## MAA6617 COURSE NOTES SPRING 2018

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## 20. Normed vector spaces

Let $\mathcal{X}$ be a vector space over a field $\mathbb{K}$ (in this course we always have either $\mathbb{K}=\mathbb{R}$ or $\mathbb{K}=\mathbb{C})$.

Definition 20.1. A norm on $\mathcal{X}$ is a function $\|\cdot\|: \mathcal{X} \rightarrow \mathbb{R}$ satisfying:
(i) (positivity) $\|x\| \geq 0$ for all $x \in \mathcal{X}$, and $\|x\|=0$ if and only if $x=0$;
(ii) (homogeneity) $\|k x\|=|k|\|x\|$ for all $x \in \mathcal{X}$ and $k \in \mathbb{K}$, and
(iii) (triangle inequality) $\|x+y\| \leq\|x\|+\|y\|$ for all $x, y \in \mathcal{X}$.

Using these three properties it is straightforward to check that the quantity

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

defines a metric on $\mathcal{X}$. The resulting topology is the norm topology. The next proposition is simple but fundamental; it says that the norm and the vector space operations are continuous in the norm topology.
normed-tvs Proposition 20.2 (Continuity of vector space operations). Let $\mathcal{X}$ be a normed vector space over $\mathbb{K}$.
a) If $\left(x_{n}\right)$ converges to $x$ in $\mathcal{X}$, then $\left(\left\|x_{n}\right\|\right)$ converges to $\|x\|$ in $\mathbb{R}$.
b) If $\left(k_{n}\right)$ converges to $k$ in $\mathbb{K}$ and $\left(x_{n}\right)$ converges to $x$ in $\mathcal{X}$, then $\left(k_{n} x_{n}\right)$ converges to $k x$ in $\mathcal{X}$.
c) If $\left(x_{n}\right)$ converges to $x$ and ( $y_{n}$ ) converges to $y$ in $\mathcal{X}$, then $\left(x_{n}+y_{n}\right)$ converges to $x+y$ in $\mathcal{X}$.

Proof. The proofs follow readily from the properties of the norm, and are left as exercises.

Two norms $\|\cdot\|_{1},\|\cdot\|_{2}$ on $\mathcal{X}$ are equivalent if there exist absolute constants $C, c>0$ such that

$$
c\|x\|_{1} \leq\|x\|_{2} \leq C\|x\|_{1} \quad \text { for all } x \in \mathcal{X} .
$$

Equivalent norms determine the same topology on $\mathcal{X}$ and the same Cauchy sequences (Problem ${ }^{\text {prob: equivalent-norms }}$ 20.2). A normed space is a Banach space if it is complete in the norm topology. It follows that if $\mathcal{X}$ is equipped with two equivalent norms $\|\cdot\|_{1},\|\cdot\|_{2}$ then it is complete (a Banach space) in one norm if and only if it is complete in the other.

The following proposition gives a convenient criterion for a normed vector space to be complete. A series $\sum_{n=1}^{\infty} x_{n}$ in $\mathcal{X}$ is absolutely convergent if $\sum_{n=1}^{\infty}\left\|x_{n}\right\|<\infty$. The series converges in $\mathcal{X}$ if the limit $\lim _{N \rightarrow \infty} \sum_{n=1}^{N} x_{n}$ exists in $\mathcal{X}$ (in the norm topology). (Quite explicitly, the series $\sum_{n=1}^{\infty} x_{n}$ converges to $x \in \mathcal{X}$ if $\lim _{N \rightarrow \infty}\left\|x-\sum_{n=1}^{N} x_{n}\right\|=0$.)

Proposition 20.3. A normed space $(\mathcal{X},\|\cdot\|)$ is complete if and only if every absolutely convergent series in $\mathcal{X}$ is convergent.

Proof. First suppose $\mathcal{X}$ is complete and $\sum_{n=1}^{\infty} x_{n}$ is absolutely convergent. Write $s_{N}=$ $\sum_{n=1}^{N} x_{n}$ for the $N^{t h}$ partial sum and let $\epsilon>0$ be given. Since $\sum_{n=1}^{\infty}\left\|x_{n}\right\|$ is convergent, there exists an $L$ such that $\sum_{n=L}^{\infty}\left\|x_{n}\right\|<\epsilon$. If $N>M \geq L$, then

$$
\left\|s_{N}-s_{M}\right\|=\left\|\sum_{n=M+1}^{N} x_{n}\right\| \leq \sum_{n=M+1}^{N}\left\|x_{n}\right\|<\epsilon
$$

Thus the sequence $\left(s_{N}\right)$ is Cauchy in $\mathcal{X}$, hence convergent by hypothesis.

Conversely, suppose every absolutely convergent series in $\mathcal{X}$ is convergent. Given a Cauchy sequence $\left(x_{n}\right)$ from $X$, choose a super-Cauchy subsequence $\left(y_{k}\right)$; i.e., $\left(y_{k}=x_{n_{k}}\right)_{k}$ and

$$
\sum_{k=1}^{\infty}\left\|y_{k+1}-y_{k}\right\|<\infty
$$

(To do this, first choose $N_{1}$ such that $\left\|x_{n}-x_{m}\right\|<2^{-1}$ for all $n, m \geq N_{1}$. Next choose $N_{2}>N_{1}$ such that $\left\|x_{n}-x_{m}\right\|<2^{-2}$ for all $n, m \geq N_{2}$. Continuing in this way recursively defines an increasing sequence of integers $\left(N_{k}\right)_{k=1}^{\infty}$ such that $\left\|x_{n}-x_{m}\right\|<2^{-k}$ for all $n, m \geq N_{k}$. Set $y_{k}=x_{N_{k}}$.) The series $\sum_{k=1}^{\infty}\left(y_{k+1}-y_{k}\right)$ is absolutely convergent and hence, by hypothesis, convergent in $\mathcal{X}$. In other words, the sequence $\left(y_{k}-y_{1}\right)_{k}$ of partial sums converges in $\mathcal{X}$ which means that $\left(x_{n}\right)$ has a convergent subsequence. The proof is finished by invoking a standard fact about convergence in metric spaces: if $\left(x_{n}\right)$ is a Cauchy sequence which has a convergent subsequence, then the full sequence converges.

### 20.1. Examples.

(a) Of course, $\mathbb{K}^{n}$ with the usual Euclidean norm $\left\|\left(x_{1}, \ldots x_{n}\right)\right\|=\left(\sum_{k=1}^{n}\left|x_{k}\right|^{2}\right)^{1 / 2}$ is a Banach space. The vector space $\mathbb{K}^{n}$ can also be equipped with the $\ell^{p}$-norms

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

for $1 \leq p<\infty$, and the $\ell^{\infty}$-norm

$$
\left\|\left(x_{1}, \ldots x_{n}\right)\right\|_{\infty}:=\max \left(\left|x_{1}\right|, \ldots\left|x_{n}\right|\right)
$$

For $1 \leq p<\infty$ and $p \neq 2$, it is not immediately obvious that $\|\cdot\|_{p}$ defines a norm. We will prove this assertion later. It is not too hard to show that all of the $\ell^{p}$ norms $(1 \leq p \leq \infty)$ are equivalent on $\mathbb{K}^{n}$ (though the constants $c, C$ depend on the dimension $n$ ). It turns out that any two norms on a finite-dimensional vector space are equivalent. As a corollary, every finite-dimensional normed space is a Banach space. See Problem prob:findim-basics
(b) (Sequence spaces) Define

$$
\begin{aligned}
c_{0} & :=\left\{f: \mathbb{N} \rightarrow \mathbb{K}\left|\lim _{m \rightarrow \infty}\right| f(m) \mid=0\right\} \\
\ell^{\infty} & :=\left\{f: \mathbb{N} \rightarrow \mathbb{K}\left|\sup _{m \in \mathbb{N}}\right| f(m) \mid<\infty\right\} \\
\ell^{1} & :=\left\{f: \mathbb{N} \rightarrow \mathbb{K}\left|\sum_{m=0}^{\infty}\right| f(m) \mid<\infty\right\} .
\end{aligned}
$$

It is a simple exercise to check that each of these is a vector space (a subspace of the vector space of all functions $f: \mathbb{N} \rightarrow \mathbb{K}$ ). Define, for functions $f: \mathbb{N} \rightarrow \mathbb{K}$,

$$
\begin{aligned}
\|f\|_{\infty} & :=\sup _{m}|f(m)| \\
\|f\|_{1} & :=\sum_{m=1}^{\infty}|f(m)|
\end{aligned}
$$

Then $\|f\|_{\infty}$ defines a norm on both $c_{0}$ and $\ell^{\infty}$, and $\|f\|_{1}$ is a norm on $\ell^{1}$. Equipped with these respective norms, each is a Banach space. We sketch the proof for $c_{0}$. Verification of the other two assertions is left as exercises (Problem prob:norm-examples

The key observation is that $\left(f_{n}\right)$ converges to $f$ in the $\|\cdot\|_{\infty}$ norm if and only if $\left(f_{n}\right)$ converges to $f$ uniformly as functions on $\mathbb{N}$. Suppose $\left(f_{n}\right)$ is a Cauchy sequence
in $c_{0}$. Then the sequence of functions $f_{n}$ is uniformly Cauchy on $\mathbb{N}$, and in particular converges pointwise to a function $f$. To this end, let $\epsilon>0$ be given. There is an $N$ so that $\left\|f_{m}-f_{n}\right\|<\epsilon$ for $m, n \geq N$. Thus, for each $n \geq N$, all $\ell \in \mathbb{N}$ and all $m \geq n \geq N,\left|f_{m}(\ell)-f_{n}(\ell)\right|<\epsilon$ and thus, $\left|f(\ell)-f_{N}(\ell)\right| \leq \epsilon$ for all $\ell$. There is an $M$ so that $\left|f_{N}(\ell)\right|<\epsilon$ for $\ell \geq M$. Hence, for such $\ell$,

$$
|f(\ell)| \leq\left|f_{N}(\ell)\right|+\epsilon<2 \epsilon
$$

Thus $f \in c_{0}$ and $\left(f_{n}\right)$ converges to $f$ in $c_{0}$.
Along with these spaces it is also helpful to consider the vector space

$$
c_{00}:=\{f: \mathbb{N} \rightarrow \mathbb{K} \mid f(n)=0 \text { for all but finitely many } n\}
$$

Notice that $c_{00}$ is a vector subspace of each of $c_{0}, \ell^{1}$ and $\ell^{\infty}$. Thus it can be equipped with either the $\|\cdot\|_{\infty}$ or $\|\cdot\|_{1}$ norms. It is not complete in either of these norms, however. What is true is that $c_{00}$ is dense in $c_{0}$ and $\ell^{1}$ (but not in $\ell^{\infty}$ ). (See

(c) ( $L^{1}$ spaces) Let $(X, \mathscr{M}, m)$ be a measure space. The quantity

$$
\|f\|_{1}:=\int_{X}|f| d m
$$

defines a norm on $L^{1}(m)$, provided we agree to identify $f$ and $g$ when $f=g$ a.e. (Indeed the chief motivation for making this identification is that it makes $\|\cdot\|_{1}$ into a norm. Note that $\ell^{1}$ from the last example is a special case (what is the measure space?))

Proposition 20.4. $L^{1}(m)$ is a Banach space.

Proof. It suffices to verify the hypotheses of Proposition $\frac{\text { prop:abs-cys-complete }}{20.3 \text {. } 1 t \sum_{n=1} f_{n} \text { is abso- }}$ lutely convergent (so that $\sum_{n=1}^{\infty}\left\|f_{n}\right\|_{1}<\infty$ ), then

$$
\sum_{n=1}^{\infty}\left|f_{n}\right| d m<\infty
$$

Thus the function $g:=\sum_{n=1}^{\infty}\left|f_{n}\right|$ belongs to $L^{1}$ and is thus finite $m$-a.e. by Tonelli. In particular the sequence of partial sums $s_{N}=\sum_{n=1}^{N} f_{n}$ is a sequence of measurable functions with $\left|s_{N}\right| \leq g$ that converges pointwise a.e. to a measurable function $f$. Hence by the DCT and its corollary, $f \in L^{1}$ and the partial sums $\left(s_{N}\right)_{N}$ converges to $f$ in $L^{1}$.
(d) ( $L^{p}$ spaces) Again let $(X, \mathscr{M}, m)$ be a measure space. For $1 \leq p<\infty$ let $L^{p}(m)$ denote the set of measurable functions $f$ for which

$$
\|f\|_{p}:=\left(\int_{X}|f|^{p} d m\right)^{1 / p}<\infty
$$

(again we identify $f$ and $g$ when $f=g$ a.e.). It turns out that this quantity is a norm on $L^{p}(m)$, and $L^{p}(m)$ is complete, though we will not prove this yet (it is not immediately obvious that the triangle inequality holds when $p>1$ ). The sequence space $\ell^{p}$ is defined analogously: it is the set of $f: \mathbb{N} \rightarrow \mathbb{K}$ for which

$$
\|f\|_{p}:=\left(\sum_{n=1}^{\infty}|f(n)|^{p}\right)^{1 / p}<\infty
$$

and this quantity is a norm making $\ell^{p}$ into a Banach space.

When $p=\infty$, we define $L^{\infty}(m)$ to be the set of all functions $f: X \rightarrow \mathbb{K}$ with the following property: there exists $M>0$ such that

$$
\begin{equation*}
|f(x)| \leq M \quad \text { for } m-\text { a.e. } x \in X \tag{1}
\end{equation*}
$$

as for the other $L^{p}$ spaces we identify $f$ and $g$ when there are equal a.e. When $f \in L^{\infty}$, let $\|f\|_{\infty}$ be the smallest $M$ for which (egn:L-infty-def $\|\cdot\|_{\infty}$ is a norm making $L^{\infty}(m)$ into a Banach space.
(e) $(C(X)$ spaces) Let $X$ be a compact metric space and let $C(X)$ denote the set of continuous functions $f: X \rightarrow \mathbb{K}$. It is a standard fact from advanced calculus that the quantity $\|f\|_{\infty}:=\sup _{x \in X}|f(x)|$ is a norm on $C(X)$. A sequence is Cauchy in this norm if and only if it is uniformly Cauchy. It is thus also a standard fact that $C(X)$ is complete in this norm-completeness just means that a uniformly Cauchy sequence of continuous functions on $X$ converges uniformly to a continuous function.

This example can be generalized somewhat: let $X$ be a locally compact metric space. Say a function $f: X \rightarrow \mathbb{K}$ vanishes at infinity if for every $\epsilon>0$, there exists a compact set $K \subset X$ such that $\sup _{x \notin K}|f(x)|<\epsilon$. Let $C_{0}(X)$ denote the set of continuous functions $f: X \rightarrow \mathbb{K}$ that vanish at infinity. Then $C_{0}(X)$ is a vector space, the quantity $\|f\|_{\infty}:=\sup _{x \in X}|f(x)|$ is a norm on $C_{0}(X)$, and $C_{0}(X)$ is complete in this norm. (Note that $c_{0}$ from above is a special case.)
(f) (Subspaces and direct sums) If $(\mathcal{X},\|\cdot\|)$ is a normed vector space and $\mathcal{Y} \subset \mathcal{X}$ is a vector subspace, then the restriction of $\|\cdot\|$ to $\mathcal{Y}$ is clearly a norm on $\mathcal{Y}$. If $\mathcal{X}$ is a

Banach space, then $(\mathcal{Y},\|\cdot\|)$ is a Banach space if and only if $\mathcal{Y}$ is closed in the norm topology of $\mathcal{X}$. (This is just a standard fact about metric spaces-a subspace of a complete metric space is complete in the restricted metric if and only if it is closed.)

If $\mathcal{X}, \mathcal{Y}$ are vector spaces then the algebraic direct sum is the vector space of ordered pairs

$$
\mathcal{X} \oplus \mathcal{Y}:=\{(x, y): x \in \mathcal{X}, y \in \mathcal{Y}\}
$$

with entrywise operations. If $\mathcal{X}, \mathcal{Y}$ are equipped with norms $\|\cdot\|_{\mathcal{X}},\|\cdot\|_{\mathcal{Y}}$, then each of the quantities

$$
\begin{aligned}
\|(x, y)\|_{\infty} & :=\max \left(\|x\|_{\mathcal{X}},\|y\|_{\mathcal{Y}}\right) \\
\|(x, y)\|_{1} & :=\|x\|_{\mathcal{X}}+\|y\|_{\mathcal{Y}}
\end{aligned}
$$

is a norm on $\mathcal{X} \oplus \mathcal{Y}$. These two norms are equivalent; indeed it follows from the definitions that

$$
\|(x, y)\|_{\infty} \leq\|(x, y)\|_{1} \leq 2\|(x, y)\|_{\infty}
$$

If $\mathcal{X}$ and $\mathcal{Y}$ are both complete, then $\mathcal{X} \oplus \mathcal{Y}$ is complete in both of these norms. The resulting Banach spaces are denoted $\mathcal{X} \oplus_{\infty} \mathcal{Y}, \mathcal{X} \oplus_{1} \mathcal{Y}$ respectively.
(g) (Quotient spaces) If $\mathcal{X}$ is a normed vector space and $\mathcal{M}$ is a proper subspace, then one can form the algebraic quotient $\mathcal{X} / \mathcal{M}$, defined as the collection of distinct cosets $\{x+\mathcal{M}: x \in \mathcal{X}\}$. From linear algebra, $\mathcal{X} / \mathcal{M}$ is a vector space under the standard
operations. If $\mathcal{M}$ is a closed subspace of $\mathcal{X}$, then the quantity

$$
\|x+\mathcal{M}\|:=\inf _{y \in \mathcal{M}}\|x-y\|
$$

is a norm on $\mathcal{X} / \mathcal{M}$, called the quotient norm. (Geometrically, $\|x+\mathcal{M}\|$ is the distance in $\mathcal{X}$ from $x$ to the closed set $\mathcal{M}$.) It turns out that if $\mathcal{X}$ is complete, so is $\mathcal{X} / \mathcal{M}$. See Problem $\frac{\text { prob:quotient-norm }}{20.20 .}$

More examples are given in the exercises and further examples will appear after the development of some theory.

### 20.2. Linear transformations between normed spaces.

Definition 20.5. Let $\mathcal{X}, \mathcal{Y}$ be normed vector spaces. A linear transformation $T: \mathcal{X} \rightarrow \mathcal{Y}$ is bounded if there exists a constant $C>0$ such that $\|T x\|_{\mathcal{Y}} \leq C\|x\|_{\mathcal{X}}$ for all $x \in \mathcal{X}$. Remark 20.6. Note that in this definition it would suffice to require that $\|T x\|_{\mathcal{Y}} \leq$ $C\|x\|_{\mathcal{X}}$ just for all $x \neq 0$, or for all $x$ with $\|x\|_{\mathcal{X}}=1$ (why?)

The importance of boundedness and the following simple proposition is hard to overstate. Recall, a mapping $f: X \rightarrow Y$ between metric spaces is Lipschitz continuous if there is a constant $C>0$ such that $d(f(x), f(y)) \leq C d(x, y)$ for all $x, y \in X$. A simple exercise shows Lipschitz continuity implies continuity.
dd-iff-cns Proposition 20.7. If $T: \mathcal{X} \rightarrow \mathcal{Y}$ is a linear transformation between normed spaces, then the following are equivalent:
(i) $T$ is bounded.
(ii) $T$ is Lipschitz continuous.
(iii) $T$ is uniformly continuous.
(iv) $T$ is continuous.
(v) $T$ is continuous at 0 .

Moreover, in this case,

$$
\begin{aligned}
\|T\| & :=\sup \{\|T x\|:\|x\|=1\} \\
& =\sup \left\{\frac{\|T x\|}{\|x\|}: x \neq 0\right\} \\
& =\inf \{C:\|T x\| \leq C\|x\| \text { for all } x \in \mathcal{X}\}
\end{aligned}
$$

and $\|T\|$ is the smallest number (the infimum is attained) such that

$$
\begin{equation*}
\|T x\| \leq\|T\|\|x\| \tag{2}
\end{equation*}
$$

for all $x \in \mathcal{X}$.

Proof. Suppose $T$ is bounded so that there exists a $C>0$ such that $\|T x\| \leq C\|x\|$ for all $x \in \mathcal{X}$. Thus, if $x, y \in \mathcal{X}$, then, $\|T x-T y\|=\|T(x-y)\| \leq C\|x-y\|$ by linearity of $T$. Hence (i) implies (ii). The implications (ii) implies (iii) implies (iv) implies (v) are evident. The proof of (v) implies (i) exploits the homogeneity of the norm and the linearity of $T$. By hypothesis, with $\epsilon=1$ there exists $\delta>0$ such that if $\|x\|<\delta$, then $\|T x\|<1$. Fix a nonzero vector $x \in \mathcal{X}$ and a real number $0<\lambda<\delta$. The vector $\lambda x /\|x\|$ has norm less than $\delta$, so

$$
\left\|T\left(\frac{\lambda x}{\|x\|}\right)\right\|=\lambda \frac{\|T x\|}{\|x\|}<1 .
$$

Rearranging this we find $\|T x\| \leq(1 / \lambda)\|x\|$ for all $x \neq 0$, which shows $T$ is bounded; in fact we can take $C=\frac{1}{\delta}$.

The rest of the proof is left as an exercise.

The set of all bounded linear operators from $\mathcal{X}$ to $\mathcal{Y}$ is denoted $B(\mathcal{X}, \mathcal{Y})$. It is a vector space under the operations of pointwise addition and scalar multiplication. The quantity $\|T\|$ is called the operator norm of $T$.

Proposition 20.8. For normed vector spaces $\mathcal{X}$ and $\mathcal{Y}$, the operator norm makes $B(\mathcal{X}, \mathcal{Y})$ into a normed vector space that is complete if $\mathcal{Y}$ is complete.

Proof. That $B(\mathcal{X}, \mathcal{Y})$ is a normed vector space follows readily from the definitions and is left as an exercise. Suppose now $\mathcal{Y}$ is complete, and let $T_{n}$ be a Cauchy sequence in $B(\mathcal{X}, \mathcal{Y})$. For each $x \in \mathcal{X}$, we have

$$
\begin{equation*}
\left\|T_{n} x-T_{m} x\right\|=\left\|\left(T_{n}-T_{m}\right) x\right\| \leq\left\|T_{n}-T_{m}\right\|\|x\| \tag{3}
\end{equation*}
$$

eq: $\operatorname{TnTm}$
which shows that $\left(T_{n} x\right)$ is a Cauchy sequence in $\mathcal{Y}$. By hypothesis, $T_{n} x$ converges in $\mathcal{Y}$. Define $T: \mathcal{X} \rightarrow \mathcal{Y}$ by setting $T x:=y$. It is straightforward to check that $T$ is linear.

Let $B$ denote the closed unit ball in $\mathcal{X}$. The sequence $\left(\left.T_{n}\right|_{B}\right)$ is uniformly Cauchy by equation (k) and ( TnTm converges pointwise to $\left.T\right|_{B}$. Hence $\left.T\right|_{B}$ is continuous and $\left(\left.T_{n}\right|_{B}\right)$ converges uniformly to $\left.T\right|_{B}$. It follows that $T$ is continuous at 0 and therefore $T$ is continuous. Since $\left\|T_{n}-T\right\|=\sup \left\{\left\|\left(T_{n}-T\right) x\right\|: x \in B\right\}$ and since $\left(\left.T_{n}\right|_{B}\right)$ converges to $\left.T\right|_{B}$ uniformly,
${ }^{* *}$ explicitly, given $\epsilon>0$ there is an $N$ such that if $n \geq N$, then $\left\|\left.\left(T_{n}\right)\right|_{B} x-T_{B} x\right\| \leq \epsilon$;
thus for $|x| \leq 1,\left\|T_{n} x-T x\right\| \leq \epsilon^{* *}$
it follows that $\left(T_{n}\right)$ converges to $T$ in $B(\mathcal{X}, \mathcal{Y})$.

If $T \in B(\mathcal{X}, \mathcal{Y})$ and $S \in B(\mathcal{Y}, \mathcal{Z})$, then two applications of the the inequality (文)
gives, for $x \in \mathcal{X}$,

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

and it follows that $S T \in B(\mathcal{X}, \mathcal{Z})$ and $\|S T\| \leq\|S\|\|T\|$. In the special case that $\mathcal{Y}=\mathcal{X}$ is complete, $B(\mathcal{X}):=B(\mathcal{X}, \mathcal{X})$ is an example of a Banach algebra.

The following proposition is very useful in constructing bounded operators-at least when the codomain is complete. Namely, it suffices to define the operator (and show that it is bounded) on a dense subspace.

Proposition 20.9 (Extending bounded operators). Let $\mathcal{X}, \mathcal{Y}$ be normed vector spaces with $\mathcal{Y}$ complete, and $\mathcal{E} \subset \mathcal{X}$ a dense linear subspace. If $T: \mathcal{E} \rightarrow \mathcal{Y}$ is a bounded linear operator, then there exists a unique bounded linear operator $\widetilde{T}: \mathcal{X} \rightarrow \mathcal{Y}$ extending $T$ (so $\left.\left.\widetilde{T}\right|_{\mathcal{E}}=T\right)$. Further $\|\widetilde{T}\|=\|T\|$.

Proof. Recall, if $X, Y$ are metric spaces, $Y$ is complete, $D \subset X$ is dense and $f: D \rightarrow Y$ is uniformly continuous, then $f$ has a unique continuous extension $\tilde{f}: X \rightarrow Y$. Moreover, this extension can be defined as follows. Given $x \in X$, choose a sequence $\left(x_{n}\right)$ from $D$
converging to $x$ and let $\tilde{f}(x)=\lim f\left(x_{n}\right)$ (that the sequence $f\left(x_{n}\right)$ is Cauchy follows from uniform continuity; that it converges from the assumption that $\mathcal{Y}$ is complete and finally it is an exercise to show $\tilde{f}(x)$ is well defined independent of the choice of $\left.\left(x_{n}\right)\right)$. Thus, it only remains to verify that the extension $\tilde{T}$ of $T$ is in fact linear and $\|T\|=\|\tilde{T}\|$. Both are routine exercises.

Remark: The completeness of $\mathcal{Y}$ is essential in the above proposition; Problem $\frac{\text { prob:extending-bounded-operators }}{20.11}$ suggests a counterexample.

A bounded linear transformation $T \in B(\mathcal{X}, \mathcal{Y})$ is said to be invertible if it is bijective (being bijective automatically $T^{-1}$ exists and is a linear transformation) and $T^{-1}$ is bounded from $\mathcal{Y}$ to $\mathcal{X}$. Two normed spaces $\mathcal{X}, \mathcal{Y}$ are said to be (boundedly) isomorphic if there exists an invertible linear transformation $T: \mathcal{X} \rightarrow \mathcal{Y}$. As an example, given equivalent norms $\|\cdot\|_{1}$ and $\|\cdot\|_{2}$ on a vector space $\mathcal{X}$, the identity mapping $\iota:\left(\mathcal{X},\|\cdot\|_{1}\right) \rightarrow$ $\left(\mathcal{X},\|\cdot\|_{2}\right)$ is (boundedly) invertible and witnesses the fact that these two normed vector spaces are boundedly isomorphic.

An operator $T: \mathcal{X} \rightarrow \mathcal{Y}$ such that $\|T x\|=\|x\|$ for all $x \in \mathcal{X}$ is an isometry. Note that an isometry is automatically injective and if it is also surjective then it is automatically invertible and $T^{-1}$ is also an isometry. An isometry need not be surjective, however. The normed vector spaces are isometrically isomorphic if there is an invertible isometry $T: \mathcal{X} \rightarrow \mathcal{Y}$.

### 20.3. Examples.

(a) If $\mathcal{X}$ is a finite-dimensional normed space and $\mathcal{Y}$ is any normed space, then every linear transformation $T: \mathcal{X} \rightarrow \mathcal{Y}$ is bounded. See Problem $\frac{\text { prob:finitedimbounded }}{20.14}$
(b) Let $\mathcal{X}$ denote $c_{00}$ equipped with the $\|\cdot\|_{1}$ norm, and $\mathcal{Y}$ denote $c_{00}$ equipped with the $\|\cdot\|_{\infty}$ norm. Then the identity map $i d_{\mathcal{X}, \mathcal{Y}}: \mathcal{X} \rightarrow \mathcal{Y}$ is bounded as an operator (in fact its norm is equal to 1 ), but its inverse, the identity map $\iota \mathcal{y}, \mathcal{X}: \mathcal{Y} \rightarrow \mathcal{X}$ is unbounded. For positive integers $n$, let $f_{n}$ denote the element of $c_{00}(1, \ldots, 1,0, \ldots)$. Now $\left\|\iota_{\mathcal{y}, \mathcal{X}}\left(f_{n}\right)\right\|_{1}=n$, but $\left\|f_{n}\right\|_{\infty}=1$.
(c) Consider $c_{00}$ with the $\|\cdot\|_{\infty}$ norm. Let $a: \mathbb{N} \rightarrow \mathbb{K}$ be any function and define a linear transformation $T_{a}: c_{00} \rightarrow c_{00}$ by

$$
\begin{equation*}
T_{a} f(n)=a(n) f(n) \tag{4}
\end{equation*}
$$

The mapping $T_{a}$ is bounded if and only if $M=\sup _{n \in \mathbb{N}}|a(n)|<\infty$, in which case $\left\|T_{a}\right\|=M$. In this case, $T_{a}$ extends uniquely to a bounded operator from $c_{0}$ to $c_{0}$, and one may check that the formula (保; defa-def the extension. All of these claims remain true if we use the $\|\cdot\|_{1}$ norm instead of the $\|\cdot\|_{\infty}$ norm. In this case, we get a bounded operator from $\ell^{1}$ to itself.
(d) Define $S: \ell^{1} \rightarrow \ell^{1}$ as follows given the sequence $(f(n))_{n}$ from $\ell^{1}$ let $S f(1)=0$ and $S f(n)=f(n-1)$ for $n>1$. (Viewing $f$ as a sequence, $S$ shifts the sequence one place to the right and fills in a 0 in the first position). This $S$ is an isometry, but is not surjective. In contrast, if $\mathcal{X}$ is finite-dimensional, then the rank-nullity theorem
from linear algebra guarantees that every injective linear map $T: \mathcal{X} \rightarrow \mathcal{X}$ is also surjective.
(e) Let $C^{\infty}([0,1])$ denote the vector space of functions on $[0,1]$ with continuous derivatives of all orders. The differentiation map $D: C^{\infty}([0,1]) \rightarrow C^{\infty}([0,1])$ defined by $D f=\frac{d f}{d x}$ is a linear transformation. Since, for $t \in \mathbb{R}$, we have $D e^{t x}=t e^{t x}$, it follows that there is no norm on $C^{\infty}([0,1])$ such that $\frac{d}{d x}$ is bounded.

### 20.4. Problems.

Problem 20.1. Prove Proposition prop:normed-tvs

Problem 20.2. Prove equivalent norms define the same topology and the same Cauchy sequences.

Problem 20.3. (a) Prove all norms on a finite dimensional vector space $\mathcal{X}$ are equivalent. Suggestion: Fix a basis $e_{1}, \ldots e_{n}$ for $\mathcal{X}$ and define $\left\|\sum a_{k} e_{k}\right\|_{1}:=\sum\left|a_{k}\right|$. It is routine to check that $\|\cdot\|_{1}$ is a norm on $\mathcal{X}$. Now complete the following outline.
(i) Let $\|\cdot\|$ be the given norm on $\mathcal{X}$. Show there is an $M$ such that $\|x\| \leq M\|x\|_{1}$. Conclude that the mapping $\iota:\left(\mathcal{X},\|\cdot\|_{1}\right) \rightarrow(\mathcal{X},\|\cdot\|)$ defined by $\iota(x)=x$ is continuous;
(ii) Show that the unit sphere $S=\left\{x \in \mathcal{X}:\|x\|_{1}=1\right\}$ in $\left(\mathcal{X},\|\cdot\|_{1}\right)$ is compact in the $\|\cdot\|_{1}$ topology;
(iii) Show that the mapping $f: S \rightarrow(\mathcal{X},\|\cdot\|)$ given by $f(x)=\|x\|$ is continuous and hence attains its infimum. Show this infimum is not 0 and finish the proof.
(b) Combine the result of part (a) with the result of Problem prob: equivalent-norms 20.2 to conclude that every finite-dimensional normed vector space is complete.
(c) Let $\mathcal{X}$ be a normed vector space and $\mathcal{M} \subset \mathcal{X}$ a finite-dimensional subspace. Prove $\mathcal{M}$ is closed in $\mathcal{X}$.


Problem 20.5. A function $f:[0,1] \rightarrow \mathbb{K}$ is called Lipschitz continuous if there exists a constant $C$ such that

$$
|f(x)-f(y)| \leq C|x-y|
$$

for all $x, y \in[0,1]$. Define $\|f\|_{L i p}$ to be the best possible constant in this inequality. That is,

$$
\|f\|_{L i p}:=\sup _{x \neq y} \frac{|f(x)-f(y)|}{|x-y|}
$$

Let $\operatorname{Lip}[0,1]$ denote the set of all Lipschitz continuous functions on $[0,1]$. Prove $\|f\|:=$ $|f(0)|+\|f\|_{\text {Lip }}$ is a norm on $\operatorname{Lip}[0,1]$, and that $\operatorname{Lip}[0,1]$ is complete in this norm.

Problem 20.6. Let $C^{1}([0,1])$ denote the space of all functions $f:[0,1] \rightarrow \mathbb{R}$ such that $f$ is differentiable in $(0,1)$ and $f^{\prime}$ extends continuously to $[0,1]$. Prove

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

is a norm on $C^{1}([0,1])$ and that $C^{1}$ is complete in this norm. Do the same for the norm $\|f\|:=|f(0)|+\left\|f^{\prime}\right\|_{\infty} .\left(\right.$ Is $\left\|f^{\prime}\right\|_{\infty}$ a norm on $\left.C^{1} ?\right)$

Problem 20.7. Let $(X, \mathscr{M})$ be a measurable space. Let $M(X)$ denote the (real) vector space of all signed measures on $(X, \mathscr{M})$. Prove the total variation norm $\|\mu\|:=|\mu|(X)$ is a norm on $M(X)$, and $M(X)$ is complete in this norm.

Problem 20.8. Prove, if $\mathcal{X}, \mathcal{Y}$ are normed spaces, then the operator norm is a norm on $B(\mathcal{X}, \mathcal{Y})$.
prob:c00 Problem 20.9. Prove $c_{00}$ is dense in $c_{0}$ and $\ell^{1}$. (That is, given $f \in c_{0}$ there is a sequence $f_{n}$ in $c_{00}$ such that $\left\|f_{n}-f\right\|_{\infty} \rightarrow 0$, and the analogous statement for $\ell^{1}$.) Using these facts, or otherwise, prove that $c_{00}$ is not dense in $\ell^{\infty}$. (In fact there exists $f \in \ell^{\infty}$ with $\|f\|_{\infty}=1$ such that $\|f-g\|_{\infty} \geq 1$ for all $g \in c_{00}$. )

Problem 20.10. Prove $c_{00}$ is not complete in the $\|\cdot\|_{1}$ or $\|\cdot\|_{\infty}$ norms. (After we have studied the Baire Category theorem, you will be asked to prove that there is no norm on $c_{00}$ making it complete.)

Problem 20.11. Consider $c_{0}$ and $c_{00}$ equipped with the $\|\cdot\|_{\infty}$ norm. Prove there is no bounded operator $T: c_{0} \rightarrow c_{00}$ such that $\left.T\right|_{c_{00}}$ is the identity map. (Thus the conclusion of Proposition $\frac{\text { prop:extending-bounded-operators }}{20.9 \text { can fail if } \mathcal{Y} \text { is not complete.) }}$

Problem 20.12. Prove the $\|\cdot\|_{1}$ and $\|\cdot\|_{\infty}$ norms on $c_{00}$ are not equivalent. Conclude from your proof that the identity map on $c_{00}$ is bounded from the $\|\cdot\|_{1}$ norm to the $\|\cdot\|_{\infty}$ norm, but not the other way around.
prob:c0rn) Problem 20.13. a) Prove $f \in C_{0}\left(\mathbb{R}^{n}\right)$ if and only if $f$ is continuous and $\lim _{|x| \rightarrow \infty}|f(x)|=$ 0 . b) Let $C_{c}\left(\mathbb{R}^{n}\right)$ denote the set of continuous, compactly supported functions on $\mathbb{R}^{n}$. Prove $C_{c}\left(\mathbb{R}^{n}\right)$ is dense in $C_{0}\left(\mathbb{R}^{n}\right)$ (where $C_{0}\left(\mathbb{R}^{n}\right)$ is equipped with sup norm).
dimbounded Problem 20.14. Prove, if $\mathcal{X}, \mathcal{Y}$ are normed spaces and $\mathcal{X}$ is finite dimensional, then every linear transformation $T: \mathcal{X} \rightarrow \mathcal{Y}$ is bounded. Suggestion: Let $d$ denote the dimension of $X$ and let $\left\{e_{1}, \ldots, e_{d}\right\}$ denote a basis. The function $\|\cdot\|_{1}$ on $\mathcal{X}$ defined by $\left\|\sum x_{j} e_{j}\right\|_{1}=\sum\left|x_{j}\right|$ is a norm. Apply Problem prob:findim-basics $\left\|\sum x_{j} e_{j}\right\|_{1}=\sum\left|x_{j}\right|$ is a norm. Apply Problem 10.3 .

Problem 20.15 Prover:bounded-operator-examples
Problem 20.15. Prove the claims in Example $20.3(\mathrm{c})$.

Problem 20.16. Let $g: \mathbb{R} \rightarrow \mathbb{K}$ be a (Lebesgue) measurable function. The map $M g: f \rightarrow g f$ is a linear transformation on the space of measurable functions. Prove, if $g \notin L^{\infty}(\mathbb{R})$, then there is an $f \in L^{1}(\mathbb{R})$ such that $g f \notin L^{1}(\mathbb{R})$. Conversely, show if $g \in L^{\infty}(\mathbb{R})$, then $M_{g}$ is bounded from $L^{1}(\mathbb{R})$ to itself and $\left\|M_{g}\right\|=\|g\|_{\infty}$.
20.17 sec:norm-examples

Problem 20.17. Prove the claims about direct sums in Example 20.1(f).
iesz-lemma Problem 20.18. Let $\mathcal{X}$ be a normed vector space and $\mathcal{M}$ a proper closed subspace. Prove for every $\epsilon>0$, there exists $x \in \mathcal{X}$ such that $\|x\|=1$ and $\inf _{y \in \mathcal{M}}\|x-y\|>1-\epsilon$.
(Hint: take any $u \in \mathcal{X} \backslash \mathcal{M}$ and let $a=\inf _{y \in \mathcal{M}}\|u-y\|$. Choose $\delta>0$ small enough so that $\frac{a}{a+\delta}>1-\epsilon$, and then choose $v \in \mathcal{M}$ so that $\|u-v\|<a+\delta$. Finally let $x=\frac{u-v}{\|u-v\|}$.)

Note that the distance to a (closed) subspace need not be attained. Here is an example. Consider the Banach space $C([0,1])$ (with the sup norm of course and either real or complex valued functions) and the closed subspace

$$
T=\left\{f \in C([0,1]): f(0)=0=\int_{0}^{1} f d t\right\}
$$

Using machinery in the next section it will be evident that $T$ is a closed subspace of $C([0,1])$. For now, it can be easily verified directly. Let $g$ denote the function $g(t)=t$. Verify that, for $f \in T$, that

$$
\frac{1}{2}=\int g d t=\int(g-f) d t \leq\|g-f\|_{\infty}
$$

In particular, the distance from $g$ to $T$ is at least $\frac{1}{2}$.
Note that the function $h=x-\frac{1}{2}$, while not in $T$, satisfies $\|g-h\|_{\infty}=\frac{1}{2}$.
On the other hand, for any $\epsilon>0$ there is an $f \in T$ so that $\|g-f\|_{\infty} \leq \frac{1}{2}+\epsilon$ (simply modify $h$ appropriately). Thus, the distance from $g$ to $T$ is $\frac{1}{2}$. Now verify, using the inequality above, that $h$ is the only element of $C([0,1])$ such that $\int h d t=0$ and $\|g-h\|_{\infty}=\frac{1}{2}$.

Problem 20.19. Prove, if $\mathcal{X}$ is an infinite-dimensional normed space, then the unit ball $\operatorname{ball}(\mathcal{X}):=\{x \in \mathcal{X}:\|x\| \leq 1\}$ is not compact in the norm topology. (Hint: use the
 $\left\|x_{n}\right\|=1$ for all $n$ and $\left\|x_{n}-x_{m}\right\| \geq \frac{1}{2}$ for all $m<n$.)

Problem 20.20. (The quotient norm) Let $\mathcal{X}$ be a normed space and $\mathcal{M}$ a proper closed subspace.
a) Prove the quotient norm is a norm (see Example $\frac{\mid \text { sec:norm-examples }}{Z 01(\mathrm{~g})) \text {. }}$
b) Show that the quotient map $x \rightarrow x+\mathcal{M}$ has norm 1. (Use Problem $\frac{\text { prob:FRiesz-lemma }}{20.18 .)}$
c) Prove, if $\mathcal{X}$ is complete, so is $\mathcal{X} / \mathcal{M}$.

Problem 20.21. A normed vector space $\mathcal{X}$ is called separable if it is separable as a metric space (that is, there is a countable subset of $\mathcal{X}$ which is dense in the norm topology). Prove $c_{0}$ and $\ell^{1}$ are separable, but $\ell^{\infty}$ is not. (Hint: for $\ell^{\infty}$, show that there is an uncountable collection of elements $\left\{f_{\alpha}\right\}$ such that $\left\|f_{\alpha}-f_{\beta}\right\|=1$ for $\alpha \neq \beta$.)

## 21. Linear functionals and the Hahn-Banach theorem

If there is a "fundamental theorem of functional analysis," it is the Hahn-Banach theorem. The theorem is somewhat abstract-looking at first, but its importance will be clear after studying some of its corollaries.

Let $\mathcal{X}$ be a normed vector space over the field $\mathbb{K}$. A linear functional on $\mathcal{X}$ is a linear map $L: \mathcal{X} \rightarrow \mathbb{K}$. As one might expect, we are especially interested in bounded linear functionals. Since $\mathbb{K}=\mathbb{R}$ or $\mathbb{C}$ is complete, the vector space of bounded linear functionals $B(\mathcal{X}, \mathbb{K})$ is itself a Banach space (complete normed vector space). This space is called the dual space of $\mathcal{X}$ and is denoted $\mathcal{X}^{*}$. It is not yet obvious that $\mathcal{X}^{*}$ need be non-trivial (that is, that there are any bounded linear functionals on $\mathcal{X}$ besides 0 ). One corollary of the Hahn-Banach theorem is there exist enough bounded linear functionals on $\mathcal{X}$ to separate points.
21.1. Examples. For each of the sequence spaces $c_{0}, \ell^{1}, \ell^{\infty}$, for each $n$ the map $f \rightarrow$ $f(n)$ is a bounded linear functional. If we fix $g \in \ell^{1}$, then the functional $L_{g}: c_{0} \rightarrow \mathbb{K}$ defined by

$$
L_{g}(f):=\sum_{n=0}^{\infty} f(n) g(n)
$$

is bounded, since

$$
\left|L_{g}(f)\right| \leq \sum_{n=0}^{\infty}|f(n) g(n)| \leq\|f\|_{\infty} \sum_{n=0}^{\infty}|g(n)|=\|g\|_{1}\|f\|_{\infty}
$$

This inequality shows that $\left\|L_{g}\right\| \leq\|g\|_{1}$. In fact, equality holds, and every bounded linear functional on $c_{0}$ is of this form:
dual-of-c0 Proposition 21.1. The map $\Phi: \ell^{1} \rightarrow c_{0}^{*}$ defined by $\Phi(g)=L_{g}$ is an isometric isomorphism from $\ell^{1}$ onto the dual space $c_{0}^{*}$.

Proof. We have already seen that each $g \in \ell^{1}$ gives rise to a bounded linear functional $L_{g} \in c_{0}^{*}$ via

$$
L_{g}(f):=\sum_{n=0}^{\infty} g(n) f(n)
$$

and that $\left\|L_{g}\right\| \leq\|g\|_{1}$. We will prove simultaneously that this map is onto and that $\left\|L_{g}\right\| \geq\|g\|_{1}$.

