

where $o$ depends on $x_{1}$ and $\lim _{x_{2} \rightarrow x_{1}} \frac{o\left(x_{2}-x_{1}\right)}{\left\|x_{2}-x_{1}\right\|}=0$. We have $x_{1}=f^{-1}\left(y_{1}\right)$ and $x_{2}=$ $f^{-1}\left(y_{2}\right)$ if we put $y_{i}:=f\left(x_{i}\right)$. Hence, the above identity becomes

$$
y_{2}=y_{1}+f^{\prime}\left(f^{-1}\left(y_{1}\right)\right)\left(f^{-1}\left(y_{2}\right)-f^{-1}\left(y_{1}\right)\right)+o\left(f^{-1}\left(y_{2}\right)-f^{-1}\left(y_{1}\right)\right) .
$$

To this identity, we apply the inverse operator $\left(f^{\prime}\left(f^{-1}\left(y_{1}\right)\right)\right)^{-1}$ and we obtain
$f^{-1}\left(y_{2}\right)=f^{-1}\left(y_{1}\right)+\left(f^{\prime}\left(f^{-1}\left(y_{1}\right)\right)\right)^{-1}\left(y_{2}-y_{1}\right)-\left(f^{\prime}\left(f^{-1}\left(y_{1}\right)\right)\right)^{-1} o\left(f^{-1}\left(y_{2}\right)-f^{-1}\left(y_{1}\right)\right)$.
Since $f^{-1}$ is continuous, the last term on the right-hand side of the last equality is sublinear. Hence, $f^{-1}$ is differentiable and

$$
\left(f^{-1}\right)^{\prime}\left(y_{1}\right)=\left(f^{\prime}\left(f^{-1}\left(y_{1}\right)\right)\right)^{-1} .
$$

From this identity (using that $f^{-1}$ and $f^{\prime}$ are continuous) we obtain that $f^{-1}$ is continuously differentiable. The claim is proved.

Proof (Proof of the local inverse function theorem). Consider the function

$$
\begin{aligned}
g: U & \rightarrow X, \\
x & \mapsto f^{\prime}(\bar{x})^{-1} f(x) .
\end{aligned}
$$

It suffices to show that $g: V \rightarrow W$ is a $C^{1}$ diffeomorphism for appropriate neighbourhoods $V$ of $\bar{x}$ and $W$ of $g(\bar{x})$.

Consider also the function

$$
\begin{aligned}
\varphi: U & \rightarrow X, \\
x & \mapsto x-g(x) .
\end{aligned}
$$

This function $\varphi$ is continuously differentiable and $\varphi^{\prime}(x)=I-f^{\prime}(\bar{x})^{-1} f^{\prime}(x)$ for every $x \in U$. In particular, $\varphi^{\prime}(\bar{x})=0$. By continuity of $\varphi^{\prime}$, there exists $r>0$ and $L<1$ such that $\left\|\varphi^{\prime}(x)\right\| \leq L$ for every $x \in \bar{B}(\bar{x}, r) \subseteq U$. Hence,

$$
\left\|\varphi\left(x_{1}\right)-\varphi\left(x_{2}\right)\right\| \leq L\left\|x_{1}-x_{2}\right\| \quad \text { for every } x_{1}, x_{2} \in \bar{B}(\bar{x}, r)
$$

By the definition of $\varphi$, this implies

$$
\begin{align*}
\left\|g\left(x_{1}\right)-g\left(x_{2}\right)\right\| & =\left\|x_{1}-x_{2}-\left(\varphi\left(x_{1}\right)-\varphi\left(x_{2}\right)\right)\right\|  \tag{8.2}\\
& \geq\left\|x_{1}-x_{2}\right\|-L\left\|x_{1}-x_{2}\right\| \\
& =(1-L)\left\|x_{1}-x_{2}\right\| .
\end{align*}
$$

We claim that for every $y \in \bar{B}(g(\bar{x}),(1-L) r)$ there exists a unique $x \in \bar{B}(\bar{x}, r)$ such that $g(x)=y$.

The uniqueness follows from (8.2).
In order to prove existence, let $x_{0}=\bar{x}$, and then define recursively $x_{n+1}=y+$ $\varphi\left(x_{n}\right)=y+x_{n}-f^{\prime}(\bar{x})^{-1} f\left(x_{n}\right)$ for every $n \geq 0$. Then

$$
\begin{aligned}
\left\|x_{n}-\bar{x}\right\| & =\left\|\sum_{k=0}^{n-1} x_{k+1}-x_{k}\right\| \\
& \leq\left\|x_{1}-x_{0}\right\|+\sum_{k=1}^{n-1}\left\|\varphi\left(x_{k}\right)-\varphi\left(x_{k-1}\right)\right\| \\
& \leq \sum_{k=0}^{n-1} L^{k}\left\|x_{1}-x_{0}\right\| \\
& =\frac{1-L^{n}}{1-L}\|y-g(\bar{x})\| \\
& \leq\left(1-L^{n}\right) r \leq r
\end{aligned}
$$

which implies $x_{n} \in \bar{B}(\bar{x}, r)$ for every $n \geq 0$. Similarly, for every $n \geq m \geq 0$,

$$
\left\|x_{n}-x_{m}\right\| \leq \sum_{k=m}^{n-1} L^{k}\|y-g(\bar{x})\|,
$$

so that the sequence $\left(x_{n}\right)$ is a Cauchy sequence in $\bar{B}(\bar{x}, r)$. Since $\bar{B}(\bar{x}, r)$ is complete, there exists $\lim _{n \rightarrow \infty} x_{n}=: x \in \bar{B}(\bar{x}, r)$. By continuity,

$$
x=y+\varphi(x)=y+x-g(x),
$$

or

$$
g(x)=y .
$$

This proves the above claim, that is, $g$ is locally invertible. It remains to show that $g^{-1}$ is continuous (then $g$ is a homeomorphism, and therefore a $C^{1}$ diffeomorphism by Lemma 8.4). Contiunity of the inverse function, however, is a direct consequence of (8.2) (which even implies Lipschitz continuity).

Remark 8.5. The iteration formula

$$
x_{n+1}=y+x_{n}-f^{\prime}(\bar{x})^{-1} f\left(x_{n}\right)
$$

used in the proof of the local inverse theorem in order to find a solution of $g(x)=$ $f^{\prime}(\bar{x})^{-1} f(x)=y$ should be compared to the discrete Newton iteration

$$
x_{n+1}=y+x_{n}-f^{\prime}\left(x_{n}\right)^{-1} f\left(x_{n}\right) ;
$$

see Theorem 8.8 below.
Proof (Proof of the implicit function theorem). Consider the function

$$
\begin{aligned}
F: U & \rightarrow X_{1} \times Y, \\
\left(x_{1}, x_{2}\right) & \mapsto\left(x_{1}, f\left(x_{1}, x_{2}\right)\right) .
\end{aligned}
$$

Then $F$ is continuously differentiable and

$$
F^{\prime}(\bar{x})\left(h_{1}, h_{2}\right)=\left(h_{1}, \frac{\partial f}{\partial x_{1}}(\bar{x}) h_{1}+\frac{\partial f}{\partial x_{2}}(\bar{x}) h_{2}\right) .
$$

In particular, by the assumption, $F^{\prime}(\bar{x})$ is locally invertible with inverse

$$
F^{\prime}(\bar{x})^{-1}\left(y_{1}, y_{2}\right)=\left(y_{1},\left(\frac{\partial f}{\partial x_{2}}(\bar{x})\right)^{-1}\left(y_{2}-\frac{\partial f}{\partial x_{1}}(\bar{x}) y_{1}\right)\right)
$$

By the local inverse theorem (Theorem 8.2), there exists a neighbourhood $U_{1}$ of $\bar{x}_{1}$, a neighbourhood $U_{2}$ of $\bar{x}_{2}$ and a neighbourhood $V$ of $\left(\bar{x}_{1}, f(\bar{x})\right)=F(\bar{x})$ such that $F: U_{1} \times U_{2} \rightarrow V$ is a $C^{1}$ diffeomorphism. The inverse is of the form

$$
F^{-1}\left(y_{1}, y_{2}\right)=\left(y_{1}, h_{2}\left(y_{1}, y_{2}\right)\right)
$$

where $h_{2}$ is a function such that $f\left(y_{1}, h_{2}\left(y_{1}, y_{2}\right)\right)=y_{2}$. Let

$$
\tilde{U}_{1}:=\left\{x_{1} \in U_{1}:\left(x_{1}, f(\bar{x})\right) \in V\right\} .
$$

Then $\tilde{U}_{1}$ is open by continuity of the function $x_{1} \mapsto\left(x_{1}, f(\bar{x})\right)$, and $\bar{x}_{1} \in \tilde{U}_{1}$. We restrict $F$ to $\tilde{U}_{1} \times U_{2}$, and we define

$$
\begin{align*}
g: \tilde{U}_{1} & \rightarrow X_{2}  \tag{8.3}\\
& x_{1} \mapsto g\left(x_{1}\right)=F^{-1}\left(x_{1}, f(\bar{x})\right)_{2},
\end{align*}
$$

where $F^{-1}(\cdot)_{2}$ denotes the second component of $F^{-1}(\cdot)$. Then $g$ is continuously differentiable, $g\left(\tilde{U}_{1}\right) \subseteq U_{2}$ and $g$ satisfies the required property of the implicit function.

Lemma 8.6 (Higher regularity of the local inverse). Let $f \in C^{k}(U ; Y)$ for some $k \geq 1$ and assum that $f: U \rightarrow f(U)$ is a $C^{1}$ diffeomorphism. Then $f$ is a $C^{k}$ diffeomorphism, that is, $f^{-1}$ is $k$ times continuously differentiable.

Proof. For every $y \in f(U)$ we have

$$
\left(f^{-1}\right)^{\prime}(y)=f^{\prime}\left(f^{-1}(y)\right)^{-1}
$$

The proof therefore follows by induction on $k$.
Lemma 8.7 (Higher regularity of the implicit function). If, in the implicit function theorem (Theorem 8.3), the function $f$ is $k$ times continuously differentiable, then the implicit function $g$ is also $k$ times continuously differentiable.

Proof. This follows from the previous lemma (Lemma 8.6) and the definition of the implicit function in the proof of the implicit function theorem.

## 8.3 * Newton's method

Theorem 8.8 (Newton's method). Let $X$ and $Y$ be two Banach spaces, $U \subseteq X$ an open set. Let $f \in C^{1}(U ; Y)$ and assume that there exists $\bar{x} \in U$ such that (i) $f(\bar{x})=0$ and (ii) $f^{\prime}(\bar{x}) \in \mathscr{L}(X, Y)$ is an isomorphism. Then there exists a neighbourhood $V \subseteq U$ of $\bar{x}$ such that for every $x_{0} \in V$ the operator $f^{\prime}\left(x_{0}\right)$ is an isomorphism, the sequence $\left(x_{n}\right)$ defined iteratively by

$$
\begin{equation*}
x_{n+1}=x_{n}-f^{\prime}\left(x_{n}\right)^{-1} f\left(x_{n}\right), \quad n \geq 0 \tag{8.4}
\end{equation*}
$$

remains in $V$ and $\lim _{n \rightarrow \infty} x_{n}=\bar{x}$.
Proof. By Corollary 1.35 and continuity, there exists a neighbourhood $\tilde{V} \subseteq U$ of $\bar{x}$ such that $f^{\prime}(x)$ is isomorphic for all $x \in \tilde{V}$. Next, it will be useful to define the auxiliary function $\varphi: \tilde{V} \rightarrow X$ by

$$
\varphi(x):=x-f^{\prime}(x)^{-1} f(x), \quad x \in \tilde{V}
$$

Since $f(\bar{x})=0$, we find that for every $x \in \tilde{V}$

$$
\begin{aligned}
\varphi(x)-\varphi(\bar{x}) & =x-f^{\prime}(x)^{-1}(f(x)-f(\bar{x}))-\bar{x} \\
& =x-\bar{x}-f^{\prime}(x)^{-1}\left(f^{\prime}(\bar{x})(x-\bar{x})+r(x-\bar{x})\right)
\end{aligned}
$$

so that by the continuity of $f^{\prime}(\cdot)^{-1}$

$$
\lim _{x \rightarrow \bar{x}} \frac{\|\varphi(x)-\varphi(\bar{x})\|}{\|x-\bar{x}\|}=0
$$

Hence, there exists $r>0$ such that $V:=B(\bar{x}, r) \subseteq \tilde{V} \subseteq U$ and such that for every $x \in V$

$$
\|\varphi(x)-\bar{x}\|=\|\varphi(x)-\varphi(\bar{x})\| \leq \frac{1}{2}\|x-\bar{x}\|
$$

This implies that for every $x_{0} \in V$ one has $\varphi\left(x_{0}\right) \in V$ and if we define iteratively $x_{n+1}=\varphi\left(x_{n}\right)=\varphi^{n+1}\left(x_{0}\right)$, then

$$
\left\|x_{n}-\bar{x}\right\| \leq\left(\frac{1}{2}\right)^{n}\left\|x_{0}-\bar{x}\right\| \rightarrow 0 \text { as } n \rightarrow \infty
$$

## Chapter 9 <br> Sobolev spaces

### 9.1 Test functions, convolution and regularization

Let $\Omega \subseteq \mathbb{R}^{d}$ be an open set. For every continuous function $\varphi \in C(\Omega)$ we define the support

$$
\operatorname{supp} \varphi:=\overline{\{x \in \Omega: \varphi(x) \neq 0\}},
$$

where the closure is to be understood in $\mathbb{R}^{d}$. Thus, the support is by definition always closed in $\mathbb{R}^{d}$, but it is not necessarily a subset of $\Omega$. Next we let

$$
\mathscr{D}(\Omega):=C_{c}^{\infty}(\Omega):=\left\{\varphi \in C^{\infty}(\Omega): \operatorname{supp} \varphi \subseteq \Omega \text { is compact }\right\}
$$

be the space of test functions on $\Omega$, and

$$
L_{l o c}^{1}(\Omega):=\left\{f: \Omega \rightarrow \mathbb{K} \text { measurable }: \int_{K}|f|<\infty \forall K \subseteq \Omega \text { compact }\right\}
$$

the space of locally integrable functions on $\Omega$. For every $f \in L_{l o c}^{1}\left(\mathbb{R}^{d}\right)$ and every $\varphi \in \mathscr{D}\left(\mathbb{R}^{d}\right)$ we define the convolution $f * \varphi$ by

$$
\begin{aligned}
f * \varphi(x) & :=\int_{\mathbb{R}^{d}} f(x-y) \varphi(y) \mathrm{d} y \\
& =\int_{\mathbb{R}^{d}} f(y) \varphi(x-y) \mathrm{d} y
\end{aligned}
$$

Lemma 9.1. For every $f \in L_{\text {loc }}^{1}\left(\mathbb{R}^{d}\right)$ and every $\varphi \in \mathscr{D}\left(\mathbb{R}^{d}\right)$ one has $f * \varphi \in C^{\infty}\left(\mathbb{R}^{d}\right)$ and for every $1 \leq i \leq d$,

$$
\frac{\partial}{\partial x_{i}}(f * \varphi)=f * \frac{\partial \varphi}{\partial x_{i}} .
$$

Proof. Let $e_{i} \in \mathbb{R}^{d}$ be the $i$-th unit vector. Then

$$
\lim _{h \rightarrow 0} \frac{1}{h}\left(\varphi\left(x+h e_{i}\right)-\varphi(x)\right)=\frac{\partial \varphi}{\partial x_{i}}(x)
$$

uniformly in $x \in \mathbb{R}^{d}$ (note that $\varphi$ has compact support). Hence, for every $x \in \mathbb{R}^{d}$

$$
\begin{aligned}
& \frac{1}{h}\left(f * \varphi\left(x+h e_{i}\right)-f * \varphi(x)\right) \\
& =\frac{1}{h} \int_{\mathbb{R}^{d}} f(y)\left(\varphi\left(x+h e_{i}-y\right)-\varphi(x-y)\right) \mathrm{d} y \\
& \rightarrow \int_{\mathbb{R}^{d}} f(y) \frac{\partial \varphi}{\partial x_{i}}(x-y) \mathrm{d} y
\end{aligned}
$$

The following theorem is proved in courses on measure theory. We omit the proof.