Let $L \in c_{0}^{*}$. We will first show that there is unique $g \in \ell^{1}$ so that $L=L_{g}$. Let $e_{n} \in c_{0}$ be the indicator function of $n$, that is

$$
e_{n}(m)=\delta_{n m}
$$

Define a function $g: \mathbb{N} \rightarrow \mathbb{K}$ by

$$
g(n)=L\left(e_{n}\right)
$$

We claim that $g \in \ell^{1}$ and $L=L_{g}$. To see this, fix an integer $N$ and let $h \in c_{00}$ be the function

$$
h(n)= \begin{cases}\overline{g(n)} /|g(n)| & \text { if } n \leq N \text { and } g(n) \neq 0 \\ 0 & \text { otherwise }\end{cases}
$$

By definition $h \in c_{00}$ and $\|h\|_{\infty} \leq 1$. Note that $h=\sum_{n=0}^{N} h(n) e_{n}$. Now

$$
\sum_{n=0}^{N}|g(n)|=\sum_{n=0}^{N} h(n) g(n)=L(h)=|L(h)| \leq\|L\|\|h\| \leq\|L\|
$$

It follows that $g \in \ell^{1}$ and $\|g\|_{1} \leq\|L\|$. Moreover, the same calculation shows that $L=L_{g}$ when restricted to $c_{00}$, so by the uniqueness of extensions of bounded operators, $L=L_{g}$. Thus the map $g \rightarrow L_{g}$ is onto and

$$
\|g\|_{1} \leq\|L\|=\left\|L_{g}\right\| \leq\|g\|_{1} .
$$

al-of-ell1 Proposition 21.2. $\left(\ell^{1}\right)^{*}$ is isometrically isomorphic to $\ell^{\infty}$.

Proof. The proof follows the same lines as the proof of the previous proposition; the details are left as an exercise.

The same mapping $g \rightarrow L_{g}$ also shows that every $g \in \ell^{1}$ gives a bounded linear functional on $\ell^{\infty}$, but it turns out these do not exhaust $\left(\ell^{\infty}\right)^{*}$ (see Problem $\frac{\text { prob:ell-infty-dual }}{11.11 \text { ). }}$

If $f \in L^{1}(m)$ and $g$ is a bounded measurable function with $\sup _{x \in X}|g(x)|=M$, then the map

$$
L_{g}(f):=\int_{X} f g d m
$$

is a bounded linear functional of norm at most $M$. We will prove in Section $\frac{\sec : \mathrm{Lp}}{24}$ that the norm is in fact equal to $M$, and every bounded linear functional on $L^{1}(m)$ is of this type (at least when $m$ is $\sigma$-finite).

If $X$ is a compact metric space and $\mu$ is a finite, signed Borel measure on $X$, then

$$
L_{\mu}(f):=\int_{X} f d \mu
$$

is a bounded linear functional on $C_{\mathbb{C}}(X)$ with norm $\|\mu\|=|\mu|(X)$ (see Problem $\frac{\text { prob:normC (X) star }}{21.8) \text {. }}$

A version of the Riesz Markov Theorem says the converse is true too.
thm:RM Theorem 21.3 (Riesz-Markov). Suppose $X$ is a compact Hausdorff space. If $\lambda \in C(X)^{*}$, then there exists a unique regular Borel measure $\sigma$ such that, for $f \in C(X)$,

$$
\lambda(f)=\int_{X} f d \sigma
$$

The result is true with both real and complex scalars. Focusing on the case of real scalars, the strategy is to write $\lambda$ as the difference of positive linear functionals on $C(X)$ and apply the Riesz-Markov Theorem (twice). (For complex scalars, write $\lambda$ in terms of its real and imaginary parts and apply the result in the real case (twice)).

A function $f \in C(X)$ is positive (really nonnegative) if $f(x) \geq 0$ for all $x \in X$, written $f \geq 0$. Let $C(X)^{+}$denote the positive elements of $C(X)$. Given linear functionals $\lambda, \rho \in C(X)^{*}$, the inequality $\lambda \leq \rho$ means that $\lambda(f) \leq \rho(f)$ for all $f \in C(X)^{+}$.
21.2. The Hahn-Banach Extension Theorem. To state and prove the Hahn-Banach Extension Theorem, we first work in the setting $\mathbb{K}=\mathbb{R}$, then extend the results to the complex case.

Definition 21.4. Let $\mathcal{X}$ be a real vector space. A Minkowski functional is a function $p: \mathcal{X} \rightarrow \mathbb{R}$ such that $p(x+y) \leq p(x)+p(y)$ and $p(\lambda x)=\lambda p(x)$ for all $x, y \in \mathcal{X}$ and nonnegative $\lambda \in \mathbb{R}$.

For example, if $L: \mathcal{X} \rightarrow \mathbb{R}$ is any linear functional, then the function $p: \mathcal{X} \rightarrow \mathbb{R}$ defined by $p(x):=|L(x)|$ is a Minkowski functional. More generally if $\|\cdot\|$ is a seminorm on $\mathcal{X}$, then $p: \mathcal{X} \rightarrow \mathbb{R}$ defined by $p(x)=\|x\|$ is a Minkowski functional.
thm:r-hb Theorem 21.5 (The Hahn-Banach Theorem, real version). Let $\mathcal{X}$ be a vector space over $\mathbb{R}$, p a Minkowski functional on $\mathcal{X}$, and $\mathcal{M}$ a subspace of $\mathcal{X}$. If $L$ a linear functional on $\mathcal{M}$ such that $L(x) \leq p(x)$ for all $x \in \mathcal{M}$, then there exists a linear functional $L^{\prime}$ on $\mathcal{X}$ such that
(i) $\left.L^{\prime}\right|_{\mathcal{M}}=L$ ( $L^{\prime}$ extends $\left.L\right)$
(ii) $L^{\prime}(x) \leq p(x)$ for all $x \in \mathcal{X}\left(L^{\prime}\right.$ is dominated by $\left.p\right)$.

The proof will invoke Zorns Lemma, a result that is equivalent to the axiom of choice (as well as the well ordering principal and the Hausdorff maximality principal). A partial order $\preceq$ on a set is a relation that is reflexive, symmetric and transitive; that is $x \preceq x$ for all $x \in S$; for $x, y \in S$ if $x \preceq y$ and $y \preceq x$, then $x=y$; and for $x, y, z \in \S$, if $x \preceq y$ and $y \preceq z$, then $x \preceq z$. We call $S$, or more precisely, $(S, \preceq)$ a partially ordered set or poset. A subset $T$ of $S$ is totally ordered, if for each $x, y \in T$ either $x \preceq y$ or $y \preceq x$. A totally ordered subset $T$ is often called a chain. An upper bound $z$ for a chain $T$ is an element $z \in S$ such that $t \preceq z$ for all $t \in T$. A maximal element for $S$ is $w \in S$ that has no successor; that is there does not exist an $s \in S$ such that $s \neq w$ and $w \preceq s$.

Theorem 21.6 (Zorn's Lemma). Suppose $S$ is a partially ordered set. If every chain in $S$ has an upper bound, then $S$ has a maximal element.

Proof. The idea is to show that the extension can be done one dimension at a time and then infer the existence of an extension to the whole space by appeal to Zorn's lemma. We may of course assume $\mathcal{M} \neq \mathcal{X}$. So, fix a vector $x \in \mathcal{X} \backslash \mathcal{M}$ and consider the subspace $\mathcal{M}+\mathbb{R} x \subset \mathcal{X}$. For any $m_{1}, m_{2} \in \mathcal{M}$, by hypothesis,

$$
L\left(m_{1}\right)+L\left(m_{2}\right)=L\left(m_{1}+m_{2}\right) \leq p\left(m_{1}+m_{2}\right) \leq p\left(m_{1}-x\right)+p\left(m_{2}+x\right)
$$

Rearranging gives, for $m_{1}, m_{2} \in \mathcal{M}$,

$$
L\left(m_{1}\right)-p\left(m_{1}-x\right) \leq p\left(m_{2}+x\right)-L\left(m_{2}\right)
$$

and thus

$$
\sup _{m \in \mathcal{M}}\{L(m)-p(m-x)\} \leq \inf _{m \in \mathcal{M}}\{p(m+x)-L(m)\}
$$

Now choose any real number $\lambda$ satisfying

$$
\sup _{m \in \mathcal{M}}\{L(m)-p(m-x)\} \leq \lambda \leq \inf _{m \in \mathcal{M}}\{p(m+x)-L(m)\} .
$$

In particular, for $m \in \mathcal{M}$,

$$
\begin{align*}
& L(m)-\lambda \leq p(m-x) \\
& L(m)+\lambda \leq p(m+x) \tag{5}
\end{align*}
$$

Let $\mathcal{N}=\mathcal{M}+\mathbb{R} x$ and define $L^{\prime}: \mathcal{N} \rightarrow \mathbb{R}$ by $L^{\prime}(m+t x)=L(m)+t \lambda$ for $m \in \mathcal{M}$ and $t \in \mathbb{R}$. Thus $L^{\prime}$ is linear and agrees with $L$ on $\mathcal{M}$ by definition. We now check that $L^{\prime}(y) \leq p(y)$ for all $y \in \mathcal{M}+\mathbb{R} x$. Accordingly, suppose $m \in \mathcal{M}, t \in \mathbb{R}$ and let
$y=m+t x$. If $t=0$ there is nothing to prove. If $t>0$, then, in view of equation (eq):choosel

$$
L^{\prime}(y)=L^{\prime}(m+t x)=t\left(L\left(\frac{m}{t}\right)+\lambda\right) \leq t p\left(\frac{m}{t}+x\right)=p(m+t x)=p(y)
$$

and a similar estimate shows that $L^{\prime}(m+t x) \leq p(m+t x)$ for $t<0$.

We have thus successfully extended $L$ to $\mathcal{M}+\mathbb{R} x$. To finish the proof, let $\mathcal{L}$ denote the set of pairs $\left(L^{\prime}, \mathcal{N}\right)$ where $\mathcal{N}$ is a subspace of $\mathcal{X}$ containing $\mathcal{M}$, and $L^{\prime}$ is an extension of $L$ to $\mathcal{N}$ obeying $L^{\prime}(y) \leq p(y)$ on $\mathcal{N}$. Declare $\left(L_{1}^{\prime}, \mathcal{N}_{1}\right) \preceq\left(L_{2}^{\prime}, \mathcal{N}_{2}\right)$ if $\mathcal{N}_{1} \subset \mathcal{N}_{2}$ and $\left.L_{2}^{\prime}\right|_{\mathcal{N}_{1}}=L_{1}^{\prime}$. This relation $\preceq$ is a partial order on $\mathcal{L}$. An exercise shows, given any increasing chain $\left(L_{\alpha}^{\prime}, \mathcal{N}_{\alpha}\right)$ in $\mathcal{L}$, it has as an upper bound $\left(L^{\prime}, \mathcal{N}\right)$ in $\mathcal{L}$, where $\mathcal{N}:=\bigcup_{\alpha} \mathcal{N}_{\alpha}$ and $L\left(n_{\alpha}\right):=L_{\alpha}^{\prime}\left(n_{\alpha}\right)$ for $n_{\alpha} \in \mathcal{N}_{\alpha}$. By Zorn's lemma the collection $\mathcal{L}$ has a maximal element $\left(L^{\prime}, \mathcal{N}\right)$ with respect to the order $\preceq$. Since it always possible to extend to a strictly larger subspace, the maximal element must have $\mathcal{N}=\mathcal{X}$, and the proof is finished.

The proof is a typical application of Zorn's lemma - one knows how to carry out a construction one step a time, but there is no clear way to do it all at once.

In the special case that $p$ is a seminorm, since $L(-x)=-L(x)$ and $p(-x)=p(x)$ the inequality $L \leq p$ is equivalent to $|L| \leq p$.

Corollary 21.7. Suppose $\mathcal{X}$ is a normed vector space over $\mathbb{R}, \mathcal{M}$ is a subspace, and $L$ is a bounded linear functional on $\mathcal{M}$. If $C \geq 0$ and $|L(x)| \leq C\|x\|$ for all $x \in \mathcal{M}$, then there exists a bounded linear functional $L^{\prime}$ on $\mathcal{X}$ extending $L$ such that $\left\|L^{\prime}\right\| \leq C$.

Proof. Apply the Hahn-Banach theorem with the Minkowski functional $p(x)=C\|x\|$.

Before obtaining further corollaries, we extend these results to the complex case. First, if $\mathcal{X}$ is a vector space over $\mathbb{C}$, then trivially it is also a vector space over $\mathbb{R}$, and there is a simple relationship between the $\mathbb{R}$ - and $\mathbb{C}$-linear functionals.
complexify Proposition 21.8. Let $\mathcal{X}$ be a vector space over $\mathbb{C}$. If $L: \mathcal{X} \rightarrow \mathbb{C}$ is a $\mathbb{C}$-linear functional, then $u(x)=\operatorname{Re} L(x)$ defines an $\mathbb{R}$-linear functional on $\mathcal{X}$ and $L(x)=u(x)-$ $i u(i x)$. Conversely, if $u: \mathcal{X} \rightarrow \mathbb{R}$ is $\mathbb{R}$-linear then $L(x):=u(x)-i u(i x)$ is $\mathbb{C}$-linear. If in addition $p: \mathcal{X} \rightarrow \mathbb{R}$ is a seminorm, then $|u(x)| \leq p(x)$ for all $x \in \mathcal{X}$ if and only if $|L(x)| \leq p(x)$ for all $x \in \mathcal{X}$.

Proof. Problem $\frac{\text { prob:complexify }}{21.5 .}$

To prove the last statement, it is immediate that $|u(x)| \leq|L(x)|$ for all $x \in \mathcal{X}$.

Conversely, given $x$ there is a unimodular $\alpha$ such that $\alpha L(x)=|L(x)|$. Hence,

$$
|L(x)|=L(\alpha x)=|u(\alpha x)| \leq p(\alpha x)=|\alpha| p(x)=p(x) .
$$

thm:c-hb Theorem 21.9 (The Hahn-Banach Theorem, complex version). Let $\mathcal{X}$ be a vector space over $\mathbb{C}$, $p$ a seminorm on $\mathcal{X}$, and $\mathcal{M}$ a subspace of $\mathcal{X}$. If $L: \mathcal{M} \rightarrow \mathbb{C}$ is a $\mathbb{C}$-linear functional satisfying $|L(x)| \leq p(x)$ for all $x \in \mathcal{M}$, then there exists a $\mathbb{C}$-linear functional $L^{\prime}: \mathcal{X} \rightarrow \mathbb{C}$ such that
(i) $\left.L^{\prime}\right|_{\mathcal{M}}=L$ and
(ii) $\left|L^{\prime}(x)\right| \leq p(x)$ for all $x \in \mathcal{X}$.

Proof. The proof consists of applying the real Hahn-Banach theorem to extend the $\mathbb{R}$ linear functional $u=\operatorname{Re} L$ to a functional $u^{\prime}: \mathcal{X} \rightarrow \mathbb{R}$ and then defining $L^{\prime}$ from $u^{\prime}$ as in Proposition prop: complexify 21.8 . The details are left as an exercise.

The following corollaries are quite important, and when the Hahn-Banach theorem is applied it is usually in one of the following forms:

## Corollary 21.10. Let $\mathcal{X}$ be a normed vector space.

it: $\mathrm{HB} \quad$ (i) If $\mathcal{M} \subset \mathcal{X}$ is a subspace and $L: \mathscr{M} \rightarrow \mathbb{K}$ is a bounded linear functional, then there exists a bounded linear functional $L^{\prime}: \mathcal{X} \rightarrow \mathbb{K}$ such that $\left.L^{\prime}\right|_{\mathscr{M}}=L$ and $\left\|L^{\prime}\right\|=\|L\|$.
(ii) (Linear functionals detect norms) If $x \in \mathcal{X}$ is nonzero, there exists $L \in \mathcal{X}^{*}$ with $\|L\|=1$ such that $L(x)=\|x\|$.
(iii) (Linear functionals separate points) If $x \neq y$ in $\mathcal{X}$, there exists $L \in \mathcal{X}^{*}$ such that $L(x) \neq L(y)$.
(iv) (Linear functionals detect distance to subspaces) If $\mathcal{M} \subset \mathcal{X}$ is a closed subspace and $x \in \mathcal{X} \backslash \mathcal{M}$, there exists $L \in \mathcal{X}^{*}$ such that
(a) $\left.L\right|_{\mathcal{M}}=0$;
(b) $\|L\|=1$; and
(c) $L(x)=\operatorname{dist}(x, \mathcal{M})=\inf _{y \in \mathcal{M}}\|x-y\|>0$.

Proof. (i): Consider the (semi)norm $p(x)=\|L\|\|x\|$. By construction, $|L(x)| \leq p(x)$ for $x \in \mathcal{M}$. Hence, there is a linear functional $L^{\prime}$ on $\mathcal{X}$ such that $\left.L^{\prime}\right|_{\mathcal{M}}=L$ and $\left|L^{\prime}(x)\right| \leq p(x)$ for all $x \in \mathcal{X}$. In particular, $\left\|L^{\prime}\right\| \leq\|L\|$. On the other hand, $\left\|L^{\prime}\right\| \geq\|L\|$ since $L^{\prime}$ agrees with $L$ on $\mathcal{M}$.
(ii): Let $\mathcal{M}$ be the one-dimensional subspace of $\mathcal{X}$ spanned by $x$. Define a functional $L: \mathcal{M} \rightarrow \mathbb{K}$ by $L\left(t \frac{x}{\|x\|}\right)=t$. In particular, $|L(y)|=\|y\|$ for $y \in \mathcal{M}$ and thus $\|L\|=1$. By (i), the functional $L$ extends to a functional (still denoted $L$ ) on $\mathcal{X}$ such that $\|L\|=1$.
(iii): Apply (ii) to the vector $x-y$.
(iv): Let $\delta=\operatorname{dist}(x, \mathcal{M})$. Since $\mathcal{M}$ is closed, $\delta>0$. Define a functional $L$ : $\mathcal{M}+\mathbb{K} x \rightarrow \mathbb{K}$ by $L(y+t x)=t \delta$. Since for $t \neq 0$ and $y \in \mathcal{M}$,

$$
\|y+t x\|=|t|\left\|t^{-1} y+x\right\| \geq|t| \delta=|L(y+t x)|
$$

by Hahn-Banach we can extend $L$ to a functional $L \in \mathcal{X}^{*}$ with $\|L\| \leq 1$.

Needless to say, the proof of the Hahn-Banach theorem is thoroughly non-constructive, and in general it is an important (and often difficult) problem, given a normed space $\mathcal{X}$, to find some concrete description of the dual space $\mathcal{X}^{*}$. Usually doing so means finding a Banach space $\mathcal{Y}$ and a bounded (or, better, isometric) isomorphism $T: \mathcal{Y} \rightarrow \mathcal{X}^{*}$.

Note that since $\mathcal{X}^{*}$ is a normed space, we can form its dual, denoted $\mathcal{X}^{* *}$, and called the bidual or double dual of $\mathcal{X}$. There is a canonical relationship between $\mathcal{X}$ and $\mathcal{X}^{* *}$.

Each fixed $x \in \mathcal{X}$ gives rise to a linear functional $\hat{x}: \mathcal{X}^{*} \rightarrow \mathbb{K}$ via evaluation,

$$
\hat{x}(L):=L(x) .
$$

Since $|\hat{x}(L)|=|L(x)| \leq\|L\|\|x\|$, the linear functional $\hat{x}$ is in $\mathcal{X}^{* *}$ and $\|\hat{x}\| \leq\|x\|$.

Corollary 21.11. (Embedding in the bidual) The $\operatorname{map} x \rightarrow \hat{x}$ is an isometric linear map from $\mathcal{X}$ into $\mathcal{X}^{* *}$.

Proof. First, from the definition we see that

$$
|\hat{x}(L)|=|L(x)| \leq\|L\|\|x\|
$$

so $\hat{x} \in \mathcal{X}^{* *}$ and $\|\hat{x}\| \leq\|x\|$. It is straightforward to check (recalling that the L's are linear) that the map $x \rightarrow \hat{x}$ is linear. Finally, to show that $\|\hat{x}\|=\|x\|$, fix a nonzero $x \in \mathcal{X}$. From Corollary $\frac{\text { cor: } \frac{\text { hbjeitif }}{21 . i(1)(1)} \text { there exists }}{L} \in \mathcal{X}^{*}$ with $\|L\|=1$ and $L(x)=\|x\|$. But then for this $x$ and $L$, we have $|\hat{x}(L)|=|L(x)|=\|x\|$ so $\|\hat{x}\| \geq\|x\|$, and the proof is complete.

Definition 21.12. A Banach space $\mathcal{X}$ is called reflexive if the map ${ }^{\wedge}: \mathcal{X} \rightarrow \mathcal{X}^{* *}$ is surjective.

In other words, $\mathcal{X}$ is reflexive if the map^ is an (isometric) isomorphism of $\mathcal{X}$ with
 Reflexive spaces often have nice properties. For instance, the distance from a point to a (closed) subspace is attained. On the other hand, by Propositions $\frac{\text { prop:dual proppcaual-of-ell1 }}{21.1 \text { and } 12.2, c_{0}^{*} \text { is }}$
isometrically isomorphic to $\ell^{\infty}$. In Problem $\frac{\text { prob: badc0 }}{21.7 \text { you will show that } c_{0} \text { is not isometrically }}$ isomorphic to $\ell^{\infty}$ and so $c_{0}$ is not reflexive. After we have studied the $L^{p}$ and $\ell^{p}$ spaces in more detail, we will see that $L^{p}$ is reflexive for $1<p<\infty$.

The embedding into the bidual has many applications; one of the most basic is the following.

Proposition 21.13 (Completion of normed spaces). If $\mathcal{X}$ is a normed vector space, then there is a Banach space $\overline{\mathcal{X}}$ and in isometric map $\iota: \mathcal{X} \rightarrow \overline{\mathcal{X}}$ such that the image $\iota(\mathcal{X})$ is dense in $\overline{\mathcal{X}}$.

Proof. Embed $\mathcal{X}$ into $\mathcal{X}^{* *}$ via the map $x \rightarrow \hat{x}$ and let $\overline{\mathcal{X}}$ be the closure of the image of $\mathcal{X}$ in $\mathcal{X}^{* *}$. Since $\overline{\mathcal{X}}$ is a closed subspace of a complete space, it is complete.

The space $\overline{\mathcal{X}}$ is called the completion of $\mathcal{X}$. It is unique in the sense that if $\mathcal{Y}$ is another Banach space and $j: \mathcal{X} \rightarrow \mathcal{Y}$ embeds $\mathcal{X}$ isometrically as a dense subspace of $\mathcal{Y}$, then $\mathcal{Y}$ is isometrically isomorphic to $\overline{\mathcal{X}}$. The proof of this fact is left as an exercise.
21.3. Dual spaces and adjoint operators. Let $\mathcal{X}, \mathcal{Y}$ be normed spaces with duals $\mathcal{X}^{*}, \mathcal{Y}^{*}$. If $T: \mathcal{X} \rightarrow \mathcal{Y}$ is a linear transformation and $f: \mathcal{Y} \rightarrow \mathbb{K}$ is a linear functional, then $T^{*} f: \mathcal{X} \rightarrow \mathbb{K}$ defined by

$$
\begin{equation*}
\left(T^{*} f\right)(x)=f(T x) \tag{6}
\end{equation*}
$$

is a linear functional on $\mathcal{X}$. If $T$ and $f$ are both continuous (that is, bounded) then the composition $T^{*} f$ is bounded, and more is true:
adjointmap Theorem 21.14. Let $T: \mathcal{X} \rightarrow \mathcal{Y}$ be a bounded linear transformation. For $f \in \mathcal{Y}^{*}$, define $T^{*} f$ by the formula (legn: adjoint-def
it:Tstarf
(i) $T^{*} f$ belongs to $\mathcal{X}^{*}$, and $T^{*}$ is a linear map from $\mathcal{Y}^{*}$ into $\mathcal{X}^{*}$.
ii) $T^{*}: \mathcal{Y}^{*} \rightarrow \mathcal{X}^{*}$ is bounded and $\left\|T^{*}\right\|=\|T\|$.

Proof. Since $T$ is assumed bounded, for a fixed $f \in \mathcal{Y}^{*}$ and all $x \in \mathcal{X}$

$$
\left|T^{*} f(x)\right|=|f(T x)| \leq\|f\|\|T x\| \leq\|f\|\|T\|\|x\| .
$$

It follows that $T^{*} f$ is bounded on $\mathcal{X}$ (thus, belongs to $\mathcal{X}^{*}$ ) and

$$
\begin{equation*}
\left\|T^{*} f\right\| \leq\|f\|\|T\| . \tag{7}
\end{equation*}
$$

Thus $T^{*}$ maps $\mathcal{Y}^{*}$ into $\mathcal{X}^{*}$ and it is straightforward to verify that $T^{*}$ is linear, which
 and $\left\|T^{*}\right\| \leq\|T\|$.

It remains to show $\left\|T^{*}\right\| \geq\|T\|$. Toward this end, let $0<\epsilon<1$ be given and choose $x \in \mathcal{X}$ with $\|x\|=1$ and $\|T x\|>(1-\epsilon)\|T\|$. Now consider $T x$. By the Hahn-Banach
 For this $f$,

$$
\left\|T^{*}\right\| \geq\left\|T^{*} f\right\| \geq\left|T^{*} f(x)\right|=|f(T x)|=\|T x\|>(1-\epsilon)\|T\|
$$

Hence, $\left\|T^{*}\right\| \geq(1-\epsilon)\|T\|$. Since $\epsilon$ was arbitrary, $\left\|T^{*}\right\| \geq\|T\|$.
21.4. Duality for Sub and Quotient Spaces. The Hahn-Banach Theorem allows for the identification of the duals of subspaces and quotients of Banach spaces. Informally,
the dual of a subspace is a quotient and the dual of a quotient is a subspace. The precise results are stated below for complex scalars, but they hold also for real scalars.

Given a (closed) subspace $\mathcal{M}$ of the Banach space $\mathcal{X}$, let $\pi$ denote the map from $\mathcal{X}$ to the quotient $\mathcal{X} / \mathcal{M}$. Recall (see Problem $\frac{\text { prob:quotient-norm }}{20.20) \text {, the quotient }}$ is a Banach space with the norm,

$$
\|z\|=\inf \{\|y\|: \pi(y)=z\}
$$

In particular, if $x \in \mathcal{X}$, then

$$
\|\pi(x)\|=\inf \{\|x-m\|: m \in \mathcal{M}\}
$$

It is evident from the construction that $\pi$ is continuous and $\|\pi\| \leq 1$. Further, by
 $\|\pi\|=1$. In particular, $\pi^{*}:(\mathcal{X} / \mathcal{M})^{*} \rightarrow \mathcal{X}^{*}$ (defined by $\left.\pi^{*} \lambda=\lambda \circ \pi\right)$ is also continuous. Moreover, if $x \in \mathcal{M}$, then

$$
\pi^{*} \lambda(x)=0
$$

Let

$$
\mathcal{M}^{\perp}=\left\{f \in \mathcal{X}^{*}: f(x)=0 \text { for all } x \in \mathcal{M}\right\}
$$

$\left(\mathcal{M}^{\perp}\right.$ is called the annihilator of $\mathcal{M}$ in $\mathcal{X}^{*}$.) Recall, given $x \in \mathcal{X}$, the element $\hat{x} \in \mathcal{X}^{* *}$ is defined by $\hat{x}(\tau)=\tau(x)$, for $\tau \in \mathcal{X}^{*}$. In particular,

$$
\mathcal{M}^{\perp}=\cap_{x \in \mathcal{M}} \operatorname{ker}(\hat{x})
$$

and thus $\mathcal{M}^{\perp}$ is a closed subspace of $\mathcal{X}^{*}$. Further, if $\lambda \in(\mathcal{X} / \mathcal{M})^{*}$, then $\pi^{*} \lambda \in \mathcal{M}^{\perp}$.
op:pi-star Proposition 21.15 (The dual of a quotient). The mapping $\psi:(\mathcal{X} / \mathcal{M})^{*} \rightarrow \mathcal{M}^{\perp} d e$ fined by

$$
\psi(\lambda)=\pi^{*} \lambda
$$

is an isometric isomorphism; i.e., the mapping $\pi^{*}:(\mathcal{X} / \mathcal{M})^{*} \rightarrow \mathcal{X}^{*}$ is an isometric isomorphism onto $\mathcal{M}^{\perp}$.

Informally, the proposition is expressed as $(\mathcal{X} / \mathcal{M})^{*}=\mathcal{M}^{\perp}$.

Proof. The linearity of $\psi$ follows from Theorem $\left.\frac{\text { thm: adjointmap }}{21.14 \text { as does }} \right\rvert\, \psi\|=\| \pi \| \leq 1$. To prove that $\psi$ is isometric, let $\lambda \in(\mathcal{X} / \mathcal{M})^{*}$ be given. Automatically, $\|\psi(\lambda)\| \leq\|\lambda\|$. To prove the reverse inequality, fix $r>1$. Let $q \in \mathcal{X} / \mathcal{M}$ with $\|q\|=1$ be given. There exists an $x \in X$ such that $\|x\|<r$ and $\pi(x)=q$. Hence,

$$
|\lambda(q)|=\mid \lambda(\pi(x))\|=\| \psi(\lambda)(x)\|\leq\| \psi(\lambda)\| \| x\|<r\| \psi(\lambda) \| .
$$

Taking the supremum over such $q$ shows $\|\lambda\| \leq r\|\psi(\lambda)\|$. Finally, since $1<r$ is arbitrary, $\|\lambda\| \leq\|\psi(\lambda)\|$.

To prove that $\psi$ is onto, and complete the proof, let $\tau \in \mathcal{M}^{\perp}$ be given. Fix $q \in \mathcal{X} / \mathcal{M}$. If $x, y \in \mathcal{X}$ and $\pi(x)=q=\pi(y)$, then $\tau(x)=\tau(y)$. Hence, the mapping $\lambda: \mathcal{X} / \mathcal{M} \rightarrow \mathbb{C}$ defined by $\lambda(q)=\tau(x)$ is well defined. That $\lambda$ is linear is left as an exercise. To see that $\lambda$ is continuous, observe that

$$
|\lambda(q)|=|\tau(x)| \leq\|\tau\|\|x\|,
$$

for each $x \in \mathcal{X}$ such that $\pi(x)=q$. Taking the infimum over such $x$ gives shows

$$
|\lambda(q)| \leq\|\tau\|\|q\| .
$$

Finally, by construction $\psi(\lambda)=\tau$.

Since $\mathcal{M}^{\perp}$ is closed in $\mathcal{X}^{*}$, the quotient space $\mathcal{X}^{*} / \mathcal{M}^{\perp}$ is a Banach space. Let $\rho: \mathcal{X}^{*} \rightarrow \mathcal{X}^{*} / \mathcal{M}^{\perp}$ denote the quotient mapping. Suppose $\lambda \in \mathcal{M}^{*}$. By Corollary $\begin{aligned} & \text { leor: } \mathrm{hb}-\mathrm{cor} \\ & 21.10,\end{aligned}$ there is an $f \in \mathcal{X}^{*}$ such that $\left.f\right|_{\mathcal{M}}=\lambda$; that is $f$ is a bounded extension of $\lambda$ (and indeed $f$ can be chosen such that $\|f\|=\|\lambda\|)$. If $f$ and $g$ are two extensions of $\lambda$ to bounded linear functionals on $\mathcal{X}^{*}$, then $f(x)-g(x)=0$ for $x \in \mathcal{M}$. Hence $f-g \in \mathcal{M}^{\perp}$ or equivalently, $\rho(f)=\rho(g)$. Consequently, the mapping $\varphi: \mathcal{M}^{*} \rightarrow \mathcal{X}^{*} / \mathcal{M}^{\perp}$ defined by $\varphi(\lambda)=\rho(f)$ (where $f$ is any bounded extension of $\lambda$ to $\mathcal{X}$ ) is well defined. It is easily verified that $\varphi$ is linear. Further, given $q \in \mathcal{X}^{*} / m m^{\perp}$, there is an $f \in \mathcal{X}^{*}$ such that $\rho(f)=q$. In particular, with $\left.\lambda=\left.f\right|_{\mathcal{M}}\right) \mathrm{i}$ we have $\varphi(\lambda)=\rho(f)$. Therefore $\varphi$ is onto.

Proposition 21.16 (The dual of a subspace). The mapping $\varphi: \mathcal{M}^{*} \rightarrow \mathcal{X}^{*} / \mathcal{M}^{\perp}$ is an isometric isomorphism.

Proof. It remains to show that $\varphi$ is an isometry, a fact that is an easy consequence of the Hahn-Banach Theorem. Fix $\lambda \in \mathcal{M}^{*}$ and let $q=\varphi(\lambda)$. If $f$ is any bounded extension
of $\lambda$ to $\mathcal{X}^{*}$, then $\|f\| \geq\|\lambda\|$. Hence,

$$
\begin{aligned}
\|\varphi(\lambda)\| & =\|q\| \\
& =\inf \left\{\|f\|: f \in \mathcal{X}^{*}, \quad \rho(f)=q\right\} \\
& =\inf \left\{\|f\|: f \in \mathcal{X}^{*},\left.\quad f\right|_{\mathcal{M}}=\lambda\right\} \\
& \geq\|\lambda\|
\end{aligned}
$$

On the other hand, by the Hahn-Banach Theorem there is a bounded extension $g$ of $\lambda$ with $\|g\|=\|\lambda\|$. Thus $\|\lambda\| \leq\|q\|$.

A special case of the following useful fact was used in the proofs above. If $\mathcal{X}, \mathcal{Y}$ are vector spaces and $T: \mathcal{X} \rightarrow \mathcal{Y}$ is linear and $\mathcal{M}$ is a subspace of the kernel of $T$, then $T$ induces a linear map $\tilde{T}: \mathcal{X} / \mathcal{M} \rightarrow \mathcal{Y}$. A canonical choice is $\mathcal{M}=\operatorname{ker}(T)$ in which case $\tilde{T}$ is one-one. If $\mathcal{X}$ is a Banach space, $\mathcal{Y}$ is a normed vector space and $\mathcal{M}$ is closed, then $\mathcal{X} / \mathcal{M}$ is a Banach space. space and $T: \mathcal{X} \rightarrow \mathcal{Y}$ is continuous, then the mapping $\tilde{T}$ is bounded and $\|\tilde{T}\|=\|T\|$.

Proof. Let $\pi: \mathcal{X} \rightarrow \mathcal{X} / \mathcal{M}$ denote the quotient map and observe that $\tilde{T} \pi=T$. Fix $q \in \mathcal{X} / \mathcal{M}$. For any $x \in \mathcal{X}$ such that $\pi(x)=q$,

$$
\|\tilde{T} q\|=\|\tilde{T}(\pi(x))\|=\|T x\| \leq\|T\|\|x\|
$$

Taking the infimum on $x$ such that $\pi(x)=q$ gives,

$$
\|\tilde{T} q\| \leq\|T\|\|q\|
$$

### 21.5. Problems.

Problem 21.1. Prove, if $\mathcal{X}$ is any normed vector space, $\left\{x_{1}, \ldots x_{n}\right\}$ is a linearly independent set in $\mathcal{X}$, and $\alpha_{1}, \ldots \alpha_{n}$ are scalars, then there exists a bounded linear functional $f$ on $\mathcal{X}$ such that $f\left(x_{j}\right)=\alpha_{j}$ for $j=1, \ldots n$. (Recall linear maps from a finite dimensional normed vector space to a normed vector space are bounded.)

Problem 21.2. Let $\mathcal{X}, \mathcal{Y}$ be normed spaces and $T: \mathcal{X} \rightarrow \mathcal{Y}$ a linear transformation. Prove $T$ is bounded if and only if there exists a constant $C$ such that for all $x \in \mathcal{X}$ and $f \in \mathcal{Y}^{*}$,

$$
\begin{equation*}
|f(T x)| \leq C\|f\|\|x\| ; \tag{8}
\end{equation*}
$$

in which case $\|T\|$ is equal to the best possible $C$ in (eqn:bilinear-bdd-prob

Problem 21.3. Let $\mathcal{X}$ be a normed vector space. Show that if $\mathcal{M}$ is a closed subspace of $\mathcal{X}$ and $x \notin \mathcal{M}$, then $\mathcal{M}+\mathbb{K} x$ is closed. Use this result to give another proof that every finite-dimensional subspace of $\mathcal{X}$ is closed.

Problem 21.4. Prove, if $\mathcal{M}$ is a finite-dimensional subspace of a Banach space $\mathcal{X}$, then there exists a closed subspace $\mathcal{N} \subset \mathcal{X}$ such that $\mathcal{M} \cap \mathcal{N}=\{0\}$ and $\mathcal{M}+\mathcal{N}=\mathcal{X}$. (In other words, every $x \in \mathcal{X}$ can be written uniquely as $x=y+z$ with $y \in \mathcal{M}, z \in \mathcal{N}$.) Hint: Choose a basis $x_{1}, \ldots x_{n}$ for $\mathcal{M}$ and construct, using Problem prob:dualbasis 1.1 and the HahnBanach Theorem, bounded linear functionals $f_{1}, \ldots f_{n}$ on $\mathcal{X}$ such that $f_{i}\left(x_{j}\right)=\delta_{i j}$. Now
let $\mathcal{N}=\cap_{i=1}^{n} \operatorname{ker} f_{i}$. (Warning: this conclusion can fail badly if $\mathcal{M}$ is not assumed finite dimensional, even if $\mathcal{M}$ is still assumed closed. Perhaps the first known example is that $c_{0}$ is not complemented in $\ell^{\infty}$, though it is nontrivial to prove.)

## Problem 21.5. Prove Proposition $\frac{\text { prop:complexify }}{21.8 \text {. }}$

-reflexive Problem 21.6. Prove every finite-dimensional Banach space is reflexive.
prob:badc0 Problem 21.7. Let $B$ denote the subset of $\ell^{\infty}$ consisting of sequences which take values in $\{-1,1\}$. Show that any two (distinct) points of $B$ are a distance 2 apart. Show, if $C$ is a countable subset of $\ell^{\infty}$, then there exists a $b \in B$ such that $\|b-c\| \geq 1$ for all $c \in C$. Conclude $\ell^{\infty}$ is not separable. Prove there is no isometric isomorphism $\Lambda: c_{0} \rightarrow \ell^{\infty}$. As a corollary, conclude that $c_{0}$ is not reflexive. (Of course, saying $c_{0} \neq \ell^{\infty}$ via the canonical embedding of Corollary $\begin{aligned} & \text { cor: embedinXss } \\ & 21.111 \text { is much }\end{aligned}$ weaker than saying there is no isometric isomorphism between $c_{0}$ and $\ell^{\infty}$.)
rmC(X)star Problem 21.8. Prove, if $\mu$ is a finite regular (signed) Borel measure on a compact Hausdorff space, then the linear function $L_{\mu}: C(X) \rightarrow \mathbb{R}$ defined by

$$
L_{\mu}(f)=\int_{X} f d \mu
$$

is bounded (continuous) and $\left\|L_{\mu}\right\|=\|\mu\|:=|\mu|(X)$. (See the Riesz-Markov Theorem for positive linear functionals.)

Problem 21.9. Let $\mathcal{X}$ and $\mathcal{Y}$ be normed vector spaces and $T \in L(\mathcal{X}, \mathcal{Y})$.
a) Consider $T^{* *}: \mathcal{X}^{* *} \rightarrow \mathcal{Y}^{* *}$. Identifying $\mathcal{X}, \mathcal{Y}$ with their images in $\mathcal{X}^{* *}$ and $\mathcal{Y}^{* *}$, show that $\left.T^{* *}\right|_{\mathcal{X}}=T$.
b) Prove $T^{*}$ is injective if and only if the range of $T$ is dense in $\mathcal{Y}$.
c) Prove that if the range of $T^{*}$ is dense in $\mathcal{X}^{*}$, then $T$ is injective; if $\mathcal{X}$ is reflexive then the converse is true.
rable-dual Problem 21.10. Prove, if $\mathcal{X}$ is a Banach space and $\mathcal{X}^{*}$ is separable, then $\mathcal{X}$ is separable.
[Hint: let $\left\{f_{n}\right\}$ be a countable dense subset of $\mathcal{X}^{*}$. For each $n$ choose $x_{n}$ such that $\left\|x_{n}\right\|=1$ and $\left|f_{n}\left(x_{n}\right)\right| \geq \frac{1}{2}\left\|f_{n}\right\|$. Show that the set of $\mathbb{Q}$-linear combinations of $\left\{x_{n}\right\}$ is dense in $\mathcal{X}$.]
infty-dual Problem 21.11. a) Prove there exists a bounded linear functional $L \in\left(\ell^{\infty}\right)^{*}$ with the following property: whenever $f \in \ell^{\infty}$ and $\lim _{n \rightarrow \infty} f(n)$ exists, then $L(f)$ is equal to this limit. (Hint: first show that the set of such $f$ forms a subspace $\left.\mathcal{M} \subset \ell^{\infty}\right)$.
b) Show that such a functional $L$ is not equal to $L_{g}$ for any $g \in \ell^{1}$; thus the map $T: \ell^{1} \rightarrow\left(\ell^{\infty}\right)^{*}$ given by $T(g)=L_{g}$ is not surjective.
c) Give another proof that $T$ is not surjective, using Problem $\begin{aligned} & \text { prob:sep } \\ & 21.10 .\end{aligned}$

## 22. The Baire Category Theorem and applications

Recall, a set $D$ in a metric space $X$ is dense if $\bar{D}=X$. Thus $D$ is dense if and only if $D^{c}$ does not contain a nonempty open set if and only if it has nontrivial intersection with every nonempty open set. A topological space $X$ is called a Baire space if it has the following property: if $\left(U_{n}\right)_{n=1}^{\infty}$ is a countable sequence of open dense subsets of $X$, then the intersection $\cap_{n=1}^{\infty} U_{n}$ is dense in $X$.
thm:baire Theorem 22.1 (The Baire Category Theorem). Every complete metric space $X$ is a Baire space. In other words, if $X$ is a complete metric space and if $\left(U_{n}\right)_{n=1}^{\infty}$ is a sequence of open dense subsets of $X$, then $\cap_{n=1}^{\infty} U_{n}$ is dense in $X$.
thm: baire
Theorem $\frac{\text { thm: baire }}{22.1}$ is true if $X$ is a locally compact Hausdorff space and there are connections between the Baire Category Theorem and the axiom of choice.

A subset $E \subset X$ is nowhere dense if its closure has empty interior. Equivalently, $\bar{E}^{c}$ is open and dense. A set $F$ in a metric space $X$ is first category (or meager) if it can be expressed as the countable union of nowhere dense sets. In particular, a countable union of first category sets is first category. A set $G$ is second category if it is not first category. For applications, the following corollary often suffices.
reinaction Corollary 22.2. If $X$ is a complete metric space, then $X$ is not a countable union of nowhere dense sets; i.e., $X$ is of second category in itself.

Proof. Take complements and apply Theorem $\frac{\text { thm:baire }}{22.1 .}$ ME? Prove the contrapositive.
Suppose $X=\cup E_{n}$ where the $E_{n}$ are nowhere dense. Thus each $(\bar{E})^{\circ}$ is open and dense and $X=\cup \bar{E}_{n}$. Taking complements,

$$
\emptyset=\cap\left(\bar{E}_{n}\right)^{\circ} .
$$

Hence $X$ is not complete by Theorem $\begin{aligned} & \text { thm:baire } \\ & 22.1 . \text { END ME?? }\end{aligned}$

Thus, the Baire property is used as a kind of pigeonhole principle: the "thick" Baire space $X$ cannot be expressed as a countable union of the "thin" nowhere dense sets $E_{n}$. Equivalently, if $X$ is Baire and $X=\bigcup_{n} E_{n}$, then at least one of the $E_{n}$ is somewhere dense.

The following lemma should be familiar from advanced calculus.
em:weakfip Lemma 22.3. Let $X$ be a complete metric space and suppose $\left(C_{n}\right)$ is a sequence of subsets of $X$. If
(i) each $C_{n}$ is nonempty;
(ii) $\left(C_{n}\right)$ is nested decreasing;
(iii) each $C_{n}$ is closed; and
(iv) $\left(\operatorname{diam}\left(C_{n}\right)\right)$ converges to 0 ,
then there is an $x \in X$ such that

$$
\{x\}=\cap C_{n} .
$$

Moreover, if $x_{n} \in C_{n}$, then $\left(x_{n}\right)$ converges to some $x$.

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Proof of Theorem $\mid$ thm:baire $\left(U_{n}\right)_{n=1}^{\infty}$ be a sequence of open dense sets in $X$ and let $I=$ $\cap U_{n}$. To prove $I$ is dense, it suffices to show that $I$ has nontrivial intersection with every nonempty open set $W$. Fix such a $W$. Since $U_{1}$ is dense, there is a point $x_{1} \in W \cap U_{1}$. Since $U_{1}$ and $W$ are open, there is a radius $0<r_{1}<1$ such that the $\overline{B\left(x_{1}, r_{1}\right)}$ is contained in $W \cap U_{1}$. Similarly, since $U_{2}$ is dense and open there is a point $x_{2} \in B\left(x_{1}, r_{1}\right) \cap U_{2}$ and a radius $0<r_{2}<\frac{1}{2}$ such that

$$
\overline{B\left(x_{2}, r_{2}\right)} \subset B\left(x_{1}, r_{1}\right) \cap U_{2} \subset W \cap U_{1} \cap U_{2}
$$

Continuing inductively, since each $U_{n}$ is dense and open there is a sequence of points $\left(x_{n}\right)_{n=1}^{\infty}$ and radii $0<r_{n}<\frac{1}{n}$ such that

$$
\overline{B\left(x_{n}, r_{n}\right)} \subset B\left(x_{n-1}, r_{n-1}\right) \cap U_{n} \subset W \cap\left(\cap_{j=1}^{n} U_{n}\right)
$$

The sequence of sets $\left(\overline{B\left(x_{n}, r_{n}\right)}\right)$ satisfies the hypothesis of Lemma $\frac{1 \text { em:weakf } \overline{22} .3 \text { and } X \text { is com- }}{}$ pact. Hence there is an $x \in X$ such that

$$
x \in \cap_{n} \overline{B\left(x_{n}, r_{n}\right)} \subset W \cap I
$$

We now give three important applications of the Baire category theorem in functional analysis. These are the Principle of Uniform boundedness (also known as the Banach-Steinhaus theorem), the Open Mapping Theorem, and the Closed Graph Theorem. (In learning these theorems, keep careful track of what completeness hypotheses are needed.)

Theorem 22.4 (The Principle of Uniform Boundedness (PUB)). Suppose $\mathcal{X}, \mathcal{Y}$ are normed spaces and $\left\{T_{\alpha}: \alpha \in A\right\} \subset B(\mathcal{X}, \mathcal{Y})$ is a collection of bounded linear transformations from $\mathcal{X}$ to $\mathcal{Y}$. Let $B$ denote the set

$$
\begin{equation*}
B:=\left\{x \in X: M(x):=\sup _{\alpha}\left\|T_{\alpha} x\right\|<\infty\right\} . \tag{9}
\end{equation*}
$$

If $B$ is of the second category (thus not a countable union of nowhere dense sets) in $X$, then

$$
\sup _{\alpha}\left\|T_{\alpha}\right\|<\infty .
$$

In particular, if $X$ is complete and if the collection $\left\{T_{\alpha}: \alpha \in A\right\}$ is pointwise bounded, then it is uniformly bounded.

Proof. For each integer $n \geq 1$ consider the set

$$
V_{n}:=\{x \in \mathcal{X}: M(x)>n\} .
$$

Since each $T_{\alpha}$ is bounded, the sets $V_{n}$ are open. (Indeed, for each $\alpha$ the map $x \rightarrow\left\|T_{\alpha} x\right\|$ is continuous from $\mathcal{X}$ to $\mathbb{R}$, so if $\left\|T_{\alpha} x\right\|>n$ for some $\alpha$ then also $\left\|T_{\alpha} y\right\|>n$ for all $y$ sufficiently close to $x$.) Let $E_{n}$ denote the complement of $V_{n}$ and observe that $B=\cup_{n=1}^{\infty} E_{n}$. Since $B$ is assumed to be of the second category, there is an $N$ such that $\left(\bar{E}_{N}\right)^{\circ}$ is not empty. Since $E_{N}$ is closed, it follows that $E_{N}$ has nonempty interior; i.e., there is an $x_{0} \in E_{N}$ and $r>0$ so that $x_{0}-x \in E_{N}$ for all $\|x\|<r$. Thus, for every $\alpha$ and every $\|x\|<r$,

$$
\left\|T_{\alpha} x\right\| \leq\left\|T_{\alpha}\left(x-x_{0}\right)\right\|+\left\|T_{\alpha} x_{0}\right\| \leq N+N
$$

That is, if $\|x\|<r$, then $M(x) \leq 2 N$. By rescaling we conclude that if $\|x\|<1$, then $\left\|T_{\alpha} x\right\| \leq 2 N / r$ for all $\alpha$ and thus $\sup _{\alpha}\left\|T_{\alpha}\right\| \leq 2 N / r<\infty$.

Given a subset $B$ of a vector space $\mathcal{X}$ and a scalar $s \in \mathbb{K}$, let $s B=\{s b: b \in B\}$. Similarly, for $x \in \mathcal{X}$, let $B-x=\{b-x: b \in B\}$. Let $\mathcal{X}, \mathcal{Y}$ be normed vector spaces and suppose $T: \mathcal{X} \rightarrow \mathcal{Y}$ is linear. If $B \subset \mathcal{X}$ and $s \in \mathbb{K}$ is nonzero, then $T(s B)=s T(B)$ and further, an easy argument shows $\overline{T(s B)}=s \overline{T(B)}$. It is also immediate that if $B$ is open, then so is $B-x$.

Recall that if $X, Y$ are topological spaces, a mapping $f: X \rightarrow Y$ is called open if $f(U)$ is open in $Y$ whenever $U$ is open in $X$. In particular, if $f$ is a bijection, then $f$ is open if and only if $f^{-1}$ is continuous. In the case of normed linear spaces the condition that a linear map be open can be refined somewhat.
ate dilate Lemma 22.5 (Translation and Dilation lemma). Let $\mathcal{X}, \mathcal{Y}$ be normed vector spaces, let $B$ denote the open unit ball of $\mathcal{X}$, and let $T: \mathcal{X} \rightarrow \mathcal{Y}$ be a linear map. The following are equivalent.
(i) The map $T$ is open;
(ii) $T(B)$ contains an open ball centered to 0 ;
(iii) there is an $s>0$ such that $T(s B)$ contains an open ball centered to 0 ; and
(iv) $T(s B)$ contains an open ball centered to 0 for each $s>0$.

Proof. This result is more or less immediate from the fact, for fixed $z_{0}$ and $r \in \mathbb{K}$, that the translation map $z \rightarrow z+z_{0}$ and the dilation map $z \rightarrow r z$ are continuous in a normed vector space. The implication (i) implies (ii) is immediate. The fact that $T(s B)=s T(B)$ for $s>0$ readily shows (ii), (iii) and (iv) are equivalent.

To finish the proof it suffice to show (iv) implies (i). Accordingly, suppose (iv) holds and let $U \subset X$ be a given open set. To prove that $T(U)$ is open, let $y \in T(U)$ be given. There is an $x \in U$ such that $T(x)=y$. There is an $s>0$ such that the ball $B(x, s)$ lies in $U$; that is $B(x, s) \subset U$. The ball $s B=B(0, s)=B(x, r)-x$ is an open ball centered to 0 . By hypothesis there is an $r>0$ such that $B^{\mathcal{Y}}(0, r) \subset T(B(0, s))$. (Here we use $B^{\mathcal{Y}}$ to emphasize this ball is a subset of $\mathcal{Y}$.) By linearity of $T$,

$$
\begin{aligned}
B^{\mathcal{Y}}(y, r) & =B^{\mathcal{Y}}(0, r)+y \subset T(B(0, s))+y \\
& =T(B(0, s))+T(x)=T(B(0, s)+x)=T(B(x, s)) \subset T(U) .
\end{aligned}
$$

Thus $T(U)$ is open.

Theorem 22.6 (Open Mapping). Suppose that $\mathcal{X}$ is a Banach space, $\mathcal{Y}$ is a normed vector space and $T: \mathcal{X} \rightarrow \mathcal{Y}$ is bounded. If the range of $T$ is of second category, then
(i) $T(\mathcal{X})=\mathcal{Y}$;
(ii) $\mathcal{Y}$ is complete (so a Banach space); and
(iii) $T$ is open.

In particular, if $\mathcal{X}, \mathcal{Y}$ are Banach spaces, and $T: \mathcal{X} \rightarrow \mathcal{Y}$ is bounded and onto, then $T$ is an open map.

Proof. Observe that (i) follows immediately from (iii). To prove (iii), let $B(x, r)$ denote the open ball of radius $r$ centered at $x$ in $\mathcal{X}$. Trivially $\mathcal{X}=\bigcup_{n=1}^{\infty} B(0, n)$ and thus $T(X)=\bigcup_{n=1}^{\infty} T(B(0, n))$. Since the range of $T$ is assumed second category, there is an $N$ such that $T(B(0, N))$ is second category and hence not nowhere dense. In other words, $\overline{T(B(0, N))}$ has nonempty interior. By scaling (see Lemma llem:trantate dilate nonempty interior. Hence, there exists $p \in \mathcal{Y}$ and $r>0$ such that $\overline{T(B(0,1))}$ contains the open ball $B^{\mathcal{Y}}(p, r)$. (Here the superscript $\mathcal{Y}$ is used to emphasize this ball is in $\mathcal{Y}$.) It follows that for all $\|y\|<r$,

$$
y=-p+(y+p) \in \overline{T(B(0,2))}
$$

In other words,

$$
B^{\mathcal{Y}}(0, r) \subset \overline{T(B(0,2))}
$$

By scaling, it follows that, for $n \in \mathbb{N}$,

$$
B^{\mathcal{Y}}\left(0, \frac{r}{2^{n+1}}\right) \subset \overline{T\left(B\left(0, \frac{1}{2^{n}}\right)\right)} .
$$

We will use the hypothesis that $\mathcal{X}$ is complete to prove $B^{\mathcal{Y}}\left(0, \frac{r}{4}\right) \subset T(B(0,1))$. Accordingly let $y$ such that $\|y\|<\frac{r}{4}$ be given. Since $y$ is in the closure of $T\left(B\left(0, \frac{1}{2}\right)\right)$, there is a $y_{1} \in T\left(B\left(0, \frac{1}{2}\right)\right)$ such that $\left\|y-y_{1}\right\|<\frac{r}{8}$. Since $y-y_{1} \in B^{\mathcal{Y}}\left(0, \frac{r}{8}\right)$ it is is in the closure of $T\left(B\left(0, \frac{1}{4}\right)\right)$. Thus there is a $y_{2} \in T\left(B\left(0, \frac{1}{4}\right)\right)$ such that $\left\|\left(y-y_{1}\right)-y_{2}\right\|<\frac{r}{16}$. Continuing in this fashion produces a sequence $\left(y_{j}\right)_{j=1}^{\infty}$ from $\mathcal{Y}$ such that,
(a) $\left\|y-\sum_{j=1}^{n} y_{j}\right\| \leq \frac{r}{2^{n+2}}$; and
(b) $y_{n} \in T\left(B\left(0, \frac{1}{2^{n}}\right)\right.$
for all $n$. It follows that $\sum_{j=1}^{\infty} y_{j}$ converges to $y$. Further, for each $j$ there is an $x_{j} \in$ $B\left(0, \frac{1}{2^{j}}\right)$ such that $y_{j}=T x_{j}$. Since

$$
\|x\| \leq \sum_{j=1}^{\infty}\left\|x_{j}\right\|<\sum_{k=1}^{\infty} 2^{-k}=1
$$

the series $\sum_{j=1}^{\infty} x_{j}$ converges to some $x \in B(0,1)$. It follows that $y=T x$ by continuity of $T$. Consequently $y \in T(B(0,1))$ and the proof of (iii) is complete.

To prove (ii), let $\mathcal{M}$ denote the kernel of $T$ and $\tilde{T}$ the mapping $\tilde{T}: \mathcal{X} / \mathcal{M} \rightarrow \mathcal{Y}$ determined by $\tilde{T} \pi=T$. By Lemma $\frac{1 \text { lem:descend to quotient }}{21.17, T \text { is continuous and one-one. Further its range }}$ is the same as the range of $T$, namely $\mathcal{Y}$, and is thus second category. Hence, by what has already been proved, $\tilde{T}$ is an open map. and consequently $\tilde{T}^{-1}$ is continuous. Hence $\mathcal{X} / \mathcal{M}$ and $\mathcal{Y}$ are isomorphic (though of course not necessarily isometrically isomorphic) as normed vector spaces. Therefore, since $\mathcal{X} / \mathcal{M}$ is complete, so is $\mathcal{Y}$.

Note that the proof of item (ii) in the Open Mapping Theorem shows, in the case that in the case that $T$ is one-one and its range is of second category, that $T$ is onto and its inverse is continuous. In particular, if $T: \mathcal{X} \rightarrow \mathcal{Y}$ is a continuous bijection and $\mathcal{Y}$ is a Banach space (so the range of $T$ is second category), then $T^{-1}$ is continuous.
somorphism Corollary 22.7 (The Banach Isomorphism Theorem). If $\mathcal{X}, \mathcal{Y}$ are Banach spaces and $T: \mathcal{X} \rightarrow \mathcal{Y}$ is a bounded bijection, then $T^{-1}$ is also bounded (hence, $T$ is an isomorphism).

To state the final result of this section, we need a few more definitions. Let $\mathcal{X}, \mathcal{Y}$ be normed spaces. The Cartesian product $\mathcal{X} \times \mathcal{Y}$ is a topological space in the product topology. A set is open in the product topology if and only if it can be written as a union of products of open sets. Alternately, a set $O$ is open if and only if for each $z=(x, y) \in O$ there exists open sets $U \subset \mathcal{X}$ and $V \subset \mathcal{Y}$ such that $z \in U \times V \subset O$. It is not too hard to show that $\mathcal{X} \times \mathcal{Y}$ is metrizable (in fact the product topology can be realized by norming $\mathcal{X} \times \mathcal{Y}$, e.g. with the norm $\|(x, y)\|:=\max (\|x\|,\|y\|))$. It is easy to see that a sequence $z_{n}=\left(x_{n}, y_{n}\right)$ converges in the product topology if and only if both $\left(x_{n}\right)$ and $\left(y_{n}\right)$ converge. Further, if $\mathcal{X}, \mathcal{Y}$ are both Banach spaces (complete), then $\mathcal{X} \times \mathcal{Y}$ is also complete and hence a Banach space. The space $\mathcal{X} \times \mathcal{Y}$ is equipped with the coordinate projections $\pi_{\mathcal{X}}(x, y)=x, \pi_{\mathcal{Y}}(x, y)=y$. It is clear from the definition of the product topology that these maps are continuous. (In fact the product topology is the coarsest topology such that the coordinate projections are continuous.)

Given a linear map $T: \mathcal{X} \rightarrow \mathcal{Y}$, its graph is the set

$$
G(T):=\{(x, y) \in \mathcal{X} \times \mathcal{Y}: y=T x\}
$$

Observe that since $T$ is a linear map, $G(T)$ is a linear subspace of $\mathcal{X} \times \mathcal{Y}$. The transformation $T$ is closed if $G(T)$ is a closed subset of $\mathcal{X} \times \mathcal{Y}$. It is an easy exercise to show that $G(T)$ is closed if and only if whenever $\left(x_{n}, T x_{n}\right)$ converges to $(x, y)$, we have $y=T x$.

other hand, if $\mathcal{X}, \mathcal{Y}$ are complete (Banach spaces), then $G(T)$ is closed if and only if $T$ is continuous.

Theorem 22.8 (The Closed Graph Theorem). If $\mathcal{X}, \mathcal{Y}$ are Banach spaces and $T: \mathcal{X} \rightarrow$ $\mathcal{Y}$ is closed, then $T$ is bounded.

Proof. Let $\pi_{1}, \pi_{2}$ be the coordinate projections $\pi_{\mathcal{X}}, \pi_{\mathcal{Y}}$ restricted to $G(T)$; explicitly $\pi_{1}(x, T x)=x$ and $\pi_{2}(x, T x)=T x$. Note that $\pi_{1}$ is a bijection between $G(T)$ and $\mathcal{X}$ and in particular $\pi_{1}^{-1}(x)=(x, T x)$. By hypothesis $G(T)$ is a closed subset of a Banach space and hence a Banach space. Thus $\pi_{1}$ is a bounded linear bijection between Banach spaces
 $\pi_{2} \circ \pi_{1}^{-1}: X \rightarrow Y$ is continuous. To finish the proof, observe $\pi_{2} \circ \pi_{1}^{-1}(x)=\pi_{2}(x, T x)=$ $T x$.

### 22.1. Problems.

Problem 22.1. Show that there exists a sequence of open, dense subsets $U_{n} \subset \mathbb{R}$ such that $m\left(\bigcap_{n=1}^{\infty} U_{n}\right)=0$.
ot-bounded Problem 22.2. Consider the linear subspace $\mathcal{D} \subset c_{0}$ defined by

$$
\mathcal{D}=\left\{f \in c_{0}: \lim _{n \rightarrow \infty}|n f(n)|=0\right\}
$$

and the linear transformation $T: \mathcal{D} \rightarrow c_{0}$ defined by $(T f)(n)=n f(n)$.
a) Prove $T$ is closed, but not bounded. b) Prove $T$ is bijective and $T^{-1}: c_{0} \rightarrow \mathcal{D}$ is bounded (and surjective), but not open. c) What can be said of $\mathcal{D}$ as a subset of $c_{0}$ ?

Problem 22.3. Suppose $\mathcal{X}$ is a yector space equipped with two norms $\|\cdot\|_{1},\|\cdot\|_{2}$ such that $\|\cdot\|_{1} \leq\|\cdot\|_{2}$. Prove that if $\mathcal{X}$ is complete in both norms, then the two norms are equivalent.

Problem 22.4. Let $\mathcal{X}, \mathcal{Y}$ be Banach spaces. Provisionally, say that a linear transformation $T: \mathcal{X} \rightarrow \mathcal{Y}$ is weakly bounded if $f \circ T \in \mathcal{X}^{*}$ whenever $f \in \mathcal{Y}^{*}$. Prove, if $T$ is weakly bounded, then $T$ is bounded.
b:TntoTPUB Problem 22.5. Let $\mathcal{X}, \mathcal{Y}$ be Banach spaces. Suppose $\left(T_{n}\right)$ is a sequence in $B(\mathcal{X}, \mathcal{Y})$ and $\lim _{n} T_{n} x$ exists for every $x \in \mathcal{X}$. Prove, if $T$ is defined by $T x=\lim _{n} T_{n} x$, then $T$ is bounded.
amel-basis Problem 22.6. Suppose that $\mathcal{X}$ is a vector space with a countably infinite basis. (That is, there is a linearly independent set $\left\{x_{n}\right\} \subset \mathcal{X}$ such that every vector $x \in \mathcal{X}$ is expressed uniquely as a finite linear combination of the $x_{n}$ 's.) Prove there is no norm on $\mathcal{X}$ under which it is complete. (Hint: consider the finite-dimensional subspaces $\left.\mathcal{X}_{n}:=\operatorname{span}\left\{x_{1}, \ldots x_{n}\right\}.\right)$

Problem 22.7. The Baire Category Theorem can be used to prove the existence of (very many!) continuous, nowhere differentiable functions on $[0,1]$. To see this, let $E_{n}$ denote the set of all functions $f \in C[0,1]$ for which there exists $x_{0} \in[0,1]$ (which may depend on $f$ ) such that $\left|f(x)-f\left(x_{0}\right)\right| \leq n\left|x-x_{0}\right|$ for all $x \in[0,1]$. Prove the sets $E_{n}$ are nowhere dense in $C[0,1]$; the Baire Category Theorem then shows that the set of nowhere differentiable functions is second category. (To see that $E_{n}$ is nowhere dense, approximate an arbitrary continuous function $f$ uniformly by piecewise linear functions $g$, whose pieces have slopes greater than $2 n$ in absolute value. Any function sufficiently close to such a $g$ will not lie in $E_{n}$.)

2firstinL1 Problem 22.8. Let $L^{2}([0,1])$ denote the Lebesgue measurable functions $f:[0,1] \rightarrow \mathbb{C}$ such that $|f|^{2}$ is in $L^{1}([0,1])$. It turns out, as we will see later, that $L^{2}([0,1])$ is a linear manifold (subspace of the vector space $L^{1}([0,1])$ ), though this fact is not needed for this problem.

Let $g_{n}:[0,1] \rightarrow \mathbb{R}$ denote the function which takes the value $n$ on $\left[0, \frac{1}{n^{3}}\right]$ and 0 elsewhere. Show,
(i) if $f \in L^{2}([0,1])$, then $\lim _{n \rightarrow \infty} \int g_{n} f d m=0$;
(ii) $L_{n}: L^{1}([0,1]) \rightarrow \mathbb{C}$ defined by $L_{n}(f)=\int g_{n} f d m$ is bounded, and $\left\|L_{g}\right\|=n$;
(iii) conclude $L^{2}([0,1])$ is of the first category in $L^{1}([0,1])$.
h-analytic Problem 22.9. A Banach space of functions on a set $X$ is a vector subspace $B$ of the space of complex-valued functions on $X$ with a norm $\|\cdot\|$ making $B$ a Banach space such that, for each $x \in X$, the mapping $E_{x}: B \rightarrow \mathbb{C}$ defined by $E_{x}(f)=f(x)$ is continuous (bounded) and if $f(x)=0$ for all $x \in X$, then $f=0$.

Suppose $g: X \rightarrow \mathbb{C}$. Show, if $g f \in B$ for each $f \in B$, then the linear map $M_{g}: B \rightarrow B$ defined by $M_{g} f=g f$ is bounded.
my solution: By the closed graph theorem, it suffices to show that the graph of $M_{g}$ is closed. To this end, suppose $\left(f_{n}, g f_{n}\right)$ is a convergent sequence from $G\left(M_{g}\right)$; that is that there exist $f, h \in B$ such that $\left(f_{n}, g f_{n}\right)$ converges to $(f, g)$. In particular, $\left(f_{n}\right)$ converges to $f$ and $\left(g_{n} f\right)$ converges to $h$ in $B$. By continuity, for $x \in X$, we have $\left(f_{n}(x)=E_{x}\left(f_{n}\right)\right)$ converges to $f(x)=E_{x}(f)$. Similarly, $\left(g(x) f_{n}(x)\right)$ converges to $h(x)$. On the other hand, $\left(g(x) f_{n}(x)\right)$ also converges to $g(x) f(x)$. Hence $h(x)=g(x) f(x)$ and $h=M_{g} f$. Thus $(f, h)=\left(f, M_{g} f\right) \in G\left(M_{g}\right)$ and the proof is complete.
ment-v-bdd Problem 22.10. Suppose $\mathcal{X}$ is a Banach space and $\mathcal{M}$ and $\mathcal{N}$ are closed subspaces. Show, if for each $x \in \mathcal{X}$ there exist unique $m \in \mathcal{M}$ and $n \in \mathcal{N}$ such that

$$
x=m+n
$$

then the mapping $P: \mathcal{X} \rightarrow \mathcal{M}$ defined by $P x=m$ is bounded.
anachLimit Problem 22.11. Here, for definiteness we take the scalar field $\mathbb{R}$.
Let $T: \ell^{\infty} \rightarrow \ell^{\infty}$ denote the backward shift operator defined by $T f(n)=f(n+1)$

A bounded linear functionals $\lambda: \ell^{\infty} \rightarrow \mathbb{R}$ satisfying,
(i) if $f \in \ell^{\infty}$ and $(f(n))$ converges, then $\lambda(f)=\lim _{n \rightarrow \infty} f(n)$; and
(ii) $\lambda(T f)=\lambda(f)$
is a Banach Limit.

Prove
(a) Banach limits exists.
(b) If $\lambda$ is a Banach limit and $f \in \ell^{\infty}$, then

$$
\liminf f(n) \leq \lambda(f) \limsup f(n)
$$

A sequence $f \in \ell^{\infty}$ for which $(f(n))$ does not converge, but $\lambda(f)=\mu(f)$ for all Banach limits $\lambda$ and $\mu$ is almost convergent. Show that $g$ defined by $g(n)=(-1)^{n}$ is almost convergent. (Suggestion: given a Banach limit $\lambda$, consider $\lambda(g+T g)$.

## 23. Hilbert spaces

23.1. Inner products. Let $\mathbb{K}$ denote either $\mathbb{C}$ or $\mathbb{R}$.

Definition 23.1. Let $X$ be a vector space over $\mathbb{K}$. An inner product on $X$ is a function $u: X \times X \rightarrow \mathbb{K}$ such that, for all $x, y, z \in X$ and all $\alpha, \beta \in \mathbb{K}$,
it:psd
it:antisym it:linear
(i) $u(x, x) \geq 0$ and $u(x, x)=0$ if and only if $x=0$.
(ii) $u(x, y)=\overline{u(y, x)}$
(iii) $u(\alpha x+\beta y, z)=\alpha u(x, z)+\beta u(y, z)$.

it:more
(iv) $u(x, \alpha y+\beta z)=\bar{\alpha} u(x, y)+\bar{\beta} u(x, z)$.

Remark 23.2. A function $u$ satisfying only items (it:linearit:more (iii) and (23.1) is called a bilinear form (when $\mathbb{K}=\mathbb{R})$ or a sesquilinear form (when $\mathbb{K}=\mathbb{C}$ ). In this case, if (it:antisym satisfied then $u$ is called symmetric $(\mathbb{R})$ or $\operatorname{Hermitian}(\mathbb{C})$. A Hermitian or symmetric form satisfying $u(x, x) \geq 0$ for all $x$ is called positive semidefinite or a pre-inner product.

Typically, $u$ is written $\langle\cdot, \cdot\rangle$ so that $u(x, y)=\langle x, y\rangle$.

Finally, observe if $u$ is a bilinear (resp. sesquilinear) form, then each $z \in X$ induces a linear functional on $X$ defined by $x \mapsto u(x, z)$.
thm:cs Theorem 23.3 (The Cauchy-Schwarz inequality). Suppose $\langle\cdot, \cdot\rangle$ is a pre-inner product on the vector space $X$.
it:CS (i) For $x, y \in X$,

$$
\begin{equation*}
|\langle x, y\rangle|^{2} \leq\langle x, x\rangle\langle y, y\rangle \tag{10}
\end{equation*}
$$

eqn:cs
(ii) If $x, z \in X$ and $\langle z, z\rangle=0$, then $\langle x, z\rangle=0$.
(iii) The set

$$
M=\{x \in X:\langle x, x\rangle=0\}
$$

is a subspace of $X$.
(iv) Let $Y=X / M$ and let $\pi: X \rightarrow Y$ denote the quotient map. The form $[p, q]=\langle x, y\rangle$ where $x, y \in X$ are any choices of vectors such that $\pi(x)=p$ and $\pi(y)=q$ is well defined and an inner product on $Y$.

Item (iit: (is known as the Cauchy-Schwarz inequality.

Proof. Fix $x, y \in X$. For $t \in \mathbb{R}$, let $\lambda=t\langle x, y\rangle$ and compute, using the nonnegativity assumption

$$
\begin{aligned}
0 & \leq\langle x-\lambda y, x-\lambda y\rangle \\
& =\langle x, x\rangle-2 t|\langle x, y\rangle|^{2}+t^{2}|\langle x, y\rangle|^{2}\langle y, y\rangle \\
& =: P(t)
\end{aligned}
$$

Since $P(t)$ is a nonnegative quadratic, its discriminant is nonpositive; i.e.,

$$
|\langle x, y\rangle|^{4} \leq\langle x, x\rangle\langle y, y\rangle|\langle x, y\rangle|^{2}
$$

and the Cauchy-Schwarz inequality follows.

If $\langle z, z\rangle=0$ and $x \in X$, then (i) immediately implies $\langle x, z\rangle=0$, which proves (ii). In particular, if both $x, y \in M$ and $c \in \mathbb{C}$, then

$$
\langle x+c y, x+c y\rangle=\langle x, x\rangle+c\langle y, x\rangle+\bar{c}\langle x, y\rangle+|c|^{2}\langle y, y\rangle=0
$$

and so $M$ is a subpace.

Item (iv) is an exercise in definition chasing.

### 23.2. Examples.

$\mathbb{K}^{n}$ : It is easy to check that the standard scalar product on $\mathbb{R}^{n}$ is an inner product; it is defined as usual by

$$
\begin{equation*}
\langle x, y\rangle=\sum_{j=1}^{n} x_{j} y_{j} \tag{11}
\end{equation*}
$$

where we have written $x=\left(x_{1}, \ldots x_{n}\right) ; y=\left(y_{1}, \ldots y_{n}\right)$. Similarly, the standard inner product of vectors $z=\left(z_{1}, \ldots z_{n}\right), w=\left(w_{1}, \ldots w_{n}\right)$ in $\mathbb{C}^{n}$ is given by

$$
\begin{equation*}
\langle z, w\rangle=\sum_{j=1}^{n} z_{j} \overline{w_{j}} . \tag{12}
\end{equation*}
$$

```
eqn:rn_inner_p
```

(Note that it is necessary to take complex conjugates of the w's to obtain positive definiteness.)
$\ell^{2}(\mathbb{N}):$ Let

$$
\ell^{2}(\mathbb{N})=\left\{\left.\left(a_{1}, a_{2}, \ldots a_{n}, \ldots\right)\left|a_{n} \in \mathbb{K}, \sum_{j=1}^{\infty}\right| a_{n}\right|^{2}<\infty\right\}
$$

We may view $\ell^{2}$ as a subset of the vector space $\mathscr{S}$ of all sequence (with domain $\mathbb{N})$ with entrywise addition and scalar multiplication. Define, for sequences $a=$

$$
\left(a_{1}, a_{2}, \ldots\right) \text { and } b=\left(b_{1}, b_{2}, \ldots\right) \text { in } \ell^{2}
$$

$$
\begin{equation*}
\sum_{n=1}^{\infty} a_{n} \overline{b_{n}} \tag{13}
\end{equation*}
$$

is seen to converge absolutely using the comparison test and the inequality $2\left|a_{n} b_{n}\right| \leq\left|a_{n}\right|^{2}+\left|b_{n}\right|^{2}$. From here it is not hard to prove that $\ell^{2}$ is closed under the vector space operations of $\mathscr{S}$ and is hence a vector space; and further that (|egn:12innerprod 113 defines an inner product, $\langle a, b\rangle=\sum a_{n} \overline{b_{n}}$, called the standard inner product on $\ell^{2}$.
$L^{2}(\mu)$ : Generalizing the previous example, let $(X, \mathscr{M}, \mu)$ be a measure space. Consider the set of all measurable functions $f: X \rightarrow \mathbb{K}$ such that

$$
\int_{X}|f|^{2} d \mu<\infty
$$

The space $L^{2}(\mu)$ is defined to be this set, modulo the equivalence relation which declares $f$ equivalent to $g$ if $f=g$ almost everywhere. From the inequality $2|f \bar{g}| \leq|f|^{2}+|g|^{2}$ it follows that $L^{2}(\mu)$ is a vector space and that we can define the inner product on $L^{2}(\mu)$ by

$$
\begin{equation*}
\langle f, g\rangle=\int_{X} f \bar{g} d \mu . \tag{14}
\end{equation*}
$$

```
eqn:L2_inner_p
```

23.3. Norms. Given a vector space $X$ over $\mathbb{K}$ and a semi-inner product $\langle\cdot, \cdot\rangle$, define for each $x \in X$

$$
\begin{equation*}
\|x\|:=\sqrt{\langle x, x\rangle} \tag{15}
\end{equation*}
$$

This quantity should act something like a "length" of the vector $x$. Clearly $\|x\| \geq 0$ for all $x$, and moreover we have:
erprodnorm
Theorem 23.4. Let $X$ be a semi-inner product space over $\mathbb{K}$, with $\|\cdot\|$ defined by equation (11). Then for all $x, y \in X$ and $\alpha \in \mathbb{K}$,
(a) $\|x+y\| \leq\|x\|+\|y\|$
(b) $\|\alpha x\|=|\alpha|\|x\|$.

Thus $\|\cdot\|$ is a seminorm on $X$.

If $\langle\cdot, \cdot\rangle$ is an inner product, then also
(c) $\|x\|=0$ if and only if $x=0$,
and thus $\|\cdot\|$ is a norm on $X$.

Proof. For all $x, y \in X$ we have

$$
\begin{align*}
\|x+y\|^{2} & =\langle x+y, x+y\rangle \\
& =\|x\|^{2}+2 \operatorname{Re}\langle x, y\rangle+\|y\|^{2} \\
& \leq\|x\|^{2}+2|\langle x, y\rangle|+\|y\|^{2} \\
& \leq\|x\|^{2}+2\|x\|\|y\|+\|y\|^{2}  \tag{16}\\
& =(\|x\|+\|y\|)^{2}
\end{align*}
$$

where we have used the Cauchy-Schwarz inequality in (leqn:normproofcs 10 . Taking square roots finishes lit:triangleit:scalest:definitely
the proof of item (it:triang1eit:scalest: Items (b) and (23.4) are left as exercises.

When $X$ is an inner product space, the quantity $\|x\|$ will be called the norm of $x$. Item ( (it:triangle will be referred to as the triangle inequality. On $\mathbb{R}^{n}$,

$$
\|x\|=\left(x_{1}^{2}+\cdots+x_{n}^{2}\right)^{1 / 2}
$$

is the usual Euclidean norm.
continuity Lemma 23.5. Let $H$ be an inner product space equipped with the norm topology. If $\left(x_{n}\right)$ converges to $x$ and $\left(y_{n}\right)$ converges to $y$ in $H$, then $\left(\left\langle x_{n}, y_{n}\right\rangle\right)$ converges to $\langle x, y\rangle$.

Proof. By Cauchy-Schwarz,

$$
\left|\left\langle x_{n}, y_{n}\right\rangle-\langle x, y\rangle\right| \leq\left|\left\langle x_{n}, y_{n}-y\right\rangle\right|+\left|\left\langle x_{n}-x, y\right\rangle\right| \leq\left\|x_{n}\right\|\left\|y_{n}-y\right\|+\left\|x_{n}-x\right\|\|y\| \rightarrow 0
$$

since $\left\|x_{n}-x\right\|,\left\|y_{n}-y\right\| \rightarrow 0$ and the sequence $\left\|x_{n}\right\|$ is bounded.
23.4. Orthogonality. In this section we show that many of the basic features of the Euclidean geometry of $\mathbb{K}^{n}$ extend naturally to the setting of an inner product space.

Definition 23.6. Let $H$ be an inner product space.
(i) Two vectors $x, y \in H$ are orthogonal if $\langle x, y\rangle=0$, written $x \perp y$.
(ii) Two subsets $A, B$ of $H$ are orthogonal if $x \perp y$ for all $x \in A$ and $y \in B$, written $A \perp B$.
(iii) A subset $A$ of $H$ is orthogonal if $x \perp y$ for each $x, y \in A$ with $x \neq y$ and is orthonormal if also $\langle x, x\rangle=1$ for all $x \in A$.
(iv) The orthogonal complement of a subset $E$ of $H$ is

$$
E^{\perp}=\{x \in H:\langle x, e\rangle=0 \text { for all } e \in E\} .
$$

The proof of the following lemma is an easy exercise. Indeed, the first item follows

perpbasics Lemma 23.7. If $E$ is a subset of an inner product space $H$, then
(i) $E^{\perp}$ is a closed subspace of $H$;
(ii) $E \cap E^{\perp} \subset(0)$; and
(iii) $E \subset\left(E^{\perp}\right)^{\perp}=E^{\perp \perp}$.

Theorem 23.8 (The Pythagorean Theorem). If $H$ is an inner product space and $f_{1}, \ldots f_{n}$ are mutually orthogonal vectors in $H$, then

$$
\left\|f_{1}+\cdots+f_{n}\right\|^{2}=\left\|f_{1}\right\|^{2}+\cdots+\left\|f_{n}\right\|^{2}
$$

Proof. When $n=2$, we have

$$
\begin{aligned}
\left\|f_{1}+f_{2}\right\|^{2} & =\left\|f_{1}\right\|^{2}+\left\langle f_{1}, f_{2}\right\rangle+\left\langle f_{2}, f_{1}\right\rangle+\left\|f_{2}\right\|^{2} \\
& =\left\|f_{1}\right\|^{2}+\left\|f_{2}\right\|^{2}
\end{aligned}
$$

The general case follows by induction.

Theorem 23.9 (The Parallelogram Law). If $H$ is an inner product space and $f, g \in H$, then

$$
\begin{equation*}
\|f+g\|^{2}+\|f-g\|^{2}=2\left(\|f\|^{2}+\|g\|^{2}\right) . \tag{17}
\end{equation*}
$$

```
eqn:parallelog
```

Proof. From the definition of the norm coming from the inner product,

$$
\|f \pm g\|^{2}=\|f\|^{2}+\|g\|^{2} \pm 2 \operatorname{Re}\langle f, g\rangle .
$$

Now add.

Subtracting, instead of adding, in the proof of the Parallelogram Law gives the polarization identity

$$
\|f+g\|^{2}-\|f-g\|^{2}=4 \operatorname{Re}\langle f, g\rangle
$$

in the case $\mathbb{K}=\mathbb{R}$.
larization Theorem 23.10 (The Polarization identity). If $H$ is an inner product space over $\mathbb{R}$, then

$$
\begin{equation*}
\langle f, g\rangle=\frac{1}{4}\left(\|f+g\|^{2}-\|f-g\|^{2}\right) \tag{18}
\end{equation*}
$$

eqn:polar

If $H$ is a complex Hilbert space, then

$$
\begin{equation*}
\langle f, g\rangle=\frac{1}{4}\left[\|f+g\|^{2}-\|f-g\|^{2}+\|f-i g\|^{2}-\|f+i g\|^{2}\right] \tag{19}
\end{equation*}
$$

Remark: An elementary (but tricky) theorem of von Neumann says, in the real case, that if $H$ is any vector space equipped with a norm $\|\cdot\|$ such that the parallelogram law (eqn: parallelogram I ) holds for all $f, g \in H$, then $H$ is an inner product space with inner product
given by formula (egn:polar (18) in the case of real scalars and formula (eg;polarizationC 19 in the case of complex
 check that it is indeed an inner product.)

### 23.5. Completeness.

Definition 23.11. A Hilbert space over $\mathbb{K}$ is an inner product space $X$ over $\mathbb{K}$ that is complete in the metric $d(x, y)=\|x-y\|$. (Here, as usual, $\mathbb{K}$ is either $\mathbb{C}$ or $\mathbb{R}$. ) $\triangleleft$

That the inner product spaces $\mathbb{K}^{n}$ are complete (and hence Hilbert spaces) is known from elementary analysis. (Note that the complex case follows from the real case, since the Euclidean norms are equal under the natural isomorphism $\mathbb{C}^{n} \cong \mathbb{R}^{2 n}$.)

## Theorem 23.12. $L^{2}(\mu)$ is complete.

Proof. We use Proposition prop:abs-cvg-complete 120. Accordingly suppose $\left(f_{k}\right)$ is a sequence in $L^{2}(\mu)$ and $\sum_{k=1}^{\infty}\left\|f_{k}\right\|=B<\infty$. Define

$$
G_{n}=\sum_{k=1}^{n}\left|f_{k}\right| \quad \text { and } \quad G=\sum_{k=1}^{\infty}\left|f_{k}\right| .
$$

By the triangle inequality, $\left\|G_{n}\right\| \leq \sum_{k=1}^{n}\left\|f_{k}\right\| \leq B$ for all $n$. Thus, by the Monotone Convergence Theorem and the Cauchy-Schwarz inequality,

$$
\begin{equation*}
\int_{X} G^{2} d \mu=\lim _{n \rightarrow \infty} \int_{X} G_{n}^{2} d \mu \leq B^{2} \tag{20}
\end{equation*}
$$

Thus $G$ belongs to $L^{2}(\mu)$ and in particular $G(x)<\infty$ for almost every $x$. Hence, by the definition of $G$, the sum

$$
\sum_{k=1}^{\infty} f_{k}(x)
$$

converges absolutely for almost every $x$. Hence there is a measurable function $f$ such that this sum converges a.e. to $f$. By construction, $|f| \leq G$ and thus $f \in L^{2}(\mu)$. Moreover, for all $n$ we have

$$
\left|f-\sum_{k=1}^{n} f_{k}\right|^{2} \leq(2 G)^{2}
$$

Equation (eqn: G2integrable says that $G^{2}$ is integrable, so we can apply the Dominated Convergence Theorem to obtain

$$
\lim _{n \rightarrow \infty}\left\|f-\sum_{k=1}^{n} f_{k}\right\|^{2}=\lim _{n \rightarrow \infty} \int_{X} f-\left.\sum_{k=1}^{n} f_{k}\right|^{2} d \mu=0
$$

23.6. Best approximation. The results of the preceding subsection used only the inner product, but to go further we will need to invoke completeness. From now on, then, we work only with Hilbert spaces. We begin with a fundamental approximation theorem. Recall that if $X$ is a vector space over $\mathbb{K}$, a subset $K \subseteq X$ is called convex if whenever $a, b \in K$ and $0 \leq t \leq 1$, we have $(1-t) a+t b \in K$ as well. (Geometrically, this means that when $a, b$ lie in $K$, so does the line segment joining them.)

Theorem 23.13. Suppose $H$ is a Hilbert space. If $K \subseteq H$ is a closed, convex, nonempty set, and $h \in H$, then there exists a unique vector $k_{0} \in K$ such that

$$
\left\|h-k_{0}\right\|=\operatorname{dist}(h, K):=\inf \{\|h-k\|: k \in K\}
$$

Proof. Let $d=\operatorname{dist}(h, K)=\inf _{k \in K}\|h-k\|$. First observe, if $x, y \in K$, then, by convexity, so is $v=\frac{x+y}{2}$ and in particular, $\|h-v\|^{2} \geq d^{2}$. Hence, by the parallelogram law,

$$
\begin{align*}
\left\|\frac{x-y}{2}\right\|^{2} & =\frac{1}{2}\left(\|x-h\|^{2}+\|y-h\|^{2}\right)-\left\|\frac{x+y}{2}-h\right\|^{2}  \tag{21}\\
& \leq \frac{1}{2}\left(\|x-h\|^{2}+\|y-h\|^{2}\right)-d^{2}
\end{align*}
$$

By assumption, there exists a sequence $\left(k_{n}\right)$ in $K$ so that $\left(\left\|k_{n}-h\right\|\right)$ converges to $d$.
 $m, n \geq N$ then

$$
\left\|\frac{k_{m}-k_{n}}{2}\right\|^{2}<\frac{1}{2}\left(2 d^{2}+\frac{1}{2} \epsilon^{2}\right)-d^{2}=\frac{1}{4} \epsilon^{2} .
$$

Consequently $\left\|k_{m}-k_{n}\right\|<\epsilon$ for $m, n \geq N$ and $\left(k_{n}\right)$ is a Cauchy sequence. Since $H$ is complete, $\left(k_{n}\right)$ converges to a limit $k_{0}$, and since $K$ is closed, $k_{0} \in K$. Since $\left(k_{n}-h\right)$ converges to $\left(k_{0}-h\right)$ and $\left\|k_{n}-h\right\|$ converges to $d$ it follows, by continuity of the norm, that $\left\|k_{0}-h\right\|=d$.

It remains to show that $k_{0}$ is the unique element of $K$ with this property. If $k^{\prime} \in K$ and $\left\|k^{\prime}-h\right\|=d$, then another application of $v=\left(k_{0}+k^{\prime}\right) / 2$ belongs to $K$, and $\|v-h\| \geq d$. By the parallelogram law again, equation (eqn: (21) gives

$$
0 \leq\left\|\frac{k_{0}-k^{\prime}}{2}\right\|^{2} \leq \frac{1}{2}\left(\left\|k_{0}-h\right\|^{2}+\left\|k^{\prime}-h\right\|^{2}\right)-d^{2}=\frac{1}{2}\left(d^{2}+d^{2}\right)-d^{2}=0
$$

Hence $k_{0}=k^{\prime}$.

The most important application of the preceding approximation theorem is in the case when $K=M$ is a closed subspace of the Hilbert space $H$. (Note that a subspace is always convex). What is significant is that in the case of a subspace, the minimizer $k_{0}$ has an elegant geometric description, namely, it is obtained by "dropping a perpendicular" from $h$ to $M$. This is the content of the next theorem.

Since we will use the notation often, let us write $M \leq H$ to mean that $M$ is a closed subspace of $H$. element of $M$ such that $\left\|h-f_{0}\right\|=\operatorname{dist}(h, M)$, then $\left(h-f_{0}\right) \perp M$. Conversely, if $f_{0} \in M$ and $\left(h-f_{0}\right) \perp M$, then $\left\|h-f_{0}\right\|=\operatorname{dist}(h, M)$.

Proof. Let $f_{0} \in M$ with $\left\|h-f_{0}\right\|=\operatorname{dist}(h, M)$ be given. Given $f \in M$, for $t \in \mathbb{R}$, let $\lambda=t \overline{\left\langle h-f_{0}, f\right\rangle}$. Since $f_{0}+\lambda f \in M$,

$$
\begin{aligned}
& 0 \leq\left\|h-\left(f_{0}+\lambda f\right)\right\|^{2}-\left\|h-f_{0}\right\|^{2} \\
& =\left\|\left(h-f_{0}\right)+\lambda f\right\|^{2}-\left\|h-f_{0}\right\|^{2} \\
& =2 \Re \lambda\left\langle h-f_{0}, f\right\rangle+|\lambda|^{2}\|f\|^{2} \\
& =\left[2 t+t^{2}\right]\|f\|^{2}\left|\left\langle h-f_{0}, f\right\rangle\right|^{2}
\end{aligned}
$$

for all $t$. Thus $\left\langle h-f_{0}, f\right\rangle=0$.