Theorem 9.2 (Young's inequality). Let $f \in L^{p}\left(\mathbb{R}^{d}\right)$ and $\varphi \in \mathscr{D}\left(\mathbb{R}^{d}\right)$. Then $f * \varphi \in$ $L^{p}\left(\mathbb{R}^{d}\right)$ and

$$
\|f * \varphi\|_{p} \leq\|f\|_{p}\|\varphi\|_{1}
$$

Theorem 9.3. For every $1 \leq p<\infty$ and every open $\Omega \subseteq \mathbb{R}^{d}$ the space $\mathscr{D}(\Omega)$ is dense in $L^{p}(\Omega)$.

Proof. The technique of this proof (regularization and truncation) is important in the theory of partial differential equations, distributions and Sobolev spaces. The first step (regularization) is based on Lemma 9.1. The truncation step is in this case relatively easy.

Regularization. Let $\varphi \in \mathscr{D}\left(\mathbb{R}^{d}\right)$ be a positive function such that $\|\varphi\|_{1}=\int_{\mathbb{R}^{d}} \varphi=$ 1. One may take for example the function

$$
\varphi(x):= \begin{cases}c e^{1 /\left(1-|x|^{2}\right)} & \text { if }|x|<1  \tag{9.1}\\ 0 & \text { otherwise }\end{cases}
$$

with an appropriate constant $c>0$. Then let $\varphi_{n}(x):=n^{d} \varphi(n x)$, so that $\left\|\varphi_{n}\right\|_{1}=$ $\int_{\mathbb{R}^{d}} \varphi_{n}=1$ for every $n \in \mathbb{N}$.

Let $f \in L^{p}\left(\mathbb{R}^{d}\right)$. By Lemma 9.1 and Young's inequality (Theorem 9.2), for every $n \in \mathbb{N}, f_{n}:=f * \varphi_{n} \in C^{\infty}\left(\mathbb{R}^{d}\right) \cap L^{p}\left(\mathbb{R}^{d}\right)$ and $\left\|f_{n}\right\|_{p} \leq\|f\|_{p}$. Hence, for every $n \in \mathbb{N}$ the operator $T_{n}: L^{p}\left(\mathbb{R}^{d}\right) \rightarrow L^{p}\left(\mathbb{R}^{d}\right), f \mapsto f * \varphi_{n}$ is linear and bounded and $\left\|T_{n}\right\| \leq 1$. Moreover, if $f=1_{I}$ for some bounded interval $I=\left(a_{1}, b_{1}\right) \times \cdots \times\left(a_{d}, b_{d}\right) \subseteq \Omega$, then

$$
\begin{aligned}
\left\|f_{n}-f\right\|_{p}^{p} & =\int_{\mathbb{R}^{d}}\left|\int_{\mathbb{R}^{d}} f(x-y) \varphi(n y) n^{d} \mathrm{~d} y-f(x)\right|^{p} \mathrm{~d} x \\
& =\int_{\mathbb{R}^{d}}\left|\int_{\mathbb{R}^{d}}\left(f\left(x-\frac{y}{n}\right)-f(x)\right) \varphi(y) \mathrm{d} y\right|^{p} \mathrm{~d} x \\
& \leq \int_{\mathbb{R}^{d}}\left(\int_{\mathbb{R}^{d}}\left|f\left(x-\frac{y}{n}\right)-f(x)\right| \varphi(y) \mathrm{d} y\right)^{p} \mathrm{~d} x \rightarrow 0
\end{aligned}
$$

as $n \rightarrow \infty$ by Lebesgue's dominated convergence theorem. In other words, $\lim _{n \rightarrow \infty}\left\|T_{n} f-f\right\|_{p}=0$ for every $f=1_{I}$ with $I$ as above. Since $\operatorname{span}\left\{1_{I}: I \subseteq \mathbb{R}^{d}\right.$
bounded interval $\}$ is dense in $L^{p}\left(\mathbb{R}^{d}\right)$, we find that $\lim _{n \rightarrow \infty}\left\|T_{n} f-f\right\|_{p}=0$ for every $f$ from a dense subset $M$ of $L^{p}\left(\mathbb{R}^{d}\right)$. Since the $T_{n}$ are bounded, we conclude from Lemma 2.48 that $T_{n} f \rightarrow f$ in $L^{p}\left(\mathbb{R}^{d}\right)$ for every $f \in L^{p}\left(\mathbb{R}^{d}\right)$. This proves that $L^{p} \cap C^{\infty}\left(\mathbb{R}^{d}\right)$ is dense in $L^{p}\left(\mathbb{R}^{d}\right)$.

Truncation. Now we consider a general open set $\Omega \subseteq \mathbb{R}^{d}$ and prove the claim. Let $\varphi \in \mathscr{D}\left(\mathbb{R}^{d}\right)$ be a positive test function such that $\operatorname{supp} \varphi \subseteq \overline{B(0,1)}$ and $\int_{\mathbb{R}^{d}} \varphi=1$ (one may take for example the function from (9.1)). Then let $\varphi_{n}(x):=n^{d} \varphi(n x)$.

For every $n \in \mathbb{N}$ we let

$$
K_{n}:=\left\{x \in \Omega: \operatorname{dist}(x, \partial \Omega) \geq \frac{1}{n}\right\} \cap \overline{B(0, n)},
$$

so that $K_{n} \subseteq \Omega$ is compact for every $n \in \mathbb{N}$.
Now let $f \in L^{p}(\Omega) \subseteq L^{p}\left(\mathbb{R}^{d}\right)$ and $\varepsilon>0$. Let

$$
f 1_{K_{n}}(x)= \begin{cases}f(x) & \text { if } x \in K_{n} \\ 0 & \text { if } x \in \Omega \backslash K_{n}\end{cases}
$$

By Lebesgue's dominated convergence theorem (since $\bigcup_{n} K_{n}=\Omega$ ),

$$
\left\|f-f 1_{K_{n}}\right\|_{p}^{p}=\int_{\Omega}|f|^{p}\left(1-1_{K_{n}}\right)^{p} \rightarrow 0 \text { as } n \rightarrow \infty
$$

In particular, there exists $n \in \mathbb{N}$ such that $\left\|f-f 1_{K_{n}}\right\|_{p} \leq \varepsilon$.
For every $m \geq 4 n$ we define $g_{m}:=\left(f 1_{K_{n}}\right) * \varphi_{m} \in L^{p} \cap C^{\infty}\left(\mathbb{R}^{d}\right)$; note that we here consider $L^{p}(\Omega)$ as a subspace of $L^{p}\left(\mathbb{R}^{d}\right)$ by extending functions in $L^{p}(\Omega)$ by 0 outside $\Omega$. However, since $g_{m}=0$ outside $K_{2 n}$, we find that actually $g_{m} \in \mathscr{D}(\Omega)$. By the first step (regularisation), there exists $m \geq 4 n$ so large that $\left\|g_{m}-f 1_{K_{n}}\right\|_{p} \leq \varepsilon$. For such $m$ we have $\left\|f-g_{m}\right\|_{p} \leq 2 \varepsilon$, and the claim is proved.
Lemma 9.4. Let $f \in L_{l o c}^{1}(\Omega)$ be such that

$$
\int_{\Omega} f \varphi=0 \quad \text { for every } \varphi \in \mathscr{D}(\Omega)
$$

Then $f=0$.
Proof. We first assume that $f \in L^{1}(\Omega)$ is real and that $\Omega$ has finite measure. By Theorem 9.3, for every $\varepsilon>0$ there exists $g \in \mathscr{D}(\Omega)$ such that $\|f-g\|_{1} \leq \varepsilon$. By assumption, this implies

$$
\left|\int_{\Omega} g \varphi\right|=\left|\int_{\Omega}(f-g) \varphi\right| \leq \varepsilon\|\varphi\|_{\infty} \quad \forall \varphi \in \mathscr{D}(\Omega) .
$$

Let $K_{1}:=\{x \in \Omega: g(x) \geq \varepsilon\}$ and $K_{2}:=\{x \in \Omega: g(x) \leq-\varepsilon\}$. Since $g$ is a test function, the sets $K_{1}, K_{2}$ are compact. Since they are disjoint and do not touch the boundary of $\Omega$,

$$
\inf \left\{|x-y|,|x-z|,|y-z|: x \in K_{1}, y \in K_{2}, z \in \partial \Omega\right\}=: \delta>0 .
$$

Let $K_{i}^{\delta}:=\left\{x \in \Omega: \operatorname{dist}\left(x, K_{i}\right) \leq \delta / 4\right\}(i=1,2)$. Then $K_{1}^{\delta}$ and $K_{2}^{\delta}$ are two compact disjoint subsets of $\Omega$. Let

$$
h(x):= \begin{cases}1 & \text { if } x \in K_{1}^{\delta} \\ -1 & \text { if } x \in K_{2}^{\delta} \\ 0 & \text { else }\end{cases}
$$

choose a positive test function $\varphi \in \mathscr{D}\left(\mathbb{R}^{d}\right)$ such that $\int_{\mathbb{R}^{d}} \varphi=1$ and $\operatorname{supp} \varphi \subseteq$ $B(0, \delta / 8)$, and let $\psi:=h * \varphi$. Then $\psi \in \mathscr{D}(\Omega),-1 \leq \psi \leq 1, \psi=1$ on $K_{1}$ and $\psi=-1$ on $K_{2}$. Let $K:=K_{1} \cup K_{2}$. Then

$$
\int_{K}|g|=\int_{K} g \psi \leq \varepsilon+\int_{\Omega \backslash K}|g \psi| \leq \varepsilon+\int_{\Omega \backslash K}|g| .
$$

Hence,

$$
\int_{\Omega}|g|=\int_{K}|g|+\int_{\Omega \backslash K}|g| \leq \varepsilon+2 \int_{\Omega \backslash K}|g| \leq \varepsilon(1+2|\Omega|),
$$

which implies

$$
\int_{\Omega}|f| \leq \int_{\Omega}|f-g|+\int_{\Omega}|g| \leq 2 \varepsilon(1+|\Omega|)
$$

Since $\varepsilon>0$ was arbitrary, we find that $f=0$.
The general case can be obtained from the particular case ( $f \in L^{1}$ and $|\Omega|<\infty$ ) by considering first real and imaginary part of $f$ separately, and then by considering $f 1_{B}$ for all closed (compact) balls $B \subseteq \Omega$.

### 9.2 Sobolev spaces in one dimension

Recall the fundamental rule of partial integration: if $f, g \in C^{1}([a, b])$ on some compact interval $[a, b]$, then

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

In particular, for every $f \in C^{1}([a, b])$ and every $\varphi \in \mathscr{D}(a, b)$

$$
\begin{equation*}
\int_{a}^{b} f \varphi^{\prime}=-\int_{a}^{b} f^{\prime} \varphi \tag{9.2}
\end{equation*}
$$

since $\varphi(a)=\varphi(b)=0$.
Let $-\infty \leq a<b \leq \infty$ and $1 \leq p \leq \infty$. We define

$$
W^{1, p}(a, b):=\left\{u \in L^{p}(a, b): \exists g \in L^{p}(a, b) \forall \varphi \in \mathscr{D}(a, b): \int_{a}^{b} u \varphi^{\prime}=-\int_{a}^{b} g \varphi\right\}
$$

The space $W^{1, p}(a, b)$ is called (first) Sobolev space. If $p=2$, then we also write $H^{1}(a, b):=W^{1,2}(a, b)$.

By Lemma 9.4, the function $g \in L^{p}(a, b)$ is uniquely determined if it exists. In the following, we will write $u^{\prime}:=g$, in accordance with (9.2). We equip $W^{1, p}(a, b)$ with the norm

$$
\|u\|_{W^{1, p}}:=\|u\|_{p}+\left\|u^{\prime}\right\|_{p}
$$

and if $p=2$, then we define the inner product

$$
\langle u, v\rangle_{H^{1}}:=\int_{a}^{b} u v+\int_{a}^{b} u^{\prime} v^{\prime},
$$

which actually yields the norm $\|u\|_{H^{1}}=\left(\|u\|_{2}^{2}+\left\|u^{\prime}\right\|_{2}^{2}\right)^{\frac{1}{2}}$ (which is equivalent to $\left.\|\cdot\|_{W^{1,2}}\right)$.
Lemma 9.5. The Sobolev spaces $W^{1, p}(a, b)$ are Banach spaces, which are separable if $p \neq \infty$. The space $H^{1}(a, b)$ is a separable Hilbert space.

Proof. The fact that the $W^{1, p}$ are Banach spaces, or that $H^{1}$ is a Hilbert space, is an exercise. Recall that $L^{p}(a, b)$ is separable (Remark 2.37). Hence, the product space $L^{p}(a, b) \times L^{p}(a, b)$ is separable, and also every subspace of this product space is separable. Now consider the linear mapping

$$
T: W^{1, p}(a, b) \rightarrow L^{p}(a, b) \times L^{p}(a, b), \quad u \mapsto\left(u, u^{\prime}\right)
$$

which is bounded and even isometric. Hence, $W^{1, p}$ is isometrically isomomorphic to a subspace of $L^{p} \times L^{p}$ which is separable. Hence $W^{1, p}$ is separable.
Lemma 9.6. Let $u \in W^{1, p}(a, b)$ be such that $u^{\prime}=0$. Then $u$ is constant.
Proof. Choose $\psi \in \mathscr{D}(a, b)$ such that $\int_{a}^{b} \psi=1$. Then, for every $\varphi \in \mathscr{D}(a, b)$, the function $\varphi-\left(\int_{a}^{b} \varphi\right) \psi$ is the derivative of a test function since $\int_{a}^{b}\left(\varphi-\left(\int_{a}^{b} \varphi\right) \psi\right)=0$. Hence, by definition,

$$
0=\int_{a}^{b} u\left(\varphi-\left(\int_{a}^{b} \varphi\right) \psi\right),
$$

or, with $c=\int_{a}^{b} u \psi=$ const,

$$
\int_{a}^{b}(u-c) \varphi=0 \quad \forall \varphi \in \mathscr{D}(a, b) .
$$

By Lemma 9.4, $u=c$ almost everywhere.
Lemma 9.7. Let $-\infty<a<b<\infty$ and let $t_{0} \in[a, b]$. Let $g \in L^{p}(a, b)$ and define

$$
u(t):=\int_{t_{0}}^{t} g(s) \mathrm{d} s, \quad t \in[a, b] .
$$

Then $u \in W^{1, p}(a, b)$ and $u^{\prime}=g$.

Proof. Let $\varphi \in \mathscr{D}(a, b)$. Then, by Fubini's theorem,

$$
\begin{aligned}
\int_{a}^{b} u \varphi^{\prime} & =\int_{a}^{b} \int_{t_{0}}^{t} g(s) \mathrm{d} s \varphi^{\prime}(t) \mathrm{d} t \\
& =\int_{a}^{t_{0}} \int_{t_{0}}^{t} g(s) \mathrm{d} s \varphi^{\prime}(t) \mathrm{d} t+\int_{t_{0}}^{b} \int_{t_{0}}^{t} g(s) \mathrm{d} s \varphi^{\prime}(t) \mathrm{d} t \\
& =-\int_{a}^{t_{0}} \int_{a}^{s} \varphi^{\prime}(t) \mathrm{d} t g(s) \mathrm{d} s+\int_{t_{0}}^{b} \int_{s}^{b} \varphi^{\prime}(t) \mathrm{d} t g(s) \mathrm{d} s \\
& =-\int_{a}^{t_{0}} \varphi(s) g(s) \mathrm{d} s-\int_{t_{0}}^{b} \varphi(s) g(s) \mathrm{d} s \\
& =-\int_{a}^{b} g \varphi
\end{aligned}
$$

Theorem 9.8. Let $u \in W^{1, p}(a, b)$ (bounded or unbounded interval). Then there exists $\tilde{u} \in C((a, b))$ which is continuous up to the boundary of $(a, b)$, which coincides with $u$ almost everywhere and such that for every $s, t \in(a, b)$

$$
\tilde{u}(t)-\tilde{u}(s)=\int_{s}^{t} u^{\prime}(r) \mathrm{d} r .
$$

Proof. Fix $t_{0} \in(a, b)$ and define $v(t):=\int_{t_{0}}^{t} u^{\prime}(s) \mathrm{d} s(t \in \overline{(a, b)})$. Clearly, the function $v$ is continuous. By Lemma 9.7, $v \in W^{1, p}(c, d)$ for every bounded interval $(c, d) \subseteq$ $(a, b)$, and $v^{\prime}=u^{\prime}$. By Lemma 9.6, $u-v=C$ for some constant $C$ which clearly does not depend on the choice of the interval $(c, d)$. This proves that $u$ coincides almost everywhere with the continuous function $\tilde{u}=v+C$. By Lemma 9.7,

$$
\tilde{u}(t)-\tilde{u}(s)=v(t)-v(s)=\int_{s}^{t} u^{\prime}(r) \mathrm{d} r
$$

Remark 9.9. By Theorem 9.8, we will identify every function $u \in W^{1, p}(a, b)$ with its continuous representant, and we say that every function in $W^{1, p}(a, b)$ is continuous.

Lemma 9.10 (Extension lemma). Let $u \in W^{1, p}(a, b)$. Then there exists $\tilde{u} \in W^{1, p}(\mathbb{R})$ such that $\tilde{u}=u$ on $(a, b)$.

Proof. Assume first that $a$ and $b$ are finite and define

$$
g(t):= \begin{cases}u^{\prime}(t) & \text { if } t \in[a, b] \\ u(a) & \text { if } t \in[a-1, a) \\ -u(b) & \text { if } t \in(b, b+1] \\ 0 & \text { else }\end{cases}
$$

Then $g \in L^{p}(\mathbb{R})$. Let $\tilde{u}(t):=\int_{-\infty}^{t} g(s) \mathrm{d} s$, so that $\tilde{u}=u$ on ( $a, b$ ). By Lemma 9.7, $\tilde{u} \in W^{1, p}(c, d)$ for every bounded interval $(c, d) \in \mathbb{R}$. However, $\tilde{u}=0$ outside $(a-$ $1, b+1)$ which implies that $\tilde{u} \in W^{1, p}(\mathbb{R})$.

The case of $a=-\infty$ or $b=\infty$ is treated similarly.
Lemma 9.11. For every $1 \leq p<\infty$, the space $\mathscr{D}(\mathbb{R})$ is dense in $W^{1, p}(\mathbb{R})$.
Proof. Let $u \in W^{1, p}(\mathbb{R})$.
Regularization: Choose a positive test function $\varphi \in \mathscr{D}(\mathbb{R})$ such that $\int_{\mathbb{R}} \varphi=1$ and put $\varphi_{n}(x)=n \varphi(n x)$. Then $u_{n}:=u * \varphi_{n} \in C^{\infty} \cap L^{p}(\mathbb{R}), u_{n}^{\prime}=u^{\prime} * \varphi_{n} \in L^{p}(\mathbb{R})$ and

$$
\begin{aligned}
& \lim _{n \rightarrow \infty}\left\|u-u_{n}\right\|_{p}=0 \text { and } \\
& \lim _{n \rightarrow \infty}\left\|u^{\prime}-u_{n}^{\prime}\right\|_{p}=0,
\end{aligned}
$$

so that $\lim _{n \rightarrow \infty}\left\|u-u_{n}\right\|_{W^{1, p}}=0$. This proves that $W^{1, p}(\mathbb{R}) \cap C^{\infty}(\mathbb{R})$ is dense in $W^{1, p}(\mathbb{R})$.

Truncation: Choose a sequence $\left(\psi_{n}\right) \subseteq \mathscr{D}(\mathbb{R})$ such that $0 \leq \psi_{n} \leq 1, \psi_{n}=1$ on $[-n, n]$ and $\left\|\psi_{n}^{\prime}\right\|_{\infty} \leq C$ for all $n \in \mathbb{N}$. Let $\varepsilon>0$. Choose $v \in C^{\infty} \cap W^{1, p}(\mathbb{R})$ such that $\|u-v\|_{W^{1, p}} \leq \varepsilon$ (regularization step). For every $n \in \mathbb{N}$, one has $v \psi_{n} \in \mathscr{D}(\mathbb{R})$ and it is easy to check that for all $n$ large enough, $\left\|v-v \psi_{n}\right\|_{W^{1, p}} \leq \varepsilon$. The claim is proved.

Corollary 9.12. For every $u \in W^{1, p}(a, b)$ (bounded or unbounded interval, $1 \leq p<$ $\infty)$ and every $\varepsilon>0$, there exists $v \in \mathscr{D}(\mathbb{R})$ such that $\left\|u-\left.v\right|_{(a, b)}\right\|_{W^{1, p}} \leq \varepsilon$.
Proof. Given $u \in W^{1, p}(a, b)$, we first choose an extension $\tilde{u} \in W^{1, p}(\mathbb{R})$ (extension lemma 9.10) and then a test function $v \in \mathscr{D}(\mathbb{R})$ such that $\|\tilde{u}-v\|_{W^{1, p}(\mathbb{R})} \leq \varepsilon$ (Lemma 9.11). Then $\|\tilde{u}-v\|_{W^{1, p}(a, b)}=\|u-v\|_{W^{1, p}(a, b)} \leq \varepsilon$.

Corollary 9.13 (Sobolev embedding theorem). Every function $u \in W^{1, p}(a, b)$ is continuous and bounded and there exists a constant $C \geq 0$ such that

$$
\|u\|_{\infty} \leq C\|u\|_{W^{1, p}} \quad \text { for every } u \in W^{1, p}(a, b) .
$$

Proof. If $p=\infty$, there is nothing to prove. We first prove the claim for the case $(a, b)=\mathbb{R}$.

So let $1 \leq p<\infty$ and let $v \in \mathscr{D}(\mathbb{R})$. Then $G(v):=|v|^{p-1} v \in C_{c}^{1}(\mathbb{R})$ and $G(v)^{\prime}=$ $p|v|^{p-1} v^{\prime}$. By Hölder's inequality,

$$
|G(v)(x)|=\left.p\left|\int_{-\infty}^{x}\right| v\right|^{p-1} v^{\prime} \mid \leq p\|v\|_{p}^{p-1}\left\|v^{\prime}\right\|_{p}
$$

so that by Young's inequality ( $\left.a b \leq \frac{1}{p} a^{p}+\frac{1}{p^{\prime}} b^{p^{\prime}}\right)$

$$
\|v\|_{\infty}=\|G(v)\|_{\infty}^{1 / p} \leq C\|v\|_{W^{1, p}} .
$$

Since $\mathscr{D}(\mathbb{R})$ is dense in $W^{1, p}(\mathbb{R})$ by Lemma 9.11, the claim for $(a, b)=\mathbb{R}$ follows by an approximation argument.

The case $(a, b) \neq \mathbb{R}$ is an exercise.

Theorem 9.14 (Product rule, partial integration). Let $u, v \in W^{1, p}(a, b)(1 \leq p \leq$ $\infty)$. Then:
(i) (Product rule). The product uv belongs to $W^{1, p}(a, b)$ and

$$
(u v)^{\prime}=u^{\prime} v+u v^{\prime} .
$$

(ii) (Partial integration). If $-\infty<a<b<\infty$, then

$$
\int_{a}^{b} u^{\prime} v=u(b) v(b)-u(a) v(a)-\int_{a}^{b} u v^{\prime}
$$

Proof. Since every function in $W^{1, p}(a, b)$ is bounded, we find that $u v, u^{\prime} v+u v^{\prime} \in$ $L^{p}(a, b)$. Choose sequences $\left(u_{n}\right),\left(v_{n}\right) \subseteq \mathscr{D}(\mathbb{R})$ such that $\left.\lim _{n \rightarrow \infty} u_{n}\right|_{(a, b)}=u$ and $\left.\lim _{n \rightarrow \infty} v_{n}\right|_{(a, b)}=v$ in $W^{1, p}(a, b)$ (Corollary 9.12). By Corollary 9.13, this implies also $\lim _{n \rightarrow \infty}\left\|\left.u_{n}\right|_{(a, b)}-u\right\|_{\infty}=\lim _{n \rightarrow \infty}\left\|\left.v_{n}\right|_{(a, b)}-v\right\|_{\infty}=0$. The classical product rule implies

$$
\left(u_{n} v_{n}\right)^{\prime}=u_{n}^{\prime} v_{n}+u_{n} v_{n}^{\prime} \text { for every } n \in \mathbb{N}
$$

and the classical rule of partial integration implies

$$
\int_{a}^{b} u_{n}^{\prime} v_{n}=u_{n}(b) v_{n}(b)-u_{n}(a) v_{n}(a)-\int_{a}^{b} u_{n} v_{n}^{\prime} \text { for every } n \in \mathbb{N}
$$

The claim follows upon letting $n$ tend to $\infty$.
For every $1 \leq p \leq \infty$ and every $k \geq 2$ we define inductively the Sobolev spaces

$$
W^{k, p}(a, b):=\left\{u \in W^{1, p}(a, b): u^{\prime} \in W^{k-1, p}(a, b)\right\},
$$

which are Banach spaces for the norms

$$
\|u\|_{W^{k, p}}:=\sum_{j=0}^{k}\left\|u^{(j)}\right\|_{p}
$$

We denote $H^{k}(a, b):=W^{k, 2}(a, b)$ which is a Hilbert space for the scalar product

$$
\langle u, v\rangle_{H^{k}}:=\sum_{j=0}^{k} u^{(j)} v^{(j)} L^{2} .
$$

Finally, we define

$$
W_{0}^{k, p}(a, b):=\overline{\mathscr{D}(a, b)} \|^{\|\cdot\|_{W^{k, p}},}
$$

that is, $W_{0}^{k, p}(a, b)$ is the closure of the test functions in $W^{k, p}(a, b)$, and we put $H_{0}^{k}(a, b):=W_{0}^{k, 2}(a, b)$.

Theorem 9.15. Let $-\infty<a<b<\infty$. A function $u \in W_{0}^{1, p}(a, b)$ if and only if $u \in$ $W^{1, p}(a, b)$ and $u(a)=u(b)=0$.

Theorem 9.16 (Poincaré inequality). Let $-\infty<a<b<\infty$ and $1 \leq p<\infty$. Then there exists a constant $\lambda>0$ such that

$$
\lambda \int_{a}^{b}|u|^{p} \leq \int_{a}^{b}\left|u^{\prime}\right|^{p} \quad \text { for every } u \in W_{0}^{1, p}(a, b) .
$$

Proof. Let $u \in W^{1, p}(a, b)$. Then

$$
\begin{aligned}
\int_{a}^{b}|u(x)|^{p} \mathrm{~d} x & =\int_{a}^{b}\left|\int_{a}^{x} u^{\prime}(y) \mathrm{d} y\right|^{p} \mathrm{~d} x \\
& \leq \int_{a}^{b}\left(\int_{a}^{b}\left|u^{\prime}(y)\right| \mathrm{d} y\right)^{p} \mathrm{~d} x \\
& \leq \int_{a}^{b}(b-a)^{p-1} \int_{a}^{b}\left|u^{\prime}(y)\right|^{p} \mathrm{~d} y \mathrm{~d} x \\
& =(b-a)^{p} \int_{a}^{b}\left|u^{\prime}(y)\right|^{p} \mathrm{~d} y .
\end{aligned}
$$

Between the first and the second line, we have used the assumption that $u(a)=0$, while in the following inequality we applied Hölder's inequality.

Theorem 9.17. Let $-\infty<a<b<\infty$. For every $f \in L^{2}(a, b)$ there exists a unique function $u \in H_{0}^{1}(a, b) \cap H^{2}(a, b)$ such that

$$
\left\{\begin{array}{l}
u-u^{\prime \prime}=f \quad \text { and }  \tag{9.3}\\
u(a)=u(b)=0
\end{array}\right.
$$

Proof. We first note that if $u \in H_{0}^{1}(a, b) \cap H^{2}(a, b)$ is a solution, then, by partial integration (Theorem 9.14), for every $v \in H_{0}^{1}(a, b)$

$$
\begin{equation*}
\int_{a}^{b}\left(u v+u^{\prime} v^{\prime}\right)=(u, v)_{H_{0}^{1}}=\int_{a}^{b} f v . \tag{9.4}
\end{equation*}
$$

By the Cauchy-Schwarz inequality, the linear functional $\varphi \in H_{0}^{1}(a, b)^{\prime}$ defined by $\varphi(v)=\int_{a}^{b} f v$ is bounded:

$$
|\varphi(v)| \leq\|f\|_{2}\|v\|_{2} \leq\|f\|_{2}\|v\|_{H_{0}^{1}} .
$$

By the theorem of Riesz-Fréchet, there exists a unique $u \in H_{0}^{1}(a, b)$ such that (9.4)holds true for all $v \in H_{0}^{1}(a, b)$. This proves uniqueness of a solution of (9.3), and if we prove that in addition $u \in H^{2}(a, b)$, then we prove existence, too. However, (9.4)holds in particular for all $v \in \mathscr{D}(a, b)$, i.e.

$$
\int_{a}^{b} u^{\prime} v^{\prime}=-\int_{a}^{b}(u-f) v \quad \forall v \in \mathscr{D}(a, b)
$$

and $u-f \in L^{2}(a, b)$ by assumption. Hence, by definition, $u^{\prime} \in H^{1}(a, b)$, i.e. $u \in$ $H^{2}(a, b)$ and $u^{\prime \prime}=u-f$. Using also Theorem 9.15, the claim is proved.

### 9.3 Sobolev spaces in several dimensions

In order to motivate Sobolev spaces in several space dimensions, we have to recall the partial integration rule in this case.

Theorem 9.18 (Gauß). Let $\Omega \subseteq \mathbb{R}^{d}$ be open and bounded such that $\partial \Omega$ is of class $C^{1}$. Then there exists a unique Borel measure $\sigma$ on $\partial \Omega$ such that for every $u, v \in$ $C^{1}(\bar{\Omega})$ and every $1 \leq i \leq d$

$$
\int_{\Omega} u \frac{\partial v}{\partial x_{i}}=\int_{\partial \Omega} u v n_{i} \mathrm{~d} \sigma-\int_{\Omega} \frac{\partial u}{\partial x_{i}} v,
$$

where $n(x)=\left(n_{i}(x)\right)_{1 \leq i \leq d}$ denotes the outer normal vector at a point $x \in \partial \Omega$.
In particular, if $u \in C^{1}(\bar{\Omega})$ and $\varphi \in \mathscr{D}(\Omega)$, then

$$
\int_{\Omega} u \frac{\partial \varphi}{\partial x_{i}}=-\int_{\Omega} \frac{\partial u}{\partial x_{i}} \varphi .
$$

Let $\Omega \subseteq \mathbb{R}^{d}$ be any open set and $1 \leq p \leq \infty$. We define

$$
\begin{aligned}
W^{1, p}(\Omega):=\left\{u \in L^{p}(\Omega): \forall 1 \leq i \leq d \exists g_{i}\right. & \in L^{p}(\Omega) \\
& \left.\forall \varphi \in \mathscr{D}(\Omega): \int_{\Omega} u \frac{\partial \varphi}{\partial x_{i}}=-\int_{\Omega} g_{i} \varphi\right\} .
\end{aligned}
$$

The space $W^{1, p}(\Omega)$ is called (first) Sobolev space. If $p=2$, then we also write $H^{1}(\Omega):=W^{1,2}(\Omega)$.