Conversely, suppose $f_{0} \in M$ and $\left(h-f_{0}\right) \perp M$. In particular, we have $\left(h-f_{0}\right) \perp$ $\left(f_{0}-f\right)$ for all $f \in M$. Therefore, for all $f \in M$

$$
\begin{align*}
\|h-f\|^{2} & =\left\|\left(h-f_{0}\right)+\left(f_{0}-f\right)\right\|^{2}  \tag{22}\\
& =\left\|h-f_{0}\right\|^{2}+\left\|f_{0}-f\right\|^{2} \quad(\text { why? })  \tag{23}\\
& \geq\left\|h-f_{0}\right\|^{2} . \tag{24}
\end{align*}
$$

Thus $\left\|h-f_{0}\right\|=\operatorname{dist}(h, M)$.

Proof. By Lemma $\frac{\text { Rem: } 23.7, ~ p e r p b a s i c s}{M} \subset\left(M^{\perp}\right)^{\perp}$. Now suppose that $x \in\left(M^{\perp}\right)^{\perp}$. By Theorem $\frac{\text { thm: dropaperp }}{23.14}$ applied to $x$ and $M$, there exists $m \in M$ such that $x-m \in M^{\perp}$. On the other hand, both $x$ and $m$ are in $\left(M^{\perp}\right)^{\perp}$ and thus by Lemma $\frac{\text { lem:perpbasics }}{23 . f,} x-m \in\left(M^{\perp}\right)^{\perp}$. Thus $x-m=0$


If $E$ is a subset of the Banach space $X$, and $\mathscr{E}$ is the collection of all closed subspaces $\mathcal{N}$ of $X$ such that $E \subset \mathcal{N}$, then

$$
\mathcal{M}=\cap_{\mathcal{N} \in \mathscr{E}} \mathcal{N}
$$

is the smallest closed subspace containing $E$.
_perp_perp Corollary 23.16. If $E$ is a subset of $H$, then $\left(E^{\perp}\right)^{\perp}$ is equal to the smallest closed subspace of $H$ containing $E$.

Proof. The proof uses Lemma $\frac{1 \text { lem:perpbasics }}{Z 3.1}$ freely. Evidently $E \subset\left(E^{\perp}\right)^{\perp}$. Further $\left(E^{\perp}\right)^{\perp}$ is a closed subspace. If $M$ is a closed subspace containing $E$, then $E^{\perp} \supset M^{\perp}$ and hence $\left(E^{\perp}\right)^{\perp} \subset\left(M^{\perp}\right)^{\perp}=M$ by Corollary $\begin{gathered}\text { cor:m_perp_perp } \\ 23.15 .\end{gathered}$

Corollary 23.17. A vector subspace $E$ of a Hilbert space $H$ is dense in $H$ if and only if $E^{\perp}=(0)$.

Lemma 23.18. Suppose $M, N \leq H$. If $M$ and $N$ are orthogonal, then $M+N$ is closed.

Given subspaces $M, N \leq H$ of a Hilbert space $H$, the notation $M \oplus N$ is used for $M+N$ in the case $M$ and $N$ are closed subspaces and $M \perp N$. Hence, $M \oplus N$ indicates that $M, N$ are orthogonal closed subspaces of $H$.

Proof. It suffices to prove that $M+N$ is complete. Accordingly suppose $\left(m_{k}+n_{k}\right)$ is a Cauchy sequence from $M+N$. From orthogonality, for $k, \ell \in \mathbb{N}$,

$$
\left\|m_{k}-m_{\ell}\right\|^{2}+\left\|n_{k}-n_{\ell}\right\|^{2}=\left\|\left(m_{k}+n_{k}\right)-\left(m_{\ell}+n_{\ell}\right)\right\|^{2}
$$

and hence $\left(m_{k}\right)$ and $\left(n_{k}\right)$ are both Cauchy. Since $H$ is complete and $M, N$ are closed, $M$ and $N$ are each complete. Thus $\left(m_{k}\right)$ converges to some $m \in M$ and $\left(n_{k}\right)$ converges to some $n \in N$ and thus $\left(m_{k}+n_{k}\right)$ converges to $m+n \in M+N$.

MplusMperp Corollary 23.19. If $M \leq H$, then $H=M \oplus M^{\perp}$.

Proof. Given $x \in H$, there exists $m \in M$ such that $x-m \in M^{\perp}$ by Theorem 1 thm:dropaperp Hence $x=m+(x-m) \in M \oplus M^{\perp}$.
23.7. The Riesz Representation Theorem. In this section we investigate the dual $H^{*}$ of a Hilbert space $H$. One way to construct bounded linear functionals on Hilbert space is as follows. Given a vector $g \in H$ define,

$$
L_{g}(h)=\langle h, g\rangle .
$$

Indeed, linearity of $L$ is just the linearity of the inner product in the first entry, and the boundedness of $L$ follows from the Cauchy-Schwarz inequality,

$$
\left|L_{g}(h)\right|=|\langle h, g\rangle| \leq\|g\|\|h\|
$$

So $\left\|L_{g}\right\| \leq\|g\|$, but in fact it is easy to see that $\left\|L_{g}\right\|=\|g\|$, just apply $L_{g}$ to the unit vector $g /\|g\|$. Hence, $L: H \rightarrow H^{*}$ defined by $g \mapsto L_{g}$ is a conjugate linear isometry (thus linear in the case of real scalars).

In fact, it is clear from linear algebra that every linear functional on $\mathbb{K}^{n}$ takes the form $L_{g}$. More generally, every bounded linear functional on a Hilbert space has the form just described.
:riesz_rep Theorem 23.20 (The Riesz RepresentationTheorem). If $H$ is a Hilbert space and $\lambda$ : $H \rightarrow \mathbb{K}$ is a bounded linear functional, then there exists a unique vector $g \in H$ such that $\lambda=L_{g}$. Thus the conjugate linear mapping $L$ is isometric and onto.

Proof. It has already been established that $L$ is isometric and in particular one-one. Thus it only remains to show $L$ is onto. Accordingly, let $\lambda \in H^{*}$ be given. If $\lambda=0$, then
$\lambda=L_{0}$. So, assume $\lambda \neq 0$. Since $\lambda$ is continuous by Proposition $\frac{\text { prop:bdd-iff-cns }}{Z 0 . /, \operatorname{ker} \lambda=\lambda}(\{0\})$
is a proper, closed subspace of $H$. Thus, by Theorem $\frac{\text { thm:dropaperp }}{23.14 \text { (or Corollary } \frac{\text { cor:MplusMperp }}{\text { 23.19) there }}}$
exists a nonzero vector $f \in(\operatorname{ker} \lambda)^{\perp}$ and by rescaling we may assume $\lambda(f)=1$.

Given $h \in H$, observe

$$
\lambda(h-\lambda(h) f)=\lambda(h)-\lambda(h) \lambda(f)=0 .
$$

Thus $h-\lambda(h) f \in \operatorname{ker} \lambda$ and consequently,

$$
\begin{aligned}
0 & =\langle h-\lambda(h) f, f\rangle \\
& =\langle h, g\rangle-\lambda(h)\langle f, f\rangle .
\end{aligned}
$$

Thus $\lambda=L_{g}$, where $g=\frac{f}{\|f\|^{2}}$ and the proof is complete.
23.8. Duality for Hilbert space. In the case $\mathbb{K}=\mathbb{R}$ the Riesz representation theorem identifies $H^{*}$ with $H$. In the case $\mathbb{K}=\mathbb{C}$, the mapping sending $\lambda \in H^{*}$ to the vector $h_{0}$ is conjugate linear and thus $H^{*}$ is not exactly $H$ (under this map). However, it is customary when working in complex Hilbert space not to make this distinction. In particular, it is a simple matter to identify the adjoint of a bounded operator $T: H \rightarrow H$
 follows. Observe that the mapping $\lambda: H \rightarrow \mathbb{C}$ defined by $\lambda(f)=\langle T f, h\rangle$ is (linear and) continuous. Hence, there is a vector $T^{*} h$ such that

$$
\langle T f, h\rangle=\lambda(f)=\left\langle f, T^{*} h\right\rangle .
$$

Conversely, if $S: H \rightarrow H$ is linear and

$$
\langle T f, h\rangle=\langle f, S h\rangle
$$

for all $f, h \in H$, then $S=T^{*}$. In particular $T^{*}$ is completely determined by $\langle T f, h\rangle=$ $\left\langle f, T^{*} h\right\rangle$ for all $f, h \in H$. Further properties of the adjoints on Hilbert space appear in prob:Hilbyadjoint
Problem $\frac{\text { prob: }}{23.2 .}$

A bounded operator $T$ on a Hilbert space $H$ is self-adjoint or hermitian if $T^{*}=T$. The proof of the following lemma uses the convenient fact that if $T$ is abounded operator on $H$ and if $\langle T h, g\rangle=0$ for all $h, g \in H$, then $T=0$.
:polarizeT Proposition 23.21. Suppose $T$ is a bounded self-adjoint operator on a Hilbert space $H$. If $\langle T h, h\rangle=0$ for all $h \in H$, then $T=0$.

Proof. The proof involves a polarization argument. Given $h, g \in H$, we have

$$
2 \Re\langle T h, g\rangle=\langle T(h+g), h+g\rangle=0 .
$$

Similarly,

$$
2 i \Im\langle T h, g\rangle=\langle T(h-i g), h-i g\rangle=0 .
$$

Hence $\langle T h, g\rangle=0$ and thus $T=0$.
 such that $\left(h-f_{0}\right) \perp M$. We thus obtain a well-defined function $P: H \rightarrow H$ (or, we could write $P: H \rightarrow M)$ defined by

$$
\begin{equation*}
P h=f_{0} . \tag{25}
\end{equation*}
$$

That is, $P h$ is characterized by $P h \in M$ and $(h-P h) \perp M$. If the space $M$ needs to be emphasized we will write $P_{M}$ for $P$.

A bounded operator $Q$ on a Hilbert space $H$ (meaning $Q: H \rightarrow H$ is linear and bounded) is a projection if $Q^{*}=Q$ and $Q^{2}=Q$. The following Theorem says if $Q$ is a projection, then $Q=P_{N}$, where $N$ is the range of $Q$; that is, $Q$ is uniquely determined by its range.

Theorem 23.22. Suppose $M \leq H$. The mapping $P=P_{M}$ is a projection with range M. Moreover, if $Q$ is a projection with range $N$, then
(i) if $h \in N$, then $Q h=h$;
(ii) $\|Q h\| \leq\|h\|$ for all $h \in H$;
(iii) $N \leq H$;
(iv) $N^{\perp}$ is the kernel of $Q$;
(v) $I-Q$ is a projection with range $N^{\perp}$; and
(vi) $Q=P_{N}$.

The mapping $P$ is called the orthogonal projection of $H$ on $M$ and, for $h \in H$, the vector $P h$ is the orthogonal projection of $h$ onto $M$.

Proof. In view of Corollary $\begin{aligned} & \text { cor:MplusMperp } \\ & 23.19, M+M^{\perp}\end{aligned}=H$ and $M \cap M^{\perp}=(0)$ from which it follows readily that $P$ is a linear map.

Evidently $P$ maps into $M$ and if $f \in M$, then $P f=f$ and hence $P$ maps onto $M$ and $\operatorname{PPf}=P f$ (and so $P^{2}=P$ ). Note that $P h=0$ if and only if $h \in M^{\perp}$. Thus $\operatorname{ker}(P)=M^{\perp}$. Further, $P_{M^{\perp}}=I-P$. In particular, the ranges of $P$ and $I-P$ are orthogonal.

If $h \in H$, then $h=P h+(h-P h)$. But $(h-P h) \in M^{\perp}$ and $P h \in M$, and thus, by the Pythagorean Theorem

$$
\|h\|^{2}=\|h-P h\|^{2}+\|P h\|^{2} .
$$

Hence $\|P h\| \leq\|h\|$. In particular, $P$ is a bounded operator on $H$. (See also Problem prob: complement-v-bdd

## [22.10.)

Given $g, f \in H$, using orthogonality of the ranges of $P$ and $I-P$,

$$
\begin{aligned}
\langle P f, P g\rangle & =\langle P f,(I-P) g+P g\rangle \\
& =\langle P f, g\rangle
\end{aligned}
$$

On the other hand, by the same reasoning

$$
\begin{aligned}
\langle P f, P g\rangle & =\langle P f+(I-P) f, P g\rangle \\
& =\langle f, P g\rangle
\end{aligned}
$$

Hence $P^{*}=P$ and all the claims about $P$ have now been proved. It is also immediate that $I-P$ is the projection onto $M^{\perp}$.

Turning to $Q$, suppose $Q$ is a projection and let $N$ denote the range of $Q$. Since $Q^{2}=Q$ it follows that $Q h=h$ for $h \in N$ (the range of $Q$ ).

An easy computation shows that $Q(I-Q)=0$. Thus if $h, f \in H$, then

$$
\langle Q h,(I-Q) f\rangle=\langle h, Q(I-Q) f\rangle=0 .
$$

Choosing $f=h$, it follows that $h=Q h+(I-Q) h$ is an orthogonal decomposition and hence $\|Q h\| \leq\|h\|$.

If $\left(h_{n}\right)$ is a sequence from the range of $Q$ which converges to $h \in H$, then, by continuity of $Q$, the sequence $\left(h_{n}=Q h_{n}\right)$ converges to $Q h$ and thus $h=Q h$ so that the range of $Q$ is closed.

Next, $f \in N^{\perp}$ if and only if

$$
0=\langle Q h, f\rangle=\langle h, Q f\rangle
$$

for every $h \in H$; if and only if $Q f=0$. Thus $N^{\perp}=\operatorname{ker}(Q)$.

Finally, an easy argument shows $I-Q$ is a projection too. In particular, $f$ is in the range of $I-Q$ if and only if $(I-Q) f=f$. On the other hand $(I-Q) f=f$ if and only if $Q f=0$. Thus the range of $I-Q$ is the kernel of $Q$. Finally, given $h \in H$, since $h-Q h=(I-Q) h \in \operatorname{ker}(Q)=N^{\perp}$, it follows that $Q h=P_{N} h$.
23.9. Orthonormal Sets and Bases. Recall, a subset $E$ of a Hilbert space $H$ is orthonormal if $\|e\|=1$ for all $e \in E$, and if $e, f \in E$ and $e \neq f$, then $e \perp f$.

Definition 23.23. An orthonormal set is maximal if it is not contained in any larger orthonormal set. A maximal orthonormal set is called an (orthonormal) basis for H. $\triangleleft$

Proposition 23.24. An orthonormal set $E$ is maximal if and only if the only vector orthogonal to $E$ is the zero vector. Equivalently, an orthonormal set $E$ is maximal if and only if the span of $E$ is dense in $H$.

To prove the proposition use $H=\bar{E} \oplus E^{\perp}$.

Remark 23.25. It must be stressed that a basis in the above sense need not be a basis in the sense of linear algebra; that is, a basis for $H$ as a vector space. In particular, it is always true that an orthonormal set is linearly independent (Exercise: prove this statement), but in general an orthonormal basis need not span $H$. In fact, if $E$ is an infinite orthonormal subset of $H$, then $E$ does not span $H$. See Problem $\frac{\text { prob:Hamel-basis }}{22.6}$

Example 23.26. Here are some common examples of orthonormal bases.
(a) Of course the standard basis $\left\{e_{1}, \ldots, e_{n}\right\}$ is an orthonormal basis of $\mathbb{K}^{n}$.
(b) In much the same way we get a orthonormal basis of $\ell^{2}(\mathbb{N})$; for each $n$ define

$$
e_{n}(k)= \begin{cases}1 & \text { if } k=n \\ 0 & \text { if } k \neq n\end{cases}
$$

It is straightforward to check that the set $E=\left\{e_{n}\right\}_{n=1}^{\infty}$ is orthonormal. In fact, it is a basis. To see this, notice that if $h: \mathbb{N} \rightarrow \mathbb{K}$ belongs to $\ell^{2}(\mathbb{N})$, then $\left\langle h, e_{n}\right\rangle=h(n)$, and hence if $h \perp E$, we have $h(n)=0$ for all $n$, so $h=0$.
(c) Let $H=L^{2}[0,1]$. Consider for $n \in \mathbb{Z}$ the functions $e_{n}(x)=e^{2 i \pi n x}$. An easy exercise shows this set is orthonormal. Though not obvious, it is in fact a basis. (See Problem prob:basisL201 23.7.)
23.10. Basis expansions. Our ultimate goal in this section is to show that vectors in Hilbert space admit expansions as (possibly infinite) linear combinations of basis vectors.

Theorem 23.27. Let $\left\{e_{1}, \ldots e_{n}\right\}$ be an orthonormal set in $H$, and let $M=\operatorname{span}\left\{e_{1}, \ldots e_{n}\right\}$.
The orthogonal projection $P=P_{M}$ onto $M$ is given by, for $h \in H$,

$$
\begin{equation*}
P h=\sum_{j=1}^{n}\left\langle h, e_{j}\right\rangle e_{j} . \tag{26}
\end{equation*}
$$

Proof. Given $h \in H$, let $g=\sum_{j=1}^{n}\left\langle h, e_{j}\right\rangle e_{j}$. Since $g \in M$, it suffices to show $(h-g) \perp M$.
For $1 \leq m \leq n$,

$$
\begin{aligned}
\left\langle h-g, e_{m}\right\rangle= & \left\langle h, e_{m}\right\rangle-\left\langle\sum_{j=1}^{n}\left\langle h, e_{j}\right\rangle e_{j}, e_{m}\right\rangle \\
& =\left\langle h, e_{m}\right\rangle-\sum_{j=1}^{n}\left\langle h, e_{j}\right\rangle\left\langle e_{j}, e_{m}\right\rangle \\
& =\left\langle h, e_{m}\right\rangle-\left\langle h, e_{m}\right\rangle=0 .
\end{aligned}
$$

It follows that $h-g$ is orthogonal to $\left\{e_{1}, \ldots, e_{n}\right\}$ and hence to $M$.

Theorem 23.28 (Gram-Schmidt process). If $\left(f_{n}\right)_{n=1}^{\infty}$ is a linearly independent sequence in $H$, then there exists an orthonormal sequence $\left(e_{n}\right)_{n=1}^{\infty}$ such that for each $n$, $\operatorname{span}\left\{f_{1}, \ldots f_{n}\right\}=$ $\operatorname{span}\left\{e_{1}, \ldots e_{n}\right\}$.

Proof. Put $e_{1}=f_{1} /\left\|f_{1}\right\|$. Assuming $e_{1}, \ldots e_{n}$ have been constructed satisfying the conditions of the theorem, let $g_{n+1}=\sum_{j=1}^{n}\left\langle f_{n+1}, e_{j}\right\rangle e_{j}$, the orthogonal projection of $f_{n+1}$ onto $M_{n}=\operatorname{span}\left\{e_{1}, \ldots, e_{n}\right\}$. Thus $g_{n+1}$ is orthogonal to $M_{n}$ and not 0 . Let $e_{n+1}=\frac{g_{n}}{\left\|g_{n+1}\right\|}$.

FIX It follows from the Gram-Schmidt process that $H$ is finite dimensional as a Hilbert space if and only if $H$ is finite dimensional as a vector space (and these dimensions agree).

Theorem 23.29 (Bessel's inequality). If $\left\{e_{1}, e_{2}, \ldots\right\}$ is an orthonormal sequence in $H$, then for all $h \in H$

$$
\begin{equation*}
\sum_{n=1}^{\infty}\left|\left\langle h, e_{n}\right\rangle\right|^{2} \leq\|h\|^{2} \tag{27}
\end{equation*}
$$

Proof. For $N \in \mathbb{N}^{+}$, let $P_{N}$ denote the projection onto $M_{N}=\operatorname{span}\left(\left\{e_{1}, \ldots, e_{N}\right\}\right)$. Given $h \in H$, Theorem thm:fd_proj
$h \in H$, Theorem 23.27 and orthogonality gives,

$$
\begin{aligned}
& \|h\|^{2}=\left\|P_{N} h+\left(I-P_{N}\right) h\right\|^{2} \\
& \quad=\left\|P_{n} h\right\|^{2}+\left\|\left(I-P_{N}\right) h\right\|^{2} \\
& \quad \geq\left\|P_{N} h\right\|^{2} \\
& \quad=\sum_{j=1}^{N}\left|\left\langle h, e_{j}\right\rangle\right|^{2} .
\end{aligned}
$$

Since the inequality holds for all $N$, the proof is complete.
le_support Corollary 23.30. If $E \subset H$ is an orthonormal set and $h \in H$, then $\langle h, e\rangle$ is nonzero for at most countably many $e \in E$.

Proof. Fix $h \in H$ and a positive integer $N$, and define

$$
E_{N}=\left\{e \in E:|\langle h, e\rangle| \geq \frac{1}{N}\right\} .
$$

We claim that $E_{N}$ is finite. If not, then it contains a countably infinite subset $\left\{e_{1}, e_{2}, \ldots\right\}$. Applying Bessel's inequality to $h$ and this subset, we get

$$
\|h\|^{2} \geq \sum_{n=1}^{\infty}\left|\left\langle h, e_{n}\right\rangle\right|^{2} \geq \sum_{n=1}^{\infty} \frac{1}{N}=+\infty,
$$

a contradiction. Hence,

$$
\{e \in E:\langle h, e\rangle \neq 0\}=\bigcup_{N=1}^{\infty} E_{N}
$$

is a countable union of finite sets, and therefore countable.
ral_bessel Corollary 23.31. Suppose $E \subset H$ is an orthonormal set and let $\mathcal{F}$ denote the collection of finite subsets of $E$. If $h \in H$, then

$$
\begin{equation*}
\sup \left\{\sum_{e \in F}|\langle h, e\rangle|^{2}: F \in \mathcal{F}\right\} \leq\|h\|^{2} \tag{28}
\end{equation*}
$$

At this point we pause to discuss convergence of infinite series in Hilbert space. We have already encountered ordinary convergence and absolute convergence in our discussion of completeness: recall that the series $\sum_{n=1}^{\infty} h_{n}$ converges if $\lim _{N \rightarrow \infty} \sum_{n=1}^{N} h_{n}$ exists; its limit $h$ is called the sum of the series. The series converges absolutely if $\sum_{n=1}^{\infty}\left\|h_{n}\right\|<\infty$ and absolute convergence implies convergence.

The series $\sum_{n=1}^{\infty} h_{n}$ is unconditionally convergent if there exists an $h \in H$ such that for each bijection $\varphi: \mathbb{N} \rightarrow \mathbb{N}$ the series $\sum_{n=1}^{\infty} h_{\varphi(n)}$ converges to $h$. (In other words, every reordering of the series converges, and to the same sum.) Of course absolute
convergence implies unconditional convergence. For ordinary scalar series, or in a finite dimensional Hilbert space such as $\mathbb{K}^{n}$, unconditional convergence implies absolute convergence; however in infinite dimensional Hilbert space unconditional convergence need not imply absolute convergence as the example following Theorem $\frac{\operatorname{thm}: u c c}{23.32}$ shows.
thm:ucc Theorem 23.32. Suppose $E=\left\{e_{1}, e_{2}, \ldots\right\} \subset H$ is a countable orthonormal set and $\left(a_{n}\right)$ is a sequence of complex numbers. The following are equivalent.
(i) the series $\sum_{j=1}^{\infty} a_{j} e_{j}$ converges;
(ii) $\sum_{j=1}^{\infty}\left|a_{j}\right|^{2}$ converges; and
(iii) the series $\sum_{j=1}^{\infty} a_{j} e_{j}$ converges unconditionally.

Further, if $h \in H$, then the series

$$
\begin{equation*}
\sum_{j=1}^{\infty}\left\langle h, e_{j}\right\rangle e_{j} \tag{29}
\end{equation*}
$$

is unconditionally convergent and, letting g denote the (unconditional) sum,

$$
\left\langle g, e_{j}\right\rangle=\left\langle h, e_{j}\right\rangle
$$

Suppose $\left\{e_{1}, e_{2}, \ldots\right\}$ is a countable orthonormal set in a Hilbert space $H$. The series

$$
\sum_{j=1}^{\infty} \frac{1}{j} e_{j}
$$

is Cauchy (verify this as an exercise) and hence converges to some $h \in H$. From Theorem $\frac{\text { thm:ucc }}{23.32 \text { it }}$ follows that $\left\langle h, e_{j}\right\rangle=\frac{1}{j}$ and the series above converges unconditionally to $h$. On the other hand, this series does not converge absolutely and hence unconditional convergence does not imply absolute convergence.

Proof. Let $s_{n}$ denote the partial sums of the series $\sum_{j=1}^{\infty} a_{j} e_{j}$,

$$
s_{n}=\sum_{j=1}^{n} a_{j} e_{j}
$$

Since $H$ is complete, the series $\sum_{j=1}^{\infty} a_{j} e_{j}$ converges if and only if for each $\epsilon>0$ there is an $N$ so that for all $m \geq n \geq N$,

$$
\begin{equation*}
\left\|s_{m}-s_{n}\right\|^{2}=\sum_{j=n+1}^{m}\left|a_{j}\right|^{2}<\epsilon \tag{30}
\end{equation*}
$$

if and only if the series $\sum_{j=N+1}^{m}\left|a_{j}\right|^{2}<\epsilon$ converges. Hence items (iit:convjitpequares. (ii1) are equivalent.

Now suppose $\sum_{j=1}^{\infty}\left|a_{j}\right|^{2}$ converges and let $\varphi: \mathbb{N} \rightarrow \mathbb{N}$. For numerical series, absolute convergence implies conditional convergence. Hence $\sum_{j=1}^{\infty}\left|a_{\varphi(j)}\right|^{2}$ converges and therefore, using the equivalence between item (iit:squares (iit: implies (iit it follows that the series

$$
\sum_{k=1}^{\infty} a_{\varphi(k)} e_{\varphi(k)}
$$

converges to some $g^{\prime}$, the limit of the partial sums

$$
s_{n}^{\prime}=\sum_{j=1}^{n} a_{\varphi(j)} e_{\varphi(j)} .
$$

It remains to show $g^{\prime}=g$.
Given $\epsilon>0$, choose $N$ so that ( $\left(\frac{\mathrm{eq} \text { : } \mathrm{sj}) \text { hog }}{}\right.$ holds. Now choose $M \geq N$ so that

$$
\{1,2, \ldots N\} \subseteq\{\varphi(1), \varphi(2), \ldots \varphi(M)\}
$$

For $n \geq M$, let $G_{n}$ be the symmetric difference of the sets $\{1, \ldots n\}$ and $\{\varphi(1), \ldots \varphi(n)\}$ (that is, their union minus their intersection). Since $n \geq M$, the set $G_{n} \subset\{N+1, N+$
$2, \ldots\}$. It follows that

$$
\begin{align*}
\left\|s_{n}-s_{n}^{\prime}\right\|^{2} & =\left\|\sum_{k \in G_{n}} \pm a_{k} e_{k}\right\|^{2}  \tag{31}\\
= & \sum_{k \in G_{n}}\left|a_{k}\right|^{2}  \tag{32}\\
& \leq \sum_{N+1}^{\infty}\left|a_{j}\right|^{2}  \tag{33}\\
& <\epsilon \tag{34}
\end{align*}
$$

Thus $g^{\prime}=g$. Hence item (iit: squares (iil) implies item (iit:uncond (ii1) and the proof of the first part of the theorem is complete.

For $h \in H$ Bessel's inequality implies the convergence of $\sum\left|\left\langle h, e_{j}\right\rangle\right|^{2}$ and thus, by what has already been proved, the series $\sum\left\langle h, e_{j}\right\rangle e_{j}$ converges unconditionally to some $g \in H$. To complete the proof, given $\epsilon>0$, choose $N$ so that if $n \geq N$, then

$$
\left\|g-\sum_{j=1}^{n}\left\langle h, e_{j}\right\rangle e_{j}\right\|<\epsilon
$$

It follows that, by the Cauchy-Schwarz inequality, for $m \leq n$,

$$
\left|\left\langle g-\sum_{j=1}^{n}\left\langle h, e_{j}\right\rangle e_{j}, e_{m}\right\rangle\right|^{2} \leq\left\|g-\sum_{j=1}^{n}\left\langle h, e_{j}\right\rangle e_{j}\right\|\left\|e_{m}\right\|<\epsilon
$$

On the other hand,

$$
\left\langle g-\sum_{j=1}^{n}\left\langle h, e_{j}\right\rangle e_{j}, e_{m}\right\rangle=\left\langle g, e_{m}\right\rangle-\left\langle h, e_{m}\right\rangle
$$

and the desired conclusion follows.

There is another notion of convergence in Hilbert space. Let $I$ be an index set and let $\mathcal{F}$ denote the collection of finite subsets of $I$. Given $\left\{h_{i}: i \in I\right\}$, a collection of
elements of $H$, the series

$$
\sum_{i \in I} h_{i}
$$

converges as a net if there exists $h \in H$ such that for every $\epsilon>0$ there exists an $F \in \mathcal{F}$ such that for every $F \subset G \in \mathcal{F}$,

$$
\left\|\sum_{i \in G} h_{i}-h\right\|<\epsilon .
$$

prop:net Proposition 23.33. If $E$ is an orthonormal subset of a Hilbert space $H$ and $h \in H$, then

$$
\sum_{e \in E}\langle h, e\rangle e
$$

converges (as a net). Moreover, if $g$ is the limit (as a net), then, for each $e \in E$,

$$
\langle g, e\rangle=\langle h, e\rangle .
$$

If $\left(h_{j}\right)$ is a sequence from $H$ and

$$
\sum_{j \in \mathbb{N}^{+}} h_{j}
$$

converges (as a net) to some $h \in H$, then

$$
\sum_{j=1}^{\infty} h_{j}
$$

converges unconditionally to $h$.

Proof. Let $E_{0}=\{e \in E:\langle h, e\rangle \neq 0\}$. From Corollary $\frac{\text { cor: countable_support }}{23.30, E_{0} \text { is at most countable. }}$
Suppose $E_{0}$ is countable and choose an enumeration, $E_{0}=\left\{e_{1}, e_{2}, \ldots\right\}$. By Theorem
thm:ucc

$$
\sum_{j=1}^{\infty}\left\langle h, e_{j}\right\rangle e_{j}
$$

converges unconditionally to some $g \in H$ and moreover $\left\langle g, e_{j}\right\rangle=\left\langle h, e_{j}\right\rangle$ for all $j$. Given $\epsilon>0$, there is an $N$ so that if $n \geq N$, then

$$
\left|g-\sum_{j=1}^{n}\left\langle h, e_{j}\right\rangle e_{j}\right|<\epsilon
$$

and, from Bessel's inequality,

$$
\sum_{j=N}^{\infty}\left|\left\langle g, e_{j}\right\rangle\right|^{2}<\epsilon^{2}
$$

Let $F=\left\{e_{1}, \ldots, e_{N}\right\}$. If $G \subset E$ is finite and $F \subset G$, then

$$
\left\|g-\sum_{e \in G}\langle g, e\rangle\right\| \leq\left\|g-\sum_{j=1}^{N}\left\langle g, e_{j}\right\rangle\right\|+\left\|\sum_{e \in G \backslash F}\langle g, e\rangle e\right\|
$$



Hence $\sum_{e \in E}\langle h, e\rangle$ converges as a net to $g$. Further, by construction, $\langle g, e\rangle=\langle h, e\rangle$ for $e \in E_{0}$. On the other hand, if $e \notin E_{0}$, then, for each $n$,

$$
\left\langle\sum_{j=1}^{n}\left\langle h, e_{j}\right\rangle e_{j}, e\right\rangle=0
$$

and thus $\langle g, e\rangle=0=\langle h, e\rangle$.

The proof of the last part of the proposition is left as a (challenging) exercise. See Problem prob:netvuncond
Problem 23.14.
m:parseval Theorem 23.34. If $E \subset H$ is an orthonormal set, then the following are equivalent:
(a) $E$ is a (orthonormal) basis for $H$;
it:basis (b) $h=\sum_{e \in E}\langle h, e\rangle e$ for each $h \in H$;
(c) $\langle g, h\rangle=\sum_{e \in E}\langle g, e\rangle\langle e, h\rangle$ for each $g, h \in H$; and
(d) $\|h\|^{2}=\sum_{e \in E}|\langle h, e\rangle|^{2}$ for each $h \in H$.

Proof. Suppose $E$ is an orthonormal set in $H$ and $h \in H$. In this case,

$$
\sum_{e \in E}\langle h, e\rangle e
$$

converges (as a net) to some $g \in H$ and moreover $\langle g, e\rangle=\langle h, e\rangle$ for all $e \in E$. Suppose $g \neq h$ and let $f=\frac{g-h}{\|g-h\|}$ so that $f$ is a unit vector. If $e \in E$, then

$$
\langle f, e\rangle=\frac{1}{\|g-h\|}\langle g-h, e\rangle=0
$$

and thus $E$ is not maximal. Hence (a) implies (b).

Now suppose (b) holds and let $h, g \in H$ be given. Given $\epsilon$, choose a finite subset $F$ of $E$ such that if $F \subset G \subset E$, then

$$
\left\|h-\sum_{e \in G}\langle h, e\rangle e\right\|,\left\|g-\sum_{e \in G}\langle g, e\rangle e\right\|<\sqrt{\epsilon}
$$

and observe, using the Cauchy-Schwarz inequality,

$$
\begin{aligned}
\epsilon & >\mid\left\langle h-\sum_{e \in G}\langle h, e\rangle e, g-\sum_{f \in G}\langle g, e\rangle e\right| \\
& =\left|\langle h, g\rangle-\sum_{e \in G}\langle h, e\rangle\langle e, g\rangle\right|
\end{aligned}
$$

Hence (b) implies (c).
Item (d) follows from (c) by choosing $g=h$. Finally, suppose that (a) does not hold. In that case there exists a unit vector $h \in H$ such that $h$ is orthogonal to $E$. In
particular,

$$
\sum_{e \in E}|\langle h, e\rangle|=0
$$

and (d) does not hold.

Theorem 23.35. Every Hilbert space $H \neq(0)$ has an orthonormal basis.

Proof. The proof is essentially the same as the Zorn's lemma proof that every vector space has a basis. Let $H$ be a Hilbert space and $\mathcal{E}$ the collection of orthonormal subsets of $H$, partially ordered by inclusion. Since $H \neq(0)$, the collection $\mathcal{E}$ is not empty. If $\left(E_{\alpha}\right)$ is an ascending chain in $\mathcal{E}$, then it is straightforward to verify that $\cup_{\alpha} E_{\alpha}$ is an orthonormal set, and is an upper bound for $\left(E_{\alpha}\right)$. Thus by Zorn's lemma, $\mathcal{E}$ has a maximal element. This maximal element is then a basis (maximal orthonormal set).

Remark 23.36. If $H$ has a finite orthonormal basis $E=\left\{e_{1}, \ldots, e_{n}\right\}$, then by Theorem thm: parseba is
$23.34(\mathrm{fb}), \mathrm{E}$ spans (in the sense of linear algebra) and is therefore a vector space (Hamel) basis for $H$. Hence $H$ has dimension $n$ as a vector space and further every orthonormal basis of $H$ has exactly $n$ elements.

On the other hand, if $H$ has an infinite orthonormal basis $E$, then it contains an infinite linearly independent set (the basis $E$ ) and so has infinite dimension as a vector space.

Theorem 23.37. Any two bases of a Hilbert space $H$ have the same cardinality.

In the proof we let $|S|$ denote the cardinality of the set $S$.

Proof. Suppose $E, F$ are orthonormal bases for $H$. If $E$ is finite, then $E$ is a basis in the vector space sense and thus $H$ is finite dimensional as a vector space. Since $F$ is orthonormal, it is linearly independent and hence $|F| \leq|E|$. Thus $F$ is also a vector space basis for $H$ and so $|F|=|E|$. By symmetry, either both $E$ and $F$ are finite and have the same cardinality or both are infinite. Accordingly suppose both are infinite.

Fix $e \in E$ and consider the set

$$
F_{e}=\{f \in F \mid\langle f, e\rangle \neq 0\} .
$$

Since $F$ is orthonormal, each $F_{e}$ is at most countable by Corollary $\frac{\text { cor: countable_support }}{23.30, \text { and since } E \text { is }}$ a basis, each $f \in F$ belongs to at least one $F_{e}$. Thus $\bigcup_{e \in E} F_{e}=F$, and

$$
|F|=\left|\bigcup_{e \in E} F_{e}\right| \leq|E| \cdot \aleph_{0}=|E|
$$

where the last equality holds since $E$ is infinite.

By symmetry, $|F| \leq|E|$ and the proof is complete.

In light of this theorem, we make the following definition.

Definition 23.38. The (orthogonal) dimension of a Hilbert space $H$ is the cardinality of any orthonormal basis, and is denoted $\operatorname{dim} H$. If $\operatorname{dim} H$ is finite or countable, $H$ is separable or separable Hilbert space.
23.11. Weak convergence. In addition to the norm topology, Hilbert spaces carry another topology called the weak topology. In these notes we will stick to the seperable case and just study weakly convergent sequences.

Definition 23.39. Let $H$ be a seperable Hilbert space. A sequence $\left(h_{n}\right)$ in $H$ converges weakly to $h \in H$ if for all $g \in H$,

$$
\left\langle h_{n}, g\right\rangle \rightarrow\langle h, g\rangle .
$$

The Cauchy-Schwarz inequality implies if $\left(h_{n}\right)$ converges to $h$ in norm, then $\left(h_{n}\right)$ converges weakly to $h$. However, when $H$ is infinite-dimensional, the converse can fail. For instance, let $\left\{e_{n}\right\}_{n=1}^{\infty}$ be an orthonormal basis for $H$. Then $\left(e_{n}\right)$ converges to 0 weakly. (The proof is an exercise, see Problem prob:hilbert-weak 23.9 ). On the other hand, the sequence $\left(e_{n}\right)$ is not norm convergent, since it is not Cauchy. In this section weak convergence is characterized as "bounded coordinate-wise convergence" and it is shown that the unit ball of a separable Hilbert space is weakly sequentially compact.
akconverge Proposition 23.40. Let $H$ be a Hilbert' space with orthonormal basis $\left\{e_{j}\right\}_{j=1}^{\infty}$. A sequence $\left(h_{n}\right)$ in $H$ is weakly convergent if and only if
i) $\sup _{n}\left\|h_{n}\right\|<\infty$, and
ii) $\lim _{n}\left\langle h_{n}, e_{j}\right\rangle$ exists for each $j$.

Proof. Suppose $\left(h_{n}\right)$ converges to $h$ weakly. For each $n$

$$
L_{n}(g)=\left\langle g, h_{n}\right\rangle
$$

is a bounded linear functional on $H$. Since, for fixed $g$, the sequence $\left|L_{n}(g)\right|$ converges, it is bounded. Thus, the family of linear functionals $\left(L_{n}\right)$ is pointwise bounded and hence,
by the Principle of Uniform boundedness, $\sup \left\|h_{n}\right\|=\sup \left\|L_{n}\right\|<\infty$, showing (i) holds. Item (ii) is immediate from the definition of weak convergence.

Conversely, suppose (i) and (ii) hold, let $M=\sup \left\|h_{n}\right\|$. Define

$$
\hat{h}_{j}=\lim \left\langle h_{n}, e_{j}\right\rangle
$$

We will show that $\sum_{j}\left|\hat{h}_{j}\right|^{2} \leq M$ (so that the series $\sum \hat{h}_{j} e_{j}$ is norm convergent in $H$ ); we then define $h$ to be the sum of this series and show that $h_{n} \rightarrow h$ weakly.

For positive integers $J$ and all $n$,

$$
\sum_{j=1}^{J}\left|\left\langle h_{n}, e_{j}\right\rangle\right|^{2} \leq\left\|h_{n}\right\|^{2} \leq M^{2}
$$

by Bessel's inequality. Thus,

$$
\sum_{j=1}^{J}\left|\hat{h}_{j}\right|^{2}=\sum_{j=1}^{J} \lim _{n}\left|\left\langle h_{n}, e_{j}\right\rangle\right|^{2}=\lim _{n} \sum_{j=1}^{J}\left|\left\langle h_{n}, e_{j}\right\rangle\right|^{2} \leq M^{2} .
$$

Thus $\sum_{j}\left|\hat{h}_{j}\right|^{2} \leq M^{2}$ and therefore the series $\sum_{j} \hat{h}_{j} e_{j}$ is norm convergent to some $h \in H$ such that $\left\langle h, e_{j}\right\rangle=\hat{h}_{j}$ by Theorem $\frac{\text { thm:ucc }}{23.32 .}$ By Theorem $\frac{\text { thm: parseval }}{23.34,\|h\| \leq} M$.

Now we prove that $\left(h_{n}\right)$ converges to $h$ weakly. Fix $g \in H$ and let $\epsilon>0$ be given. Since $g=\sum_{j}\left\langle g, e_{j}\right\rangle e_{j}$ (where the series is norm convergent) there exists an positive integer $J$ large enough so that

$$
\left\|g-\sum_{j=1}^{J}\left\langle g, e_{j}\right\rangle e_{j}\right\|=\left\|\sum_{j=J+1}^{\infty}\left\langle g, e_{j}\right\rangle e_{j}\right\|<\epsilon
$$

Let $g_{0}=\sum_{j=1}^{J}\left\langle g, e_{j}\right\rangle e_{j}$, write $g=g_{0}+g_{1}$, observe $\left\|g_{1}\right\|<\epsilon$ and estimate,

$$
\left|\left\langle h_{n}-h, g\right\rangle\right| \leq\left|\left\langle h_{n}-h, g_{0}\right\rangle\right|+\left|\left\langle h_{n}-h, g_{1}\right\rangle\right| .
$$

By (ii), the first term on the right hand side goes to 0 with $n$, since $g_{0}$ is a finite sum of $e_{j}$ 's. By Cauchy-Schwarz, the second term is bounded by $2 M \epsilon$. As $\epsilon$ was arbitrary, we see that the left-hand side goes to 0 with $n$.

It turns out, if $\left(h_{n}\right)$ converges to $h$ weakly, then $\|h\| \leq \lim \inf \left\|h_{n}\right\|$ and further, still assuming $\left(h_{n}\right)$ converges weakly to $h,\|h\|=\lim \left\|h_{n}\right\|$ if and only if $\left(h_{n}\right)$ converges to $h$ in norm. See Problem $\frac{\text { prob:hilbert-weak }}{23.9 \text {. }}$

Theorem 23.41 (Weak compactness of the unit ball in Hilbert space). If $\left(h_{n}\right)$ is a bounded sequence in a separable Hilbert space $H$, then $\left(h_{n}\right)$ has a weakly convergent subsequence.

Theorem $\frac{\text { thm:weak-compactness }}{23.41 \text { holds without the separability hypothesis, but the proof is much }}$ simpler with the hypothesis.

Proof. Using the previous proposition, it suffices to fix an orthonromal basis $\left(e_{j}\right)$ and produce a subsequence $\left(h_{n_{k}}\right)_{k}$ such that $\left\langle h_{n_{k}}, e_{j}\right\rangle$ converges for each $j$. This is a standard "diagonalization" argument, and the details are left as an exercise (Problem prob;weak-compactness

### 23.12. Problems.

Problem 23.1. Prove the complex form of the polarization identity: if $H$ is a Hilbert space over $\mathbb{C}$, then for all $g, h \in H$

$$
\langle g, h\rangle=\frac{1}{4}\left(\|g+h\|^{2}-\|g-h\|^{2}+i\|g+i h\|^{2}-i\|g-i h\|^{2}\right)
$$

lbyadjoint Problem 23.2. (Adjoint operators) Let $H$ be a Hilbert space and $T: H \rightarrow H$ a bounded linear operator.
a) Prove there is a unique bounded operator $T^{*}: H \rightarrow H$ satisfying $\langle T g, h\rangle=$ $\left\langle g, T^{*} h\right\rangle$ for all $g, h \in H$, and $\left\|T^{*}\right\|=\|T\|$.
b) Prove, if $S, T \in B(H)$, then $(a S+T)^{*}=\bar{a} S^{*}+T^{*}$ for all $a \in \mathbb{K}$, and that $T^{* *}=T$.
c) Prove $\left\|T^{*} T\right\|=\|T\|^{2}$.
d) Prove $\operatorname{ker} T$ is a closed subspace of $H, \overline{(\operatorname{ranT} T)}=\left(\operatorname{ker} T^{*}\right)^{\perp}$ and $\operatorname{ker} T^{*}=(\operatorname{ran} T)^{\perp}$.
isometries Problem 23.3. Let $H, K$ be Hilbert spaces. A linear transformation $T: H \rightarrow K$ is called isometric if $\|T h\|=\|h\|$ for all $h \in H$, and unitary if it is a surjective isometry. Prove the following:
a) $T$ is an isometry if and only if $\langle T g, T h\rangle=\langle g, h\rangle$ for all $g, h \in H$, if and only if $T^{*} T=I$ (here $I$ denotes the identity operator on $H$ ).
b) $T$ is unitary if and only if $T$ is invertible and $T^{-1}=T^{*}$, if and only if $T^{*} T=$ $T T^{*}=I$.
c) Prove, if $E \subset H$ is an orthonormal set and $T$ is an isometry, then $T(E)$ is an orthonormal set in $K$.
d) Prove, if $H$ is finite-dimensional, then every isometry $T: H \rightarrow H$ is unitary.
e) Consider the shift operator $S \in B\left(\ell^{2}(\mathbb{N})\right)$ defined by

$$
\begin{equation*}
S\left(a_{0}, a_{1}, a_{2}, \ldots\right)=\left(0, a_{0}, a_{1}, \ldots\right) \tag{35}
\end{equation*}
$$

Prove $S$ is an isometry, but not unitary. Compute $S^{*}$ and $S S^{*}$.

Problem 23.4. For any set $J$, let $\ell^{2}(J)$ denote the set of all functions $f: J \rightarrow \mathbb{K}$ such that $\sum_{j \in J}|f(j)|^{2}<\infty$. Then $\ell^{2}(J)$ is a Hilbert space.
a) Prove $\ell^{2}(I)$ is isometrically isomorphic to $\ell^{2}(J)$ if and only if $I$ and $J$ have the same cardinality. (Hint: use Problem $\frac{\text { prob;hilby-isometries }}{23.3(\mathrm{c}) \text { ) }}$
b) Prove, if $H$ is any Hilbert space, then $H$ is isometrically isomorphic to $\ell^{2}(J)$ for some set $J$.
ense-in-L2 Problem 23.5. Let $(X, \mathscr{M}, \mu)$ be a $\sigma$-finite measure space. Prove the simple functions that belong to $L^{2}(\mu)$ are dense in $L^{2}(\mu)$.

Problem 23.6. (The Fourier basis) Prove the set $E=\left\{e_{n}(t):=e^{2 \pi i n t} \mid n \in \mathbb{Z}\right\}$ is an orthonormal basis for $L^{2}[0,1]$. (Hint: use the Stone-Weierstrass theorem to prove that the set of trigonometric polynomials $P=\left\{\sum_{n=-M}^{N} c_{n} e^{2 \pi i n t}\right\}$ is uniformly dense in the space of continuous functions $f$ on $[0,1]$ that satisfy $f(0)=f(1)$. Then show that this space of continuous functions is dense in $L^{2}[0,1]$. Finally show that if $f_{n}$ is a sequence in $L^{2}[0,1]$ and $f_{n} \rightarrow f$ uniformly, then also $f_{n} \rightarrow f$ in the $L^{2}$ norm.)

Problem 23.7. Let $\left(g_{n}\right)_{n \in \mathbb{N}}$ be an orthonormal basis for $L^{2}[0,1]$, and extend each function to $\mathbb{R}$ by declaring it to be 0 off of $[0,1]$. Prove the functions $h_{m n}(x):=$ $\mathbf{1}_{[m, m+1]}(x) g_{n}(x-m), n \in \mathbb{N}, m \in \mathbb{Z}$ form an orthonormal basis for $L^{2}(\mathbb{R})$. (Thus $L^{2}(\mathbb{R})$ is separable.)

Problem 23.8. Let $(X, \mathscr{M}, \mu),(Y, \mathscr{N}, \nu)$ are $\sigma$-finite measure spaces, and let $\mu \times \nu$ denote the product measure. Prove, if $\left(f_{m}\right)$ and $\left(g_{n}\right)$ are orthonormal bases for $L^{2}(\mu), L^{2}(\nu)$ respectively, then the collection of functions $\left\{h_{m n}(x, y)=f_{m}(x) g_{n}(y)\right\}$ is an orthonromal basis for $L^{2}(\mu \times \nu)$. Use this result to construct an orthonormal basis for $L^{2}\left(\mathbb{R}^{n}\right)$, and conclude that $L^{2}\left(\mathbb{R}^{n}\right)$ is separable.

Problem 23.9. (Weak Convergence)
a) Prove, if $\left(h_{n}\right)$ converges to $h$ in norm, then also $\left(h_{n}\right)$ converges to $h$ weakly. (Hint: Cauchy-Schwarz.)
b) Prove, if $H$ is infinite-dimensional, and $\left(e_{n}\right)$ is an orthonormal sequence in $H$, then $e_{n} \rightarrow 0$ weakly, but $e_{n} \nrightarrow 0$ in norm. (Thus weak convergence does not imply norm convergence.)
c) Prove $\left(h_{n}\right)$ converges to $h$ in norm if and only if $\left(h_{n}\right)$ converges to $h$ weakly and

$$
\left\|h_{n}\right\| \rightarrow\|h\| .
$$

d) Prove if $\left(h_{n}\right)$ converges to $h$ weakly, then $\|h\| \leq \lim \inf \left\|h_{n}\right\|$.

Problem 23.10. Suppose $H$ is countably infinite-dimensional (separable Hilbert space).
Prove, if $h \in H$ and $\|h\| \leq 1$, then there is a sequence $h_{n}$ in $H$ with $\left\|h_{n}\right\|=1$ for all $n$, and $\left(h_{n}\right)$ converges to $h$ weakly, but $h_{n}$ does not converge to $h$ strongly.

Problem 23.12. Prove, if $\left(a_{n}\right)$ is a sequence of complex numbers, then the following are equivalent.
(1) $\sum_{n \in \mathbb{N}} a_{n}$ converges as a net;
(2) $\sum_{n=1}^{\infty} a_{n}$ converges unconditionally;
(3) $\sum_{n=1}^{\infty} a_{n}$ converges absolutely.

Problem 23.13. Suppose $\left(h_{n}\right)$ is a sequence from a Hilbert space $H$. Show, if $\sum_{n=1}^{\infty} h_{n}$ converges absolutely, then $\sum_{n=1}^{\infty} h_{n}$ converges unconditionally and as anet.
netvuncond Problem 23.14. Suppose $H$ is a Hilbert space and $\left(h_{j}\right)$ is a sequence from $H$. Show, $\sum_{j=1}^{\infty} h_{j}$ converges unconditionally if and only if $\sum_{j \in \mathbb{N}} h_{j}$ converges as a net. (Warning: showing unconditional convergence implies convergence as a net is challenging.) Suppose $\sum h_{j}$ converges unconditionally to $h$, but does not converge to $h$ as a net. Thus, there is a $\delta>0$ such that if $F \subset \mathbb{N}$ is a finite set, then there is a finite set $G \supset F$ such that

$$
\left\|\sum_{j \in G} h_{j}-h\right\| \geq \delta
$$

On the other hand, there is an $N$ such that if $n \geq N$, then

$$
\left\|\sum_{j=1}^{n} h_{j}-h\right\|<\frac{\delta}{2}
$$

Let $I_{1}=\{1,2, \ldots, N\}$. There is a finite set $J_{1} \supset I_{1}$ such that

$$
\left\|\sum_{j \in J_{1}} h_{j}-h\right\| \geq \delta
$$

Let $I_{2}=\left\{1, \ldots, \max \left(J_{1}\right)\right\}$. There is a finite set $J_{2} \supset I_{1}$ such that

$$
\left\|\sum_{j \in J_{2}} h_{j}-h\right\| \geq \delta
$$

Continuing in this fashion gives sequence $\left(I_{k}\right)$ and $\left(J_{k}\right)$ of finite subsets of $\mathbb{N}$ such that $I_{k} \subset J_{k} \subset I_{k+1}$ and

$$
\begin{array}{r}
\left\|\sum_{j \in J_{k}} h_{j}-h\right\| \geq \delta \\
\left\|\sum_{j \in I_{k}} h_{j}-h\right\| \leq \frac{\delta}{2} .
\end{array}
$$

Choose a permutation $\pi$ of $\mathbb{N}$ compatible with the ordering $I_{1}, J_{1} \backslash I_{1}, I_{2} \backslash J_{1}, \ldots$ To complete the proof observe that $\sum_{j} h_{\pi(j)}$ does not converge to $h$ since this sum has as some of its partial sum each of $\sum_{j \in J_{k}} h_{j}$.

Definition 24.1. Let $(X, \mathscr{M}, \mu)$ be a measure space. For $0<p<\infty$, let $L^{p}(\mu)$ denote the space of measurable functions $f: X \rightarrow \mathbb{C}$ which satisfy

$$
\|f\|_{p}:=\left(\int_{X}|f|^{p} d \mu\right)^{1 / p}<\infty
$$

where $f$ and $g$ are identified when $f=g$ a.e.

The inequality

$$
|f+g|^{p} \leq(2 \max (|f|,|g|))^{p} \leq 2^{p}\left(|f|^{p}+|g|^{p}\right)
$$

and monotonicity of the integral imply that $L^{p}$ is a vector space for all $0<p<\infty$.

Proposition 24.2. If $1 \leq p<\infty$, then $\|f\|_{p}$ is a norm on $L^{p}$.

Proof. Trivially $\|f\|_{p} \geq 0$, and if $\|f\|_{p}=0$ then $f=0$ a.e., which means $f=0$ since we have identified functions that agree almost everywhere. The homogoeneity $\|c f\|_{p}=|c|\|f\|_{p}$ is evident from the definition. To prove the triangle inequality, let $f, g \in L^{p}$ and make the following reductions. First, since $|f+g|^{p} \leq(|f|+|g|)^{p}$, we can assume $f, g \geq 0$. Using homogeneity, scale $f$ and $g$ by the same constant factor so that $\|f\|_{p}+\|g\|_{p}=1$. Choose $t=\|f\|_{p}$. Thus $f=t F, g=(1-t) G$ with $F, G \in L^{p}$ and $\|F\|_{p}=\|G\|_{p}=1$. Proving the triangle inequality has now been reduced to proving

$$
\begin{equation*}
\int_{X}|t F+(1-t) G|^{p} d \mu \leq 1 \tag{36}
\end{equation*}
$$

The function $x \rightarrow|x|^{p}$ is convex on $[0,+\infty)$, that is,

$$
\begin{equation*}
\left|t x_{1}+(1-t) x_{2}\right|^{p} \leq t\left|x_{1}\right|^{p}+(1-t)\left|x_{2}\right|^{p} \tag{37}
\end{equation*}
$$

for all $x_{1}, x_{2} \geq 0$ Applying the inequality of Equation (eqn :pbigger1conveleqn:lp-convexity
for all $x_{1}, x_{2} \geq 0$. Applying the inequality of Equation (37) to Equation (36) gives

$$
\int_{X}|t F+(1-t) G|^{p} d \mu \leq \int_{X} t|F|^{p}+(1-t)|G|^{p} d \mu=1
$$

as desired.

Remark 24.3. Note that the proof of the triangle inequality breaks down when $p<1$ because the function $x \rightarrow|x|^{p}$ is not convex for these $p$. In fact, one can use the nonconvexity to show that the triangle inequality fails in this range. Indeed since $x \rightarrow|x|^{p}$ is concave, $a^{p}+b^{p}>(a+b)^{p}$ for all $a, b>0$. Thus, for disjoint sets $E, F$ of finite positive measure,

$$
\left\|\mathbf{1}_{E}+\mathbf{1}_{F}\right\|_{p}=(\mu(E)+\mu(F))^{1 / p}>\mu(E)^{1 / p}+\mu(F)^{1 / p}=\left\|\mathbf{1}_{E}\right\|_{p}+\left\|\mathbf{1}_{F}\right\|_{p}
$$

When $p=\infty$, recall that $L^{\infty}(\mu)$ is the space of all essentially bounded measurable functions, again identifying functions that agree almost everywhere. Recall too, a function is essentially bounded if there exists a number $M<\infty$ such that $|f(x)| \leq M$ for a.e. $x \in X$ (equivalently, $\mu(\{|f|>M\})=0$ ) and $\|f\|_{\infty}$ is the smallest $M$ with this property. In particular we have

$$
\|f\|_{\infty}=\inf \{M:|f(x)| \leq M \text { a.e. }\}=\sup \{M: \mu(\{x:|f(x)|>M\})>0\}
$$

It is straightforward to check that $\|f\|_{\infty}$ is a norm on $L^{\infty}(\mu)$, and that $f_{n} \rightarrow f$ in $L^{\infty}$ if and only if $f_{n}$ converges to $f$ essentially uniformly, from which it follows that $L^{\infty}$ is complete. (Problem prob:Linfty-basics

Example 24.4. If $f=A 1_{E}$ with $A>0$ and $0<\mu(E)<\infty$, then $f \in L^{p}$ for every $0<p \leq \infty$. In fact $\|f\|_{p}=A \mu(E)^{1 / p}$ for finite $p$, and $\|f\|_{\infty}=A$. Thus $\|f\|_{\infty}=\lim _{p \rightarrow \infty}\|f\|_{p}$. This result is true more generally (see Problem [prob:lim-lp-norms

The proof that $L^{p}$ is complete for $1 \leq p<\infty$ is essentially the same as the ones already given in the case of $L^{1}$ and $L^{2}$ and is left as an exercise.

Theorem 24.5. For $1 \leq p \leq \infty, L^{p}$ is a Banach space.

For $p<\infty$, a function $f$ is in $L^{p}$ if and only if $|f|^{p} \in L^{1}$. Thus, an application of Markov's inequality to $|f|^{p}$ gives the following result.
:chebyshev Proposition 24.6 (Chebyshev's inequality). Suppose $f \in L^{p}(\mu)$ and $t>0$. Then

$$
\begin{equation*}
\mu(\{x:|f|>t\}) \leq \frac{1}{t^{p}} \int_{X}|f|^{p} d \mu=\left(\frac{\|f\|_{p}}{t}\right)^{p} \tag{38}
\end{equation*}
$$

e-dense-Lp Proposition 24.7. Simple functions are dense in $L^{p}$ for all $1 \leq p \leq \infty$.

Proof. We prove the $1 \leq p<\infty$ case and leave $p=\infty$ as an exercise. Let $f \in L^{p}$. It is straightforward to check that $\operatorname{Re} f, \operatorname{Im} f$ are in $L^{p}$, as are the positive and negative parts when $f$ is real valued. Thus, because $L^{p}$ is a normed space, it may be assumed that $f$ is
unsigned. Next, by dominated convergence we see that $f$ can be approximated in $L^{p}$ by bounded functions (take $f_{n}=\min (f, n)$ and note that $0 \leq\left(f-f_{n}\right)^{p} \leq f^{p}$ and converges pointwise to 0 ), so we can assume $f$ is bounded; and again by dominated convergence we can approximate $f$ in $L^{p}$ by functions supported on sets of finite measure (take $f_{n}$ to be $1_{E_{n}} f$, where $E_{n}=\left\{|f|>\frac{1}{n}\right\}$ ). Thus we may assume $f$ is nonnegative, bounded, and supported on a set of finite measure. But in this case $f$ can be approximated essentially uniformly by simple functions, and it is easy to verify that if $f_{n} \rightarrow f$ essentially uniformly on a finite measure space, then $\left(f_{n}\right)$ converges to $f$ in $L^{p}$ also (the proof is the same as in the $L^{1}$ case).

When we write $L^{p}\left(\mathbb{R}^{n}\right)$ we always refer to Lebesgue measure. We then have the following density result for $L^{p}\left(\mathbb{R}^{n}\right)$; the proof is essentially the same as the $L^{1}$ case.

Proposition 24.8. For $1 \leq p<\infty$ the space of continuous functions of compact support $C_{c}\left(\mathbb{R}^{n}\right)$ is dense in $L^{p}\left(\mathbb{R}^{n}\right)$.

It should be clear that $C_{c}\left(\mathbb{R}^{n}\right)$ is not dense in $L^{\infty}\left(\mathbb{R}^{n}\right)$ (why?)

Proof Sketch. Consider the case $n=1$ and suppose $f \in L^{p}(\mathbb{R})$. Given $\epsilon>0$, there is an $L^{1}$ simple function $\psi$ such $\|\psi-f\|_{1}<\epsilon$ by Proposition $\frac{\text { prop: simple-dense-Lp }}{24.7 \text {. It suftices (by finearity) to }}$ assume $\psi=\mathbf{1}_{E}$ for a set $E$ with $m(E)<\infty$. By Littlewood's first principle, we can find a set $A$, a finite union of disjoint open intervals $A=\bigcup_{j=1}^{n}\left(a_{j}, b_{j}\right)$, such that $m(A \Delta E)<\epsilon^{p}$. It follows that $\left\|\mathbf{1}_{A}-\mathbf{1}_{E}\right\|_{p}=\left\|\mathbf{1}_{A \Delta E}\right\|_{p}<\epsilon$. Let $g_{j}$ denote the a continuous, piecewise
linear function such that $\left|g_{j}(x)\right| \leq 1$ for all $x$ and is equal to 1 for $x \in\left(a_{j}, b_{j}\right)$ and 0 for $x<a_{j}-\frac{1}{2} \frac{\epsilon^{p}}{2^{j p}}$ and $x>b_{j}+\frac{1}{2} \frac{\epsilon^{p}}{2^{j p}}$. (Draw a picture.) Then for each interval $I_{j}=\left(a_{j}, b_{j}\right)$, we have $\left\|g_{j}-\mathbf{1}_{I_{j}}\right\|_{p}<\epsilon 2^{-j}$. Let $g=\sum_{j=1}^{n} g_{j}$. Then $g$ is continuous and belongs to $L^{1}$, and from the triangle inequality we have $\left\|g-\mathbf{1}_{A}\right\|_{p} \leq \sum_{j=1}^{n}\left\|g_{j}-\mathbf{1}_{I_{j}}\right\|_{p}<\epsilon$. It follows that $\left\|g-\mathbf{1}_{E}\right\|_{p}<2 \epsilon$ and hence

$$
\|f-g\|_{p} \leq\left\|f-\mathbf{1}_{E}\right\|_{p}+\left\|\mathbf{1}_{E}-\mathbf{1}_{A}\right\|_{p}+\left\|\mathbf{1}_{A}-g\right\|_{p}<3 \epsilon
$$

and the proof is complete in the case $n=1$.

In higher dimensions, the same approximation scheme works; it suffices (using linearity, Littlewood's first principle, and the $\epsilon / 2^{n}$ trick as before) to approximate the indicator function of a box $B=I_{1} \times \cdots \times I_{n}$; again a piecewise linear function which is 1 on the box and 0 outside a suitably small neighborhood of the box suffices. The details are left as an exercise.
24.1. Duality in $L^{p}$ spaces. In this section it is shown that, for $1 \leq p<\infty$, the dual of the Banach space $L^{p}$ is $L^{q}$, where $\frac{1}{p}+\frac{1}{q}=1$ (interpret $1 / \infty=0$ ). The starting point is Hölder's inequality.
thm:holder Theorem 24.9 (Hölder's inequality). Suppose $1 \leq p, q \leq \infty$ and $\frac{1}{p}+\frac{1}{q}=1$ (interpret $1 / \infty=0)$. If $f \in L^{p}$ and $g \in L^{q}$, then $f g \in L^{1}$, and

$$
\|f g\|_{1} \leq\|f\|_{p}\|g\|_{q}
$$

Further, for $1<p, q<\infty$, equality holds if and only if $\|f\|_{p}^{p}|g|^{q}=\|g\|_{q}^{q}|f|^{p}$; and for $p=1$ and $q=\infty$, equality holds for $g=1$.

For $1<p<\infty$ and $f \in L^{p}$, let $g=|f|^{p} f^{-1}$. A computation shows $g \in L^{q}$ and $\|g\|_{q}=\|f\|_{p}^{p-1}$. Thus,

$$
\int f g d u=\|f g\|_{1}=\|f\|_{p}^{p}=\|f\|_{p}\|f\|_{p}^{p-1}=\|f\|_{p}\|g\|_{q} .
$$

Thus for this choice of $g$, equality holds in Hölder's inequality. On the other hand, consider the function $g:[0,1] \rightarrow[0,1]$ defined by

$$
g(x)=\sum_{j=1}^{\infty} \frac{1}{2^{j}} \mathbf{1}_{\left(\frac{1}{2^{j}}, \frac{1}{2^{j-1}}\right.} .
$$

Thus $g \in L^{\infty}([0,1])$ and $\|g\|_{\infty}=1$. On the other hand, if $f \in L^{1}([0,1])$ and $\|f\|_{1}=1$, then

$$
\left|\int f g d \mu\right|<1
$$

and hence, for this $g \in L^{\infty}([0,1])$, equality never holds.

The proof of Hölder's inequality uses Young's inequality.
lem:young Lemma 24.10 (Young's inequality). If $a, b$ are nonnegative numbers and $1<p, q<\infty$ satisfy $\frac{1}{p}+\frac{1}{q}=1$, then

$$
a b \leq \frac{a^{p}}{p}+\frac{b^{q}}{q}
$$

and equality holds if and only if $b^{q}=a^{p}$.

Proof. If $a$ or $b$ is 0 the result is evident. So suppose $a, b>0$. Define $\psi:\left[-\frac{1}{p}, \infty\right) \rightarrow \mathbb{R}$ by $\psi(t)=(1+t)^{p}-(1+p t)$. A calculus exercise shows $\psi(0)=0$ and $\psi(t)>0$ for $t \neq 0$.

Let $u=a^{p}$ and $v=b^{q}$ and choose $t$ so that $1+p t=\frac{u}{v}>0$. Now,

$$
\begin{aligned}
a b & =v(1+p t)^{\frac{1}{p}} \\
& \leq v(1+t) \\
& =v\left[1+\frac{\frac{u}{v}-1}{p}\right] \\
& =\frac{u}{p}+v\left(1-\frac{1}{p}\right) \\
& =\frac{a^{p}}{p}+\frac{b^{q}}{q} .
\end{aligned}
$$

Note that equality holds if and only if $t=0$; that is $a^{p}=b^{q}$
 case the other exponent is 1 ), so suppose $1<p, q<\infty$ and $f, g$ are both nonzero. By homogeneity we may normalize so that $\|f\|_{p}=\|g\|_{q}=1$. We must now show that

$$
\begin{equation*}
\int_{X}|f g| d \mu \leq 1 \tag{39}
\end{equation*}
$$

lem: young
Applying Lemma 2 [em. 240 give

$$
\begin{equation*}
|f(x) g(x)| \leq \frac{1}{p}|f(x)|^{p}+\frac{1}{q}|g(x)|^{q} \tag{40}
\end{equation*}
$$

Integrating ( $\left.(40)^{4}\right)^{4}$ with respect to $\mu$ and applying the normalizations on $p, q, f, g$ gives (egn:
Finally, observe, in the case $1<p, q<\infty$, that equality holds in Hölders in inequality if and only if equality holds a.e. $\mu$ in equation (eg; 4 if and only if $|f|^{p}=|g|^{q}$ a.e. $\mu$ by Lemma $\frac{1 \text { nem: young }}{24.10 .}$

Example 24.11. One can get a more intuitive feel for what Hölder's inequality says by examining it in the case of step functions. Let $E, F$ be sets of finite, positive measure
and put $f=\mathbf{1}_{E}, g=\mathbf{1}_{F}$. Then $\|f g\|_{1}=\mu(E \cap F)$ and

$$
\|f\|_{p}\|q\|_{q}=\mu(E)^{1 / p} \mu(F)^{1 / q}
$$

so Hölder's inequality can be proved easily in this case using the relation $\frac{1}{p}+\frac{1}{q}=1$ and the fact that $\mu(E \cap F) \leq \min (\mu(E), \mu(F))$.

Remark 24.12. A more general version of Hölder's inequality says the following: let $1 \leq p, q \leq \infty$ and $\frac{1}{r}=\frac{1}{p}+\frac{1}{q}$. If $f \in L^{p}$ and $g \in L^{q}$, then $f g \in L^{r}$, and

$$
\begin{equation*}
\|f g\|_{r} \leq\|f\|_{p}\|g\|_{q} \tag{41}
\end{equation*}
$$

prob:generalized-holder
See Problem $\frac{\text { prob: }}{24.1 .}$

Numbers $1 \leq p, q \leq \infty$ such that $\frac{1}{p}+\frac{1}{q}=1$ are conjugate exponents. Given $p$, its conjugate index $q$ is of course uniquely determined, $q=\frac{p}{p-1}$. There are numerous relations between $p$ and $q$. For instance $p(q-1)=q$ and $(q-1)(p-1)=1$. Hölder's inequality implies that each $g \in L^{q}$ determines a bounded linear functional $L_{g}: L^{p}(\mu) \rightarrow$ $\mathbb{C}$ by

$$
\lambda_{g}(f)=\int_{X} f g d \mu
$$

and moreover $\left\|\lambda_{g}\right\| \leq\|g\|_{q}$. One of the most important facts about $L^{p}$ spaces is that the converse is true for a $\sigma$-finite measure $\mu$ and $1 \leq p<\infty$.
hm:Lp-dual Theorem 24.13. Suppose $1 \leq p<\infty, \frac{1}{p}+\frac{1}{q}=1$ and $\mu$ is a $\sigma$-finite measure. If $\lambda: L^{p}(\mu) \rightarrow \mathbb{C}$ is a bounded linear functional, then there exists a unique $g \in L^{q}(\mu)$ such that $\lambda=\lambda_{g}$. Moreover $\|\lambda\|=\|g\|_{q}$.

Thus $\Lambda: L^{q}(\mu) \rightarrow L^{p}(\mu)^{*}$ defined by $\Lambda(g)=\lambda_{g}$ is an isometric isomomorphism. It is in this sense that we write $L^{p}(\mu)^{*}=L^{q}(\mu)\left(\right.$ for $1 \leq p<\infty$ and $\left.\frac{1}{p}+\frac{1}{q}=1\right)$.

The following two lemmas will be used to pass from the case of a finite measure to that of a $\sigma$-finite measure in the proof of Theorem thm:Lp-dual that of a $\sigma$-finite measure in the proof of Theorem 24.13.
lem:wsig1 Lemma 24.14. If $\mu$ is a $\sigma$-finite measure on a measurable space $(X, \mathscr{M})$, then there exists a measurable function $w \in L^{1}(\mu)$ such that $0<w(x)<1$ for all $x$.

Proof. Write $X=\cup_{n=1}^{\infty} X_{n}$, a countable union of disjoint measurable sets of finite measure. Let $w_{n}=\frac{1}{2^{n}\left(\mu\left(X_{n}\right)+1\right)} \mathbf{1}_{X_{n}}$ and $w=\sum_{n=1}^{\infty} w_{n}$.
lem:wsig2 Lemma 24.15. Suppose $\mu$ is a $\sigma$-finite measure, $w \in L^{1}(\mu)$ and $0<w(x)<1$ for all $x$. Let $\tau$ denote the measure $w d \mu$.

For $1<p<\infty$, a measurable function $f$ is in $L^{p}(\mu)$ if and only if $g=w^{\frac{1}{p}} f \in L^{p}(\tau)$ and in this case $\|f\|_{p}=\|g\|_{p}$; that is, the mapping $\Phi_{p}: L^{p}(\tau) \rightarrow L^{p}(\mu)$ defined by

$$
\Phi_{p}(f)=w^{\frac{1}{p}} f
$$

is a (linear) isometric isomorphism.

Proof. It is easy to check that $\Phi_{p}$ is isometric with inverse $\Psi$ defined by $\Psi f=w^{-\frac{1}{p}} f$.
 then, for any measurable set $E$ of finite measure, $\lambda_{g-h}\left(\mathbf{1}_{E}\right)=0$ and it thus follows from the $\sigma$-finiteness assumption, that $g=h \mu$-a.e.

The idea of the proof is to use the linear functional $\lambda$ to define a set function $\mathscr{M} \ni E \rightarrow \lambda\left(\mathbf{1}_{E}\right)$, prove this set function is a measure absolutely continuous to $\mu$ and choose $g$ via the Radon-Nikodym theorem. Finally it must be shown that this $g$ belongs to $L^{q}$. The details follow.

Suppose for now that $\mu$ is finite; i.e., $(X, \mathscr{M}, \mu)$ is a finite measure space. In this case, if $E \in \mathscr{M}$, then $\mathbf{1}_{E} \in L^{p}(\mu)$. Let $u=\operatorname{Re}(\lambda)$. Since $\lambda$ is bounded, $u$ is continuous. Define $\nu: \mathscr{M} \rightarrow \mathbb{R}$ by

$$
\nu(E):=u\left(\mathbf{1}_{E}\right)
$$

To prove that $\nu$ is a countably additive, suppose $\left(E_{n}\right)_{n=1}^{\infty}$ is a sequence of disjoint measurable sets and let $E=\cup_{j=1}^{\infty} E_{j}$. Let $s_{n}=\sum_{j=1}^{n} \mathbf{1}_{E_{j}}$. In particular, $\left(s_{n}\right)$ increases pointwise with limit $s=\mathbf{1}_{E}$. Further, $0 \leq\left(s-s_{n}\right)^{p} \leq s$ and thus, by dominated convergence, $\left(s_{n}\right)$ converges to $s$ in $L^{p}(\mu)$. Using continuity and (real) linearity of $u$,

$$
\nu(E)=u(s)=\lim _{n \rightarrow \infty} u\left(s_{n}\right)=\lim _{n \rightarrow \infty} \sum_{j=1}^{n} \nu\left(E_{j}\right)=\sum_{j=1}^{\infty} \nu\left(E_{j}\right) .
$$

Evidently $\nu(\emptyset)=0$ and hence $\nu$ defines a finite signed measure on $(X, \mathscr{M})$.

If $\mu(E)=0$, then $\mathbf{1}_{E}=0$ in $L^{p}$ so $\nu(E)=0$. Thus, the measure $\nu$ is absolutely continuous with respect to $\mu$. Let $g_{1}$ denote the Radon-Nikodym derivative $d \nu / d \mu$. Since $\nu$ is finite, the Radon-Nikodym theorem gives $g_{1} \in L^{1}(\mu)$. Applying the same arguments to $\operatorname{Im}(\lambda)$, we define $g_{2}$ similarly, and put $g=g_{1}+i g_{2}$. In particular, for $E \in \mathscr{M}$,

$$
\begin{equation*}
\lambda\left(\mathbf{1}_{E}\right)=\int_{E} g d \mu . \tag{42}
\end{equation*}
$$

eq: GE

Temporarily, view $\lambda_{g}$ as defined (and continuous) on $L^{\infty}(\mu)$.

Now suppose $f$ is a bounded unsigned measurable function. Since $\mu$ is a finite measure, $f$ is in $L^{p}(\mu)$ as well as $L^{\infty}(\mu)$. Hence, both $\lambda_{g}(f)$ and $\lambda(f)$ are defined. If $s$ is a measurable simple function, then by equation (eq:GE

$$
\lambda(s)=\int_{X} s g d \mu
$$

There exists a sequence $\left(s_{n}\right)$ of measurable simple functions $0 \leq s_{n} \leq f$ such that $\left(s_{n}\right)$ converges to $f$ uniformly and therefore in both $L^{p}(\mu)$ and $L^{\infty}(\mu)$. It follows

$$
\lambda(f)=\lim \lambda\left(s_{n}\right)=\lambda_{g}\left(s_{n}\right)=\lambda_{g}(f)
$$

It now follows that if $f$ is bounded and measurable, then $\lambda(f)=\lambda_{g}(f)$.

To prove $g \in L^{q}$, first assume $p>1$. For positive integers $N$, let $E_{N}=\{|g| \leq N\}$ and let $g_{N}=g \mathbf{1}_{E_{n}}$. Thus $g_{N}$ is bounded and so is $f_{N}={\overline{g_{N}}}^{\frac{q}{2}} g_{N}^{\frac{q}{2}-1}$. By $\left(L^{p}(\mu)\right)$ continuity of $\lambda$,

$$
\begin{equation*}
\left\|f_{N}\right\|_{p}\|\lambda\| \geq\left|\lambda\left(f_{N}\right)\right|=\left|\lambda_{g}\left(f_{N}\right)\right|=\int_{X} f_{N} g d \mu=\int_{X}\left|g_{N}\right|^{q} d \mu=\left\|g_{N}\right\|_{q}^{q} \tag{43}
\end{equation*}
$$

On the other hand, $\left\|f_{N}\right\|_{p}=\left\|g_{N}\right\|_{q}^{q-1}$, and combining this equality with (egn: dual-lpthat $\left\|g_{N}\right\|_{q} \leq\|\lambda\|$. By monotone congvergence, $\|g\|_{q} \leq\|\lambda\|<\infty$. Now that we know $g \in L^{q}$, it follows that $\lambda_{g}$ is continuous. It also agrees with $\lambda$ on unsigned bounded functions and therefore on simple functions. Since simple functions are dense in $L^{q}(\mu)$, the conclusion $\lambda=\lambda_{g}$ follows. Further, $\left\|\lambda_{g}\right\|=\|\lambda\| \geq\|g\|_{q}$. and since Hölder gives the reverse inequality, $\left\|\lambda_{g}\right\|=\|g\|_{q}$.

In the $p=1$ case, put $E_{t}=\{|g|>t\}$ and let $f_{t}=\frac{\bar{g}}{g \mu\left(E_{t}\right)} \mathbf{1}_{E_{t}}$. Thus $\left\|f_{t}\right\|_{1}=\mu\left(E_{t}\right)$ for all $t$, and, since $f \in L^{\infty}(\mu)$,

$$
\begin{equation*}
\mu\left(E_{t}\right)\|\lambda\|=\|\lambda\|\left\|f_{t}\right\|_{1} \geq\left|\lambda\left(f_{t}\right)\right|=\left|\lambda_{g}\left(f_{t}\right)\right|=\int_{X} f_{t} g d \mu=\int_{E_{t}} g d \mu \geq t \mu\left(E_{t}\right) \tag{44}
\end{equation*}
$$

Hence $\mu\left(E_{t}\right)=0$ for $t>\|\lambda\|$ and thus $g \in L^{\infty}(\mu)$ and in fact $\|\lambda\| \geq\|g\|_{\infty}$.

For the $\sigma$-finite case with $p>1$, suppose $X=\cup_{n=1}^{\infty} X_{n}$ where the $X_{n}$ are measureable and of finite measure. For $1<p<\infty$, let $w$ be as in Lemma 24.14. Likewise, let
 are linear isometric isomorphisms for $1<r<\infty$. Thus, $\psi=\lambda \Phi_{p}$ is a bounded linear functional on $L^{p}(\tau)$. Since $\tau$ is a finite measure, by what is already proved, there is a $G \in L^{q}(\tau)$ such that $\psi=\lambda_{G}$. Let $g=\Phi_{q} G=w^{\frac{1}{q}} G$. Thus $g \in L^{q}(\mu)$ and $\|g\|_{q}=\|G\|_{q}$. Moreover, if $f \in L^{p}(\mu)$, then $h:=\Phi_{p}^{-1} f=w^{-\frac{1}{p}} f \in L^{p}(\tau)$ and

$$
\begin{aligned}
\lambda(f) & =\psi(h)=\lambda_{G}(h) \\
& =\int h G d \tau=\int h G w d \mu \\
& =\int w^{\frac{1}{p}} h\left(w^{\frac{1}{q}} G\right) d \mu=\int f g d \mu=\lambda_{g}(f) .
\end{aligned}
$$

Extending from finite to $\sigma$-finite in the case $p=1$ is left to the gentle reader.
rem:testLp Remark 24.16. In the case $(X, \mathscr{M}, \mu)$ is $\sigma$-finite and $1<p<\infty$, Theorem $\frac{\text { thm:Lp-dual }}{24.13 \text { can }}$ be used to test whether a given measurable function $f: X \rightarrow \mathbb{C}$ is in $L^{p}$. Namely, letting $q$ denote the conjugate exponent to $p$, if for each $g \in L^{q}(\mu)$ the function $f g$ is integrable and the linear functional $\lambda: L^{q}(\mu) \rightarrow \mathbb{C}$ defined by $\lambda(g)=\int f g d \mu$ is bounded,
then there is an $h \in L^{p}(\mu)$ such that $\lambda=\lambda_{h}$. It follows that $\int_{E} f d \mu=\int_{E} h d \mu$ for all measurable sets $E$ of finite measure. Therefore $f=h$ a.e. $\mu$ and hence $f \in L^{p}(\mu)$. $\diamond$

Recall, a Banach space $X$ is reflexive if the canonical isometric inclusion of $X$ into $X^{* *}$ is onto. Recall this inclusion is given by mapping $x \in X$ to the linear functional $\hat{x}$ defined by $\hat{x}(\lambda)=\lambda(x)$ for $\lambda \in X^{*}$.
-reflexive Corollary 24.17. For $1<p<\infty, L^{p}$ is reflexive.

Proof. Let $\Phi: L^{p} \rightarrow\left(L^{p}\right)^{* *}$ denote the canonical inclusion. Thus, for $g \in L^{p}$, the functional $\Phi(g)$ is given by $\Phi(g)(\lambda)=\lambda(g)$ for $\lambda \in\left(L^{p}\right)^{*}$. It suffices to show $\Phi$ is onto. Accordingly, let $\sigma \in L^{p}(\mu)^{* *}$ be given. Let $\Lambda: L^{q} \rightarrow\left(L^{p}\right)^{*}$ denote the isometric
 continuous linear functional and hence, by Theorem $\frac{\text { thm: } \frac{\text { Lp }- \text { dual }}{24.13} \text {, there is a } g \in L^{p}(\mu) \text { such }}{}$ that $\lambda_{g}=\sigma \circ \Lambda$. Now let $\lambda \in\left(L^{p}\right)^{*}$ be given. There is an $h \in L^{q}(\mu)$ such that $\lambda=\Lambda(h)$. Thus, for $\psi \in L^{p}$,

$$
\lambda(\psi)=\int \psi h d \mu
$$

It follows that

$$
\begin{aligned}
\sigma(\lambda) & =\sigma(\Lambda(h))=\lambda_{g}(h) \\
& =\int g h d \mu=\lambda(g)=\Phi(g)(\lambda)
\end{aligned}
$$

Theorem ${ }^{\text {thm }} 24.13$ Lp-dual above fails in general when $p=1$. The canonical inclusion of $L^{1}$ into $\left(L^{\infty}\right)^{*}$ is defined by sending $g \in L^{1}$ to the bounded linear functional on $L^{\infty}$ defined by
the formula $\lambda_{g}(f)=\int_{X} f g d \mu$. Unless $\mu$ is a finite sum of atoms, there exist bounded linear functionals on $L^{\infty}$ that are not of this form. An abstract way to see this is that in when $\mu$ is $\sigma$-finite, $L^{1}(\mu)$ is separable, but if $L^{\infty}(\mu)$ is separable, then $(X, \mathscr{M}, \mu)$ is finitely atomic; i.e., $X=\cup_{j=1}^{N} A_{j}$ where $\mu\left(A_{j}\right)>0$ and if $B \subset A_{j}$ and $\mu(B)<\mu\left(A_{j}\right)$, then $\mu(B)=0$. Thus if it were the case that $\left(L^{\infty}\right)^{*} \cong L^{1}$, then $L^{\infty}$ would be separable by the result in Problem $\frac{\text { prob:separable-dual }}{21.10, ~ a ~ c o n t r a d i c t i o n . ~ P r o b l e m ~} \frac{\text { prob:dual-of-linfty }}{24.10}$ gives a somewhat more explicit argument in the case of $L^{\infty}(\mathbb{R})$. Note that in the example following the statement of Hölder's inequality, there does not exists and $f \in L^{1}$ such that $\lambda_{f}(g)=\|g\|_{\infty}$ showing that, in view of the Hahn-Banach Theorem, the dual of $L^{\infty}([0,1])$ properly contains $L^{1}([0,1])$.
24.2. Distribution functions and weak $L^{p}$. Let $(X, \mathscr{M}, \mu)$ be a measure space and $f: X \rightarrow \mathbb{C}$ a measurable function. The distribution function of $f$ is the function $\lambda_{f}:(0,+\infty) \rightarrow[0,+\infty]$ defined by

$$
\lambda_{f}(t)=\mu(\{x:|f(x)|>t\})
$$

To begin building an intuition about $\lambda_{f}$, we have the following lemma.
of-simples Lemma 24.18. Let $f=\sum_{j=1}^{n} c_{j} \mathbf{1}_{E_{j}}$ be a simple function with the $E_{j}$ disjoint measurable sets, and the $c_{j}$ ordered as $0<c_{1}<c_{2}<\cdots<c_{n}<+\infty$. Then

$$
\lambda_{f}(t)= \begin{cases}\mu\left(E_{n}\right)+\cdots+\mu\left(E_{2}\right)+\mu\left(E_{1}\right), & 0 \leq t<c_{1} \\ \mu\left(E_{n}\right)+\cdots+\mu\left(E_{2}\right), & c_{1} \leq t<c_{2} \\ \cdots & , \cdots \\ \mu\left(E_{n}\right), & c_{n-1} \leq t<c_{n} \\ 0 & t \geq c_{n}\end{cases}
$$

The basic properties of $\lambda_{f}$ are collected in the following proposition:
tion-basic Proposition 24.19. a) $\lambda_{f}$ is decreasing and right continuous.
b) [Monotonicity] If $|f| \leq|g|$ a.e., then $\lambda_{f} \leq \lambda_{g}$ everywhere.
c) [Monotone convergence] If $\left|f_{n}\right|$ increases to $|f|$ pointwise a.e., then $\lambda_{f_{n}}$ increases to $\lambda_{f}$.
d) [Subadditivity] If $f=g+h$, then $\lambda_{f}(t) \leq \lambda_{g}(t / 2)+\lambda_{h}(t / 2)$.

Proof. Let $E(t, f)=\{x:|f(x)|>t\}$. Thus $\lambda_{f}(t)=\mu(E(t, f))$. In particular, $\lambda_{f}$ is decreasing since $E(t, f) \subset E(s, f)$ when $s<t$. Since $E(t, f)$ is the increasing union of $E\left(t+\frac{1}{n}, f\right)$, monotone convergence for sets gives

$$
\lambda_{f}(E(t, f))=\mu\left(\cup_{n=1}^{\infty} E\left(t+\frac{1}{n}, f\right)\right)=\lim _{n \rightarrow \infty} E\left(t+\frac{1}{n}, f\right)=\lim _{n \rightarrow \infty} \lambda_{f}\left(t+\frac{1}{n}\right)
$$

Now, if $\left(t_{m}\right)$ is any sequence converging to $t$ for which $t \leq t_{m}$ for all $m$, then, given $\epsilon>0$ choose $N$ so that $\lambda_{f}(t)-\lambda_{f}\left(t_{N}\right)<\epsilon$. Choose $M$ such that $t+\frac{1}{M}>t_{m} \geq t$ for $m \geq M$ and note that, since $\lambda_{f}$ is decreasing,

$$
0 \leq \lambda_{f}(t)-\lambda_{f}\left(t_{m}\right) \leq \lambda_{f}(t)-\lambda_{f}\left(t_{N}\right)<\epsilon
$$

to complete the proof of item (a).

Item (b) is immediate from the definition. For (c) we have that for each fixed $t>0, E(t, f)$ is the increasing union of the $E\left(t, f_{n}\right)$. Now argue as in the proof of part (a). Finally (d) is a pigeonhole argument: if $|f(x)|>t$, then either $|g(x)|>t / 2$ or
$|h(x)|>t / 2$. Hence,

$$
E(t, f) \subset E\left(\frac{t}{2}, g\right)+E\left(\frac{t}{2}, h\right)
$$

The main use of the distribution function is to convert integrals of functions of $f$ into integrals against the measure induced by $\lambda_{f}$. Indeed, the function $\lambda_{f}(t)$ is decreasing and right continuous on $[0,+\infty]$ and hence defines a (negative) Borel measure $\nu$ on $[0,+\infty]$ by

$$
\left.\nu((a, b]):=\lambda_{f}(a)-\lambda_{f}(b)\right)
$$

and passing to the Caratheodory extension. Thus if $\varphi:[0,+\infty] \rightarrow \mathbb{C}$ is a Borel function we can consider integrals $\int \varphi d \nu=\int \varphi d \lambda_{f}$. The following formula relates these integrals to $f$ and $\mu$.

Proposition 24.20. Suppose $\lambda_{f}(t)<\infty$ for all $t>0$ and $\varphi \geq 0$ is an unsigned Borel function. Then

$$
\begin{equation*}
\int_{X} \varphi(|f|) d \mu=-\int_{0}^{\infty} \varphi(t) d \lambda_{f}(t) . \tag{45}
\end{equation*}
$$

Proof. Let $0 \leq a<b$ and $\varphi=\mathbf{1}_{(a, b]}$. Then $\varphi(|f|)=\mathbf{1}_{a<|f| \leq b}$, so

$$
\int_{X} \varphi(|f|) d \mu=\mu(a<|f| \leq b)=\lambda_{f}(b)-\lambda_{f}(a)
$$

on the other hand

$$
-\int_{0}^{\infty} \varphi(t) d \lambda_{f}(t)=-\nu((a, b])=-\left(\lambda_{f}(a)-\lambda_{f}(b)\right)
$$

so (哭: dist-integration-1emma holds when $\varphi=1_{(a, b]}$. Since both sides are linear in $\varphi$, it also holds for simple functions, and then for all unsigned Borel functions by monotone convergence.

The most important case of the above is $\varphi(t)=t^{p}$, since it will allow us to obtain very useful expressions for the $L^{p}$ integrals $\int_{X}|f|^{p} d \mu$. In fact what is most useful is not (egn: dist-integration-lemma
(45) itself but its "integrated-by-parts" form:
p-integral Proposition 24.21. If $0<p<\infty$ then

$$
\int_{X}|f|^{p} d \mu=p \int_{0}^{\infty} t^{p-1} \lambda_{f}(t) d t .
$$

Proof. This can be proved using the previous proposition and integration by parts for Lebesgue-Stieltjes measures, or directly as follows. If $\lambda_{f}(t)=+\infty$ for some $t$ then both integrals are infinite. Otherwise, first let $f$ be a simple function; then the identity can be
 $0 \leq f_{n}$ increasing to $|f|$; then the formula holds by Lemma $\frac{\text { lem:dist-of-simpleqprop:dist-function-basic }}{24.18, \text { Proposition } 24.19(c),}$ and monotone convergence.