Let $u \in W^{1, p}(\Omega)$. By Lemma 9.4, the functions $g_{i}$ are uniquely determined. We write $\frac{\partial u}{\partial x_{i}}:=g_{i}$ and call $\frac{\partial u}{\partial x_{i}}$ the partial derivative of $u$ with respect to $x_{i}$. As in the one-dimensional case, the following holds true.

Lemma 9.19. The Sobolev spaces $W^{1, p}(\Omega)$ are Banach spaces for the norms

$$
\|u\|_{W^{1, p}}:=\|u\|_{p}+\sum_{i=1}^{d}\left\|\frac{\partial u}{\partial x_{i}}\right\|_{p} \quad(1 \leq p \leq \infty)
$$

and $H^{1}(\Omega)$ is a Hilbert space for the inner product

$$
\langle u, v\rangle_{H^{1}}:=\langle u, v\rangle_{L^{2}}+\sum_{i=1}^{d}\left\langle\frac{\partial u}{\partial x_{i}}, \frac{\partial v}{\partial x_{i}}\right\rangle_{L^{2}} .
$$

Proof. Exercise.
Not all properties of Sobolev spaces on intervals carry over to Sobolev spaces on open sets $\Omega \subseteq \mathbb{R}^{d}$. For example, it is not true that every function $u \in W^{1, p}(\Omega)$ is continuous (without any further restrictions on $p$ and $\Omega$ )!

For every open $\Omega \subseteq \mathbb{R}^{d}, 1 \leq p \leq \infty$ and every $k \geq 2$ we define inductively the Sobolev spaces

$$
W^{k, p}(\Omega):=\left\{u \in W^{1, p}(\Omega): \forall 1 \leq i \leq d: \frac{\partial u}{\partial x_{i}} \in W^{k-1, p}(\Omega)\right\},
$$

which are Banach spaces for the norms

$$
\|u\|_{W^{k, p}}:=\|u\|_{p}+\sum_{i=0}^{k}\left\|\frac{\partial u}{\partial x_{i}}\right\|_{W^{k-1, p}} .
$$

We denote $H^{k}(\Omega):=W^{k, 2}(\Omega)$ which is a Hilbert space for the inner product

$$
\langle u, v\rangle_{H^{k}}:=\langle u, v\rangle_{L^{2}}+\sum_{i=0}^{k}\left\langle\frac{\partial u}{\partial x_{i}}, \frac{\partial v}{\partial x_{i}}\right\rangle_{H^{k-1}} .
$$

Finally, we define

$$
W_{0}^{k, p}(\Omega):=\overline{\mathscr{D}(\Omega)}{ }^{\|\cdot\|_{W^{k}, p}},
$$

that is, $W_{0}^{k, p}(\Omega)$ is the closure of the test functions in $W^{k, p}(\Omega)$, and we put $H_{0}^{k}(\Omega):=W_{0}^{k, 2}(\Omega)$.
Theorem 9.20 (Poincaré inequality). Let $\Omega \subseteq \mathbb{R}^{d}$ be a bounded domain, and let $1 \leq p<\infty$. Then there exists a constant $C \geq 0$ such that

$$
\int_{\Omega}|u|^{p} \leq C^{p} \int_{\Omega}|\nabla u|^{p} \quad \text { for every } u \in W_{0}^{1, p}(\Omega) .
$$

We note that the Poincaré inequality implies that

$$
\|u\|:=\left(\int_{\Omega}|\nabla u|^{p}\right)^{\frac{1}{p}}
$$

defines an equivalent norm on $W_{0}^{1, p}(\Omega)$ if $\Omega \subseteq \mathbb{R}^{d}$ is bounded. Clearly,

$$
\|u\| \leq\|u\|_{W_{0}^{1, p}} \quad \text { for every } u \in W_{0}^{1, p}
$$

by the definition of the norm in $W^{1, p}$. On the other hand,

$$
\begin{aligned}
\|u\|_{W_{0}^{1, p}} & \leq C\left(\|u\|_{L^{p}}+\|\nabla u\|_{L^{p}}\right) \\
& \leq C\|\nabla u\|_{L^{p}}=C\|u\|,
\end{aligned}
$$

by the Poincaré inequality.
We also state the following two theorems without proof.
Theorem 9.21 (Sobolev embedding theorem). Let $\Omega \subseteq \mathbb{R}^{d}$ be an open set with $C^{1}$ boundary. Let $1 \leq p \leq \infty$ and define

$$
p^{*}:= \begin{cases}\frac{d p}{d-p} & \text { if } 1 \leq p<d \\ \infty & \text { if } d<p\end{cases}
$$

and if $p=d$, then $p^{*} \in[1, \infty)$. Then, for every $p \leq q \leq p^{*}$ we have

$$
W^{1, p}(\Omega) \subseteq L^{q}(\Omega)
$$

with continuous embedding, that is, there exists $C=C(p, q) \geq 0$ such that

$$
\|u\|_{L^{q}} \leq C\|u\|_{W^{1, p}} \quad \text { for every } u \in W^{1, p}(\Omega) .
$$

Theorem 9.22 (Rellich-Kondrachov). Let $\Omega \subseteq \mathbb{R}^{d}$ be an open and bounded set with $C^{1}$ boundary. Let $1 \leq p \leq \infty$ and define $p^{*}$ as in the Sobolev embedding theorem. Then, for every $p \leq q<\infty$ the embedding

$$
W^{1, p}(\Omega) \subseteq L^{q}(\Omega)
$$

is compact, that is, every bounded sequence in $W^{1, p}(\Omega)$ has a subsequence which converges in $L^{q}(\Omega)$.

## 9.4 * Elliptic partial differential equations

Let $\Omega \subseteq \mathbb{R}^{d}$ be an open, bounded set, $f \in L^{2}(\Omega)$, and consider the elliptic partial differential equation

$$
\left\{\begin{array}{l}
u-\Delta u=f \text { in } \Omega,  \tag{9.5}\\
u=0 \quad \text { in } \partial \Omega,
\end{array}\right.
$$

where

$$
\Delta u(x):=\sum_{i=1}^{d} \frac{\partial^{2}}{\partial x_{i}^{2}} u(x)
$$

stands for the Laplace operator.
If $u \in H_{0}^{1}(\Omega) \cap H^{2}(\Omega)$ is a solution of (9.5), then, by definition of the Sobolev spaces, for every $v \in \mathscr{D}(a, b)$

$$
\begin{aligned}
\langle u, v\rangle_{H_{0}^{1}} & =\int_{\Omega}\left(u v+\sum_{i=1}^{d} \frac{\partial u}{\partial x_{i}} \frac{\partial v}{\partial x_{i}}\right) \\
& =\int_{\Omega}\left(u v-\sum_{i=1}^{d} \frac{\partial^{2} u}{\partial x_{i}^{2}} v\right) \\
& =\int_{\Omega}(u-\Delta u) v \\
& =\int_{\Omega} f v .
\end{aligned}
$$

By density of the test functions in $H_{0}^{1}(\Omega)$, this equality holds actually for all $v \in$ $H_{0}^{1}(\Omega)$. This may justify the following definition of a weak solution. A function $u \in H_{0}^{1}(\Omega)$ is called a weak solution of (9.5) if for every $v \in H_{0}^{1}(\Omega)$

$$
\begin{equation*}
\langle u, v\rangle_{H_{0}^{1}}=\int_{\Omega} u v+\int_{\Omega} \nabla u \nabla v=\int_{\Omega} f v, \tag{9.6}
\end{equation*}
$$

where $\nabla u$ is the usual, euclidean gradient of $u$.
Theorem 9.23. Let $\Omega \subseteq \mathbb{R}^{d}$ be an open, bounded set. Then, for every $f \in L^{2}(\Omega)$ there exists a unique weak solution $u \in H_{0}^{1}(\Omega)$ of the problem (9.5).

Proof. By the Cauchy-Schwarz inequality, the linear functional $\varphi \in H_{0}^{1}(\Omega)^{\prime}$ defined by $\varphi(v)=\int_{\Omega} f v$ is bounded:

$$
|\varphi(v)| \leq\|f\|_{2}\|v\|_{2} \leq\|f\|_{2}\|v\|_{H_{0}^{1}} .
$$

By the theorem of Riesz-Fréchet, there exists a unique $u \in H_{0}^{1}(\Omega)$ such that (9.6) holds true for all $v \in H_{0}^{1}(a, b)$. The claim is proved.

## Chapter 10 <br> Bochner-Lebesgue and Bochner-Sobolev spaces

### 10.1 The Bochner integral

Let $X$ and $Y$ be Banach spaces with norms denoted by $\|\cdot\|_{X}$ and $\|\cdot\|_{Y}$, respectively. If the norm is clear from the context, we simply write $\|\cdot\|$. The space of all bounded, linear operators from $X$ into $Y$ is denoted by $\mathscr{L}(X, Y)$. Let $(\Omega, \mathscr{A}, \mu)$ be a measure space. A function $f: \Omega \rightarrow X$ is called step function, if there exists a sequence $\left(A_{n}\right) \subseteq \mathscr{A}$ of mutually disjoint measurable sets and a sequence $\left(x_{n}\right) \subseteq X$ such that $f=\sum_{n} 1_{A_{n}} x_{n}$. A function $f: \Omega \rightarrow X$ is called measurable, if there exists a sequence $\left(f_{n}\right)$ of step functions $f_{n}: \Omega \rightarrow X$ such that $f_{n} \rightarrow f$ pointwise $\mu$-almost everywhere.

Remark 10.1. Note that there may be a difference to the definition of mesurability of scalar valued functions. On the one hand, measurability of a function is here depending on the measure $\mu$. However, if the measure space $(\Omega, \mathscr{A}, \mu)$ is complete in the sense that $\mu(A)=0$ and $B \subseteq A$ implies $B \in \mathscr{A}$, then the above definition of measurability and the classical definition of measurability coincide. Note that one may always consider complete measure spaces. On the other hand, measurability of a function between two measurable spaces is defined via the property that preimages of measurable sets are measurable. Although one may always equip a Banach space with the Borel- $\sigma$-algebra, this definition via preimages is not appropriate for the following purposes.

Lemma 10.2. If $f: \Omega \rightarrow X$ is measurable, then $\|f\|_{X}: \Omega \rightarrow \mathbb{R}$ is measurable. More generally, if $f: \Omega \rightarrow X$ is measurable and if $g: X \rightarrow Y$ is continuous, then $g \circ f:$ $\Omega \rightarrow Y$ is measurable.

Proof. This is an easy consequence of the definition of measurability and the continuity of $g$. Note that in particular the norm $\|\cdot\|_{X}: X \rightarrow \mathbb{R}$ is continuous.

Lemma 10.3. If $f: \Omega \rightarrow X$ and $g: \Omega \rightarrow \mathbb{K}$ are measurable, then $f g: \Omega \rightarrow X$ is measurable. Similarly, if $f: \Omega \rightarrow X$ and $g: \Omega \rightarrow X^{\prime}$ are measurable, then $\langle g, f\rangle_{X^{\prime}, X}$ : $\Omega \rightarrow \mathbb{K}$ is measurable.

Proof. For the proof it suffices to use the definition of measurability and to show that the (duality) product of two step functions is again a step function. This is, however, straightforward.

Theorem 10.4 (Pettis). A function $f: \Omega \rightarrow X$ is measurable if and only if $\left\langle x^{\prime}, f\right\rangle$ is measurable for every $x^{\prime} \in X^{\prime}$ (we say that $f$ is weakly measurable) and if there exists a $\mu$-null set $N \in \mathscr{A}$ such that $f(\Omega \backslash N)$ is separable (we say that $f$ is almost separably valued).

For the following proof of Pettis' theorem, see Hille \& Phillips [Hille and Phillips (1957)].

Proof. Sufficiency. Assume that $f$ is measurable. Then $f$ is weakly measurable by Lemma 10.2. Moreover, by definition, there exists a sequence $\left(f_{n}\right)$ of step functions and a $\mu$-null set $N \in \mathscr{A}$ such that

$$
f_{n}(t) \rightarrow f(t) \text { for all } t \in \Omega \backslash N .
$$

Hence,

$$
f(\Omega \backslash N) \subseteq \overline{\bigcup_{n} f_{n}(\Omega)}
$$

Since for every step function $f_{n}$ the range is countable, the set on the right-hand side of this inclusion is separable, and hence $f$ is almost separably valued.

Necessity. Assume that $f$ is weakly measurable and almost separably valued. We first show that $\|f\|_{X}$ is measurable. By assumption, there exists a $\mu$-null set and a sequence $\left(x_{n}\right)$ in $X$ such that $D:=\left\{x_{n}: n \in \mathbb{N}\right\}$ is dense in $f(\Omega \backslash N)$. By the Hahn-Banach theorem, there exists a sequence $\left(x_{n}^{\prime}\right)$ in $X^{\prime}$ such that $\left\|x_{n}^{\prime}\right\|_{X}=1$ and $\left\langle x_{n}^{\prime}, x_{n}\right\rangle=\left\|x_{n}\right\|_{X}$. Since $f$ is weakly measurable, $\left|\left\langle x_{n}^{\prime}, f\right\rangle\right|$ is measurable for every $n$. As a consequence, $\sup _{n}\left|\left\langle x_{n}^{\prime}, f\right\rangle\right|$ is measurable. But $\sup _{n}\left|\left\langle x_{n}^{\prime}, f\right\rangle\right|=\|f\|_{X}$ on $\Omega \backslash N$ by the choice of the sequence $\left(x_{n}^{\prime}\right)$ and the density of $D$ in the $f(\Omega \backslash N)$. Since our measure space $(\Omega, \mathscr{A}, \mu)$ is supposed to be complete, we obtain that $\|f\|_{X}$ is measurable. In a similar way, one shows that $|f-x|_{X}$ is measurable for every $x \in X$, and in particular for $x=x_{n}$.