The distribution function is used to define the so-called "weak $L^{p "}$ spaces, as follows:
first observe that if $f \in L^{p}$, then from Chebyshev's inequality we have

$$
\mu(\{|f|>t\}) \leq \frac{1}{t^{p}} \int_{X}|f|^{p} d \mu
$$

or rearranging

$$
t^{p} \lambda_{f}(t) \leq\|f\|_{p}^{p} \quad \text { for all } t>0
$$

For general $f$, say that $f$ belongs to weak $L^{p}$ if

$$
\left(\sup _{t>0} t^{p} \lambda_{f}(t)\right)^{1 / p}:=[f]_{p}<\infty
$$

From what was just said, if $f \in L^{p}$, then $f$ belongs to weak $L^{p}$, but the converse does not hold. The standard example is $f(x)=x^{-1 / p}$ on $(0, \infty)$. On the other hand, weak $L^{p}$ functions are "almost" in $L^{p}$, in the sense that if we use Proposition $\frac{\text { prop:dist-Lp-integral }}{24.21}$ we find

$$
\int_{X}|f|^{p} d \mu=p \int_{0}^{\infty} t^{p-1} \lambda_{f}(t) d t \leq[f]_{p} \int_{0}^{\infty} t^{-1} d t
$$

and the integral is just barely divergent.
24.3. The Hardy-Littlewood maximal function redux. As an illustration of the usefulness of the distribution function (and the associated idea of splitting $L^{p}$ functions into their "small" and "large" parts, we reconsider the Hardy-Littlewood maximal function. (This and the next subsection follow sections I. 1 and I. 4 of Singular Integrals and Differentiability Properties of Functions by Eli Stein.) Let $f$ be a locally integrable function on $\mathbb{R}^{n}$, then

$$
(M f)(x):=\sup _{r>0} \frac{1}{m(B(r, x))} \int_{B(r, x)}|f(y)| d y
$$

In the language of distribution functions, the Hardy-Littlewood maximal theorem (Theorem 16.4) says that if $f \in L^{1}\left(\mathbb{R}^{n}\right)$, then $M f$ belongs to weak $L^{1}$, with $[M f]_{1} \leq 3^{n}\|f\|_{1}$. We now investigate what happens for $f \in L^{p}, 1<p \leq \infty$, and find the situation is rather better. First, for $p=\infty$ it is trivial that $\|M f\|_{\infty} \leq\|f\|_{\infty}$. For finite $p$ we have:

Theorem 24.22. If $f \in L^{p}\left(\mathbb{R}^{n}\right), 1<p<\infty$, then $M f \in L^{p}$ and

$$
\|M f\|_{p} \leq 2\left(\frac{3^{n} p}{p-1}\right)^{1 / p}\|f\|_{p}
$$

Proof. Let $f \in L^{p}\left(\mathbb{R}^{n}\right)$. We fix a parameter $t>0$ and use it to cut off $f$ : let

$$
f_{1}(x):= \begin{cases}f(x) & \text { if } f(x) \geq t / 2 \\ 0 & \text { otherwise }\end{cases}
$$

Note that by vertical truncation, $f_{1} \in L^{1}$, and we have $|f(x)| \leq\left|f_{1}(x)\right|+t / 2$ and $(M f)(x) \leq\left(M f_{1}\right)(x)+t / 2$. It follows that

$$
\{x: M f(x)>t\} \subset\left\{x: M f_{1}(x)>t / 2\right\}
$$

so by the Hardy-Littlewood maximal theorem applied to the $L^{1}$ function $f_{1}$, writing $E_{t}=\{x: M f(x)>t\}$ we have

$$
\begin{equation*}
m\left(E_{t}\right) \leq \frac{2 \cdot 3^{n}}{t}\left\|f_{1}\right\|_{1}=\frac{2 \cdot 3^{n}}{t} \int_{|f|>t / 2}|f(y)| d y \tag{46}
\end{equation*}
$$

eqn:HL-p-step

Now write $g=M f$, then $m\left(E_{t}\right)$ is just the distribution function $\lambda_{g}(t)$. By Proposi-prop:dist-Lp-integral tion $\frac{\text { prop:dist-Lp-in }}{24.21 \text { we have }}$

$$
\int_{\mathbb{R}^{n}}|M f(x)|^{p} d x=p \int_{0}^{\infty} t^{p-1} \lambda_{g}(t) d t
$$

Using (egn:HL-p-step

$$
\|M f\|_{p}^{p}=p \int_{0}^{\infty} t^{p-1} m\left(E_{t}\right) d t \leq p \int_{0}^{\infty} t^{p-1}\left(\frac{2 \cdot 3^{n}}{t} \int_{|f|>t / 2}|f(y)| d y\right) d t
$$

We apply Fubini to the double integral and inegrate $d t$ first. This gives

$$
p \int_{0}^{\infty} t^{p-1}\left(2 \cdot 3^{n} \int_{|f|>t / 2}|f(y)| d y\right) d t=\int_{\mathbb{R}^{n}}|f(y)|\left(\int_{0}^{2|f(y)|} t^{p-2} d t\right) d y
$$

The inner integral is

$$
\int_{0}^{2|f(y)|} t^{p-2} d t=\frac{1}{p-1} 2^{p-1}|f(y)|^{p-1}
$$

so we have finally

$$
\|M f\|_{p}^{p} \leq \frac{2^{p} \cdot 3^{n} p}{p-1} \int_{\mathbb{R}^{n}}|f(y)|^{p} d y
$$

which finishes the proof.
24.4. The Marcinkiewicz interpolation theorem. The idea used in the proof of the $L^{p}$ boundedness of the Hardy-Littlewood maximal operator can be extended to prove a more general result, called the Marcinkiewicz interpolation theorem. We will not consider the most general version of the theorem here (it may be found in Folland) but prove a special case that is adequate for many purposes. We fix a measureable space $(X, \mathscr{M}, \mu)$; $L^{p}$ always refers to $L^{p}(\mu)$.

We need a few definitions. Let $1 \leq p, q \leq \infty$. A mapping $T: L^{p} \rightarrow L^{q}$ is of type $(p, q)$ if there is a constant $A>0$ so that

$$
\|T f\|_{q} \leq A\|f\|_{p}
$$

for all $f \in L^{p}$. The mapping is of weak type $(p, q)$ if

$$
\mu\left(\{x:|T f(x)|>t\} \leq\left(\frac{A\|f\|_{p}}{t}\right)^{q}, \quad \text { for } q<\infty\right.
$$

where the constant $A$ does not depend on $f$ or $t$. (In other words, $T$ maps $L^{p}$ into weak $L^{q}$, with $\left.[T f]_{q} \leq A\|f\|_{p}.\right)$

When $q=\infty$ weak type $(p, q)$ simply means type $(p, q)$.

We say that a transformation $T$ (defined on some space of measurable functions)
is sub-linear if $|T f(x)| \leq|T g(x)|+|T h(x)|$ for $f=g+h$, and $|T(c f)|=|c||T f|$ for all scalars $c$.

Finally we let $L^{p}+L^{q}$ denote the vector space of functions of the form $f=g+h$ where $g \in L^{p}, h \in L^{q}$. Note that if $p<r<q$, then by the truncation lemmas we see that $L^{r} \subset L^{p}+L^{q}$.

Theorem 24.23 (Marcinkiewicz interpolation theorem (special case)). Suppose that $1<r \leq \infty$. If $T$ is a sub-linear transformation from $L^{1}+L^{r}$ to the vector space of measurable functions, and $T$ is of weak type $(1,1)$ and weak type $(r, r)$, then $T$ is of type $(p, p)$ for all $1<p<r$.

Proof. The case $r=\infty$ closely parallels the proof given for the Hardy-Littlewood maximal function and is left as an exercise, so we assume $1<r<\infty$. Fix $f \in L^{p}$, we wish to estimate its distribution function $\lambda_{f}(t)$. We fix $t>0$ for the moment and use this to cut off $f$ : define

$$
\begin{aligned}
& f_{1}(x):= \begin{cases}f(x) & \text { if }|f(x)|>t \\
0 & \text { if }|f(x)| \leq t\end{cases} \\
& f_{2}(x):= \begin{cases}f(x) & \text { if }|f(x)| \leq t \\
0 & \text { if }|f(x)|>t\end{cases}
\end{aligned}
$$

so that $f=f_{1}+f_{2}, f_{1} \in L^{1}$ and $f_{2} \in L^{r}$. Since $|T f| \leq\left|T f_{1}\right|+\left|T f_{2}\right|$, we have by the subadditvity of $\lambda$

$$
\lambda_{T f}(t) \leq \lambda_{T f_{1}}(t / 2)+\lambda_{T f_{2}}(t / 2)
$$

and so by the assumption that $T$ is of weak type $(1,1)$ and $(r, r)$,

$$
\lambda_{T f}(t) \leq \frac{A_{1}}{t / 2} \int\left|f_{1}(x)\right| d \mu(x)+\frac{A_{r}^{r}}{(t / 2)^{r}} \int\left|f_{2}(x)\right|^{r} d \mu(x)
$$

for some fixed constants $A_{1}, A_{r}$ as in the definition of weak type. Because of the choice of splitting $f=f_{1}+f_{2}$, we have

$$
\begin{equation*}
\lambda_{T f}(t) \leq \frac{2 A_{1}}{t} \int_{|f|>t}|f(x)| d \mu(x)+\frac{\left(2 A_{r}\right)^{r}}{t^{r}} \int_{|f| \leq t}|f(x)|^{r} d \mu(x) \tag{47}
\end{equation*}
$$

Using the formula $\|T f\|_{p}^{p}=p \int_{0}^{\infty} t^{p-1} \lambda_{T f}(t) d t$ we multiply both sides of (egn: (marc-step and integrate $d t$. To handle first integral in (egn:marc-step

$$
\begin{align*}
\int_{0}^{\infty} t^{p-1} t^{-1} \int_{|f|>t}|f(x)| d \mu(x) d t & =\int_{X}|f| \int_{0}^{|f|} t^{p-2} d t d \mu(x)  \tag{48}\\
& =\frac{1}{p-1} \int_{X}|f||f|^{p-1} d \mu \tag{49}
\end{align*}
$$

since $p>1$, similarly for the second integral

$$
\begin{align*}
\int_{0}^{\infty} t^{p-1} t^{-r} \int_{|f| \leq t}|f(x)|^{r} d \mu(x) d t & =\int_{X}|f|^{r} \int_{|f|}^{\infty} t^{p-1-r} d t d \mu(x)  \tag{50}\\
& =\frac{1}{r-p} \int_{X}|f|^{r}|f|^{p-r} d \mu \tag{51}
\end{align*}
$$

Putting these together we find that

$$
\|T f\|_{p} \leq A_{p}\|f\|_{p}, \quad \text { with } A_{p}^{p}=\left(\frac{2^{r} A_{1}}{p-1}+\frac{\left(2 A_{r}\right)^{r}}{r-p}\right) p
$$

24.5. Some inequalities. The following should be viewed as a continuous analog of the triangle inequality for the $L^{p}$ norm. It says that "the $L^{p}$ norm of the integral is less than the integral of the $L^{p}$ norms."

Theorem 24.24 (Minkowski's inequality for integrals). Let $(X, \mathscr{M}, \mu)$ and $(Y, \mathscr{N}, \nu)$ be $\sigma$-finite measure spaces and let $f$ be a jointly measurable function on $X \times Y$.
a) If $f$ is unsigned and $1 \leq p<\infty$, then

$$
\begin{equation*}
\left[\int_{X}\left(\int_{Y} f(x, y) d \nu(y)\right)^{p} d \mu(x)\right]^{1 / p} \leq \int_{Y}\left[\int_{X} f(x, y)^{p} d \mu(x)\right]^{1 / p} d \nu(y) \tag{52}
\end{equation*}
$$

eqn:minkowski-
b) If $1 \leq p \leq \infty, f(\cdot, y) \in L^{p}(\mu)$ for a.e. $y$, and the function $\|f(\cdot, y)\|_{p}$ is in $L^{1}(\nu)$, then $f(x, \cdot) \in L^{1}(\nu)$ for a.e. $x$, the function $\int f(x, y) d \nu(y)$ is in $L^{p}(\mu)$, and

$$
\left\|\int f(\cdot, y) d \nu(y)\right\|_{p} \leq \int\|f(\cdot, y)\|_{p} d \nu(y)
$$

The $\sigma$-finite hypothesis is needed to invoke Tonelli's Theorem. The special case

$$
\left(\int\left|\sum_{j=1}^{N} f_{j}\right|^{p} d \mu\right)^{\frac{1}{p}}=\left\|\sum^{N} f_{j}\right\|_{p} \leq \sum^{N}\left\|f_{j}\right\|_{p}=\sum\left(\int\left|f_{j}\right|^{p}\right)^{\frac{1}{p}}
$$

should provide some perspective.

Proof. For $1 \leq p<\infty$, item (b) follows easily from item (a). For $p=\infty$, item (b) is immediate. To prove (a), note that the result is just Tonelli's theorem when $p=1$. For the case $1<p<\infty$, if the right-hand side of $\left(\frac{\text { egn: minkowski-a }}{52) \text { is infmite, there is nothing to prove. }}\right.$ Assuming this right hand side is finite, let $h(x)=\int f(x, y) d \nu(y)$ and fix $g \in L^{q}(\mu)$ where $q$ is the conjugate index to $p$. By Tonelli and Hölder,

$$
\begin{aligned}
\int h(x)|g(x)| d \mu(x) & =\iint f(x, y)|g(x)| d \mu(x) d \nu(y) \\
& \leq\left[\int\left[\int f(x, y)^{p} d \mu(x)\right]^{1 / p} d \nu(y)\right]\|g\|_{q}
\end{aligned}
$$

Thus the linear functional $\lambda: L^{q}(\mu) \rightarrow \mathbb{C}$ given by $\lambda(g)=\int h g d \mu$ is bounded on $L^{q}(\mu)$ with norm at most

$$
\int\left[\int f(x, y)^{p} d \mu(x)\right]^{1 / p} d \nu(y)
$$

By Theorem $\frac{\text { thm:Lp-dual }}{24.13, \text { it follows that }} h \in L^{p}(\mu)$ and obeys the estimate $\left(\frac{\text { egn }}{52}\right)$..

There are many important linear operators on $L^{p}$ spaces arise as integral operators in the sense that they are expressible in the form

$$
\begin{equation*}
T f(x)=\int K(x, y) f(y) d \nu(y) \tag{53}
\end{equation*}
$$

The next proposition gives a simple sufficient condition for the boundedness of such an operator on $L^{p}$. It is a special case of a more general criterion known as the Schur test.
young-ineq Proposition 24.25 (Young's inequality). Let $\mu, \nu$ be $\sigma$-finite positive measures on spaces $X, Y$ respectively and let $K(x, y)$ be a jointly measurable funciton on $X \times Y$. Suppose there is a constant $C$ so that

$$
\sup _{x \in X} \int_{Y}|K(x, y)| d \nu(y) \leq C, \quad \sup _{y \in Y} \int_{X}|K(x, y)| d \mu(x) \leq C .
$$

If $f \in L^{p}(\nu), 1 \leq p \leq \infty$, then, for almost every $x, K(x, y) f(y) \in L^{1}(\nu)$, the function Tf defined by (eqn: basic-integral-operator

$$
\|T f\|_{p} \leq C\|f\|_{p}
$$

Once again, the $\sigma$-finite hypothesis is used to invoke Tonelli's Theorem.

Proof. For $p=1$ or $p=\infty$, the proof is straightforward and left as an exercise. Suppose $1<p<\infty$ and let $q$ denote the conjugate index. The idea is to split $K$ as $K=K^{1 / q} K^{1 / p}$ and apply Hölder's inequality. Indeed,

$$
\begin{aligned}
\int_{Y}|K(x, y) \| f(y)| d \nu(y) & \leq\left(\int_{Y}|K(x, y)| d \nu(y)\right)^{1 / q}\left(\int_{Y}|K(x, y)||f(y)|^{p} d \nu(y)\right)^{1 / p} \\
& \leq C^{1 / q}\left(\int_{Y}|K(x, y)||f(y)|^{p} d \nu(y)\right)^{1 / p}
\end{aligned}
$$

Thus, by Tonelli there is a measurable function $g(x)$ such that $g(x)=\int K(x, y) f(y) d \nu(y)$ almost everywhere $\mu$. Further, by Tonelli's theorem,

$$
\begin{aligned}
\int_{X}|g(x)|^{p} d \mu(x) & \leq C^{p / q} \int_{X}\left(\int_{Y}|K(x, y) \| f(y)|^{p} d \nu(y)\right) d \mu(x) \\
& =C^{p / q} \int_{Y}|f(y)|^{p}\left(\int_{X}|K(x, y)| d \mu\right) d \nu(y) \\
& \leq C^{1+p / q} \int_{Y}|f(y)|^{p} d \nu(y) \\
& =C^{p}\|f\|_{p}^{p} .
\end{aligned}
$$

Hence $g \in L^{p}(\mu)$ and and taking $p^{t h}$ roots finishes the proof.

One context in which Young's inequality is often used is the following: let $X=Y=$ $\mathbb{R}^{n}$ with Lebesgue measure, and fix a function $g \in L^{1}\left(\mathbb{R}^{n}\right)$. If we put $K(x, y)=g(x-y)$, then for a measurable function $f$ on $\mathbb{R}^{n}$ the function

$$
T f(x):=\int_{\mathbb{R}^{n}} g(x-y) f(y) d y
$$

is called the convolution of $f$ and $g$, defined at each $x$ where the integrand is $L^{1}$. Usually we write $T f=g * f$. From the translation invariance of Lebesgue measure,

$$
\sup _{x \in \mathbb{R}^{n}} \int|g(x-y)| d y=\sup _{y \in \mathbb{R}^{n}} \int|g(x-y)| d x=\|g\|_{1}
$$


ineq-convo Corollary 24.26 (Young's inequality for convolutions). If $g \in L^{1}\left(\mathbb{R}^{n}\right), f \in L^{p}\left(\mathbb{R}^{n}\right)$, $1 \leq p \leq \infty$, then $g * f \in L^{p}\left(\mathbb{R}^{n}\right)$ and

$$
\|g * f\|_{p} \leq\|g\|_{1}\|f\|_{p} .
$$

### 24.6. Problems.

 $\left.|f|^{r}, G=|g|^{r}.\right)$

Problem 24.2. Suppose $f, g \geq 0$ with $f \in L^{p}, g \in L^{q}, 0<p, q<\infty$. Show that equality holds in Hölder's inequality (and the generalized Hölder inequality) if and only if one the functions $f^{p}, g^{q}$ is a scalar multiple of the other.
truncation Problem 24.3. [Truncation of $L^{p}$ functions] Suppose $f$ is an unsigned function in $L^{p}(\mu)$, $1<p<\infty$. For $t>0$ let

$$
E_{t}=\{x: f(x)>t\}
$$

a) Show that for each real number $t>0$, the horizontal truncation $\mathbf{1}_{E_{t}} f$ belongs to $L^{q}$ for all $1 \leq q \leq p$.
b) Show that for each real number $t>0$, the vertical truncation $f_{t}:=\min (f, t)$ belongs to $L^{q}$ for all $p \leq q \leq \infty$.
c) As a corollary, show that every $f \in L^{p}, 1<p<\infty$, can be decomposed as $f=g+h$ where $g \in L^{1}$ and $h \in L^{\infty}$.
m-lp-norms Problem 24.4. Suppose $f \in L^{p_{0}} \cap L^{\infty}$ for some $p_{0}<\infty$. Prove $f \in L^{p}$ for all $p_{0} \leq p \leq$ $\infty$, and $\lim _{p \rightarrow \infty}\|f\|_{p}=\|f\|_{\infty}$.
fty-basics Problem 24.5. Prove $f_{n} \rightarrow f$ in the $L^{\infty}$ norm if and only if $f_{n} \rightarrow f$ essentially uniformly, and that $L^{\infty}$ is complete.

Problem 24.6. Suppose $p_{0}<p<p_{1}$ and $f \in L^{p_{0}} \cap L^{p_{1}}$. Prove $f \in L^{p}$ and $\|f\|_{p} \leq$ $\|f\|_{p_{0}}^{1-\theta}\|f\|_{p_{1}}^{\theta}$, where $0<\theta<1$ is chosen so that $\frac{1}{p}=\frac{1-\theta}{p_{0}}+\frac{\theta}{p_{1}}$. When does equality hold?

## ontainment

Problem 24.7. [Containments of $L^{p}$ spaces] a) Show that if $\mu$ is a finite measure, then $L^{p} \subset L^{q}$ for all $p \geq q$. b) Show that $\ell^{p} \supset \ell^{q}$ for all $p \geq q$. c) More generally, show that $L^{p} \subset L^{q}$ for all $p \geq q$ if and only if $\mu$ does not admit sets of arbitrarily large finite measure, and $L^{p} \supset L^{q}$ for all $p \geq q$ if and only if $\mu$ does not admit sets of arbitrarily small positive measure.

Alternate version of the problem above. Let $(X, \mathscr{M}, \mu)$ denote a measure space. If $1 \leq r<p<\infty$ and $L^{r}(\mu) \subset L^{p}(\mu)$, then $\mu$ does not admit sets of arbitrarily small positive measure; that is, there exists a $\delta>0$ such that if $E \in \mathscr{M}$ and $\mu(E)>0$,
then $\mu(E) \geq \delta$. [Suggestion: Apply the closed graph theorem to the inclusion map $\imath: L^{r} \rightarrow L^{p}$.] Conversely, if $\mu$ does not admit sets of arbitrarily small measure, and $1<r \leq p<\infty$, then $L^{r}(\mu) \subset L^{p}(\mu)$. As an example consider the $\ell^{p}$ spaces.

If $L^{p} \subset L^{q}$ use closed graph to show the inclusion is continuous and hence bounded. Get $\left\|1_{E}\right\|_{p} \leq\|\iota\| \|\left. 1_{E}\right|_{q}$ etc. Could really be: if $p>q$ and $L^{p} \subset L^{q}$ then $\ldots$ and conversely, ... . conversely: if doesn't admit arbitrarily small: if $f i n L^{r}$, then with $E_{n}=f>n$, $\mu\left(E_{n}\right)$ converges to 0 hence is eventually 0 , say $\mu\left(E_{M}\right)=0$ and hence

$$
\int|f|^{p}=\int|f|^{r}|f|^{p-r} \leq M^{p-r}\|f\|_{r}<\infty .
$$

inments-rn Problem 24.8. Show that $L^{p}\left(\mathbb{R}^{n}\right) \not \subset L^{q}\left(\mathbb{R}^{n}\right)$ for any pair $p, q$.

Problem 24.9. [Convergence in $L^{p}$ norm] Prove, if $f_{n} \rightarrow f$ in the $L^{p}$ norm, then $f_{n} \rightarrow f$ in measure, and hence a subsequence converges to $f$ a.e. Conversely, show that if $f_{n} \rightarrow f$ in measure and there exists a $g \in L^{p}$ such that $\left|f_{n}\right| \leq g$ for all $n$, then $f_{n} \rightarrow f$ in the $L^{p}$ norm. (Hint: go back and look at the results in Section 12 in last semester's notes, especially Remark 12.18 and Corollary 12.19.)
-of-linfty Problem 24.10. Consider $L^{\infty}(\mathbb{R})$.
a) Show that $\mathcal{M}:=C_{0}(\mathbb{R})$ is a closed subspace of $L^{\infty}(\mathbb{R})$ (more precisely, that the set of $L^{\infty}$ functions that are a.e. equal to a $C_{0}$ function is closed in $L^{\infty}$ ). Prove there is a bounded linear functional $\lambda: L^{\infty} \rightarrow \mathbb{K}$ such that $\left.\lambda\right|_{\mathcal{M}}=0$ and $\lambda\left(\mathbf{1}_{\mathbb{R}}\right)=1$.
b) Prove there is no function $g \in L^{1}(\mathbb{R})$ such that $\lambda(f)=\int_{\mathbb{R}} f g d m$ for all $f \in L^{\infty}$. (Hint: look at the restriction of $\lambda$ to $C_{0}(\mathbb{R})$.)

Problem 24.11. Prove Lemma $\frac{1 \text { em:dist-of-simples }}{24.18 \text {, and use it to carry out the calculation omitted }}$ in the proof of Proposition $\frac{\text { prop:dist-Lp-integral }}{24.21 \text {. }}$

Problem 24.12. Prove, if $f \in L^{p}$ then

$$
\lim _{t \rightarrow 0} t^{p} \lambda_{f}(t)=\lim _{t \rightarrow \infty} t^{p} \lambda_{f}(t)=0
$$

(One way to proceed is to first suppose $f$ is a simple function. Another is to consider the integrals $\int_{\frac{s}{2}}^{s} t^{p-1} \lambda_{f}(t) d t$.)

Problem 24.13. Prove the $r=\infty$ case of the Marcinkiewicz interpolation theorem.

Problem 24.14. Prove the following more general form of Young's inequality: suppose $p, q, r \geq 1$ satisfy $\frac{1}{q}=\frac{1}{p}+\frac{1}{r}-1$. Suppose $K(x, y)$ satisfies

$$
\sup _{x \in X}\|K(x, \cdot)\|_{L^{r}(\nu)} \leq C, \quad \sup _{y \in Y}\|K(\cdot, y)\|_{L^{r}(\mu)} \leq C
$$

for some constant $C$. Prove, if $f \in L^{p}(\nu)$, then $T f=\int_{Y} K(x, y) f(y) d \nu(y) \in L^{q}(\mu)$, and

$$
\|T f\|_{q} \leq C\|f\|_{p}
$$

(Hint: use the same strategy of splitting $|K|=|K|^{\alpha}|K|^{\beta}$, for a suitable choice of $\alpha+\beta=$ 1.)

Deduce the following corollary for convolutions on $\mathbb{R}^{n}$ : with $p, q, r$ as above, if $f \in L^{p}\left(\mathbb{R}^{n}\right)$ and $g \in L^{r}\left(\mathbb{R}^{n}\right)$, then $g * f \in L^{q}\left(\mathbb{R}^{n}\right)$ and

$$
\|g * f\|_{q} \leq\|g\|_{r}\|f\|_{p} .
$$



## 25. The Fourier transform

We assume all functions are complex-valued unless stated otherwise.

Definition 25.1. [The Fourier transform] Let $f \in L^{1}(\mathbb{R})$. The Fourier transform of $f$ is the function $c F 0(f)=\widehat{f}: \mathbb{R} \rightarrow \mathbb{C}$ defined at each $t \in \mathbb{R}$ by

$$
\begin{equation*}
\widehat{f}(t):=\int_{-\infty}^{\infty} f(x) e^{-2 \pi i t x} d x \tag{54}
\end{equation*}
$$

> eqn:fhat-def
$\triangleleft$

Note that $\widehat{f}$ makes sense, since the integrand belongs to $L^{1}$ for each $t \in \mathbb{R}$. We sometimes also use the phrase Fourier transform for the mapping that sends $f$ to $\widehat{f}$. The basic properties of the Fourier transform listed in the following proposition stem from two basic facts: first, that Lebesgue measure is translation invariant, and second that, for each $t \in \mathbb{R}$, the function

$$
\chi_{t}: x \rightarrow \exp (2 \pi i t x)
$$

is a character of the additive group $(\mathbb{R},+)$. This means that $\chi_{t}$ is a homomorphism from $\mathbb{R}$ into the mulitplicative group of unimodular complex numbers, explicitly for all $s, t \in \mathbb{R}$

$$
\chi_{t}(x+y)=\chi_{t}(x) \chi_{t}(y)
$$

Before going further we introduce some notation: for fixed $y \in \mathbb{R}$ and a function $f$ : $\mathbb{R} \rightarrow \mathbb{C}$, define $f_{y}(x):=f(x-y)$.
rier-basic Proposition 25.2 (Basic properties of the Fourier transform). Let $f, g \in L^{1}(\mathbb{R})$ and let $\alpha \in \mathbb{R}$.
(a) (Linearity) $\widehat{c f+g}=c \widehat{f}+\widehat{g}$
(b) (Translation) $\widehat{f}_{y}(t)=e^{-2 \pi i t y} \widehat{f}(t)$
(c) (Modulation) If $g(x)=e^{2 \pi i \alpha x} f(x)$, then $\widehat{g}(t)=\widehat{f}(t-\alpha)$
(d) (Reflection) If $g(x)=\overline{f(-x)}$, then $\widehat{g}(t)=\overline{\hat{f}(t)}$.
(e) (Scaling) If $\lambda>0$ and $g(x)=f(x / \lambda)$ then $\widehat{g}(t)=\lambda \widehat{f}(\lambda t)$.

Proof. Each of these properties is verified by elementary transformations of the integral defining $\widehat{f}$; the details are left as an exercise

It is immediate from the definition that $\hat{f}$ is always a bounded function; indeed $|\widehat{f}(t)| \leq\|f\|_{1}$ for all $t$. Our next observation is:
$\overline{\text { continuity }}$ Proposition 25.3. If $f \in L^{1}(\mathbb{R})$, then $\widehat{f}$ is continuous and bounded $\left(\widehat{f} \in C_{b}(\mathbb{R})\right)$ and $\|\widehat{f}\|_{\infty} \leq\|f\|_{1}$. In particular, the mapping $L^{1}(\mathbb{R}) \ni f \mapsto \widehat{f} \in L^{\infty}(\mathbb{R})$ is a bounded linear map. Moreover if $f_{n}$ is sequence in $L^{1}$ and $f_{n} \rightarrow f$ in the $L^{1}$ norm, then $\widehat{f}_{n} \rightarrow \widehat{f}$ uniformly.

Proof. Fix $t \in \mathbb{R}$ and a sequence $t_{n} \rightarrow t$. The sequence $f(x) e^{-2 \pi i t_{n} x}$ converges to $f(x) e^{-2 \pi i t x}$ pointwise on $\mathbb{R}$, and since trivially $\left|f(x) e^{-2 \pi i t_{n} x}\right| \leq|f(x)|$ for all $n$, we have by dominated convergence

$$
\begin{aligned}
\widehat{f}(t) & =\int_{-\infty}^{\infty} f(x) e^{-2 \pi i t x} d x \\
& =\int_{-\infty}^{\infty} \lim _{n \rightarrow \infty}\left[f(x) e^{-2 \pi i t_{n} x}\right] d x \\
& =\lim _{n \rightarrow \infty} \int_{-\infty}^{\infty} f(x) e^{-2 \pi i t_{n} x} d x \\
& =\lim _{n \rightarrow \infty} \widehat{f}\left(t_{n}\right)
\end{aligned}
$$

The second statement of the theorem follows immediately from the estimate $\sup _{t \in \mathbb{R}}|\widehat{f}(t)| \leq$ $\|f\|_{1}$.

In fact, $\widehat{f}$ always belongs to $C_{0}(\mathbb{R})$, this is known as the Riemann-Lebesgue Lemma. To prove it we first need the following result, which we will apply often (recall the notation $\left.f_{y}(x):=f(x-y)\right)$ :
ranslation Lemma 25.4 (Translation is continuous on $\left.L^{p}\right)$. If $1 \leq p<\infty$ and $f \in L^{p}(\mathbb{R})$, then $\lim _{y \rightarrow 0}\left\|f_{y}-f\right\|_{p}=0$.

Proof. First observe, if $f \in L^{p}$, the $\left\|f_{y}\right\|_{p}=\|f\|_{p}$ by translation invariance of Lebesgue measure. Next, if $g$ is continuous with compact support, then $g \in L^{p}$ and is uniformly continuous from which it readily follows that $\lim _{y \rightarrow 0}\left\|g-g_{y}\right\|_{p}=0$. Continuous functions with compact support are dense in $L^{p}$. Thus, given $f \in L^{p}$ and $\epsilon>0$ there is a continuous function $g$ with compact support such that $\|f-g\|_{p}<\epsilon$. Hence,

$$
\left\|f-f_{y}\right\|_{p} \leq\|f-g\|_{p}+\left\|g-g_{y}\right\|_{p}+\left\|(g-f)_{y}\right\|_{p}<2 \epsilon+\left\|g-g_{y}\right\|_{p}
$$

and the result follows.

Note that the translation invariance of Lebesgue measure shows that the above proposition implies a more general version of itself: if $y_{n} \rightarrow y$ in $\mathbb{R}$, then $f_{y_{n}} \rightarrow f_{y}$ in $L^{p}$.
lem:RL Lemma 25.5 (The Riemann-Lebesgue Lemma). If $f \in L^{1}(\mathbb{R})$, then $\widehat{f} \in C_{0}(\mathbb{R})$.

Proof. The proof is accomplished using the continuity of translation in $L^{1}$ (Lemma $\frac{11 \mathrm{em} \text { : L1-translation }}{2.4}$ ) and a simple trick: first, since $e^{-\pi i}=-1$, we can write

$$
\widehat{f}(t)=-\int_{-\infty}^{\infty} f(x) e^{-2 \pi i t(x+(1 / 2 t))} d x=-\int_{-\infty}^{\infty} f\left(x-\frac{1}{2 t}\right) e^{-2 \pi i x t} d x
$$

Combining this identity with the usual definition of $\widehat{f}$, we have

$$
\widehat{f}(t)=\frac{1}{2} \int_{-\infty}^{\infty}\left(f(x)-f\left(x-\frac{1}{2 t}\right)\right) e^{-2 \pi i x t} d x
$$

so

$$
|\widehat{f}(t)| \leq \frac{1}{2}\left\|f-f_{\frac{1}{2 t}}\right\|_{1}
$$

But by Lemma $\frac{1 \mathrm{em}: \mathrm{L} 1-\text { translation }}{25.4, \text { we have } \| f}-f_{\frac{1}{2 t}} \|_{1} \rightarrow 0$ as $t \rightarrow \pm \infty$.

Continuing our catalog of basic properties, we see that the Fourier transform also interacts nicely with differentiation.
lt-to-diff Proposition 25.6 (Multiplication becomes differentiation). Suppose $f \in L^{1}(\mathbb{R})$. If $g(x):=$ $x f(x)$ belongs to $L^{1}$, then $\widehat{f}$ is differentiable for all $t \in \mathbb{R}$, and $\widehat{g}(t)=\frac{-1}{2 \pi i} \frac{d}{d t} \widehat{f}(t)$.

The proof uses the standard estimate,

$$
\left|1-e^{i t}\right| \leq|t|
$$

for $t$ real and dominated convergence.

Proof. Let $s \neq t$ be real numbers; from the definition of $\widehat{f}$ we have

$$
\begin{equation*}
\frac{\widehat{f}(s)-\widehat{f}(t)}{s-t}=\int_{-\infty}^{\infty} \frac{e^{-2 \pi i s x}-e^{-2 \pi i t x}}{s-t} f(x) d x \tag{55}
\end{equation*}
$$

Now the estimate

$$
\left|\frac{e^{-2 \pi i s x}-e^{-2 \pi i t x}}{s-t}\right| \leq 2 \pi|x|
$$

holds for all $s \neq t$, so by the assumption $x f(x) \in L^{1}$ we can apply dominated convergence in eqn:fourier-diff-step0
in (55) to take the fimit as $s \rightarrow t$ to obtain

$$
\begin{aligned}
\lim _{s \rightarrow t} \frac{\widehat{f}(s)-\widehat{f}(t)}{s-t} & =\lim _{s \rightarrow t} \int_{-\infty}^{\infty} \frac{e^{-2 \pi i s x}-e^{-2 \pi i t x}}{s-t} f(x) d x \\
& =\int_{\infty}^{\infty}(-2 \pi i) e^{-2 \pi i t x} x f(x) d x \\
& =-2 \pi i \widehat{g}(t)
\end{aligned}
$$

Thus $\widehat{f}$ is differentiable and the claimed formula holds.

Note that if $f \in L^{1}$ and also $g(x):=x^{n} f(x) \in L^{1}$ for some integer $n \geq 1$, then $x^{k} f(x)$ belongs to $L^{1}$ for all $0 \leq k \leq n$. The previous proposition can then be applied inductively to conclude.
iff-higher Corollary 25.7. If $f \in L^{1}$ and $g:=x^{n} f \in L^{1}$, then $\widehat{f}$ is $n$ times differentiable, and

$$
\widehat{x^{k} f}=\left(\frac{-1}{2 \pi i}\right)^{k} \widehat{f}^{(k)} \quad \text { for each } 0 \leq k \leq n
$$

One would also expect a theorem in the opposite direction: the Fourier transform should convert differentiation to multiplication by the independent variable. Under reasonable hypotheses, this is the case.
ff-to-mult Proposition 25.8. If $f \in C_{0}(\mathbb{R})$ and $f^{\prime}$ is continuous and in $L^{1}$, then

$$
c F 0\left(f^{\prime}\right)(t)=\widehat{f^{\prime}}(t)=2 \pi i t \hat{f}(t)
$$

Proof. Compute

$$
\begin{aligned}
\widehat{f}^{\prime}(t) & =\int_{-\infty}^{\infty} f^{\prime}(x) e^{-2 \pi i t x} d x \\
& =\lim _{b \rightarrow \infty} \int_{-b}^{b} f^{\prime}(x) e^{-2 \pi i t x} d x \\
& =\lim _{b \rightarrow \infty}\left(\left[f(b) e^{-2 \pi i b t}-f(-b) e^{2 \pi i b t}\right]+2 \pi i t \int_{-\infty}^{\infty} f(x) e^{-2 \pi i x t} d x\right) \\
& =2 \pi i t \widehat{f}(t)
\end{aligned}
$$

where the second equality follows from the Dominated Convergence Theorem, the third using integration by parts, and the fourth from the $C_{0}(\mathbb{R})$ assumption on $f$.

The last set of basic properties of the Fourier transform concern its interaction with convolution, which we now introduce. If $f, g$ are measurable functions on $\mathbb{R}$, the convolution of $f$ and $g$ is the function

$$
\begin{equation*}
(f * g)(x):=\int_{-\infty}^{\infty} f(x-y) g(y) d y \tag{56}
\end{equation*}
$$

defined at each $x$ for which the integral makes sense. In particular, if $f \in L^{\infty}$ and $g \in L^{1}$, then $f * g$ is defined on all of $\mathbb{R}$. Observe, using the invariance of Lebesgue measure with respect to $x \rightarrow-x$ and a simple change of variable,

$$
\begin{equation*}
f * g(x)=\int_{-\infty}^{\infty} g(x-y) f(y) d y=g * f(x) \tag{57}
\end{equation*}
$$

p:convo-L1 Proposition 25.9. If $f, g \in L^{1}(\mathbb{R})$, then
(a) $f * g$ is defined for almost every $x \in \mathbb{R}$;
(b) $f * g$ is measurable;
(c) $f * g \in L^{1}(\mathbb{R}) ;$ and
(d) $\|f * g\|_{1} \leq\|f\|_{1}\|g\|_{1}$.

Proof. Let $H(x, y)=f(x-y) g(y)$. One may check (exercise) that $H$ is jointly measurable as a function of $x$ and $y$. By Tonelli

$$
\begin{aligned}
\iint|H(x, y)| d x d y & =\int_{-\infty}^{\infty}|g(y)|\left(\int_{-\infty}^{\infty}|f(x-y)| d x\right) d y \\
& =\|f\|_{1} \int_{-\infty}^{\infty}|g(y)| d y \\
& =\|f\|_{1}\|g\|_{1}
\end{aligned}
$$

where we have used the translation invariance of Lebesgue measure in the second equality. Hence $H$ is in $L^{1}\left(\mathbb{R}^{2}\right)$. Thus by Fubini, $\int_{-\infty}^{\infty}|f(x-y) g(y)| d y=\int_{-\infty}^{\infty}|H(x, y)| d y$ is finite for almost every $x \in \mathbb{R}$ and $L^{1}(\mathbb{R})$ and further $f * g$ is defined almost everywhere and $L^{1}$ and moreover,

$$
|f * g(x)| \leq \int_{-\infty}^{\infty}|H(x, y)| d y
$$

Hence,

$$
\int_{-\infty}^{\infty}\left|\int_{-\infty}^{\infty} f(x-y) g(y) d y\right| d x \leq \iint|H(x, y)| d y d x=\|f\|_{1}\|g\|_{1}
$$

The following basic properties of convolution are immediate. Its proof is left as an exercise.
ion-basics Proposition 25.10. Let $f, g, h \in L^{1}(\mathbb{R})$.
a) (Commutativity) $f * g=g * f$.
b) (Associativity) $(f * g) * h=f *(g * h)$.
c) (Distributivity) $(f+g) * h=f * g+f * h$.
d) (Scalar multiplication) If $c \in \mathbb{C}$, then $(c f) * g=c(f * g)$.

Notice that these properties together say that, if we equip $L^{1}(\mathbb{R})$ with the usual addition of functions and treat convolution as multiplication, then $L^{1}(\mathbb{R})$ becomes a commutative ring. (In fact it has even more structure, that of a Banach algebra, but we will not pursue this direction in this course). We can now describe how convolution behaves under the Fourier transform.
$\overline{\text { convo-mult }}$ Proposition 25.11 (Convolution becomes multiplication). Let $f, g \in L^{1}(\mathbb{R})$. Then $\widehat{f * g}(t)=$ $\widehat{f}(t) \widehat{g}(t)$.
 $L^{1}\left(\mathbb{R}^{2}\right)$ and thus we can use Fubini's theorem to compute $\widehat{f * g}(t)$ :

$$
\begin{aligned}
\widehat{f * g}(t) & =\int_{-\infty}^{\infty}\left(\int_{-\infty}^{\infty} f(x-y) g(y) d y\right) e^{-2 \pi i x t} d x \\
& =\int_{-\infty}^{\infty} g(y)\left(\int_{-\infty}^{\infty} f(x-y) e^{-2 \pi i x t} d x\right) d y \\
& =\int_{-\infty}^{\infty} \widehat{f}(t) e^{-2 \pi i y t} g(y) d y \\
& =\widehat{f}(t) \widehat{g}(t)
\end{aligned}
$$

where we have used Proposition $\frac{\text { prop; fourier-basic }}{25.2(b) .}$

Given what we have proved so far, it follows that the Fourier transform is a ring homomorphism from $L^{1}(\mathbb{R})$ (with addition and convolution) to $C_{0}(\mathbb{R})$ (with pointwise addition and multiplication). We will see later that the Fourier transform is injective. It turns out that it is not surjective, however.

Let us finish this section by computing an important example. Let $a>0$ and consider the Gaussian

$$
g(x):=e^{-\pi a x^{2}}
$$

(The factor of $\pi$ will be convenient given our choice of normalization in the definition of the Fourier transform.) It should be clear that $x^{n} g(x) \in L^{1}$ for all $a>0$ and $n \geq 0$.
m-gaussian Lemma 25.12. $\widehat{g}(t)=\frac{1}{\sqrt{a}} e^{-\pi t^{2} / a}$.

Proof. Rather than computing the integral directly, we exploit Propositions prop:mult-to-diff prop:diff-to-mult
prop:diff-to-mult 25.8. We may also assume $a=1$ since the general case follows from this by scaling
(Proposition $\frac{\text { prop; fourier-basic }}{25.2(\mathrm{e})) \text {. Let } \hbar=x} \mathrm{f}$ and note that $g^{\prime}=-2 \pi h$. Since $h \in L^{1}$,

$$
\begin{align*}
(\widehat{g})^{\prime}(t) & =-2 \pi i \hat{h}(t) \\
& =-2 \pi i\left(\widehat{\left.-\frac{1}{2 \pi} g\right)^{\prime}}\right.  \tag{58}\\
& =i 2 \pi i t \hat{g}(t) \\
& =-2 \pi t \hat{g}(t),
\end{align*}
$$




$$
\frac{d}{d t}\left(e^{\pi t^{2}} \widehat{g}(t)\right)=0
$$

Hence the function $e^{\pi t^{2}} \widehat{g}(t)$ is constant. To evaluate the constant, we set $t=0$ and use the well-known Gaussian integral

$$
\widehat{g}(0)=\int_{-\infty}^{\infty} e^{-\pi x^{2}} d x=1
$$

25.1. Digression - convolution and approximate units. In this section we pause to develop further properties of convolutions; these will be necessary in order to study the Fourier inversion problem later.

A general principle is that the convolution of two functions inherits the best properties of both. We will see a number of instances of this phenomenon. A simple expression of this principle is the following:

Proposition 25.13. If $f \in L^{\infty}$ and $g \in L^{1}$, then then $f * g$ is bounded and continuous.

Proof. Since $f$ is bounded,

$$
|(f * g)(x)| \leq \int_{-\infty}^{\infty}\left|f(x-y)\|g(y) \mid d y \leq\| f\left\|_{\infty}\right\| g \|_{1}\right.
$$

To see that $f * g$ is continuous, fix $x \in \mathbb{R}$ and let $x_{n} \rightarrow x$. Let $h_{n}(y)=\left[g\left(x_{n}-y\right)-\right.$ $g(x-y)] f(y)$. Since $g \in L^{1}$ and $f \in L^{\infty}$, each $h_{n} \in L^{1}$ and

$$
\left|g * f\left(x_{n}\right)-g * f(x)\right| \leq\left\|h_{n}\right\|_{1} \leq\left\|g_{-x_{n}}-g_{-x}\right\|_{1}\|f\|_{\infty}
$$

where $g_{z}(y)=g(y-z)$. The result now follow from continuity of translation in $L^{1}$ (Lemma $\frac{1 \mathrm{lem}}{25.4)}$ ) (and commutativity of the convolution).
pt-support Proposition 25.14. If $f, g$ are $L^{1}$ functions and are both compactly supported, then so if $f * g$.

Proof. Problem ${ }^{275.3 \text { proo: }}$

The dominated convergence argument can be extended to apply to differentiability:
convo-diff Proposition 25.15. If $f$ is a compactly supported $C^{k}$ function $(1 \leq k \leq \infty)$ and $g \in L^{1}$, then $f * g \in C^{k}$.

Proof. First assume $k=1$. Then one can differentiate under the integral sign as in the proof of Proposition $\frac{\text { prop:mult-to-diff }}{25.6 \text { using uniform differentiability of } f \text {. The general case is proved }}$ by induction; the details are left as an exercise.
pprox-unit Definition 25.16. An $L^{1}$ approximate unit is a collection of functions $\phi_{\lambda} \in L^{1}(\mathbb{R})$ indexed by $\lambda>0$ such that:
a) $\phi_{\lambda}(t) \geq 0$ almost everywhere, for each $\lambda$,
b) $\int_{-\infty}^{\infty} \phi_{\lambda}(t) d t=1$ for all $\lambda$, and
c) For each fixed $\delta>0$, we have $\left\|\mathbf{1}_{|t|>\delta} \phi_{\lambda}\right\|_{1} \rightarrow 0$ as $\lambda \rightarrow 0$.

Approximate units are easy to construct: Let $\phi$ be any nonnegative, $L^{1}$ function with $\int_{-\infty}^{\infty} \phi(x) d x=1$; then the functions $\phi_{\lambda}(x):=\frac{1}{\lambda} \phi\left(\frac{x}{\lambda}\right)$ form an $L^{1}$ approximate unit (Problem ${ }^{\text {prob; scaled-approx-unit }}$ 2.5(a)). The simplest example comes from taking $\phi(x)=\mathbf{1}_{[-1 / 2,1 / 2]}$; the resulting $\phi_{\lambda}$ is known as the box kernel. (Draw a few of these for different values of $\lambda$ to see what is going on.) Of course approximate units need not be compactly supported; we will see a very important example later (the Poisson kernel). The significance of approximate units (and their name) is explained by the following theorem.
prox-units Theorem 25.17. Let $\phi_{\lambda}$ be an $L^{1}$ approximate identity. If $1 \leq p<\infty$ and $f \in L^{p}(\mathbb{R})$, then for each $\lambda$ the convolution $f * \phi_{\lambda}$ is defined almost everywhere, $f * \phi_{\lambda} \in L^{p}(\mathbb{R})$ and $\left\|\phi_{\lambda} * f-f\right\|_{p} \rightarrow 0$ as $\lambda \rightarrow 0$.

To prove the theorem we need two facts. The first is Minkowski's integral inequality (Theorem thm:minkowski matter); note where each of the properties the approximate unit is used.
pprox-unit Lemma 25.18. Suppose $\phi_{\lambda}$ is an $L^{1}$ approximate unit and $g \in L^{\infty}$. If $g$ is continuous at a point $x \in \mathbb{R}$, then

$$
\lim _{\lambda \rightarrow 0}\left(g * \phi_{\lambda}\right)(x)=g(x)
$$

Proof. Using the trick $g(x)=\int_{-\infty}^{\infty} \phi_{\lambda}(y) g(x) d y$,

$$
\left(g * \phi_{\lambda}\right)(x)-g(x)=\int_{-\infty}^{\infty}(g(x-y)-g(x)) \phi_{\lambda}(y) d y
$$

so using the positivity of $\phi_{\lambda}$

$$
\begin{equation*}
\left|\left(g * \phi_{\lambda}\right)(x)-g(x)\right| \leq \int_{-\infty}^{\infty}|g(x-y)-g(x)| \phi_{\lambda}(y) d y \tag{59}
\end{equation*}
$$

To estimate the right-hand side, let $\epsilon>0$ be given. By the continuity of $g$ at $x$, choose $\delta>0$ so that $|g(x-y)-g(x)|<\epsilon$ when $|y|<\delta$. We then split the integral in (年gn: approx-unit-lemma-ster two integrals, over the regions $|y| \leq \delta$ and $|y|>\delta$ :

$$
\int_{-\infty}^{\infty}|g(x-y)-g(x)| \phi_{\lambda}(y) d y=\int_{\{|y| \leq \delta\}}|g(x-y)-g(x)| \phi_{\lambda}(y) d y+\int_{\{|y|>\delta\}}|g(x-y)-g(x)| \phi_{\lambda}(y) d y
$$

The first integrand is bounded by $\epsilon \phi_{\lambda}$, so

$$
\int_{\{|y| \leq \delta\}}|g(x-y)-g(x)| \phi_{\lambda}(y) d y \leq \epsilon \int_{\{|y| \leq \delta\}} \phi_{\lambda}(y) d y \leq \epsilon
$$

since $\int_{-\infty}^{\infty} \phi_{\lambda}(y) d y=1$. The second integrand is bounded by $2\|g\|_{\infty} \mathbf{1}_{\{|y|>\delta\}} \phi_{\lambda}(y)$, so goes to 0 as $\lambda \rightarrow 0$ by property (c) in the definition of approximate unit. This proves the lemma.

Remark: If one assumes that the approximate unit $\phi_{\lambda}$ was constructed as $\phi_{\lambda}(y)=$ $\frac{1}{\lambda} \phi\left(\frac{y}{\lambda}\right)$ for some $\phi$ satisfying $\phi \geq 0$ and $\int \phi=1$, then the lemma has an easier proof (see Problem ${ }^{\text {prob; scaled-approx-unit }}$

Proof of Theorem $\frac{\text { thm:L1-approx-units }}{25.17 . ~ L e t ~} f \in L^{p}$ be given. By Corollary $\frac{\text { cor:young-ineq-convo }}{24.26 \text { (or a routine argu- }}$ ment), $f * \phi_{\lambda}$ is defined for almost every $x$ and is in $L^{p}$. Let $d \mu=d x$ denote Lebesgue measure and $d \nu$ the measure $\phi_{\lambda}(y) d y$. Define $F(x, y)=\left(f_{y}-f\right)(x)=f(x-y)-f(x)$. In particular, $F(\cdot, y) \in L^{p}(\mu)$ and

$$
\|F(\cdot, y)\|_{p} \leq 2\|f\|_{p}
$$

for each $x$. Since $\nu$ is a finite measure, $H(y)=\|F(\cdot, y)\|_{p}$ is in $L^{1}(\mu)$.
It follows from Theorem $\frac{\text { thm: minkowski }}{24.24 \text { that for }} \psi$-almost every $x$ the function $G_{x}(y)=$ $F(x, y)=f(x-y)-f(x)$ is in $L^{1}(\nu)$,

$$
g_{\lambda}(x)=\int_{-\infty}^{\infty} F(x, y) d \nu(y)=\int_{-\infty}^{\infty} F(x, y) \phi_{\lambda}(y) d y=\left(f_{y} * \phi_{\lambda}-f\right)(x)
$$

is in $L^{p}(\mu)$ with

$$
\left\|g_{\lambda}\right\|_{p} \leq \int_{-\infty}^{\infty}\|F(\cdot, y)\|_{p} d \nu=\int_{-\infty}^{\infty}\left\|f_{y}-f\right\|_{p} \phi_{\lambda}(y) d y
$$

Thus,

$$
\left\|f * \phi_{\lambda}-f\right\|_{p} \leq h * \phi_{\lambda}(0)
$$

where $h(y)=\left\|f_{-y}-f\right\|_{p}$. Hence

$$
\begin{equation*}
\left\|f_{y} * \phi_{\lambda}-f\right\|_{p}=\left\|g_{\lambda}\right\|_{p} \leq h * \phi_{\lambda}(0) \tag{60}
\end{equation*}
$$

The function $h$ belongs to $L^{\infty}$ (indeed $\|h\|_{\infty} \leq 2\|f\|_{p}$ ) and is continuous (by the continuity of translation in $L^{p}$ ) and $h(0)=0$. Thus by Lemma lem:continuity-approx-unit equation ( ${ }^{\text {eq }} \mathrm{f} \mathrm{f}$ ) app 11 goes to 0 as $\lambda \rightarrow 0$ concluding the proof.

It is often useful to have approximate units with additional properties, such as smoothness or compact support. In fact it is possible to construct an $L^{1}$ approximate unit $\left\{\psi_{\lambda}\right\}$ consisting of $C^{\infty}$ functions with compact support. This is accomplished via bump functions:

Definition 25.19. A bump function is a function $\psi: \mathbb{R} \rightarrow \mathbb{R}$ such that:
a) $\psi \in C^{\infty}(\mathbb{R})$,
b) $\psi$ is compactly supported,
c) $\psi \geq 0$, and
d) $\int_{-\infty}^{\infty} \psi(x) d x=1$
-functions Lemma 25.20. Bump functions exist.

Proof. The main issue is to construct a $C^{\infty}$ function with compact support. Consider the function

$$
h(x)= \begin{cases}e^{-1 / x} & \text { if } x>0 \\ 0 & \text { if } x \leq 0\end{cases}
$$

Clearly $h \geq 0$ and $h$ is differentiable for all $x \neq 0$; it is a calculus exercise to verify that $h$ is infinitely differentiable at 0 and $h^{(n)}=0$ for all $n$. It is then straightforward to
verify that $\psi(x)=c \cdot h(x+1) h(1-x)$ is a bump function (for a suitable normalizing constant $c$ ), supported on $[-1,1]$.

Note that if $\psi$ is a bump function supported in $[-1,1]$, then the functions $\psi_{\lambda}(x):=$ $\frac{1}{\lambda} \psi\left(\frac{x}{\lambda}\right)$ are also bump functions, supported in $[-\lambda, \lambda]$. (Draw a picture of what these functions look like as $\lambda \rightarrow 0$ ). Thus there exist approximate units consisting of smooth, compactly supported functions. As an application, we can prove that $C_{c}^{\infty}(\mathbb{R})$ is dense in $L^{p}(\mathbb{R})$ for all $1 \leq p<\infty$. A proof is outlined in Problem prob:smooth-dense-Lp in $L^{p}(\mathbb{R})$ for all $1 \leq p<\infty$. A proof is outlined in Problem prob:
25.2. Inversion and uniqueness. In this section we study the problem of recovering $f$ from $\widehat{f}$. Loosely, the Fourier transform can be thought of as a resolution of $f$ as a superposition of sinusoidal functions $e^{2 \pi i t x}$; the value of $\widehat{f}(t)$ measures the "amplitude" of $f$ in the "frequency" $t$. This suggests that a formula like

$$
\begin{equation*}
f(x)=\int_{-\infty}^{\infty} \widehat{f}(t) e^{2 \pi i t x} d t \tag{61}
\end{equation*}
$$

ought to hold, at least if $\widehat{f} \in L^{1}$. If we formally substitute the definition of $\widehat{f}$ and switch the order of integration, we are confronted with

$$
\int_{-\infty}^{\infty} f(u)\left(\int_{-\infty}^{\infty} e^{2 \pi i(x-u) t} d t\right) d u
$$

and the inner integral is not convergent, regardless of any assumption on $\widehat{f}$. In fact (皆) :inversion-intro does hold when $\widehat{f} \in L^{1}$, but a more delicate argument is necessary. So, the goal of this section will be to prove:
-inversion Theorem 25.21 (Fourier inversion, $L^{1}$ case). If $f$ and $\widehat{f}$ belong to $L^{1}$, then

$$
\begin{equation*}
f(x)=\int_{-\infty}^{\infty} \widehat{f}(t) e^{2 \pi i x t} d t \tag{62}
\end{equation*}
$$

eqn:L1-inversi
for almost every $x \in \mathbb{R}$.

Once we have the inversion formula, we see that $L^{1}$ functions are determined by their Fourier transforms:

Corollary 25.22. Suppose $f, g \in L^{1}$. If $\widehat{f}=\widehat{g}$, then $f=g$ a.e.

Proof. From the inversion theorem, if $f \in L^{1}$ and $\hat{f}=0$, then $f=0$. By the linearity of the Fourier transform, $\widehat{f-g}=\widehat{f}-\widehat{g}$, and the corollary follows.

So, in principle, $f$ is fully determined by $\hat{f}$, even if $\widehat{f} \notin L^{1}$; and this is often the case. For example, for $a>0$ the Fourier transform $f=\mathbf{1}_{[-a, a]}$ is

$$
\widehat{f}(t)=\frac{\sin (2 \pi a t)}{\pi t}
$$

and an exercise shows $\widehat{f}$ does not belong to $L^{1}$. To recover $f$ from $\widehat{f}$ in these cases, we turn to summability methods; in fact summability methods will already be of use in proving the inversion theorem. The idea is this: suppose we have a divergent integral

$$
\int_{-\infty}^{\infty} h(t) d t
$$

where the function $h$ is, say, locally $L^{1}$, but not $L^{1}$. We might try to make sense of the integral as

$$
\lim _{a \rightarrow+\infty} \int_{-a}^{a} h(t) d t
$$

effectively we have introduced the cutoff function $\psi_{a}(t):=\mathbf{1}_{[-a, a]}$, which is positive, integrable, and increases to 1 pointwise as $a \rightarrow \infty$. Given any family of functions $\psi_{a}$ with these three propertes, we can consider the integrals

$$
\int_{-\infty}^{\infty} h(t) \psi_{a}(t) d t
$$

The "square" cutoff $\mathbf{1}_{[-a, a]}$ has some undesirable properties; e.g., its Fourier transform is not $L^{1}$ (and not of constant sign). We will work first with smoother cutoff functions, in particular the functions $t \rightarrow \exp (-a|t|)$ (here we consider $a \rightarrow 0$ rather than $a \rightarrow \infty$, but this is not important).

The first step is to compute the (inverse) Fourier transform of

$$
\begin{equation*}
Q_{a}(t)=e^{-2 a \pi|t|} \tag{63}
\end{equation*}
$$

(the extra factor of $2 \pi$ turns out to be a convenient normalization).
omputation Lemma 25.23. For all $a>0$,

$$
\int_{-\infty}^{\infty} Q_{a}(t) e^{2 \pi i t x} d t=\frac{1}{\pi} \frac{a}{a^{2}+x^{2}}
$$

prob:poisson-computation
Proof. Problem $\frac{\text { prob: }}{25.7}$.

Let us fix the notation

$$
\begin{equation*}
P_{a}(x):=\frac{1}{\pi} \frac{a}{a^{2}+x^{2}} \tag{64}
\end{equation*}
$$

Notice that $P_{1}(x)$ is nonnegative and $\int_{-\infty}^{\infty} P_{1}(x) d x=1$. Moreover, $P_{a}(x)=\frac{1}{a} P_{1}\left(\frac{x}{a}\right)$. By the remarks following Definition $\frac{\text { def: approx-unit }}{25.16 \text {, we have: }}$
lem:Pappx Lemma 25.24. $\left\{P_{a}\right\}_{a>0}$ is an $L^{1}$ approximate unit.

The function $P_{a}(x)$ (viewed as a function of the two arguments $(a$ and $x)$ is known as the Poisson kernel. We are now able to compute the integral (eqn:inversion-intro cutoff function $Q_{a}(t)$ :
onvolution Proposition 25.25. If $f \in L^{1}$, then for all $a>0$ and all $x \in \mathbb{R}$

$$
\left(f * P_{a}\right)(x)=\int_{-\infty}^{\infty} Q_{a}(t) \widehat{f}(t) e^{2 \pi i t x} d t
$$

Proof. For $a>0$ and $x$ fixed, let $G_{x}(t, y)=Q_{a}(t) f(y) e^{-2 \pi(x-y) t}$. Observe that

$$
\int\left|G_{a}(t, y)\right| d y d t=\|f\|_{1}\left\|Q_{a}\right\|_{1}=\frac{\|f\|_{1}}{\pi a}
$$

Thus $G_{a}$ is $L^{1}\left(\mathbb{R}^{2}\right)$ and hence we can apply Fubini (explaining the role of the cut-off function $Q_{a}$ )

$$
\begin{align*}
\int_{-\infty}^{\infty} Q_{a}(t) \widehat{f}(t) e^{2 \pi i x} d t & =\int_{-\infty}^{\infty} Q_{a}(t) \int_{-\infty}^{\infty} f(y) e^{2 \pi i(x-y) t} d y d t  \tag{65}\\
& =\int_{-\infty}^{\infty} Q_{a}(t) \int_{-\infty}^{\infty} f(x-y) e^{2 \pi i y t} d y d t  \tag{66}\\
& =\int_{-\infty}^{\infty} f(x-y) \int_{-\infty}^{\infty} Q_{a}(t) e^{2 \pi i y t} d t d y  \tag{67}\\
& =\left(f * P_{a}\right)(x) \tag{68}
\end{align*}
$$

Proof of Theorem |thm:L1-inversion 25.21 . Assume $f, \widehat{f} \in L^{1}(\mathbb{R})$. Define

$$
g(x)=\int_{-\infty}^{\infty} \widehat{f}(t) e^{2 \pi i t x} d t
$$

By the Riemann-Lebesgue lemma (Lemma $\frac{1 \mathrm{em}: \mathrm{RL}}{25.5),} g \in C_{0}(\mathbb{R})$. We want to show $g=f$
a.e. From Proposition prop:poisson-convolution

$$
\begin{equation*}
\left(f * P_{a}\right)(x)=\int_{-\infty}^{\infty} Q_{a}(t) \widehat{f}(t) e^{2 \pi i t x} d t \tag{69}
\end{equation*}
$$

Fix a sequence $a_{n} \rightarrow 0$. Using the hypothesis $\widehat{f} \in L^{1}$, an application of dominated convergence shows, for all $x$, the integral in the right-hand side of (egn: inversion-step $g(x)$ as $a_{n} \rightarrow 0$. On the other hand, since $P_{a}$ is an $L^{1}$ approximate unit, we know from Theorem ${ }_{2}^{\text {thm:L1-approx-units }} 25.17$ that $f * P a_{n} \rightarrow f$ in $L^{1}$. Passing to a subsequence, we may assume that $f * P_{a_{n}} \rightarrow f$ almost everywhere, but then by (年y $)^{\text {inversion-step }}$ wave $f * P_{a_{n}} \rightarrow g$ a.e., so $f=g$ a.e. and the theorem is proved.

Remark 25.26. Observe that the above proof did not really use the explicit form of $P_{a}$; rather the point was that $Q_{a}(t)=\left\{e^{-2 a \pi|t|}\right\}_{a>0}$ was a cutoff function (uniformly bounded and converging pointwise to the constant function 1) whose Fourier transform $\left\{P_{a}\right\}$ was an $L^{1}$ approximate unit. Any other cutoff function with this property could have been used.
 sense for any $f \in L^{1}$ (that is, not assuming $\widehat{f} \in L^{1}$ ). Indeed, combining Proposition prop:poisson-convolutio and Theorem $\frac{\text { thm:L1-approx-units }}{25.17}$ we have immediately:

Corollary 25.27 (Fourier inversion in the $L^{1}$ norm). If $f \in L^{1}$, then

$$
\begin{equation*}
\int_{-\infty}^{\infty} Q_{a}(t) \widehat{f}(t) e^{2 \pi i x t} d t \tag{70}
\end{equation*}
$$

eqn:regularize
converges to $f$ in the $L^{1}$ norm as $a \rightarrow 0+$.

So, we can recover $f$ but only in the $L^{1}$ norm; the corollary does not say anything about the pointwise covergence of the reguralized integrals. In fact, it is true that the interals legn:regularized-inversion
integrals (TV) converge to $f$ a.e., but this requires a more delicate argument.
25.3. The $L^{2}$ theory. In this section we study the Fourier transform on $L^{2}$. There is an immediate problem, of course, since by Problem prob:lp-no-containments-rn (eqn: fhat-def need not be defined. However, we can observe that $L^{1} \cap L^{2}$ is dense in $L^{2}$ (why?), and start there.
plancherel Lemma 25.28. If $f \in L^{1} \cap L^{2}$, then $\widehat{f}$ belongs to $L^{2}$ and $\|\widehat{f}\|_{2}=\|f\|_{2}$.

Proof. Let $\tilde{f}(x):=\overline{f(-x)}$. Since $f, \tilde{f} \in L^{1}$, the convolution $g=f * \widetilde{f}$ is defined a.e. and $g \in L^{1}$ by Proposition prop: convo-L1

$$
g(x)=\int_{-\infty}^{\infty} f(x-y) \overline{f(-y)} d y=\int_{-\infty}^{\infty} f(x+y) \overline{f(y)} d y
$$

 $L^{2}$ and of course the vector $f$ determines a continuous linear functional. Thus $g$ is a continuous function of $x$, and $g(0)=\|f\|_{2}^{2}$. By Cauchy-Schwarz again,

$$
|g(x)| \leq\left\|f_{-x}\right\|_{2}\|f\|_{2}=\|f\|_{2}^{2}
$$

so $g$ is bounded.

Let, as before $Q_{a}(t)=\exp (-2 a \pi|t|)$. Since $g \in L^{1}$ we can apply Proposition prop:poisson-convolutio to compute

$$
\left(g * P_{a}\right)(0)=\int_{-\infty}^{\infty} Q_{a}(t) \widehat{g}(t) d t
$$

lem:continuity-approx-unit
As $g$ is continuous, by Lemma $\frac{1 \mathrm{em}: \text { co }}{25.18}$

$$
\begin{equation*}
\|f\|_{2}^{2}=g(0)=\lim _{a \rightarrow 0}\left(g * P_{a}\right)(0)=\lim _{a \rightarrow 0} \int_{-\infty}^{\infty} Q_{a}(t) \widehat{g}(t) d t . \tag{71}
\end{equation*}
$$

Let us compute the limit of this last integral in a different way. Recall that by definition $g=f * \tilde{f}$, so by Propositions $\frac{\text { prop: convolemopt fourier-basic }}{25.11 \text { and } 25 / 2(\mathrm{~d}) \text {, }}$

$$
\widehat{g}(t)=|\widehat{f}(t)|^{2}
$$

Making this substitution in the integral in (ITf) and applying the monotone convergence theorem,

$$
\|f\|_{2}^{2}=\lim _{a \rightarrow 0} \int_{-\infty}^{\infty} Q_{a}(t) \widehat{g}(t) d t=\lim _{a \rightarrow 0} \int_{-\infty}^{\infty} Q_{a}(t)|\widehat{f}(t)|^{2} d t=\|\widehat{f}\|_{2}^{2}
$$

fourier-L2 Theorem 25.29 (The Fourier transform on $L^{2}$ ). There is a unique bounded linear transformation $\mathcal{F}: L^{2} \rightarrow L^{2}$ satisfying the following conditions:
a) For all $f \in L^{1} \cap L^{2}, \mathcal{F} f=\widehat{f}$.
b) (The Plancherel theorem) $\|\mathcal{F} f\|_{2}=\|f\|_{2}$ for all $f \in L^{2}$.
c) The mapping $f \rightarrow \mathcal{F} f$ is an Hilbert space isomorphism of $L^{2}$ onto $L^{2}$.
d) (The Parseval identity) $\langle f, g\rangle=\langle\mathcal{F} f, \mathcal{F} g\rangle$ for all $f, g \in L^{2}$.

Remark: Note that when $f, g \in L^{1} \cap L^{2}$, the Parseval identity reads

$$
\int_{-\infty}^{\infty} f(x) \overline{g(x)} d x=\int_{-\infty}^{\infty} \widehat{f}(t) \overline{\widehat{g}(t)} d t
$$

 mation from a dense subspace of $L^{2}$ into $L^{2}$. Thus, since the codomain $L^{2}$ is complete,
 $\mathcal{F}: L^{2} \rightarrow L^{2}$. This proves (a), and (b) follows since $\|f\|_{2}=\|\mathcal{F} f\|_{2}$ on a dense set (namely $L^{1} \cap L^{2}$ ). (d) is also an immediate consequence of (b), by Problem ${ }^{\text {prob;hilby-isometries }} 23.3(\mathrm{a})$. It remains to prove (c); what we must show is that $\mathcal{F}$ is onto.

We show that $\mathcal{F}$ has dense range; combined with the fact that $\mathcal{F}$ is an isometry, it follws that $\mathcal{F}$ is in fact onto. (The proof of this last assertion is left as an exercise). Let $M$ denote the set of all functions $g \in L^{2}$ such that $g=\widehat{f}$ for some $f \in L^{1} \cap L^{2}$. Clearly the range of $\mathcal{F}$ contains $M$, so it will suffice to prove that $M$ is dense, or equivalently, that $M^{\perp}=\{0\}$.

Recall the cutoff functions $Q_{a}(x)=e^{-2 a \pi|x|}, a>0$ introduced in equation ( $\frac{\text { deff }: \text { Qsuba }}{63}$. The functions $e^{2 \pi i b x} e^{-2 a \pi|x|}$ belong to $L^{1} \cap L^{2}$ for all $a>0$ and $b \in \mathbb{R}$, so their Fourier transforms

$$
P_{a}(t-b)=\int_{-\infty}^{\infty} e^{2 \pi i b x} Q_{a}(x) e^{-2 \pi i t x} d x
$$

belong to $M$. So, let $h \in M^{\perp}$ be given and let $H(x)=h(-x)$. Thus,

$$
\left(P_{a} * \bar{H}\right)(-b)=\int_{-\infty}^{\infty} P_{a}(-b-t) \overline{h(-t)} d t=\int_{-\infty}^{\infty} P_{a}(t-b) \overline{h(t)} d t=0
$$

for all $b$. Since, by Lemma $\frac{1 \text { em:Pappx }}{25.24, P_{a}}$ is an approximate unit and $\bar{H} \in L^{2}$, Theorem $\frac{\mid \text { thm:L1-approx-units }}{25.17}$
implies $P_{a} * \bar{H}$ converges to $\bar{H}$ in $L^{2}$. Hence $h=0$ and consequently $M$ is dense in $L^{2}$ and the proof is finished.
-inversion Theorem 25.30 ( $L^{2}$ inversion). Let $f \in L^{2}$. Define

$$
\phi_{N}(t)=\int_{-N}^{N} f(x) e^{-2 \pi i x t} d x, \quad \psi_{N}(t)=\int_{-N}^{N}(\mathcal{F} f)(t) e^{2 \pi i x t} d t .
$$

Then $\left\|\phi_{N}-\mathcal{F} f\right\|_{2} \rightarrow 0$ and $\left\|\psi_{N}(t)-f\right\|_{2} \rightarrow 0$ as $N \rightarrow \infty$.

Proof. Let $f_{N}:=\mathbf{1}_{[-N, N]} f$. Then $f_{N} \in L^{1} \cap L^{2}$, and $\phi_{N}=\widehat{f_{N}}$. An application of dominated convergence shows that $\left(f_{N}\right)_{N}$ converges to $f$ in the $L^{2}$ norm. Hence, $\left(\phi_{N}=\right.$ $\left.\widehat{f_{N}}=\mathcal{F} f_{N}\right)_{N}$ converges in $L^{2}$ to $\mathcal{F} f$ by Theorem thm:fourier-L2

The statement for $\psi_{N}$ is proved by similar methods and is left as an exercise (Problem prob:L2-inversion lem 25.10).

It is important to note that, for a general function $f \in L^{2}$, its Fourier transform is defined only as an element of $L^{2}$. In particular it is defined only a.e., and cannot be evaluated at points. From now on we just write $\widehat{f}$ for $\mathcal{F} f$ when $f \in L^{2}$, with the understanding that the integral definition is only valid when $f \in L^{1} \cap L^{2}$.
25.4. Fourier Series. We now replace the group $(\mathbb{R},+)$ with the multiplicative group of unimodular complex numbers

$$
\mathbb{T}=\left\{e^{i \theta}:-\pi \leq \theta<\pi\right\} \subset \mathbb{C}
$$

By the properties of the exponential, this group isomorphic to the group $[-\pi, \pi)$ with addition mod $2 \pi$. (This amounts to the isomorphism of the quotient group $\mathbb{R} / 2 \pi \mathbb{Z} \cong \mathbb{T}$.) We will typically use $\theta$ to identify elements of $\mathbb{T}$. We will treat functions $f: \mathbb{T} \rightarrow \mathbb{C}$ either as functions on the unit circle in $\mathbb{C}$, or, when convenient, as functions on the interval $[-\pi, \pi)$ extended $2 \pi$-peridoically to $\mathbb{R}$. For each integer $n$, the map

$$
\chi_{n}: e^{i \theta} \rightarrow e^{i n \theta}
$$

is a homomorphism from $\mathbb{T}$ to $\mathbb{T}$. In fact, any continuous homomorphism $h: \mathbb{T} \rightarrow \mathbb{C}$ is $\chi_{n}$ for some $n \in \mathbb{Z}$.

We equip $\mathbb{T}$ with normalized arc length measure, or equivalently normalized Lebesgue measure on $[-\pi, \pi)$. We write $d m$ for this measure. Note that $\mathbb{T}$ acts on itself by rotation. For fixed $\theta$, the map $\tau_{\theta}: \mathbb{T} \rightarrow \mathbb{T}$ given by

$$
\tau_{\theta}\left(e^{i \psi}\right)=e^{i(\psi+\theta)}
$$

amounts to rotation of the circle through the angle $\theta$. Moreover, $m$ is invariant under $\tau_{\theta}$ for each $\theta$. Arguing exactly as on the line, one can prove the continuity of translation in $L^{p}(\mathbb{T})$. For fixed $\varphi \in[-\pi, \pi)$, let $f_{\varphi}(\theta)=f(\theta-\varphi)$.

## cns-circle

Proposition 25.31. If $1 \leq p<\infty$ and $f \in L^{p}(\mathbb{T})$, then the mapping $\mathbb{R} \rightarrow L^{p}(\mathbb{T})$ defined by $\varphi \mapsto f_{\varphi}$ is continuous.

Now that we have a translation invariant measure and the characters $\chi_{n}$, we can define a Fourier transform.

Definition 25.32. Let $f \in L^{1}(\mathbb{T})$. The Fourier coefficients of $f$ are the numbers

$$
\widehat{f}(n):=\int_{\mathbb{T}} f(\theta) e^{-i n \theta} d m(\theta)=\frac{1}{2 \pi} \int_{-\pi}^{\pi} f(x) e^{-i n x} d x, \quad n \in \mathbb{Z}
$$

The Fourier series of $f$ is the series

$$
\begin{equation*}
f \sim \sum_{n=-\infty}^{\infty} \widehat{f}(n) e^{i n \theta} \tag{72}
\end{equation*}
$$

The function $\widehat{f}: \mathbb{Z} \rightarrow \mathbb{C}$ is called the Fourier transform of $f$. The Fourier coefficients $\widehat{f}(n)$ are the analogs of the pointwise values of $\widehat{f}(t)$ in the case of $\mathbb{R}$. Thus, in each case the Fourier transform of a function $f$ is a function defined on the (group of) characters. Similarly, the (possibly divergent) Fourier series is the analog of the (possibly divergent) integral

$$
\int_{-\infty}^{\infty} \widehat{f}(t) e^{2 \pi i t x} d t
$$

Just as in the case of the real line, the Fourier transform behaves predictably under translation, modulation, reflection, and scaling; we leave the statements and proofs of these facts as exercises. Moreover the same kind of integral estimates show that $\widehat{f}$ is always a bounded function.
leb-circle Theorem 25.33 (Riemann-Lebesgue lemma on the circle). If $f \in L^{1}(\mathbb{T})$, then

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

This version of the Riemann-Lebesgue Lemma follows readily from the version on $\mathbb{R}$ (Lemma $\frac{1 \mathrm{lem}: \mathrm{RL}}{25.5) .}$ In other words, the Fourier transform takes $L^{1}(\mathbb{T})$ into $c_{0}(\mathbb{Z})$. As in the case of the line, this map turns out to be injective (which will follow from an appropriate inversion theorem), but not surjective. Also, as we found on the line, typically $\widehat{f} \notin \ell^{1}(\mathbb{Z})$. Thus, there is an immediate difficulty in interpreting the series (egn:fourier-series-def

As before, the basic problem is to recover $f$ from $\widehat{f}$, which in turn means finding a way to attach meaning to the (in general divergent) Fourier series. Broadly, the method is the same as in the case of $\mathbb{R}$ : we introduce a cutoff function whose Fourier series is nicely convergent to an approximate unit for $L^{1}(\mathbb{T})$. Before doing this we have a look at what can go wrong, even for nice $f$. Let us try to naively sum the Fourier series: fix $f \in L^{1}(\mathbb{T})$ and consider the partial sums

$$
s_{N}(\theta)=\sum_{n=-N}^{N} \hat{f}(n) e^{i n \theta} .
$$

Since this is a finite sum, we can expand $\widehat{f}$ as an integral and pull the sum inside:

$$
\begin{equation*}
s_{N}(\theta)=\int_{\mathbb{T}} f(\phi)\left\{\sum_{n=-N}^{N} e^{-i n \phi} e^{i n \theta}\right\} d m(\phi) \tag{73}
\end{equation*}
$$

Working with the inner sum, we consider the expression

$$
D_{N}(t):=\sum_{n=-N}^{N} e^{i n t}=\frac{\sin \left(N+\frac{1}{2}\right) t}{\sin \frac{t}{2}}
$$

Then (eqn: dirichlet-convo-step

$$
s_{N}(\theta)=\int_{\mathbb{T}} f(\phi) D_{N}(\theta-\phi) d m=\left(f * D_{N}\right)(\theta)
$$

where we have introduced convolution on the circle group $\mathbb{T}$. (Note that the difference $\theta-\phi$ is interpreted in the group $\mathbb{T}$, that is, is carried out $\bmod 2 \pi$.) Thus, the question of whether the partial sums $s_{N}$ converge to $f$ in some sense (pointwise a.e., or in $L^{1}$, etc.) reduces to the question of whether $f * D_{N}$ converges to $f$ in the same sense. Unfortunately, since $D_{N}$ is not an $L^{1}(\mathbb{T})$ approximate unit, the partial sums $s_{N}$ can be badly behaved, even for nice $f$. For example we have the following:

Theorem 25.34. There exists a continuous function $f$ on $\mathbb{T}$ such that the Fourier series for $f$ diverges at $\theta=0$.

Proof. We present an outline of the proof; the details are left as an exercise. As noted above, the $N^{t h}$ partial sum of the Fourier series of $f$ at a point $\theta$ is given by $\left(f * D_{N}\right)(\theta)$. We suppose that $\left(f * D_{N}\right)(0) \rightarrow f(0)$ for every $f \in C(\mathbb{T})$ and derive a contradiction. Now

$$
s_{N}(0)=\left(f * D_{N}\right)(0)=\int_{\mathbb{T}} f(\phi) D_{N}(\phi) d m(\phi)
$$

By the construction of $D_{N}$, it is clear that $D_{N} \in L^{1}(\mathbb{T})$ for each $N$. Thus for each $N$ the map $L_{N}: f \rightarrow \int_{\mathbb{T}} f D_{N} d m$ is a bounded linear functional on $C(\mathbb{T})$, and one can show that the norm of this functional is equal to $\left\|D_{N}\right\|_{1}$. (To see this, find a sequence of continuous functions $g_{n}$ such that $\left\|g_{n}\right\|_{\infty} \leq 1$ for all $n$ and $g_{n} \rightarrow \operatorname{sgn} D_{N}$ pointwise. Then $\left|L_{N}\left(g_{n}\right)\right| \rightarrow\left\|D_{N}\right\|_{1}$.) Next, one can show by direct estimates of the integral that $\left\|D_{N}\right\|_{1} \rightarrow \infty$ as $N \rightarrow \infty$. The proof finishes by appeal to the Principle of Uniform Boundedness: if it were the case that $L_{N}(f)=s_{N}(0) \rightarrow f(0)$ for all $f \in C(\mathbb{T})$,
then the family of linear functionals $L_{N}$ would be pointwise bounded on $C(\mathbb{T})$, hence uniformly bounded, which is a contradiction. Problem $\frac{\text { prob:fourier-divergence }}{25.17}$ gives some hints on filling the details.

Before going further, let us observe that if we assume $f$ has a certain amount of smoothness at a point, then the Fourier series for $f$ will converge to $f$ at that point. A simple result of this type is the following:

Proposition 25.35. If $f \in L^{1}(\mathbb{T})$ and $f$ is differentiable at a point $\theta_{0}$, then $s_{N}\left(\theta_{0}\right) \rightarrow$ $f\left(\theta_{0}\right)$.

Proof. By considering real and imaginary parts, we may assume $f$ is real-valued, and by replacing $f(\theta)$ by $f\left(\theta+\theta_{0}\right)-f\left(\theta_{0}\right)$ we may assume that $\theta_{0}=0$ and $f(0)=0$. As we have already observed, we have

$$
\begin{align*}
s_{N}(0)=\left(f * D_{N}\right)(0) & =\frac{1}{2 \pi} \int_{-\pi}^{\pi} f(\phi) D_{N}(\phi) d \phi  \tag{74}\\
& =\frac{1}{2 \pi} \int_{-\pi}^{\pi} f(\phi) \frac{\sin \left(N+\frac{1}{2}\right) \phi}{\sin \frac{\phi}{2}} d \phi \tag{75}
\end{align*}
$$

and we wish to prove $s_{N}(0) \rightarrow 0$ as $N \rightarrow \infty$. The key observation is that the function

$$
g(\phi)=\frac{f(\phi)}{\sin \frac{\phi}{2}}
$$

belongs to $L^{1}(\mathbb{T})$. To see this, first note the elementary estimate

$$
\left|\sin \frac{\phi}{2}\right| \leq \frac{|\phi|}{2}
$$

for $|\phi| \leq \frac{\pi}{2}$ and that there is an $0<\eta \leq \frac{\pi}{2}$ such that

$$
\left|\sin \frac{\phi}{2}\right| \geq \frac{|\phi|}{\pi}
$$

for $|\phi|<\eta$.

Now, since $f$ is differentiable at 0 and $f(0)=0$, there exist $M>0$ and $0<\delta<\eta$ such that

$$
\sup _{|\phi|<\delta}\left|\frac{f(\phi)}{\phi}\right| \leq M
$$

so

$$
|g(\phi)|=\left|\frac{f(\phi)}{\phi}\right|\left|\frac{\phi}{\sin \frac{\phi}{2}}\right| \leq \pi M
$$

for $|\phi| \leq \delta$. On the other hand, for $\delta<|\phi| \leq \pi$,

$$
\left|\frac{f(\phi)}{\sin \frac{\phi}{2}}\right| \leq \frac{\pi}{\delta}|f(\phi)|
$$

Thus, $g$ is bounded near 0 and dominated by $f$ away from 0 , hence $g \in L^{1}$. Returning to (egn: diff-summability-step

$$
s_{N}(0)=\frac{1}{2 \pi} \int_{-\pi}^{\pi} g(\phi) \sin \left(N+\frac{1}{2}\right) \phi d \phi
$$

Making the change of variable $\theta=\phi / 2$ gives

$$
\begin{equation*}
s_{N}(0)=\frac{1}{\pi} \int_{-\pi / 2}^{\pi / 2} g(2 \theta) \sin ((2 N+1) \theta) d \theta \tag{76}
\end{equation*}
$$

If we now put

$$
h(\theta)=2 \mathbf{1}_{[-\pi / 2, \pi / 2]}(\theta) g(2 \theta),
$$

then $h \in L^{1}$ and the integral in (eqn: diff-convergence-step2 is nothing but the imaginary part of $\widehat{h}(2 N+1)$, which goes to 0 as $N \rightarrow \infty$ by the Riemann-Lebesgue lemma. This finishes the proof.

So, to recover $f$ from its Fourier series, as before we need to introduce a cutoff function, but since the "square" cutoff $\mathbf{1}_{[-N, N]}$ (corresponding to ordinary partial sums) is badly behaved, we choose a smoother cutoff. It turns out that the functions

$$
n \rightarrow r^{|n|}
$$

for $0 \leq r<1$ are a good choice (analgous to $e^{-a \pi|t|}$ on the line). Thus we consider the Abel means of the Fourier series

$$
A(r, \theta):=\sum_{n=-\infty}^{\infty} \widehat{f}(n) r^{|n|} e^{i n \theta}
$$

Since $\widehat{f}$ is bounded and $r<1$, this series is absolutely convergent for all $\theta$, and unifomrly convergent on $\mathbb{T}$ for each fixed $r$. Thus, we can again expand $\widehat{f}$ as an integral, and interchange the sum and integral.

$$
A(r, \theta)=\int_{\mathbb{T}} f(\phi)\left\{\sum_{n=-\infty}^{\infty} r^{|n|} e^{i n(\theta-\phi)}\right\} d m(\phi):=\left(f * P_{r}\right)(\theta)
$$

where, by summing the geometric series, $P_{r}(\theta)$ is given by

$$
P_{r}(\theta):=\sum_{n=-\infty}^{\infty} r^{|n|} e^{i n \theta}=\frac{1-r^{2}}{1-2 r \cos \theta+r^{2}}
$$

The function $P_{r}(\theta)$ is the Poisson kernel.
rox-unit-T Lemma 25.36. The family of function $P_{r}$ has the following properties.
i) $P_{r}(\theta) \geq 0$ for all $\theta \in[-\pi, \pi]$ and all $0 \leq r<1$,
ii) For each $r, \int_{\mathbb{T}} P_{r}(\theta) d m(\theta)=1$,
iii) For each fixed $0<\delta<\pi$,

$$
\frac{1}{2 \pi} \int_{\delta \leq|\theta| \leq \pi} P_{r}(\theta) d \theta \rightarrow 0 \quad \text { as } r \rightarrow 1
$$

Proof. Exercise.

In other words, $\left\{P_{r}\right\}_{r<1}$ is an $L^{1}(\mathbb{T})$ approximate unit. Just as on the line, we have

Theorem 25.37. If $1 \leq p<\infty$ and $f \in L^{p}(\mathbb{T})$, then $\left\|f * P_{r}-f\right\|_{p} \rightarrow 0$ as $r \rightarrow 1$.

Proof. A proof can be constructed using the properties of approximate units and Minkowski's
 details.

With these results in hand, we can obtain Abel summability of Fourier series on the circle:

Corollary 25.38. If $f \in L^{p}(\mathbb{T}), 1 \leq p<\infty$, then the Abel means $A(r, \theta)$ of the Fourier series for $f$ converge to $f$ in $L^{p}$ as $r \rightarrow 1$. In particular, the set $E=\left\{e_{n}=e^{i n \theta}: n \in\right.$ $\mathbb{Z}\} \subset L^{2}([-\pi, \pi])$ is an orthonormal basis of $L^{2}([-\pi, \pi])$.

Proof. The first part of the Corollary is immediate. For the second part, note $E$ is an orthonormal set (we are using normalized Lebesgue measure on $[-\pi, \pi]$. Further, if $f$ is orthogonal to $E$, then its Abel means are 0 and hence $f=0$.

Just as with the Fourier transform, there is an inversion principle. In the context of Fourier series it is $L^{1}-\ell^{1}$ inversion. First observe, if $f \in L^{1}$ and $\widehat{f}=0$, then $A_{f}(r, \theta)=0$
converges to $f$ in $L^{1}$. Hence $f=0$. Thus, if $f, g \in L^{1}$ and $\widehat{f}=\hat{g}$, then $f=g$. Finally, if $f \in L^{1}$ and $\widehat{f} \in \ell^{1}(\mathbb{Z})$, then the series

$$
g=\sum \widehat{f}(n) e^{i n \theta}
$$

converges uniformly to a continuous (and hence $L^{1}(\mathbb{T})$ ) function $g$ and $\hat{g}(n)=\widehat{f}(n)$ by direct computation. Thus $g=f$ (as $L^{1}$ functions).

What happens when $p=\infty$ ? We have already seen that the Fourier series of a continuous function can diverge at a given point; however if we use the Abel means $A(r, \theta)$ we can do better. The reason is the following lemma, which says that the Poisson kernel $P_{r}(\theta)$ obeys a stronger condition than that of Lemma lem:poisson-approx-unit-T

Lemma 25.39. For each $0<\delta<\pi$, we have $P_{r}(\theta) \longrightarrow 0$ uniformly on $\delta \leq|\theta| \leq \pi$ as $r \rightarrow 1$.