Now fix $m \in \mathbb{N}$ and define

$$
\begin{aligned}
A_{m 1} & :=\left\{\left\|f-x_{1}\right\|_{X} \leq \inf _{1 \leq k \leq m}\left\|f-x_{k}\right\|_{X}\right\}, \\
A_{m 2} & :=\left\{\left\|f-x_{2}\right\|_{X} \leq \inf _{1 \leq k \leq m}\left\|f-x_{k}\right\|_{X}\right\} \backslash A_{m 1}, \\
A_{m 3} & :=\left\{\left\|f-x_{3}\right\|_{X} \leq \inf _{1 \leq k \leq m}\left\|f-x_{k}\right\|_{X}\right\} \backslash\left(A_{m 1} \cup A_{m 2}\right), \\
& \vdots \\
& \vdots \\
A_{m m} & :=\left\{\left\|f-x_{m}\right\|_{X} \leq \inf _{1 \leq k \leq m}\left\|f-x_{k}\right\|_{X}\right\} \backslash\left(\bigcup_{k=1}^{m-1} A_{m k}\right) .
\end{aligned}
$$

Then $\left(A_{m n}\right)_{1 \leq n \leq m}$ is a family of measurable, mutually disjoint sets such that $\bigcup_{n=1}^{m} A_{m n}=\Omega$. Define ${ }^{1}$

$$
f_{m}:=\sum_{n=1}^{m} 1_{A_{m n}} x_{n} .
$$

Then $\left(f_{m}\right)$ is a sequence of step functions, $\left(\left\|f_{m}-f\right\|_{X}\right)_{m}$ is decreasing pointwise everywhere, and since $D$ is dense in $f(\Omega \backslash N)$,

$$
\lim _{m \rightarrow \infty}\left\|f_{m}(t)-f(t)\right\|_{X}=0 \text { for every } t \in \Omega \backslash N
$$

that is, $f_{m} \rightarrow f \mu$-almost everywhere. As a consequence, $f$ is measurable.
Remark 10.5. The above proof of Pettis' theorem shows that a measurable function can always be approximated almost everywhere by a sequence of finite step functions. The proof in [Hille and Phillips (1957)] is slightly different and shows that a measurable, separably valued function can always be approximated uniformly by a sequence of step functions.

Corollary 10.6. If $\left(f_{n}\right)$ is a sequence of measurable functions $\Omega \rightarrow X$ such that $f_{n} \rightarrow f$ pointwise $\mu$-almost everywhere, then $f$ is measurable.

Proof. We assume that this corollary is known in the scalar case, that is, when $X=$ $\mathbb{K}$.

By Pettis's theorem (Theorem 10.4), for all $n$ there exists a $\mu$-null set $N_{n} \in \mathscr{A}$ such that $f_{n}\left(\Omega \backslash N_{n}\right)$ is separable. Moreover there exists a $\mu$-null set $N_{0} \in \Omega$ such that $f_{n}(t) \rightarrow f(t)$ for all $t \in \Omega \backslash N_{0}$. Let $N:=\bigcup_{n \geq 0} N_{n}$; as a countable union of $\mu$-null sets, $N$ is a $\mu$-null set.

Then $f$ (restricted to $\Omega \backslash N$ ) is the pointwise limit everywhere of the sequence $\left(f_{n}\right)$. In particular $f$ is weakly measurable. Moreover, $f(\Omega \backslash N)$ is separable since

$$
f(\Omega \backslash N) \subseteq \overline{\bigcup_{n} f_{n}(\Omega \backslash N)}
$$

and since $f_{n}(\Omega \backslash N)$ is separable. The claim follows from Pettis' theorem.
A measurable function $f: \Omega \rightarrow X$ is called integrable if $\int_{\Omega}\|f\|_{X} \mathrm{~d} \mu<\infty$.
Lemma 10.7. For every integrable step function $f: \Omega \rightarrow X, f=\sum_{n} 1_{A_{n}} x_{n}$ the series $\sum_{n} x_{n} \mu\left(A_{n}\right)$ converges absolutely and its limit is independent of the representation of $f$.

Proof. Let $f=\sum_{n} 1_{A_{n}} x_{n}$ be an integrable step function. The sets $\left(A_{n}\right) \subseteq \mathscr{A}$ are mutually disjoint and $\left(x_{n}\right) \subseteq X$. Then

$$
\sum_{n}\left\|x_{n}\right\|_{X} \mu\left(A_{n}\right)=\int_{\Omega}\|f\|_{X} \mathrm{~d} \mu<\infty .
$$

[^0]Let $f: \Omega \rightarrow X$ be an integrable step function, $f=\sum_{n} 1_{A_{n}} x_{n}$. We define the Bochner integral (for integrable step functions) by

$$
\int_{\Omega} f \mathrm{~d} \mu:=\sum_{n} x_{n} \mu\left(A_{n}\right) .
$$

Lemma 10.8. a) For every measurable function $f: \Omega \rightarrow X$ there exists a sequence $\left(f_{n}\right)$ of step functions $\Omega \rightarrow X$ such that $\left\|f_{n}\right\|_{X} \leq\|f\|_{X}$ and $f_{n} \rightarrow f$ pointwise $\mu$-almost everywhere.
b) Let $f: \Omega \rightarrow X$ be integrable. Let $\left(f_{n}\right)$ be a sequence of integrable step functions such that $\left\|f_{n}\right\|_{X} \leq\|f\|_{X}$ and $f_{n} \rightarrow f$ pointwise $\mu$-almost everywhere. Then

$$
x:=\lim _{n \rightarrow \infty} \int_{\Omega} f_{n} \mathrm{~d} \mu \text { exists }
$$

and

$$
\|x\|_{X} \leq \int_{\Omega}\|f\|_{X} \mathrm{~d} \mu
$$

Proof. (a) Let $f: \Omega \rightarrow X$ be measurable. Then $\|f\|_{X}: \Omega \rightarrow \mathbb{R}$ is measurable. Therefore there exists a sequence $\left(g_{n}\right)$ of real step functions such that $0 \leq g_{n} \leq\|f\|_{X}$ and $g_{n} \rightarrow\|f\|_{X}$ pointwise $\mu$-almost everywhere.

Since $f$ is measurable, there exists a sequence $\left(\tilde{f}_{n}\right)$ of step functions $\Omega \rightarrow X$ such that $\tilde{f}_{n} \rightarrow f$ pointwise $\mu$-almost everywhere. Put

$$
f_{n}:=\frac{\tilde{f}_{n} g_{n}}{\left\|\tilde{f}_{n}\right\|_{X}+\frac{1}{n}} .
$$

(b) For every integrable step function $g: \Omega \rightarrow X$ one has

$$
\left|\int_{\Omega} g \mathrm{~d} \mu\right|_{X} \leq \int_{\Omega}\|g\|_{X} \mathrm{~d} \mu
$$

Hence, for every $n, m$

$$
\left|\int_{\Omega} f_{n}-f_{m} \mathrm{~d} \mu\right|_{X} \leq \int_{\Omega}\left\|f_{n}-f_{m}\right\|_{X} \mathrm{~d} \mu
$$

and by Lebesgue's dominated convergence theorem the sequence $\left(\int_{\Omega} f_{n} \mathrm{~d} \mu\right)$ is a Cauchy sequence. When we put $x=\lim _{n \rightarrow \infty} \int_{\Omega} f_{n} \mathrm{~d} \mu$ then

$$
\|x\|_{X} \leq \liminf _{n \rightarrow \infty} \int_{\Omega}\left\|f_{n}\right\|_{X} \mathrm{~d} \mu=\int_{\Omega}\|f\|_{X} \mathrm{~d} \mu .
$$

Let $f: \Omega \rightarrow X$ be integrable. We define the Bochner integral

$$
\int_{\Omega} f \mathrm{~d} \mu:=\lim _{n \rightarrow \infty} \int_{\Omega} f_{n} \mathrm{~d} \mu
$$

where $\left(f_{n}\right)$ is a sequence of (integrable) step functions $\Omega \rightarrow X$ such that $\left\|f_{n}\right\|_{X} \leq$ $\|f\|_{X}$ and $f_{n} \rightarrow f$ pointwise $\mu$-almost everywhere (Lemma 10.8 (a)). The definition of the Bochner integral for integrable functions is independent of the choice of the sequence $\left(f_{n}\right)$ of step functions, by Lemma 10.8 (b). Moreover, if $f$ is a step function, then this definition of the Bochner integral and the previous definition coincide. Finally, by Lemma 10.8 (b),

$$
\begin{equation*}
\left|\int_{\Omega} f \mathrm{~d} \mu\right|_{X} \leq \int_{\Omega}\|f\|_{X} \mathrm{~d} \mu \quad \text { (triangle inequality). } \tag{10.1}
\end{equation*}
$$

Remark 10.9. We will also use the following notation for the Bochner integral:

$$
\int_{\Omega} f \text { or } \int_{\Omega} f(t) \mathrm{d} \mu(t)
$$

and if $\Omega=(a, b)$ is an interval in $\mathbb{R}$ :

$$
\int_{a}^{b} f \text { or } \int_{a}^{b} f(t) \mathrm{d} \mu(t)
$$

If $\mu=\lambda$ is the Lebesgue measure then we also write

$$
\int_{\Omega} f(t) \mathrm{d} t \text { or } \int_{a}^{b} f(t) \mathrm{d} t
$$

Lemma 10.10. Let $f: \Omega \rightarrow X$ be integrable and $T \in \mathscr{L}(X, Y)$. Then $T f: \Omega \rightarrow Y$ is integrable and

$$
\int_{\Omega} T f \mathrm{~d} \mu=T \int_{\Omega} f \mathrm{~d} \mu .
$$

Proof. Exercise.
Theorem 10.11 (Lebesgue, dominated convergence). Let $\left(f_{n}\right)$ be a sequence of integrable functions. Suppose there exists an integrable function $g: \Omega \rightarrow \mathbb{R}$ and an (integrable) measurable function $f: \Omega \rightarrow X$ such that $\left\|f_{n}\right\| \leq g$ and $f_{n} \rightarrow f$ pointwise $\mu$-almost everywhere. Then

$$
\int_{\Omega} f \mathrm{~d} \mu=\lim _{n \rightarrow \infty} \int_{\Omega} f_{n} \mathrm{~d} \mu
$$

Proof. By the triangle inequality and the classical Lebesgue dominated convergence theorem,

$$
\left|\int_{\Omega} f \mathrm{~d} \mu-\int_{\Omega} f_{n} \mathrm{~d} \mu\right|_{X} \leq \int_{\Omega}\left\|f-f_{n}\right\|_{X} \mathrm{~d} \mu \rightarrow 0 \text { as } n \rightarrow \infty
$$

### 10.2 Bochner-Lebesgue spaces

Given a measure space $(\Omega, \mathscr{A}, \mu)$ and a Banach space $X$, we define

$$
\begin{aligned}
& \mathscr{L}^{p}(\Omega ; X):=\left\{f: \Omega \rightarrow X \text { measurable }: \int_{\Omega}\|f\|_{X}^{p} \mathrm{~d} \mu<\infty\right\} \quad \text { if } 1 \leq p<\infty, \text { and } \\
& \mathscr{L}^{\infty}(\Omega ; X):=\left\{f: \Omega \rightarrow X \text { measurable }: \exists C \geq 0 \text { such that } \mu\left(\left\{\|f\|_{X} \geq C\right\}\right)=0\right\} .
\end{aligned}
$$

Similarly as in the scalar case one shows that these sets a linear spaces and that

$$
\begin{aligned}
\|f\|_{p} & :=\left(\int_{\Omega}\|f\|_{X}^{p} \mathrm{~d} \mu\right)^{1 / p} \quad(1 \leq p<\infty), \text { resp. } \\
\|f\|_{\infty} & :=\inf \left\{C \geq 0: \mu\left(\left\{\|f\|_{X} \geq C\right\}\right)=0\right\}
\end{aligned}
$$

are seminorms. Starting with these definitions, and building on the following general principle, the proof of which is left as an exercise, we define the Bochner-Lebesgue $L^{p}$ spaces.

Lemma 10.12. If $\|\cdot\|_{\mathscr{X}}$ is a seminorm on the vector space $\mathscr{X}$, then

$$
\mathscr{N}:=\left\{x \in \mathscr{X}:\|x\|_{\mathscr{X}}=0\right\}
$$

is a linear subspace. Moreover, the quotient space

$$
X:=\mathscr{X} / \mathscr{N}
$$

becomes a normed space for the norm

$$
\|[x]\|_{X}:=\|x\|_{\mathscr{X}} \quad([x]=x+\mathscr{N} \in X) .
$$

By Lemma 10.12 , for every $1 \leq p \leq \infty$

$$
\begin{aligned}
\mathscr{N}_{p} & :=\left\{f \in \mathscr{L}^{p}(\Omega ; X):\|f\|_{p}=0\right\} \\
& =\left\{f \in \mathscr{L}^{p}(\Omega ; X): f=0 \mu \text {-almost everywhere }\right\}
\end{aligned}
$$

is a linear subspace of $\mathscr{L}^{p}(\Omega ; X)$. The Bochner-Lebesgue $L^{p}$ space is then defined to be the quotient space

$$
\mathrm{L}^{p}(\Omega ; X):=\mathscr{L}^{p}(\Omega ; X) / \mathscr{N}_{p}
$$

which is the space of all equivalence classes

$$
[f]:=f+\mathscr{N}_{p}, \quad f \in \mathscr{L}^{p}(\Omega ; X) .
$$

By Lemma 10.12, it is a normed space for the norm

$$
\|[f]\|_{p}:=\|f\|_{p}
$$

Remark 10.13. As in the scalar case we will in the following identify functions $f \in \mathscr{L}^{p}(\Omega ; X)$ with their equivalence classes $[f] \in \mathrm{L}^{p}(\Omega ; X)$, and we say that $\mathrm{L}^{p}$ is a function space although we should be aware that it is only a space of equivalence classes of functions.

Remark 10.14. For $\Omega=(a, b)$ an interval in $\mathbb{R}$ and for $\mu=\lambda$ the Lebesgue measure we simply write

$$
\mathrm{L}^{p}(a, b ; X):=\mathrm{L}^{p}((a, b) ; X) .
$$

We can do so since the spaces $\mathrm{L}^{p}([a, b] ; X)$ and $\mathrm{L}^{p}((a, b) ; X)$ coincide since the end points $\{a\}$ and $\{b\}$ have Lebesgue measure zero and there is no danger of confusion.

Theorem 10.15 (Fischer-Riesz). For every $1 \leq p \leq \infty$, the space $\mathrm{L}^{p}(\Omega ; X)$ is a Banach space.

Proof. The proof follows the same lines as in the classical case, that is, when $X=\mathbb{K}$.
Lemma 10.16. For every $1 \leq p<\infty$, the set of all $p$-integrable step functions $\Omega \rightarrow$ $X$ is dense in $L^{p}(\Omega ; X)$.

Proof.
Lemma 10.17. Let the measure space $(\Omega, \mathscr{A}, \mu)$ be such that $\mathrm{L}^{p}(\Omega)$ is separable for $1 \leq p<\infty$ (for example, $\Omega \subset \mathbb{R}^{d}$ be an open set with the Lebesgue measure). Let $X$ be separable. Then $\mathrm{L}^{p}(\Omega ; X)$ is separable for $1 \leq p<\infty$.

Proof. By assumption the spaces $\mathrm{L}^{p}(\Omega)$ and $X$ are separable. Let $\left(h_{n}\right) \subseteq \mathrm{L}^{p}(\Omega ; X)$ and $\left(x_{n}\right) \subseteq X$ be two dense sequences. Then the set

$$
\mathscr{F}:=\left\{f: \Omega \rightarrow X: f=h_{n} x_{m}\right\}
$$

is countable. It suffices to shows that $\mathscr{F} \subseteq \mathrm{L}^{p}(\Omega ; X)$ is total, that is, span $\mathscr{F}$ is dense in $\mathrm{L}^{p}(\Omega ; X)$. This is an exercise.

Lemma 10.18. Let $\Omega \subset \mathbb{R}^{d}$ be open and bounded. Then $\mathrm{C}(\bar{\Omega} ; X) \subseteq \mathrm{L}^{p}(\Omega ; X)$ for every $1 \leq p \leq \infty$.

Proof. Actually, for finite measure spaces, we have the more general inclusions

$$
\mathrm{L}^{\infty}(\Omega ; X) \subseteq \mathrm{L}^{p}(\Omega ; X) \subseteq \mathrm{L}^{q}(\Omega ; X) \subseteq \mathrm{L}^{1}(\Omega ; X)
$$

if $1 \leq q \leq p \leq \infty$.
Theorem 10.19. Let $\Omega$ be as in Lemma 10.17. Let $1<p<\infty$ and assume that $X$ is reflexive. Then the space $\mathrm{L}^{p}(\Omega ; X)$ is reflexive and

$$
\mathrm{L}^{p}(\Omega ; X)^{\prime} \cong \mathrm{L}^{p^{\prime}}\left(\Omega ; X^{\prime}\right)
$$

Proof. Without proof.

### 10.3 The convolution

Theorem 10.20 (Young's inequality). Let $T \in \operatorname{L}^{1}\left(\mathbb{R}^{N} ; \mathscr{L}(X, Y)\right)$ and $f \in$ $\mathrm{L}^{p}\left(\mathbb{R}^{N} ; X\right)(1 \leq p \leq \infty)$. Then for almost every $x \in \mathbb{R}^{N}$ the integral

$$
T * f(x):=\int_{\mathbb{R}^{N}} T(x-y) f(y) \mathrm{d} y
$$

converges absolutely, and for the function $T * f$ thus defined one has

$$
\begin{aligned}
& T * f \in \mathrm{~L}^{p}\left(\mathbb{R}^{N} ; Y\right) \text { and } \\
& \|T * f\|_{L^{p}} \leq\|T\|_{\mathrm{L}^{1}}\|f\|_{L^{p}} .
\end{aligned}
$$

Proof. The case $p=\infty$ is almost trivial. Actually, the strong continuity of the shift semigroup on $\mathrm{L}^{1}$ yields continuity (and thus measurability) of $T * f$ while the boundedness of $T * f$ and Young's inequality are immediate from the triangle inequality.

Assume now that $p=1$. By Tonnelli's theorem, we have

$$
\begin{aligned}
& \int_{\mathbb{R}^{N}} \int_{\mathbb{R}^{N}}\|T(x-y)\|_{\mathscr{L}(X, Y)}\|f(y)\|_{X} \mathrm{~d} y \mathrm{~d} x \\
& =\int_{\mathbb{R}^{N}} \int_{\mathbb{R}^{N}}\|T(x-y)\|_{\mathscr{L}(X, Y)}\|f(y)\|_{X} \mathrm{~d} x \mathrm{~d} y \\
& =\|T\|_{\mathrm{L}^{1}}\|f\|_{\mathrm{L}^{1}}
\end{aligned}
$$

and from this equality follows the claim.
Assume now $1<p<\infty$. From the previous case we deduce that for almost all $x \in \mathbb{R}^{N}$

$$
\|T(x-\cdot)\|_{\mathscr{L}(X, Y)}\|f(\cdot)\|_{X}^{p} \in \mathrm{~L}^{1}\left(\mathbb{R}^{N}\right)
$$

and thus

$$
\|T(x-\cdot)\|_{\mathscr{L}(X, Y)}^{\frac{1}{p}}\|f(\cdot)\|_{X} \in \mathrm{~L}^{p}\left(\mathbb{R}^{N}\right)
$$

On the other hand, $\|T(x-\cdot)\|_{\mathscr{L}(X, Y)}^{\frac{1}{p^{\prime}}} \in \mathrm{L}^{p^{\prime}}\left(\mathbb{R}^{N}\right)$ for every $x \in \mathbb{R}^{N}$. By Hölder's inequality, for almost every $x \in \mathbb{R}^{N}$,

$$
\|T(x-\cdot)\|_{\mathscr{L}(X, Y)}\|f(\cdot)\|_{X} \in \mathrm{~L}^{1}\left(\mathbb{R}^{N}\right)
$$

and

$$
\begin{aligned}
& \int_{\mathbb{R}^{N}}\left(\int_{\mathbb{R}^{N}}\|T(x-y)\|_{\mathscr{L}(X, Y)}\|f(y)\|_{X} \mathrm{~d} y\right)^{p} \mathrm{~d} x \\
& \leq \int_{\mathbb{R}^{N}}\left(\int_{\mathbb{R}^{N}}\|T(x-y)\|_{\mathscr{L}(X, Y)} \mathrm{d} y\right)^{\frac{p}{p^{\prime}}} \int_{\mathbb{R}^{N}}\|T(x-y)\|_{\mathscr{L}(X, Y)}\|f(y)\|_{X}^{p} \mathrm{~d} y \mathrm{~d} x \\
& =\|T\|_{\mathrm{L}^{1}}^{p-1} \int_{\mathbb{R}^{N}} \int_{\mathbb{R}^{N}}\|T(x-y)\|_{\mathscr{L}(X, Y)}\|f(y)\|_{X}^{p} \mathrm{~d} x \mathrm{~d} y \\
& =\|T\|_{\mathrm{L}^{1}}^{p}\|f\|_{\mathrm{L}^{p}}^{p} \\
& <\infty
\end{aligned}
$$

For every $T \in \mathrm{~L}^{1}\left(\mathbb{R}^{N} ; \mathscr{L}(X, Y)\right)$ and every $f \in \mathrm{~L}^{1}\left(\mathbb{R}^{N} ; X\right)$ we call the function $T * f \in \mathrm{~L}^{p}\left(\mathbb{R}^{N} ; Y\right)$ the convolution of $T$ and $f$. It is a fundamental tool in harmonic analysis and the theory of partial differential equations. One first property is the following regularizing effect of the convolution. We recall that we adopt multi-index notation. For example, for every multi-index $\alpha \in \mathbb{N}_{0}^{N}$ we define

$$
\begin{aligned}
|\alpha| & :=\sum_{k=1}^{N} \alpha_{k}, \\
\alpha! & :=\prod_{k=1}^{N} \alpha_{k}!, \text { and } \\
x^{\alpha} & :=\prod_{k=1}^{N} x_{k}^{\alpha_{k}} \quad\left(x \in \mathbb{C}^{N}\right) .
\end{aligned}
$$

Moreover, we denote by $\partial_{k}$ the partial derivative operator with respect to the $k$-th variable, and define the $\alpha$-th partial derivative

$$
\partial^{\alpha}:=\partial_{1}^{\alpha_{1}} \ldots \partial_{N}^{\alpha_{N}}
$$

Let $\Omega \subseteq \mathbb{R}^{N}$ be an open set. For every function $f \in \mathrm{C}(\Omega ; X)$ we define the support

$$
\operatorname{supp} f:=\overline{\{x \in \Omega: f(x) \neq 0\}}
$$

where the closure has to be taken in $\Omega$ ! We then define for $k \in \mathbb{N}_{0} \cup\{\infty\}$

$$
\mathrm{C}_{\mathrm{c}}^{k}(\Omega ; X):=\left\{f \in \mathrm{C}^{k}(\Omega ; X): \operatorname{supp} f \text { is compact }\right\}
$$

the space of compactly supported $\mathrm{C}^{k}$-functions. In the special case $X=\mathbb{K}$ we define

$$
\mathscr{D}(\Omega):=\mathrm{C}_{\mathrm{c}}^{\infty}(\Omega) .
$$

Elements of $\mathscr{D}(\Omega)$ are called test functions.
Lemma 10.21 (Regularization). For every $f \in \mathrm{~L}_{\text {loc }}^{1}\left(\mathbb{R}^{N} ; X\right)$ and every $\varphi \in \mathrm{C}_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{N}\right)$ one has $f * \varphi \in \mathrm{C}^{\infty}\left(\mathbb{R}^{N} ; X\right)$ and

$$
\partial^{\alpha}(f * \varphi)=f * \partial^{\alpha} \varphi
$$

Proof. Let $e_{i} \in \mathbb{R}^{d}$ be the $i$-th unit vector. Then

$$
\lim _{h \rightarrow 0} \frac{1}{h}\left(\varphi\left(x+h e_{i}\right)-\varphi(x)\right)=\frac{\partial \varphi}{\partial x_{i}}(x)
$$

uniformly in $x \in \mathbb{R}^{d}$ (note that $\varphi$ has compact support). Hence, for every $x \in \mathbb{R}^{d}$

$$
\begin{aligned}
& \frac{1}{h}\left(f * \varphi\left(x+h e_{i}\right)-f * \varphi(x)\right) \\
= & \frac{1}{h} \int_{\mathbb{R}^{d}} f(y)\left(\varphi\left(x+h e_{i}-y\right)-\varphi(x-y)\right) \mathrm{d} y \\
\rightarrow & \int_{\mathbb{R}^{d}} f(y) \frac{\partial \varphi}{\partial x_{i}}(x-y) \mathrm{d} y .
\end{aligned}
$$

Proof. Let $i \in\{1, \ldots, N\}$ and let $e_{i} \in \mathbb{R}^{N}$ be the $i$-th canonical unit basis vector.
Lemma 10.22 (Strong continuity of the shift-group). For every $x \in \mathbb{R}^{N}$ and every $1 \leq p \leq \infty$ we define the shift operator $S(x) \in \mathscr{L}\left(\mathrm{L}^{p}\left(\mathbb{R}^{N} ; X\right)\right)$ by

$$
(S(x) f)(y):=f(x+y) \quad\left(f \in \mathrm{~L}^{p}\left(\mathbb{R}^{N} ; X\right), y \in \mathbb{R}^{N}\right)
$$

Then $S(x)$ is an isometric isomorphism and, if $p<\infty$,

$$
\lim _{x \rightarrow 0}\|S(x) f-f\|_{L^{p}}=0 \text { for every } f \in \mathbb{L}^{p}\left(\mathbb{R}^{N} ; X\right)
$$

Proof. The first statement about $S(x)$ being an isometric isomorphism is easy (with $S(x)^{-1}=S(-x)$ ). Next, for every simple step function $f=1_{Q} \otimes x$ with a cube $Q \subseteq$ $\mathbb{R}^{N}$, the second statement follows easily from Lebesgue's dominated convergence theorem. By linearity, the second statement holds for every $f$ in the dense subspace

$$
D:=\operatorname{span}\left\{1_{Q} \otimes x: Q \subseteq \mathbb{R}^{N} \text { a cube, } x \in X\right\} .
$$

Now fix $f \in \mathrm{~L}^{p}\left(\mathbb{R}^{N} ; X\right)$ and let $\varepsilon>0$. Then there exists $g \in D$ such that $\|f-g\|_{L^{p}}<$ $\varepsilon$. Moreover, there exists $\delta>0$ such that $\|S(x) g-g\|_{L^{p}}<\varepsilon$ for every $x \in \mathbb{R}^{N}$ with $\|x\|_{X}<\delta$. Hence, for every $x \in \mathbb{R}^{N}$ with $\|x\|_{X}<\delta$

$$
\begin{aligned}
\|S(x) f-f\|_{L^{p}} & \leq\|S(x) f-S(x) g\|_{L^{p}}+\|S(x) g-g\|_{L^{p}}+\|g-f\|_{L^{p}} \\
& \leq 2\|g-f\|_{L^{p}}+\|S(x) g-g\|_{L^{p}} \\
& <3 \varepsilon .
\end{aligned}
$$

If $\varphi \in \mathrm{L}^{1}\left(\mathbb{R}^{N}\right)$ is such that $\int_{\mathbb{R}^{N}} \varphi=1$, then we call the sequence $\left(\varphi_{n}\right)_{n}$ given by

$$
\varphi_{n}(x):=n^{N} \varphi(n x) \quad\left(x \in \mathbb{R}^{N}, n \in \mathbb{N}\right)
$$

an approximate identity or an approximate unit. The reason for this notion follows from the following lemma.

Lemma 10.23 (Property of an approximate identity). Let $f \in \mathrm{~L}^{p}\left(\mathbb{R}^{N} ; X\right)(1 \leq$ $p<\infty)$ and let $\left(\varphi_{n}\right)_{n}$ be an approximate identity. Then

$$
\lim _{n \rightarrow \infty} f * \varphi_{n}=f \text { in } \mathrm{L}^{p}\left(\mathbb{R}^{N} ; X\right)
$$

Proof. By Tonnelli's theorem, the Hölder inequality, by the strong continuity of the shift-group and by Lebesgue's dominated convergence theorem we have

$$
\begin{aligned}
\left\|f * \varphi_{n}-f\right\|_{\mathrm{L}^{p}}^{p} & =\int_{\mathbb{R}^{N}}\left|\int_{\mathbb{R}^{N}} f(x-y) \varphi_{n}(y) \mathrm{d} y-f(x)\right|^{p} \mathrm{~d} x \\
& \leq \int_{\mathbb{R}^{N}}\left(\int_{\mathbb{R}^{N}}\|f(x-y)-f(x)\|\left|\varphi_{n}(y)\right| \mathrm{d} y\right)^{p} \mathrm{~d} x \\
& \leq \int_{\mathbb{R}^{N}} \int_{\mathbb{R}^{N}}\|f(x-y)-f(x)\|^{p}\left|\varphi_{n}(y)\right| \mathrm{d} y\left\|\varphi_{n}\right\|_{\mathrm{L}^{1}}^{p-1} \mathrm{~d} x \\
& =\left\|\varphi_{n}\right\|_{\mathrm{L}^{1}}^{p-1} \int_{\mathbb{R}^{N}} \int_{\mathbb{R}^{N_{2}}}\|f(x-y)-f(x)\|^{p} \mathrm{~d} x\left|\varphi_{n}(y)\right| \mathrm{d} y \\
& =\left\|\varphi_{n}\right\|_{\mathrm{L}^{1}}^{p-1} \int_{\mathbb{R}^{N}} \int_{\mathbb{R}^{N}}\left\|f\left(x-\frac{y}{n}\right)-f(x)\right\| \mathrm{d} x \varphi(y) \mathrm{d} y \\
& \rightarrow 0 \quad(n \rightarrow \infty) .
\end{aligned}
$$

Corollary 10.24. For every $1 \leq p<\infty$ the space $\mathrm{C}_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{N} ; X\right)$ is dense in $\mathrm{L}^{p}\left(\mathbb{R}^{N} ; X\right)$.
Proof (by regularization and truncation). Let $f \in \mathrm{~L}^{p}\left(\mathbb{R}^{N} ; X\right)$. In the first step, the regularization step, we choose an approximate identity $\left(\varphi_{n}\right)$ starting with a test function $\varphi \in \mathrm{C}_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{N}\right)$. By Young's inequality, $f * \varphi_{n} \in \mathrm{~L}^{p}\left(\mathbb{R}^{N} ; X\right)$, by Lemma 10.21, $f * \varphi_{n} \in \mathrm{C}^{\infty}\left(\mathbb{R}^{N}\right)$, and by Lemma 10.23,

$$
\lim _{n \rightarrow \infty}\left\|f * \varphi_{n}-f\right\|_{L^{p}}=0
$$

In the second step, the truncation step, we choose a sequence $\left(\psi_{m}\right)_{m}$ of test functions satisfying $0 \leq \psi_{m} \leq 1$ and $\psi_{m}=1$ on the ball $B(0, m)$ (such functions can be obtained by convolving characteristic functions $\chi_{B(0,2 m)}$ with appropriate positive test functions, relying on Lemma 10.21). It is clear from Lebesgue's dominated convergence theorem, that for every $g \in \mathrm{~L}^{p}\left(\mathbb{R}^{N} ; X\right)$ one has

$$
\lim _{m \rightarrow \infty}\left\|g \psi_{m}-g\right\|_{L^{p}}=0
$$

Combining the preceding two equalities, we find a sequence $\left(m_{n}\right)_{n}$ in $\mathbb{N}$ such that

$$
\lim _{n \rightarrow \infty}\left\|\left(f * \varphi_{n}\right) \psi_{m_{n}}-f\right\|_{L^{p}}=0
$$

and since $\left(f * \varphi_{n}\right) \psi_{m_{n}} \in \mathrm{C}_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{N} ; X\right)$, the claim is proved.
Corollary 10.25. Let $f \in \mathrm{~L}_{l o c}^{1}\left(\mathbb{R}^{N} ; X\right)$ be such that

$$
\int_{\mathbb{R}^{N}} f \varphi=0 \text { for every } \varphi \in \mathscr{D}\left(\mathbb{R}^{N}\right)
$$

Then $f=0$.
Proof. The assumption implies that

$$
f * \varphi(x)=\int_{\mathbb{R}^{N}} f(y) \varphi(x-y) \mathrm{d} y=0 \text { for every } x \in \mathbb{R}^{N}, \varphi \in \mathscr{D}\left(\mathbb{R}^{N}\right),
$$

which just means that

$$
f * \varphi=0 \text { for every } \varphi \in \mathscr{D}\left(\mathbb{R}^{N}\right)
$$

The claim now follows upon choosing an approximate identity $\left(\varphi_{n}\right)$ out of a test function $\varphi$ and by applying Lemma 10.23.

### 10.4 Bochner-Sobolev spaces

Let $\Omega \subseteq \mathbb{R}^{N}$ be an open set, $1 \leq p \leq \infty$ and $k \in \mathbb{N}$. We define the Bochner-Sobolev space

$$
\begin{gathered}
W^{k, p}(\Omega ; X):=\left\{u \in \mathrm{~L}^{p}(\Omega ; X): \forall \alpha \in \mathbb{N}_{0}^{N} \exists v_{\alpha} \in \mathrm{L}^{p}(\Omega ; X) \forall \varphi \in \mathscr{D}(\Omega)\right. \\
\left.\int_{\Omega} u \partial^{\alpha} \varphi=(-1)^{|\alpha|} \int_{\Omega} v_{\alpha} \varphi\right\}
\end{gathered}
$$

The functions $v_{\alpha}$ in this definition of the space $W^{k, p}(\Omega ; X)$ are uniquely determined. We write $v_{\alpha}=: \partial^{\alpha} u$ and we call the function $\partial^{\alpha} u$ the weak $\alpha$-th partial derivative of $u$. The space $W^{k, p}(\Omega ; X)$ becomes a Banach space for the norm

$$
\|u\|_{W^{k, p}}:=\sum_{\substack{\alpha \in \mathbb{N}_{0}^{N} \\|\alpha| \leq k}}\left\|\partial^{\alpha} u\right\|_{L^{p}} .
$$

Similarly as in the case of the $\mathrm{L}^{p}$-spaces we write $W^{k, p}(a, b ; X)$ instead of $W^{k, p}((a, b) ; X)$. In the special case when $p=2$ and $X=H$ is a Hilbert space, we also write

$$
H^{k}(\Omega ; H):=W^{k, 2}(\Omega ; H)
$$

This space is a Hilbert space for the inner product

$$
\langle u, v\rangle_{H^{k}}:=\sum_{\substack{\alpha \in \mathbb{N}_{N}^{N} \\|\alpha| \leq k}}\left\langle\partial^{\alpha} u, \partial^{\alpha} v\right\rangle_{\mathrm{L}^{2}} .
$$

The resulting norm $\|\cdot\|_{H^{k}}$ is equivalent to the norm $\|\cdot\|_{W^{k, 2}}$ defined above.

The main results about Sobolev spaces of scalar-valued functions remain true for Sobolev spaces of Banach space valued functions if interpreted properly. In particular, the Sobolev embedding theorem, a version of the product rule, the integration by parts formula and Poincaré's inequality remain true. Even a version of the RellichKondrachev theorem remains true.

Lemma 10.26. For every $-\infty<a<b<\infty$ and every $1 \leq p \leq \infty$ one has $W^{1, p}(a, b ; X) \subseteq \mathrm{C}^{\mathrm{b}}(\overline{(a, b)} ; X)$. For every $u \in W^{1, p}(a, b ; X)$ and every $s, t \in(a, b)$ one has

$$
u(t)-u(s)=\int_{s}^{t} u^{\prime}(r) \mathrm{d} r
$$

Lemma 10.27. Assume that the embedding $V \hookrightarrow H$ is continuous and let $u \in$ $W^{1,2}(0, T ; H) \cap \mathrm{L}^{\infty}(0, T ; V)$. Then $u$ is weakly continuous with values in $V$, that is, for every $v \in V^{\prime}$ the function $t \mapsto\langle v, u(t)\rangle_{V^{\prime}, V}$ is continuous on $[0, T]$.

Proof. Since every function $u \in W^{1,2}(0, T ; H)$ is continuous (and hence weakly continuous) with values in $H$, the claim follows from [Temam (1984), Lemma 1.4, page 263].

Lemma 10.28. Assume that the embedding $V \hookrightarrow H$ is continuous and let $\left(u_{n}\right)$ be a sequence such that

$$
\begin{aligned}
& u_{n} \rightharpoonup u \text { in } W^{1,2}(0, T ; H) \text { and } \\
& u_{n} \xrightarrow{\mathrm{w} *} u \text { in } L^{\infty}(0, T ; V) .
\end{aligned}
$$

Then there exists a subsequence of $\left(u_{n}\right)$ (which we denote again by $\left(u_{n}\right)$ ) such that

$$
u_{n}(t) \rightharpoonup u(t) \text { in } V \text { for every } t \in[0, T]
$$

Proof. Using the fact that the point evaluation in $t \in[0, T]$ from $W^{1,2}(0, T ; H)$ into $H$ is bounded and linear, and maps weakly convergent sequences into weakly convergent sequences, the assumption implies that for every $t \in[0, T]$

$$
u_{n}(t) \rightharpoonup u(t) \text { in } H
$$

Let now $w \in H^{\prime}$ and $t \in[0, T]$. Then one has

$$
\left\langle w, u_{n}(t)-u(t)\right\rangle_{V^{\prime}, V}=\left\langle w, u_{n}(t)-u(t)\right\rangle_{H^{\prime}, H} \longrightarrow 0 .
$$

Using the fact that $H^{\prime}$ is dense in $V^{\prime}$ and that the sequence $\left(u_{n}(t)\right)$ is bounded in $V$, the claim follows from Lemma ??.

## References

[Absil and Kurdyka (2006)] Absil, P.-A., Kurdyka, K. : On the stable equilibrium points of gradient systems. Systems Control Lett. 55 (7), 2006, 573-577. URL http://dx.doi.org/10.1016/j.sysconle. 2006.01 .002
[Adams (1975)] Adams, R. A. : Sobolev Spaces. Academic Press, New York, 1975.
[Amann (1995)] Amann, H. : Linear and Quasilinear Parabolic Problems: Abstract Linear Theory. Vol. 89 of Monographs in Mathematics. Birkhäuser Verlag, Basel, 1995.
[Ambrosio et al. (2005)] Ambrosio, L., Gigli, N., Savaré, G. : Gradient Flows. Lectures in Mathematics ETH Zürich. Birkhäuser, Basel, 2005.
[Angenent (1990)] Angenent, S. B. : Nonlinear analytic semiflows. Proc. Roy. Soc. Edinburgh 115A, 1990, 91-107.
[Arendt (2004)] Arendt, W., 2004. : Semigroups and evolution equations: functional calculus, regularity and kernel estimates. In: Handbook of Differential Equations (C. M. Dafermos, E. Feireisl eds.). Elsevier/North Holland, pp. 1-85.
[Aubin (1979)] Aubin, J. : Applied Functional Analysis. Pure Applied Mathematics. John Wiley \& Sons, New York, Chichester, Brisbane, Toronto, 1979.
[Aulbach (1984)] Aulbach, B. : Continuous and discrete dynamics near manifolds of equilibria. Vol. 1058 of Lecture Notes in Mathematics. Springer-Verlag, Berlin, 1984.
[Bénilan et al. (1999)] Bénilan, P., Crandall, M. G., Pazy, A. : Nonlinear Evolution Equations Governed by Accretive Operators, 1999, book manuscript.
[Beurling and Deny (1958)] Beurling, A., Deny, J. : Espaces de Dirichlet. I. Le cas élémentaire. Acta Math. 99, 1958, 203-224.
[Beurling and Deny (1959)] Beurling, A., Deny, J. : Dirichlet spaces. Proc. Nat. Acad. Sci. U.S.A. 45, 1959, 208-215.
[Brézis (1968)] Brézis, H. : Equations et inéquations non linéaires dans les espaces vectoriels en dualité. Ann. Inst. Fourier 18, 1968, 115-175.
[Brezis (1973)] Brezis, H. : Opérateurs maximaux monotones et semi-groupes de contractions dans les espaces de Hilbert. Vol. 5 of North Holland Mathematics Studies. North-Holland, Amsterdam, London, 1973.
[Brézis (1992)] Brézis, H. : Analyse fonctionnelle. Masson, Paris, 1992.
[Capuzzo Dolcetta et al. (2002)] Capuzzo Dolcetta, I., Finzi Vita, S., March, R. : Area-preserving curve-shortening flows: from phase separation to image processing. Interfaces Free Bound. 4 (4), 2002, 325-343.
URL http://dx.doi.org/10.4171/IFB/64
[Cartan (1967)] Cartan, H. : Cours de calcul différentiel. Hermann, Paris, 1967.
[Castaing and Valadier (1977)] Castaing, C., Valadier, M. : Convex analysis and measurable multifunctions. Lecture Notes in Mathematics, Vol. 580. Springer-Verlag, Berlin, 1977.
[Cauchy (1847)] Cauchy, A. : Méthodes générale pour la résolution des systèmes d'équations simultanées. C. R. Acad. Sci. Paris 25, 1847, 536-538.
[Chill (2003)] Chill, R. : On the Łojasiewicz-Simon gradient inequality. J. Funct. Anal. 201, 2003, 572-601.
[Chill (2006)] Chill, R., 2006. : The Łojasiewicz-Simon gradient inequality on Hilbert spaces. In: Proceedings of the $5^{\text {th }}$ European-Maghrebian Workshop on Semigroup Theory, Evolution Equations and Applications (2006). p. to appear.
[Chou and Zhu (2001)] Chou, K.-S., Zhu, X.-P. : The curve shortening problem. Chapman \& Hall/CRC, Boca Raton, FL, 2001.
[Cipriani and Grillo (2003)] Cipriani, F., Grillo, G. : Nonlinear Markov semigroups, nonlinear Dirichlet forms and application to minimal surfaces. J. reine angew. Math. 562, 2003, 201235.
[Clément and Li (1994)] Clément, P., Li, S. : Abstract parabolic quasilinear problems and application to a groundwater flow problem. Adv. Math. Sci. Appl. 3, 1994, 17-32.
[Coddington and Levinson (1955)] Coddington, E. A., Levinson, N. : Theory of Ordinary Differential Equations. McGraw-Hill Book Company, New York, Toronto, London, 1955.
[Couchouron (2002)] Couchouron, J.-F. : Compactness theorems for abstract evolution problems. J. Evolution Equations 2, 2002, 151-175.
[Couchouron (2010)] Couchouron, J.-F. : Semi-groupes et applications, 2010, Lecture Notes, Paul Verlaine University - Metz.
[Crandall and Evans (1975)] Crandall, M. G., Evans, L. C. : On the relation of the operator $\partial / \partial s+\partial / \partial \tau$ to evolution governed by accretive operators. Israel J. Math. 21 (4), 1975, 261278.
[Curry (1944)] Curry, H. B. : The method of steepest descent for non-linear minimization problems. Quart. Appl. Math. 2, 1944.
[Dautray and Lions (1985)] Dautray, R., Lions, J. : Analyse mathématique et calcul numérique pour les sciences et les techniques. Vol. I. INSTN: Collection Enseignement. Masson, Paris, 1985.
[Dautray and Lions (1987)] Dautray, R., Lions, J. : Analyse mathématique et calcul numérique pour les sciences et les techniques. Vol. VIII. INSTN: Collection Enseignement. Masson, Paris, 1987.
[Dautray and Lions (1988)] Dautray, R., Lions, J. : Analyse mathématique et calcul numérique pour les sciences et les techniques. Vol. VI. INSTN: Collection Enseignement. Masson, Paris, 1988.
[Deckelnick and Dziuk (2003)] Deckelnick, K., Dziuk, G., 2003. : Mean curvature flow and related topics. In: Frontiers in numerical analysis (Durham, 2002). Universitext. Springer, Berlin, pp. 63-108.
[Denk et al. (2003)] Denk, R., Hieber, M., Prüss, J. : $\mathscr{R}$-Boundedness, Fourier Multipliers and Problems of Elliptic and Parabolic Type. Vol. 166 of Memoirs Amer. Math. Soc. Amer. Math. Soc., Providence, R.I., 2003.
[Drábek and Milota (2007)] Drábek, P., Milota, J. : Methods of nonlinear analysis. Birkhäuser Advanced Texts: Basler Lehrbücher. [Birkhäuser Advanced Texts: Basel Textbooks]. Birkhäuser Verlag, Basel, 2007.
[Ecker (2004)] Ecker, K. : Regularity Theory for Mean Curvature Flow. Vol. 57 of Progress in Nonlinear Differential Equations and Their Applications. Birkäuser, Boston, Basel, Berlin, 2004.
[Ekeland (2006)] Ekeland, I. : The best of all possible worlds. University of Chicago Press, Chicago, IL, 2006.
[Escher et al. (2003)] Escher, J., Prüss, J., Simonett, G., 2003. : A new approach to the regularity of solutions of parabolic equations. In: Evolution Equations: Proceedings in Honor of J. A. Goldstein's 60th Birthday. Marcel Dekker, New York, pp. 167-190.
[Evans (1998)] Evans, L. C. : Partial Differential Equations. Vol. 19 of Graduate Studies in Mathematics. American Mathematical Society, Providence, RI, 1998.
[Fašangová (1998)] Fašangová, E. : Asymptotic analysis for a nonlinear parabolic equation on $\mathbb{R}$. Comment Math. Univ. Carolin. 39, 1998, 525-544.
[Fašangová and Feireisl (1999)] Fašangová, E., Feireisl, E. : The long-time behaviour of solutions to parabolic problems on unbounded intervals: the influence of boundary conditions. Proc. Roy. Soc. Edinburgh 129A, 1999, 319-329.
[Feireisl and Simondon (2000)] Feireisl, E., Simondon, F. : Convergence for semilinear degenerate parabolic equations in several space dimensions. J. Dynam. Differential Equations 12, 2000, 647-673.
[Fife (2000)] Fife, P. C. : Models for phase separation and their mathematics. Electron. J. Differential Equations, 2000, No. 48, 26 pp. (electronic).
[Fife (2002)] Fife, P. C., 2002. : Pattern formation in gradient systems. In: Handbook of dynamical systems, Vol. 2. North-Holland, Amsterdam, pp. 677-722.
URL http://dx.doi.org/10.1016/S1874-575X(02) 80034-0
[Gage (1983)] Gage, M. E. : An isoperimetric inequality with applications to curve shortening. Duke Math. J. 50 (4), 1983, 1225-1229. URL http://dx.doi.org/10.1215/S0012-7094-83-05052-4
[Gage (1984)] Gage, M. E. : Curve shortening makes convex curves circular. Invent. Math. 76 (2), 1984, 357-364. URL http://dx.doi.org/10.1007/BF01388602
[Galaktionov et al. (2007)] Galaktionov, V. A., Pokhozhaev, S. I., Shishkov, A. E. : On convergence in gradient systems with a degenerate equilibrium position. Mat. Sb. 198 (6), 2007, 65-88. URL http://dx.doi.org/10.1070/SM2007v198n06ABEH003862
[Gilbarg and Trudinger (2001)] Gilbarg, D., Trudinger, N. S. : Elliptic Partial Differential Equations of Second Order. Springer Verlag, Berlin, Heidelberg, New York, 2001.
[Goldstein (1962)] Goldstein, A. A. : Cauchy's method of minimization. Numer. Math. 4, 1962, 146-150.
[Hahn (1963)] Hahn, W. : Theory and application of Liapunov's direct method. English edition prepared by Siegfried H. Lehnigk; translation by Hans H. Losenthien and Siegfried H. Lehnigk. Prentice-Hall Inc., Englewood Cliffs, N.J., 1963.
[Haraux (1990)] Haraux, A. : Systèmes dynamiques dissipatifs et applications. Masson, Paris, 1990.
[Haraux and Jendoubi (2007)] Haraux, A., Jendoubi, M. A. : On the convergence of global and bounded solutions of some evolution equations. J. Evol. Equ. 7 (3), 2007, 449-470. URL http://dx.doi.org/10.1007/s00028-007-0297-8
[Haraux et al. (2003)] Haraux, A., Jendoubi, M. A., Kavian, O. : Rate of decay to equilibrium in some semilinear parabolic equations. J. Evolution Equations 3, 2003, 463-484.
[Hille and Phillips (1957)] Hille, E., Phillips, R. S. : Functional Analysis and Semi-Groups. Amer. Math. Soc., Providence, R.I., 1957.
[Jendoubi (1998)] Jendoubi, M. A. : A simple unified approach to some convergence theorems of L. Simon. J. Funct. Anal. 153, 1998, 187-202.
[Kantorovič (1947)] Kantorovič, L. V. : On the method of steepest descent. Doklady Akad. Nauk SSSR (N. S.) 56, 1947, 233-236.
[Krantz and Parks (2002a)] Krantz, S. G., Parks, H. R. : The implicit function theorem. Birkhäuser Boston Inc., Boston, MA, 2002a.
[Krantz and Parks (2002b)] Krantz, S. G., Parks, H. R. : A primer of real analytic functions, 2nd Edition. Birkhäuser Advanced Texts: Basler Lehrbücher. [Birkhäuser Advanced Texts: Basel Textbooks]. Birkhäuser Boston Inc., Boston, MA, 2002b.
[Kunstmann and Weis (2004)] Kunstmann, P. C., Weis, L., 2004. : Maximal L ${ }^{p}$ regularity for parabolic equations, Fourier multiplier theorems and $H^{\infty}$ functional calculus. In: Levico Lectures, Proceedings of the Autumn School on Evolution Equations and Semigroups (M. Iannelli, R. Nagel, S. Piazzera eds.). Vol. 69. Springer Verlag, Heidelberg, Berlin, pp. 65-320.
[Ladyženskaja et al. (1967)] Ladyženskaja, O. A., Solonnikov, V. A., Ural'ceva, N. N. : Linear and quasilinear equations of parabolic type. Translated from the Russian by S. Smith. Translations of Mathematical Monographs, Vol. 23. American Mathematical Society, Providence, R.I., 1967.
[Laplace (1878-1912a)] Laplace, P. S. d. : Oeuvres complètes de Laplace. Tome 10 /publiées sous les auspices de l'Académie des sciences, par MM. les secrétaires perpétuels. Gauthier-Villars, Paris, 1878-1912a.
[Laplace (1878-1912b)] Laplace, P. S. d. : Oeuvres complètes de Laplace. Tome 12 /publiées sous les auspices de l'Académie des sciences, par MM. les secrétaires perpétuels. Gauthier-Villars, Paris, 1878-1912b.
[Lasalle (1962)] Lasalle, J. P., 1962. : Asymptotic stability criteria. In: Proc. Symp. Appl. Math., Vol. XIII. American Mathematical Society, Providence, R.I., pp. 299-307.
[Liapunov (1966)] Liapunov, A. M. : Stability of motion. With a contribution by V. A. Pliss and an introduction by V. P. Basov. Translated from the Russian by Flavian Abramovici and Michael Shimshoni. Mathematics in Science and Engineering, Vol. 30. Academic Press, New York, 1966.
[Lions (1969)] Lions, J. : Quelques méthodes de résolution des problèmes aux limites non linéaires. Dunod, Gauthier-Villars, Paris, 1969.
[Matano (1978)] Matano, H. : Convergence of solutions of one-dimensional semilinear heat equations. J. Math. Kyoto Univ. 18, 1978, 221-227.
[Neuberger (1997)] Neuberger, J. W. : Sobolev Gradients and Differential Equations. Vol. 1670 of Lect. Notes Math. Springer Verlag, Berlin, Heidelberg, New York, 1997.
[Neuberger (2010)] Neuberger, J. W. : Sobolev gradients and differential equations, 2nd Edition. Vol. 1670 of Lecture Notes in Mathematics. Springer-Verlag, Berlin, 2010.
[Oğuztöreli et al. (1981)] Oğuztöreli, M. N., Lakshmikantham, V., Leela, S. : An algorithm for the construction of Liapunov functions. Nonlinear Anal. 5 (11), 1981, 1195-1212.
URL http://dx.doi.org/10.1016/0362-546X(81) 90013-4
[Otto (2001)] Otto, F. : The geometry of dissipative evolution equations: the porous medium equation. Comm. Partial Differential Equations 26 (1-2), 2001, 101-174.
[Ouhabaz (1992)] Ouhabaz, E.-M. : $L^{\infty}$-contractivity of semigroups generated by sectorial forms. J. London Math. Soc. (2) 46 (3), 1992, 529-542. URL http://dx.doi.org/10.1112/jlms/s2-46.3.529
[Ouhabaz (1996)] Ouhabaz, E.-M. : Invariance of closed convex sets and domination criteria for semigroups. Potential Anal. 5 (6), 1996, 611-625. URL http://dx.doi.org/10.1007/BF00275797
[Ouhabaz (2004)] Ouhabaz, E. M. : Analysis of Heat Equations on Domains. Vol. 30 of London Mathematical Society Monographs. Princeton University Press, Princeton, 2004.
[Phelps (1993)] Phelps, R. R. : Convex functions, monotone operators and differentiability, 2nd Edition. Vol. 1364 of Lecture Notes in Mathematics. Springer-Verlag, Berlin, 1993.
[Rudin (1973)] Rudin, W. : Functional analysis. McGraw-Hill Book Co., New York, 1973.
[Rudin (1974)] Rudin, W. : Real and complex analysis, 2nd Edition. McGraw-Hill Book Co., New York, 1974.
[Siskakis (1996)] Siskakis, A. G. : Semigroups of composition operators on the Dirichlet space. Results Math. 30 (1-2), 1996, 165-173. URL http://dx.doi.org/10.1007/BF03322189
[Tarasov (2005)] Tarasov, V. E. : Fractional generalization of gradient systems. Lett. Math. Phys. 73 (1), 2005, 49-58. URL http://dx.doi.org/10.1007/s11005-005-8444-z
[Temam (1984)] Temam, R. : Navier-Stokes Equations. Vol. 2 of Studies in Mathematics and its Applications. Elsevier Science Publishers, 1984.
[Triebel (1983)] Triebel, H. : Theory of Function Spaces. Birkhäuser, Basel, 1983.
[Vázquez (2007)] Vázquez, J. L. : The porous medium equation. Oxford Mathematical Monographs. The Clarendon Press Oxford University Press, Oxford, 2007.
[Villani (2003)] Villani, C. : Topics in optimal transportation. Vol. 58 of Graduate Studies in Mathematics. American Mathematical Society, Providence, RI, 2003.
[Villani (2009)] Villani, C. : Optimal transport. Vol. 338 of Grundlehren der Mathematischen Wissenschaften [Fundamental Principles of Mathematical Sciences]. Springer-Verlag, Berlin, 2009.

URL http://dx.doi.org/10.1007/978-3-540-71050-9
[Vrabie (2004)] Vrabie, I. I. : Differential equations. World Scientific Publishing Co. Inc., River Edge, NJ, 2004.
[Werner (1997)] Werner, D. : Funktionalanalysis. Springer Verlag, Berlin, Heidelberg, New York, 1997.
[Wolfe (1969)] Wolfe, P. : Convergence conditions for ascent methods. SIAM Rev. 11, 1969, 226235.
[Zambotti (2006)] Zambotti, L. : Convergence of approximations of monotone gradient systems. J. Evol. Equ. 6 (4), 2006, 601-619. URL http://dx.doi.org/10.1007/s00028-006-0275-6
[Zelenyak (1968)] Zelenyak, T. I. : Stabilization of solutions of boundary value problems for a second-order parabolic equation with one space variable. Differ. Eq. 4, 1968, 17-22, transl. from Differ. Uravn. 4 (1968), 34-45.

## Index

absolutely convergent, 15
adjoint operator, 78, 111, 124
algebra
Banach, 97
$C^{*}, 112$
normed, 97
annihilator, 80
approximate identity, 166
approximate unit, 166
approximative point spectrum, 76
Arezlà-Ascoli, 25

Banach algebra, 97
unital, 98
Banach space, 14
Bessel inequality, 35, 37
bidual space, 54
Bochner integral, 160
Borel measure, 119
finite, 119
regular, 119
boundary, 3
bounded operator, 21
Cauchy sequence, 4
Cauchy-Schwarz inequality, 30
Cesàro mean, 42
character, 101
closed, 3
operator, 69
closure, 3
coercive, 61
compact
operator, 81
compact space, 6
sequentially, 6
complemented, 51
complete
inner product space, 31
metric space, 5
normed algebra, 97
normed space, 14
completion
inner product space, 31
metric space, 8
normed space, 17
concave, 62
strictly, 62
continuous, 7
Lipschitz, 7
sequentially, 7
spectrum, 76
uniformly, 7
convergent, 4
series, 15
unconditionally, 37
weak*, 53
weakly, 43, 58
convex, 31, 59, 61
convolution, $97,143,165$
derivative, 137
differentiable, 137
partially, 138
domain, 75
dual space, 47
elliptic problem
weak solution, 128
equation
Schrödinger, 134
equicontinuous, 25
equilibrium, 62
finite Borel measure, 119
finite rank, 82
Fourier
coefficient, 38
series, 38
transform, 38
Fredholm index, 85
Fredholm operator, 85
function
almost separably valued, 158
analytic, 72, 76, 99
coercive, 61
concave, 62
continuous, 7
convex, 61
differentiable, 137
holomorphic, 72
integrable, 159
Lipschitz continuous, 7
measurable, 157
partially differentiable, 138
sequentially continuous, 7
step function, 157
strictly convex, 61
sublinear, 47
support, 143
test function, 143, 165
uniformly continuous, 7
weakly analytic, 72
weakly holomorphic, 72
weakly measurable, 158
functional
positive, 119
Gelfand space, 102
Gelfand transform, 106
Gram-Schmidt process, 34
graph norm, 70
heat equation, 131
mild solution, 131
Hilbert space, 31
identity
parallelogram, 31
Parseval, 36
resolvent, 76, 99
inequality
Bessel, 35, 37
Cauchy-Schwarz, 30
Poincaré, 151, 153
triangle, 1,11
from below, 13
Young, 98, 144, 164
inner product space, 29
integrable function, 159
integral, 160
interior, 3
involution, 107, 112
isometry, 24
isomorphic, 24
isometrically, 24
isomorphism, 24
kernel, 75
kernel operator, 84
Laplace operator, 128, 154
Lebesgue's theorem, 161

## Lemma

Baire, 65, 66
Neumann series, 24, 99
Pythagoras, 32
Riemann-Lebesgue, 38
Riesz, 16
Zorn, 48
Lipschitz continuous, 7
maximal ideal, 101
measurable function, 157
metric, 1
discrete, 2
induced, 2
metric space, 1
completion, 8
mild solution, 131, 134
Minkowski functional, 59
multi-index, 165
multiplication operator, 22, 83
neighbourhood, 3
Neumann series, 24, 99
Newton's method, 142
norm, 11
equivalent, 14
graph, 70
normal operator, 113
normed algebra, 97
normed space, 11
completion, 17
nuclear operator, 88
numerical range, 113
open, 3
operator
adjoint, 78, 111, 124
closed, 69, 75
compact, 81
finite rank, 82
Fredholm, 85
isometry, 24
isomorphism, 24
kernel, 84
Laplace, 128
left-shift, 22
multiplication, 22, 83
normal, 113
nuclear, 88
positive semidefinite, 113
powerbounded, 89
projection, 33, 51
resolvent, 76
right-shift, 22
selfadjoint, 113, 124
symmetric, 113, 124
unitary, 36, 113
orthogonal
complement, 32
space, 32
vectors, 32
parallelogram identity, 31
Parseval identity, 36
partial derivative, 165
weak partial derivative, 168
partially differentiable, 138
Pettis' theorem, 158
Poincaré inequality, 151, 153
point spectrum, 76
approximative, 76
positive functional, 119
positive semidefinite, 113
powerbounded, 89
preannihilator, 80
product space, 2,17
projection, 33, 51
quotient space, 18
range, 75
reflexive, 54
regular Borel measure, 119
regularization, 165
residual spectrum, 76
resolvent, 76, 99
identity, 76, 99
resolvent set, 75
Riesz Lemma, 16
Riesz-Markov, 119
saddle point, 62
Schrödinger equation, 134
selfadjoint operator, 113, 124
separable, 33
sequence
Cauchy, 4
convergent, 4
weak*, 53
weakly, 43, 58
sequentially closed, 4
sequentially continuous, 7
series
absolutely convergent, 15
convergent, 15
Fourier, 38
unconditionally convergent, 37
set
boundary, 3
closed, 3
closure, 3
equicontinuous, 25
interior, 3
neighbourhood, 3
open, 3
sequentially closed, 4
shift operator, 22
shift-group, 166
Sobolev space, 147, 150, 152, 153, 168
space
Banach, 14
bidual, 54
Bochner-Lebesgue, 162
compact, 6
complemented, 51
dual, 47
Hilbert, 31
inner product, 29
isomorphic, 24
metric, 1
normed, 11
of test functions, 143
product, 2, 17
quotient, 18
reflexive, 54
separable, 33
sequentially compact, 6
Sobolev, 147, 150, 152, 153
Sobolev space, 168
spectral radius, 78,100
spectrum, 76, 99, 102
approximative point, 76
continuous, 76
point, 76
residual, 76
step function, 157
sublinear, 47
support, 143, 165
symmetric operator, 113, 124
test function, 143, 165
Theorem
Arezlà-Ascoli, 25
Banach - Alaoglu, 53
Banach-Steinhaus, 67
Bounded inverse theorem, 69
Closed graph theorem, 70
Féjer, 42
Fredholm alternative, 84
Gelfand, 106
Gelfand - Mazur, 100, 103
Hahn - Banach geometric version, 59 version of functional analysis, 49 version of linear algebra, 47, 49
Hellinger-Toeplitz, 113
Implicit function theorem, 138
Local inverse function theorem, 138
Mean ergodic theorem, $90,91,93,95$
Minimization of convex functionals, 61
Open mapping theorem, 68
Plancherel, 41
projection onto closed, convex sets, 31
Rellich-Kondrachov, 154
Riesz - Fréchet, 42

Riesz-Markov, 119
Riesz-Schauder, 84
Schauder, 84
Sobolev embedding, 149, 154
Spectral theorem, 122
Uniform boundedness principle, 67
von Neumann mean ergodic, 92
von Neumann minimax, 62
theorem
Fischer-Riesz, 163
Lebesgue, 161
Pettis, 158
Young, 164
topology, 4
local uniform convergence, 4
triangle inequality, 1,11
from below, 13
unconditionally convergent, 37
uniformly continuous, 7
unitarily equivalent, 36
unitary operator, 36,113
weak solution, 128
weakly convergent, 43,58
weakly measurable, 158
Young's inequality, 98, 144


[^0]:    ${ }^{1}$ We are grateful to Anton Claußnitzer for the idea of the definition of the sets $A_{m n}$ and the functions $f_{m}$.