Proof. Fix $\delta$. For $\delta<|\theta| \leq \pi$, we have $-1 \leq \cos \theta \leq \cos \delta<1$. Thus for such $\theta$

$$
P_{r}(\theta)=\frac{1-r^{2}}{1-2 r \cos \theta+r^{2}} \leq \frac{1-r^{2}}{1-2 r \cos \delta+r^{2}}
$$

As $r \rightarrow 1$, the numerator of this last expression goes to 0 while the denominator is bounded away from 0 , which proves the lemma.

Theorem 25.40. If $f \in C(\mathbb{T})$, then $f * P_{r}=A(r, \theta) \rightarrow f(\theta)$ uniformly as $r \rightarrow 1$.

Proof. Fix $f \in C(\mathbb{T})$ and $\epsilon>0$. Since $f$ is continuous and $\mathbb{T}$ is compact, $f$ is uniformly continuous, so there exists $\delta>0$ such that $|f(\theta)-f(\phi)|<\epsilon$ whenever $|\theta-\phi|<\delta$. Using
our usual tricks with approximate units we write $f(\theta)=\int f(\theta) P_{r}(\phi) d \phi$ to obtain

$$
\left|\left(f * P_{r}\right)(\theta)-f(\theta)\right| \leq \frac{1}{2 \pi} \int_{-\pi}^{\pi}|f(\theta-\phi)-f(\theta)| P_{r}(\phi) d \phi
$$

We split the integral as $\int_{|\phi|<\delta}+\int_{\delta<|\phi| \leq \pi}$. For the first integral, we have $|f(\theta-\phi)-f(\theta)|<$ $\epsilon$ for $|\phi|<\delta$ by uniform continuity, so

$$
\frac{1}{2 \pi} \int_{|\phi|<\delta}|f(\theta-\phi)-f(\theta)| P_{r}(\phi) d \phi<\epsilon \frac{1}{2 \pi} \int_{-\pi}^{\pi} P_{r}(\theta)=\epsilon .
$$

For the second integral, for all $r$ sufficiently large we have $P_{r}(\theta)<\epsilon$ on $\delta<|\phi| \leq \pi$ by the lemma, while $|f(\theta-\phi)-f(\theta)| \leq 2\|f\|_{\infty}$, so

$$
\frac{1}{2 \pi} \int_{\delta<|\phi| \leq \pi}|f(\theta-\phi)-f(\theta)| P_{r}(\phi) d \phi \leq 2 \epsilon\|f\|_{\infty}
$$

Thus for $r$ sufficiently close to 1 , we get $\left|f * P_{r}(\theta)-f(\theta)\right| \leq\left(1+2\|f\|_{\infty}\right) \epsilon$, so $f * P_{r} \rightarrow f$ unifomrly on $\mathbb{T}$.

From the smoothing properties of convolution, we see that if $f \in L^{1}$ then $f * P_{r}$ is continuous in $\theta$ for each $r$. Thus there is no hope that $f * P_{r} \rightarrow f$ in the $L^{\infty}$ norm when $f \in L^{\infty}$. However we do have the following weaker form of convergence:
ar-fourier Proposition 25.41. If $f \in L^{\infty}(\mathbb{T})$ and $g \in L^{1}(\mathbb{T})$, then

$$
\lim _{r \rightarrow 1} \frac{1}{2 \pi} \int_{-\pi}^{\pi}\left(f * P_{r}\right)(\theta) g(\theta) d \theta=\frac{1}{2 \pi} \int_{-\pi}^{\pi} f(\theta) g(\theta) d \theta
$$

## Proof. Problem $\frac{\text { prob:weak-star-fourier }}{25.18}$

As for the line, the Fourier transform is espcially well-behaved in the $L^{2}$ setting, though here things are somewhat simpler-we have $L^{2}(\mathbb{T}) \subset L^{1}(\mathbb{T})$ since the measure is finite. The proof of the following theorem is left as an exercise.

## -L2-circle Theorem 25.42. The Fourier transform is a unitary transformation from $L^{2}(\mathbb{T})$ onto

 $\ell^{2}(\mathbb{Z})$. In particular,i) (Plancherel theorem) If $f \in L^{2}(\mathbb{T})$, then $\|f\|_{2}^{2}=\sum_{n=-\infty}^{\infty}|\hat{f}(n)|^{2}$.
ii) (Parseval identity) If $f, g \in L^{2}(\mathbb{T})$, then

$$
\int_{\mathbb{T}} f(\theta) \overline{g(\theta)} d m(\theta)=\sum_{n=-\infty}^{\infty} \widehat{f}(n) \overline{\bar{g}(n)}
$$

Moreover, for $f \in L^{2}(\mathbb{T})$ the partial sums $S_{N}$ of the Fourier Series for $f$ converges to $f$ in $L^{2}(\mathbb{T})$.

In particular, note that the Fourier transform takes the orthonromal basis $E=$ $\left\{e^{i n \theta}\right\}_{n \in \mathbb{Z}}$ of $L^{2}(\mathbb{T})$ onto the standard orthonormal basis of $\ell^{2}(\mathbb{Z})$, and the Fourier transform of a function $f \in L^{2}(\mathbb{T})$ is just its sequence of coefficients with respect to this orthonormal basis. Indeed if we write $e_{n}(\theta)=e^{i n \theta}$, then

$$
\widehat{f}(n):=\int_{\mathbb{T}} f(\theta) e^{-i n \theta} d m(\theta)=\left\langle f, e_{n}\right\rangle_{L^{2}(\mathbb{T})}
$$

Thus, if one has already proved that the functions $\left\{e_{n}\right\}$ are an orthonormal basis for thm: fourier-L2-circle . 25.42 becomes quite simple. (Indeed, the theorem is essentially equivalent to the assertion that the characters $\left\{e_{n}\right\}$ form an orthonormal basis. Why?)
25.5. Schwartz functions and distributions. This section introduces the Schwartz space $\mathscr{S}$ and the space of tempered distritbutions $\mathscr{S}^{\prime}$. The theory of distributions allows many of the important operations of analysis (such as differentiation and the Fourier transform) to be extended to objects more singular than functions (indeed distributions are sometimes known as generalized functions). The basic idea is this: if $\psi$ is a very smooth function (say $C^{\infty}$ ) and vanishes at infinity, then if $f$ is differentiable we have the integration by parts formula

$$
\int_{\mathbb{R}} f^{\prime}(x) \psi(x) d x=-\int_{\mathbb{R}} f(x) \psi^{\prime}(x) d x
$$

However, the second integral will make sense even if the first does not (that is, even if $f^{\prime}$ does not exist). If we identify $f$ with the linear functional

$$
\psi \rightarrow \int_{\mathbb{R}} f \psi
$$

then the above calculation suggests that we can interpret " $f^{\prime \prime}$ " as the linear functional

$$
\psi \rightarrow-\int f \psi^{\prime}
$$

even if $f^{\prime}$ does not exist in the usual sense. The theory of distributions makes this heuristic precise. The first step is to carefully identify the space of smooth functions we wish to use, and topologize it appropriately so that we can speak of continuous linear functionals.

Let $C_{b}^{\infty}(\mathbb{R})$ denote the vector space of bounded $C^{\infty}$ functions on $\mathbb{R}$.

Definition 25.43. The Schwartz space $\mathscr{S}$ consists of all functions $\psi \in C_{b}^{\infty}(\mathbb{R})$ such that $x^{\alpha} \psi^{(\beta)}(x)$ is bounded for all integers $\alpha, \beta \geq 0$.

We say that a function $f$ is rapidly decreasing if $x^{\alpha} \psi(x)$ is bounded for all $\alpha \geq 0$. So $\mathscr{S}$ consists of those $\psi$ such that $\psi$ and all of its derivatives are rapidly decreasing. For example, $\psi(x)=e^{-x^{2}}$ belongs to $\mathscr{S}$. It is an important fact that $\mathscr{S}$ is closed under differentiation, and under multiplication by polynomials:

Lemma 25.44. $\mathscr{S}$ is a vector space, and if $\psi \in \mathscr{S}$ then $x \psi(x)$ and $\psi^{\prime}(x)$ belong to $\mathscr{S}$, and in fact $x^{\alpha} \psi^{(\beta)} \in \mathscr{S}$ for all $\alpha, \beta \geq 0$.

Proof. Exercise.

Definition 25.45. For integers $\alpha, \beta \geq 0$, define for $\psi \in \mathscr{S}$

$$
\|\psi\|_{\alpha, \beta}:=\left\|x^{\alpha} \psi^{(\beta)}\right\|_{\infty}
$$

Lemma 25.46. Each $\|\cdot\|_{\alpha, \beta}$ is a norm on $\mathscr{S}$.

It turns out that it is appropriate to topologize $\mathscr{S}$ not with a single norm, but with the whole family of norms $\|\cdot\|_{\alpha, \beta}$ simultaneously.

Definition 25.47. Say that a sequence $\psi_{n} \subset \mathscr{S}$ is Cauchy if it is Cauchy in each of the norms $\|\cdot\|_{\alpha, \beta}$, and say that a sequence $\psi_{n} \subset \mathscr{S}$ converges if there exists $\psi \in \mathscr{S}$ such that $\left\|\psi_{n}-\psi\right\|_{\alpha, \beta} \rightarrow 0$ for all $\alpha, \beta$.

Of course, if $\psi_{n}$ converges in $\mathscr{S}$ then the limit $\psi$ is unique. (Check this.) Further:

Proposition 25.48. $\mathscr{S}$ is complete. That is, if $\psi_{n}$ is Cauchy in $\mathscr{S}$, then there exists $\psi \in \mathscr{S}$ such that $\left\|\psi_{n}-\psi\right\|_{\alpha, \beta} \rightarrow 0$ for all $\alpha, \beta \geq 0$.

Proof. Since $C_{b}(\mathbb{R})$ is complete with respect to the $\|\cdot\|_{\infty}$ norm, by the definition of $\mathscr{S}$ and the $\|\cdot\|_{\alpha, \beta}$ norms we have that, for each $\alpha, \beta \geq 0$, there is a function $\psi_{\alpha, \beta} \in C_{b}(\mathbb{R})$ such that $x^{\alpha} \psi_{n}^{(\beta)} \rightarrow \psi_{\alpha, \beta}$ uniformly on $\mathbb{R}$. Put $\psi=\psi_{0,0}$, the proof is finished if we can show that $\psi_{\alpha, \beta}=x^{\alpha} \psi^{(\beta)}$ for all $\alpha, \beta$.

From advanced calculus we know that if $f_{n}$ converges uniformly to $f$ and $f_{n}^{\prime}$ converges uniformly to $g$, then $f$ is differentiable and $f^{\prime}=g$. Applying this fact we conclude that $\psi_{0,1}=\psi_{0,0}^{\prime}$, and applying it inductively we have $\psi_{0, \beta}=\psi^{(\beta)}$ for all $\beta \geq 0$. (In particular, $\psi_{n}^{(\beta)} \rightarrow \psi^{(\beta)}$ uniformly for all $\beta$. From this it follows that $x^{\alpha} \psi_{n}^{(\beta)} \rightarrow x^{\alpha} \psi^{(\beta)}$ pointwise for all $\alpha, \beta$, but since this sequence also converges to $\psi_{\alpha, \beta}$ uniformly we conclude that $\psi_{\alpha, \beta}=x^{\alpha} \psi^{(\beta)}$ for all $\alpha, \beta$ as desired.

We observed earlier that $\mathscr{S}$ is closed under differentiation and multiplication by polynomials; we now see that these operations are continuous:
hwartz-cns Lemma 25.49. If $\psi_{n} \rightarrow \psi$ in $\mathscr{S}$, then $x^{\alpha} \psi_{n} \rightarrow x^{\alpha} \psi$ and $\psi_{n}^{(\beta)} \rightarrow \psi^{(\beta)}$ in $\mathscr{S}$.

Proof. Exercise.

It is useful to observe (in connection with our discussion of the Fourier transfrom later) that convergence in the family of norms $\|\cdot\|_{\alpha, \beta}$ controls $L^{p}$ convergence:

Proposition 25.50. $\mathscr{S} \subset L^{p}$ for all $1 \leq p \leq \infty$, and if $\psi_{n} \rightarrow \psi$ in $\mathscr{S}$ then also $\psi_{n} \rightarrow \psi$ in $L^{p}(\mathbb{R})$.

Proof. By the previous lemma, we have $\left(1+x^{2}\right) \psi \in \mathscr{S}$ for all $\psi \in \mathscr{S}$, so in particular for all $x \in \mathbb{R}$

$$
|\psi(x)| \leq \frac{|\psi(x)|+\left|x^{2} \psi(x)\right|}{1+x^{2}} \leq \frac{\|\psi\|_{0,0}+\|\psi\|_{2,0}}{1+x^{2}}
$$

But $\left(1+x^{2}\right)^{-1}$ belongs to $L^{p}(\mathbb{R})$ for all $1 \leq p \leq \infty$, so $\psi \in L^{p}$. Applying this same estimate to $\psi_{n}-\psi$ we see that

$$
\left\|\psi_{n}-\psi\right\|_{p} \leq\left(\left\|\psi_{n}-\psi\right\|_{0,0}+\left\|\psi_{n}-\psi\right\|_{2,0}\right)\left\|\left(1+x^{2}\right)^{-1}\right\|_{p}
$$

and the right hand side goes to 0 as $\psi_{n} \rightarrow \psi$ in $\mathscr{S}$.

Definition 25.51. The space of tempered distributions $\mathscr{S}^{\prime}$ consists of all continuous linear maps $F: \mathscr{S} \rightarrow \mathbb{C}$. That is, a map $F: \mathscr{S} \rightarrow \mathbb{C}$ belongs to $\mathscr{S}^{\prime}$ if and only if it is linear and $F\left(\psi_{n}\right) \rightarrow F(\psi)$ whenever $\psi_{n} \rightarrow \psi$ in $\mathscr{S}$.

It is straightforward to check that $\mathscr{S}^{\prime}$ is a vector space. To emphasize the role of $\mathscr{S}^{\prime}$ as the dual space of $\mathscr{S}$, we will write $\langle F, \psi\rangle$ for $F(\psi)$. Tempered distributions $F$ are sometimes called generalized functions. We will topologize $\mathscr{S}^{\prime}$ as follows: say $F_{n} \rightarrow F$ in $\mathscr{S}^{\prime}$ if $\left\langle F_{n}, \psi\right\rangle \rightarrow\langle F, \psi\rangle$ for all $\psi \in \mathscr{S}$.

The following examples are fundamental; the unproved claims are left as exercises.

Example 25.52. a) (Tempered functions) A measurable function $f: \mathbb{R} \rightarrow \mathbb{C}$ is called tempered if $(1+|x|)^{-N} f \in L^{1}$ for some integer $N \geq 0$.

Each tempered function $f$ defines a tempered distribution by the formula

$$
\langle f, \psi\rangle=\int_{\mathbb{R}} f \psi
$$

(To see that $f \psi \in L^{1}$ for every $\psi \in \mathscr{S}$, write $f \psi=(1+|x|)^{-N} f(1+|x|)^{N} \psi$.) The fact that $\psi \rightarrow\langle f, \psi\rangle$ is continuous follows from dominated convergence. For examples of tempered functions, note that every $f \in L^{p}, 1 \leq p \leq \infty$ is tempered (apply Hölder's inequality to $(1+|x|)^{-2} f$ ). More generally, any polynomial times a tempered function is a tempered function. Let us also observe that if $f, g$ are tempered functions, then the associated tempered distributions are equal if and only if $f=g$ a.e. This justifies the name "generalized functions."
b) (Tempered measures) A (positive, signed, or complex) Borel measure $\mu$ on $\mathbb{R}$ is called tempered if $\int_{\mathbb{R}}(1+|x|)^{-N} d|\mu|(x)<\infty$ for some integer $N \geq 0$. Every tempered measure gives rise to a tempered distribution via the pairing

$$
\langle\mu, \psi\rangle=\int_{\mathbb{R}} \psi d \mu
$$

If $\mu$ is absolutely continuous with respect to Lebesgue measure $m$, with RadonNikodym derivative $f=\frac{d \mu}{d m}$, then $\mu$ is tempered if and only if $f$ is a tempered function, and we are back to example (a).

To give more examples, we first look at ways to obtain new tempered distributions from old ones. One elementary but important way is the following:
tributions Proposition 25.53. If $F \in \mathscr{S}^{\prime}$ and $T: \mathscr{S} \rightarrow \mathscr{S}$ is a continuous linear map, then

$$
\left\langle T^{\prime} F, \psi\right\rangle:=\langle F, T \psi\rangle
$$

defines a tempered distribution.

Proof. If $\psi_{n} \rightarrow \psi$ in $\mathscr{S}$, then

$$
\left\langle F, T \psi_{n}\right\rangle \rightarrow\langle F, T \psi\rangle
$$

so $T^{\prime} F$ defines a distribution.

Before moving on to more general classes of distributions, we consider one more special example:

Proposition 25.54 (The Principal Value integral). For each $\psi \in \mathscr{S}$, the limit

$$
\begin{equation*}
\left\langle P_{1 \mid x}, \psi\right\rangle:=\lim _{\epsilon \rightarrow 0} \int_{|x| \geq \epsilon} \frac{1}{x} \psi(x) d x \tag{77}
\end{equation*}
$$

exists, and defines a tempered distribution.

Proof. We first show that (egn:hilby-transform-distribution is well-defined on $\mathscr{S}$. Let $\psi \in \mathscr{S}$, then by changing variables in the integral on the negative half-line we get for each $\epsilon>0$

$$
\int_{|x| \geq \epsilon} \frac{1}{x} \psi(x) d x=\int_{\epsilon}^{\infty} \frac{\psi(x)-\psi(-x)}{x} d x
$$

Since $\psi$ is differentiable at 0 , the integrand is bounded in a neighborhood of 0 , and since $x \psi(x)$ is bounded, the integrand decays faster than $1 / x^{2}$ near infinity, so the integral is convergent. Thus the limit exists as $\epsilon \rightarrow 0$, and equals

$$
\int_{0}^{\infty} \frac{\psi(x)-\psi(-x)}{x} d x
$$

To see that $P_{1 / x}$ is continuous on $\mathscr{S}$, first observe that for $x>0$

$$
\begin{align*}
\left|\frac{\psi(x)-\psi(-x)}{x}\right| & =\left|\frac{1}{x} \int_{-x}^{x} \psi^{\prime}(t) d t\right|  \tag{78}\\
& \leq \frac{1}{x} \int_{-x}^{x}\left|\psi^{\prime}(t)\right| d t  \tag{79}\\
& \leq 2\left\|\psi^{\prime}\right\|_{\infty} \tag{80}
\end{align*}
$$

It follows that

$$
\begin{align*}
\left|\left\langle P_{1 / x}, \psi\right\rangle\right| & \leq \int_{0}^{1}\left|\frac{\psi(x)-\psi(-x)}{x}\right|+\int_{1}^{\infty}\left|\frac{\psi(x)-\psi(-x)}{x}\right|  \tag{81}\\
& \leq 2\left\|\psi^{\prime}\right\|_{\infty}+\int_{1}^{\infty}(|x \psi(x)|+|x \psi(-x)|) \frac{d x}{x^{2}}  \tag{82}\\
& \leq 2\left\|\psi^{\prime}\right\|_{\infty}+2\|x \psi(x)\|_{\infty}  \tag{83}\\
& =2\|\psi\|_{0,1}+2\|\psi\|_{1,0} \tag{84}
\end{align*}
$$

If we consider now $\psi_{n} \rightarrow \psi$ in $\mathscr{S}$, then the above estimate applied to $\psi_{n}-\psi$ shows that $\left\langle P_{1 / x}, \psi_{n}\right\rangle \rightarrow\left\langle P_{1 / x}, \psi\right\rangle$ and the proof is finished.

Proposition 25.55 (Differentiation of temepered distributions). For any integer $\beta \geq 0$ and any tempered distribution $F \in \mathscr{S}^{\prime}$, the map

$$
\psi \rightarrow\left\langle F,(-1)^{\beta} \psi^{(\beta)}\right\rangle
$$

defines a tempered distribution, called the $\beta^{\text {th }}$ distributional derivative of $F$, denoted $F^{(\beta)}$.
 follows by Proposition prop:T

The reason for including the sign $(-1)^{\beta}$ in the definition of the distributional derivative is so that our definition is compatible with integration by parts. In particular, if $f$ and $f^{\prime}$ are tempered functions, then the (formal) integration by parts calculation at the beginning of this section is valid, and shows that the distributional derivative $\psi \rightarrow\langle f, \psi\rangle$ is $\psi \rightarrow\left\langle f^{\prime}, \psi\right\rangle$.

Proposition 25.56 (The Heaviside function). Let $H$ be the Heaviside function $H(x)=$ $\mathbf{1}_{[0, \infty)}$. Then $H^{\prime}=\delta$ in the sense of distributions.

Proof. Let $\psi \in \mathscr{S} . H$ is a tempered function, so $H^{\prime}$ is given by

$$
\begin{aligned}
\left\langle H^{\prime}, \psi\right\rangle & =-\left\langle H, \psi^{\prime}\right\rangle \\
& =-\int_{-\infty}^{\infty} H(x) \frac{d \psi}{d x} d x \\
& =-\int_{0}^{\infty} \frac{d \psi}{d x} d x \\
& =\psi(0) \\
& =\langle\delta, \psi\rangle
\end{aligned}
$$

Notice that every distribution is infinitely differentiable in the sense of distributions. So, we can take another derivative to get $H^{\prime \prime}=\delta^{\prime}$. A quick computation shows that $\left\langle\delta^{\prime}, \psi\left\langle=-\psi^{\prime}(0)\right.\right.$. It can be shown that $\delta^{\prime}$ is not given by any tempered measure (see Problem $\frac{\text { prob: eval-derivative }}{25.22}$

Our next use of Proposition $\frac{\text { prop:T-on-distributions }}{25.53 \text { will allow us to define the convolution of a distri- }}$ bution with a Schwartz function $\phi$.

Proposition 25.57. Let $\phi \in \mathscr{S}$ and $F \in \mathscr{S}^{\prime}$. Then the map

$$
\langle\phi * F, \psi\rangle:=\langle F, \tilde{\phi} * \psi\rangle
$$

defines a tempered distribution, called the convolution of $F$ and $\phi .(\operatorname{Here} \tilde{\phi}(x)=\phi(-x)$.

Proof. Again it suffices to verify that the map $\psi \rightarrow \tilde{\phi} * \psi$ is continuous on $\mathscr{S}$. The proof is left as Problem $\frac{\text { prob:schwartz-convolution }}{25.24 \text {. }}$

If $f \in L^{1}$, one can also verify that $f * \phi$, viewed as a distribution, agrees with the distribution induced by the $L^{1}$ function $f * \phi$ defined by ordinary convolution (Problem $\frac{\text { prob:sc }}{25.24) .}$

It is instructive to revisit $L^{1}$ approximate units in the context of distributions. If $\left\{\phi_{\lambda}\right\}_{\lambda>0}$ is an $L^{1}$ approximate unit, then each $\phi_{\lambda}$ is a tempered function and hence defines a distribution.

Proposition 25.58. If $\phi_{\lambda}$ is an $L^{1}$ approximate unit, then $\phi_{\lambda} \rightarrow \delta$ in $\mathscr{S}^{\prime}$.

Proof. By definition, for any $\psi \in \mathscr{S}$

$$
\left\langle\phi_{\lambda}, \psi\right\rangle=\int_{-\infty}^{\infty} \phi_{\lambda}(y) \psi(y) d y=\left(\phi_{\lambda} * \tilde{\psi}\right)(0)
$$

where $\tilde{\psi}(x)=\psi(-x)$. As $\lambda \rightarrow 0$, by Lemma 1em:continuity-approx=unit 25.18 we have $\left(\phi_{\lambda} * \psi\right)(0) \rightarrow \psi(0)=$ $\langle\delta, \psi\rangle$.

Finally we consider the Fourier transform. The key fact is the following:
r-schwartz Lemma 25.59. If $\psi \in \mathscr{S}$, then $\widehat{\psi} \in \mathscr{S}$, and the map $\mathfrak{\imath} \rightarrow \mathscr{S}$ is continuous.

Proof. The fact that $\widehat{\psi}$ belongs to $\mathscr{S}$ follows from repeated application of Propositions $\frac{\text { prop:mult } 25.6 \text { and } 25.8 \text {. Continuity }}{}$ follows from the fact that if $\psi_{n} \rightarrow \psi$ in $\mathscr{S}$, then also $\left\|x^{\alpha} \frac{d^{\beta}}{d x^{\beta}}\left(\psi_{n}-\psi\right)\right\|_{1} \rightarrow 0$ for all $\alpha, \beta$. See Problem $\begin{aligned} & \text { prob:fourier-schwartz } \\ & 25.23\end{aligned}$

Proposition 25.60 (Fourier transforms of tempered distributions). If $\psi \in \mathscr{S}$, then $\widehat{\psi} \in \mathscr{S}$, and for any $F \in \mathscr{S}^{\prime}$ the formula

$$
\langle\widehat{F}, \psi\rangle:=\langle F, \widehat{\psi}\rangle
$$

defines a tempered distribution, called the Fourier transform of $F$.

Example 25.61. a) Let $\delta_{t}$ be the point mass at $t \in \mathbb{R}$. We can compute $\widehat{\delta_{t}}$ : for $\psi \in \mathscr{S}$ we have

$$
\begin{align*}
\left\langle\widehat{\delta}_{t}, \psi\right\rangle & =\left\langle\delta_{t}, \widehat{\psi}\right\rangle  \tag{85}\\
& =\widehat{\psi}(t)  \tag{86}\\
& =\int_{-\infty}^{\infty} e^{-2 \pi i x t} \psi(x) d x  \tag{87}\\
& =\left\langle e^{-2 \pi i x t}, \psi\right\rangle \tag{88}
\end{align*}
$$

so $\widehat{\delta_{t}}=e^{-2 \pi i x t}$.
The expected inverstion $\left(\widehat{e^{-2 \pi i x t}}\right)=\delta_{t}$ also holds; the proof is left as an exercise.
b) Consider the distribution $P_{1 / x}$ of Proposition 25.54 prop:pv-distribution . One can show that $\widehat{P_{1 / x}}$ is the tempered distribution given by the tempered function

$$
F(t)=-\pi i \operatorname{sgn}(t)
$$

### 25.6. Problems.

Problem 25.1. Prove Proposition $\frac{p r o p: f o u r i e r-b a s i c ~}{25.2}$

Problem 25.2. Complete the proof of Lemma 1em:L1-translation
uous-convo
Problem 25.4. Prove, if $E \subset[0,1]$ has positive Lebesgue measure, then the set

$$
E-E=\{\hat{x}-y: x, y \in E\}
$$

contains an interval centered at the origin. (Hint: let $-E=\{-x: x \in E\}$ consider the function $h(x)=\mathbf{1}_{-E} * \mathbf{1}_{E}$.)

Problem 25.5. Suppose $\phi$ is an unsigned $L^{1}$ function with $\int \phi=1$, and let $\phi_{\lambda}(x)=$ $\frac{1}{\lambda} \phi\left(\frac{x}{\lambda}\right)$.
a) Prove $\left\{\phi_{\lambda}\right\}_{\lambda>0}$ is an $L^{1}$ approximate unit.
 tion (59n).

## h-dense-Lp

Problem 25.6.
a) Prove, if $f \in C_{c}^{1}(\mathbb{R})$ and $g$ is a compactly supported $L^{1}$ function, then $f * g$ is $C^{1}$ with compact support. (Hint: justify differentiation under the integral sign.)
b) By induction, conclude that if $f \in C_{c}^{\infty}(\mathbb{R})$ and $g \in L^{1}$ is compactly supported, then $f * g \in C_{c}^{\infty}(\mathbb{R})$.
c) Conclude that $C_{c}^{\infty}(\mathbb{R})$ is dense in $L^{p}$ for all $1 \leq p<\infty$. (Apply Theorem $\frac{\text { thm:L1-approx-units }}{25.17}$ with $\phi$ a bump function.)
d) Construct a bump function on $\mathbb{R}^{n}$ and extend the above results to $n>1$.
omputation Problem 25.7. Compute the integral in Lemma $\frac{1 \mathrm{em} \text { :poisson-computation }}{25.23 .}$
:FTnotonto Problem 25.8. This problem gives a proof that the Fourier transform ${ }^{\wedge}: L^{1} \rightarrow C_{0}(\mathbb{R})$ is not surjective.
a) Draw a picture of $h_{n}:=\mathbf{1}_{[-n, n]} * \mathbf{1}_{[-1,1]}$ and determine its $C_{0}(\mathbb{R})$ norm.
b) Show that $h_{n}$ is, up to a multiplicative constant independent of $n$, the Fourier transform of the $L^{1}$ function

$$
f_{n}:=\frac{\sin 2 \pi x \sin 2 \pi n x}{x^{2}}
$$

(Hint: you can compute integrals, or use the $L^{1}$ inversion theorem.)
c) Show that $\left\|f_{n}\right\|_{1} \rightarrow \infty$ as $n \rightarrow \infty$. Conclude that the Fourier transform is not surjective. (Hint: if it were surjective... .) The sequence $\left(h_{n}\right)$ is not Cauchy in $C_{0}(\mathbb{R})$ and hence

Problem 25.9. Suppose that $f \in L^{1}, f$ is differentiable a.e., $f^{\prime} \in L^{1}$, and $f(x)=$ $\int_{-\infty}^{x} f^{\prime}(y) d y$ for a.e. $x \in \mathbb{R}$. Prove $\widehat{f}^{\prime}=2 \pi i t \widehat{f}(t)$.

## -inversior

Problem 25.10. Complete the proof of Theorem $\frac{\text { thm:L2-inversion }}{25.30}$

Problem 25.11. Let $\varphi_{\lambda}$ be an $L^{1}(\mathbb{T})$ approximate unit. Prove, if $f \in C(\mathbb{T})$, then $f * \varphi_{\lambda} \rightarrow f$ uniformly as $\lambda \rightarrow 0$.

Problem 25.12. State and prove an analog of Proposition $\frac{\text { prop:fourier-basic }}{25.2 \text { for fourier series. }}$

Problem 25.13. Prove Theorem $\frac{\text { thm:riemann-leb-circle }}{25.33 \text {. }}$
-L2-circle Problem 25.14. Prove Theorem thm:fourier-L2-circle ${ }^{25.42 \text {. Also prove that } L^{2}(\mathbb{T}) \text { inversion is possible in }}$ the following sense: for all $f \in L^{2}(\mathbb{T})$, we have

$$
\lim _{N \rightarrow \infty} \frac{1}{2 \pi} \int_{-\pi}^{\pi}\left|f(\theta) \rightarrow \sum_{n=-N}^{N} \widehat{f}(n) e^{i n \theta}\right|^{2} d \theta=0
$$

(In other words, the partial sums $s_{N}$ of the Fourier series converge to $f$ in the $L^{2}$ norm.)

Problem 25.15. Let $\mathcal{A}(\mathbb{T})$ denote the set of functions $f \in L^{1}(\mathbb{T})$ such that the Fourier transform $\widehat{f}$ belongs to $\ell^{1}(\mathbb{Z})$.
a) Prove, if $f, g \in \mathcal{A}(\mathbb{T})$, then their product $f g$ also belongs to $\mathcal{A}(\mathbb{T})$. (Hint: use the $\ell^{1}$ inversion theorem to write $f$ and $g$ as the sums of their Fourier series, and express the Fourier coefficients of $f g$ in terms of the coefficients of $f$ and $g$.) Thus, $\mathcal{A}(\mathbb{T})$ is a ring.
b) Prove the Fourier transform is a ring isomorphism from $\mathcal{A}(\mathbb{T})$ onto $\ell^{1}(\mathbb{Z})$ (where the multiplication on $\ell^{1}(\mathbb{Z})$ is convolution - see Example $\frac{\text { ex:3banachalgs }}{Z 26.1) \text {. }}$

Problem 25.16. Prove, if $f \in C^{k}(\mathbb{T}), k \geq 1$, then the Fourier coefficients of $f$ satisfy

$$
\lim _{n \rightarrow \pm \infty}|n|^{k}|\widehat{f}(n)|=0
$$

(Hint: first compute the Fourier transform of $f^{\prime}$ explicitly.)
divergence Problem 25.17. Fill in the details in the proof of Theorem $\begin{gathered}\text { thm:fourier-divergence } \\ 25.34 \text {. To show that }\end{gathered}$ $\left\|D_{N}\right\|_{1} \rightarrow \infty$, fix $N$, and for each $0 \leq k \leq 2 N$ let $I_{k}$ denote the interval

$$
\left[\frac{1}{2}\left(\frac{k \pi}{N+\frac{1}{2}}\right), \frac{1}{2}\left(\frac{(k+1) \pi}{N+\frac{1}{2}}\right)\right]
$$

(These are intervals on which $D_{N}$ has constant sign.) Then

$$
\int_{-\pi}^{\pi}\left|D_{N}(t)\right| d t=2 \sum_{k=0}^{2 N} \int_{I_{k}}\left|\frac{\sin \left(N+\frac{1}{2}\right) t}{\sin \frac{t}{2}}\right| d t
$$

To estimate the integral over $I_{k}$, first show that there is a universal constant $C>0$ such that

$$
\left|\frac{1}{\sin \frac{t}{2}}\right| \geq C \frac{n+\frac{1}{2}}{k} \quad \text { for all } t \in I_{k}
$$

for each $k>0$.
ar-fourier Problem 25.18. Prove Proposition prop:weak-star-fourier
Problem 25.18. Prove Proposition 25.41.

Problem 25.19. [Fourier transforms of measures] Let $\mu$ be a finite (signed or complex)
Borel measure on $\mathbb{T}$. The Fourier transform of $\mu$ is the function $\widehat{\mu}: \mathbb{Z} \rightarrow \mathbb{C}$ defined by

$$
\widehat{\mu}(n):=\int_{\mathbb{T}} e^{-i n \theta} d \mu(\theta)
$$

a) Prove $\widehat{\mu}$ is bounded. Give an example of a measure $\mu$ such that $\widehat{\mu} \notin c_{0}(\mathbb{Z})$.
b) For fixed $\mu$, define for each $0 \leq r<1$

$$
A(r, \theta):=\sum_{n=-\infty}^{\infty} \widehat{\mu}(n) r^{|n|} e^{i n \theta}
$$

Prove the measures $\mu_{r}:=A(r, \theta) d m(\theta)$ converge to $\mu$ as $r \rightarrow 1$, in the following sense: for every continuous function $f$ on $\mathbb{T}$,

$$
\lim _{r \rightarrow 1} \int_{\mathbb{T}} f(\theta) A(r, \theta) d m(\theta)=\int_{\mathbb{T}} f(\theta) d \mu(\theta) .
$$

Problem 25.20. [The Fejér kernel] Consider the cutoff function on $\mathbb{Z}$

$$
\psi_{N}(k)= \begin{cases}0 & \text { if }|k|>N \\ 1-\frac{|k|}{N} & \text { if }|k| \leq N\end{cases}
$$

Find a closed form expression for

$$
F_{N}(\theta)=\sum_{k=-N}^{N} \psi_{N}(k) e^{i k \theta}
$$

and show that the family $\left\{F_{N}(\theta)\right\}_{N \geq 1}$ is an $L^{1}(\mathbb{T})$ approximate unit.

Problem 25.21. Prove $\mathscr{S}$ is dense in $L^{p}(\mathbb{R})$ for $1 \leq p<\infty$.
derivative Problem 25.22. Prove there is no finte Borel measure $\mu$ on $[-1,1]$ such that $\int_{-1}^{1} f d \mu=$ $f^{\prime}(0)$ for all $f \in C^{1}[-1,1]$.
c) Prove, if $\phi, \psi \in \mathscr{S}$ then

$$
\int_{\mathbb{R}} \widehat{\phi} \psi=\int_{\mathbb{R}} \phi \widehat{\psi}
$$

(This justifies the definition of $\widehat{F}$.)
b) Prove Fourier inversion holds in $\mathscr{S}$; that is, $(\widehat{\psi})^{-}=\psi$.
c) State and prove a Fourier inversion theorem for the Fourier transform $F \rightarrow \widehat{F}$ on $\mathscr{S}$.
onvolution Problem 25.24. a) Prove, if $\phi \in \mathscr{S}$ and $\psi_{n} \rightarrow \psi$ in $\mathscr{S}$, then $\phi * \psi_{n} \rightarrow \phi * \psi$ in $\mathscr{S}$. (Here convolution means oridnary convolution of functions.)
b) Let $f \in L^{1}$ and $\phi \in \mathscr{S}$. Prove the tempered distribution $f * \phi$ coincides with the distribution defined by the tempered function $f * \phi$.
c) Show that $\delta * \phi=\phi$ in the sense of distributions.
d) Let $\phi \in \mathscr{S}$ be a nonnegative function with $\int \phi=1$, and let $\phi_{\lambda}(x):=\frac{x}{\lambda} \phi\left(\frac{x}{\lambda}\right)$ the corresponding $L^{1}$ approximate unit. Prove for any $F \in \mathscr{S}^{\prime}, F * \phi_{\lambda} \rightarrow F$ in $\mathscr{S}^{\prime}$ as $\lambda \rightarrow 0$.

Problem 25.25. Show if $f: \mathbb{R} \rightarrow \mathbb{R}$ is twice continuously differentiable and has compact support, then $\widehat{f} \in L^{1}(\mathbb{R})$. Now show the Fourier transform $c F 0: L^{1}(\mathbb{R}) \rightarrow C_{0}(\mathbb{R})$ has dense range. (It is not onto by Problem $\frac{\text { prob:FTnotonto }}{25.8 \text {.) }}$

A Banach Algebra $\mathscr{A}$ is a complete normed (associative) algebra over $\mathbb{C}$ such that

$$
\|a b\| \leq\|a\|\|b\|
$$

for $a, b \in \mathscr{A}$ and in the case $\mathscr{A}$ is unital (with unit 1), $\|1\|=1$. (In the unital case there is always an equivalent norm in which the unit has norm 1.)
banachalgs Example 26.1. Here are three examples of unital Banach algebras.
(i) For a Hilbert space $H$, the algebra $B(H)$ is a unital Banach algebra.
(ii) Let $X$ be a compact metric space. The Banach space $C(X)$ is a unital Banach algebra.
(iii) Given $f, g \in \ell^{1}(\mathbb{Z})$ and $n \in \mathbb{Z}$ observe that

$$
\sum_{n=-\infty}^{\infty} \sum_{j=-\infty}^{\infty}|f(j) g(n-j)|=\sum_{j}\left[\sum_{n}|g(n-j)|\right]|f(j)|=\|f\|_{1}\|g\|_{1}
$$

Define $f * g: \mathbb{Z} \rightarrow \mathbb{C}$ by

$$
f * g(n)=\sum_{j=-\infty}^{\infty} f(j) g(n-j)
$$

The computation above shows this sum converges (absolutely) and further that $f * g \in \ell^{1}(\mathbb{Z})$. It also verifies that $\ell^{1}(\mathbb{Z})$ under the multiplication $*$ is a Banach algebra. The operation $*$ is the convolution product on $\ell^{1}(\mathbb{Z})$.

### 26.1. Ideals and quotients.

prop:IQs Proposition 26.2. Suppose $\mathscr{A}$ is a Banach algebra with unit 1 and let $\mathscr{G}$ denote the set of its invertible elements.
it:geom (i) if $\|a\|<1$, then $1-a \in \mathscr{G}$ and moreover,

$$
(1-a)^{-1}=\sum_{j=0}^{\infty} a^{j}
$$

and

$$
\left\|(1-a)^{-1}\right\| \leq \frac{1}{1-\|a\|}
$$

(ii) the set $\mathscr{G}$ is open;
(iii) the mapping $\mathscr{G} \rightarrow \mathscr{G}$ given by $a \rightarrow a^{-1}$ is continuous;
(iv) if $\mathscr{I} \subset \mathscr{A}$ is a proper ideal, then so is its closure;
(v) if $\mathscr{M} \subset \mathscr{A}$ is a maximal ideal, then $\mathscr{M}$ is closed;
(vi) every proper ideal is contained in a maximal ideal;
(vii) if $\mathscr{I}$ is a closed ideal, then the quotient $\mathscr{A} / \mathscr{I}$ is a Banach algebra with unit and is commutative if $\mathscr{A}$ is.

Proof. If $\|a\|<1$, then, using $\left\|a^{j}\right\| \leq\|a\|^{j}$, it follows that the partial sums

$$
s_{n}=\sum_{j=0}^{n} a_{j}
$$

are Cauchy and hence converge (as $\mathscr{A}$ is complete) to some $s$. Since

$$
(1-a) s_{n}=1-a_{n+1}
$$

and the right hands side tends to 1 ,

$$
(1-a) s=1
$$

Moreover, $\left\|(1-a)^{-1}\right\|<\frac{1}{1-\|a\|}$.
Suppose $a$ is invertible. Given $c$ such that $\|c-a\|<\frac{1}{2}\left\|a^{-1}\right\|^{-1}$,

$$
\left\|c a^{-1}-1\right\|=\left\|(c-a) a^{-1}\right\|<\frac{1}{2}
$$

and therefore $1-\left(1-c a^{-1}\right)=c a^{-1}$ is invertible and, thus so is $c$ and we have shown $\mathscr{G}$ is open. For future reference, note

$$
\left\|a c^{-1}\right\| \leq \frac{1}{1-\left\|1-c a^{-1}\right\|}<2
$$

Hence,

$$
\left\|c^{-1}\right\|=\left\|a^{-1} a c^{-1}\right\| \leq 2\left\|a^{-1}\right\|
$$

As for continuity, if $\|c-a\|<\frac{1}{2}\left\|a^{-1}\right\|^{-1}$, then

$$
\left\|a^{-1}-c^{-1}\right\|=\left\|a^{-1}(c-a) c^{-1}\right\| \leq\left\|a^{-1}\right\|\left\|c^{-1}\right\|\|c-a\|<2\left\|a^{-1}\right\|\|c-a\|
$$

Let $\mathscr{G}$ denote the set of invertible elements of $\mathscr{A}$. Thus $\mathscr{G}$ is open. If $\mathscr{I}$ is a proper ideal, then $\mathscr{I} \subset \mathscr{G}^{c}$ and hence $\overline{\mathscr{I}} \subset \mathscr{G}^{c}$. An easy argument, based upon continuity of multiplication and addition, shows $\overline{\mathscr{I}}$ is an ideal and thus a proper ideal.

Let $\pi: \mathscr{A} \rightarrow \mathscr{A} / \mathscr{I}$ denote the quotient map. That $\mathscr{A} / \mathscr{I}$ is again a Banach space with the quotient norm

$$
\|\pi(a)\|=\inf \{\|a-b\|: b \in \mathscr{I} \|\}=\inf \{\|x\|: \pi(x)=\pi(a)\} .
$$

has long since been proved. What remains to show is that the quotient norm is a Banach algebra norm. To this end, let $a, b \in \mathscr{A}$ and $\epsilon>0$ be given. There exists $a^{\prime}, b^{\prime} \in \mathscr{A}$ such
that $\pi\left(a^{\prime}\right)=a, \pi\left(b^{\prime}\right)=b$ and

$$
\left\|a^{\prime}\right\|<\|\pi(a)\|+\epsilon,\left\|b^{\prime}\right\|<\pi(b)+\epsilon .
$$

Thus, $\pi\left(a^{\prime} b^{\prime}\right)=\pi(a b)$ and

$$
\|\pi(a b)\| \leq\left\|\left(a^{\prime} b^{\prime}\right)\right\| \leq(\|\pi(a)\|+\epsilon)(\|\pi(b)\|+\epsilon)
$$

Thus multiplication in the quotient is sub-multiplicative. In particular, $0 \neq\|\pi(1)\|=$ $\left\|\pi(1)^{2}\right\| \leq\|\pi(1)\|^{2}$ and therefore $\|\pi(1)\| \geq 1$. On the other hand, by the definition of the quotient norm, $\|\pi(1)\| \leq\|1\|=1$. Alternately, note if $\|1-m\|<1$, then $m=1-(1-m)$ is invertible and hence $m$ does not lie in a proper ideal
26.2. The spectrum. Suppose $\mathscr{A}$ is a Banach algebra with unit. The spectrum of an element $a \in \mathscr{A}$, denoted $\sigma(a)$, is given by

$$
\sigma(a)=\sigma_{A}(a)=\{z \in \mathbb{C}: a-z 1 \text { is not invertible }\} .
$$

The complement of $\sigma$ is the resolvent set. Often, we write $a-z$ instead of $a-z 1$.

Theorem 26.3. The spectrum $\sigma(a)$ is not empty and compact.

Proof. Arguing by contradiction, suppose $a-z 1$ is invertible for all $z \in \mathbb{C}$. Define $f: \mathbb{C} \rightarrow \mathscr{A}$ by $f(z)=(a-z 1)^{-1}$. An easy argument using item (itii) of Proposition $\frac{\text { prop }}{26} \cdot 2$ shows $f$ is continuous and

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

and therefore $f$ is bounded. Given a linear function $\lambda \in \mathscr{A}^{*}$, let $g: \mathbb{C} \rightarrow \mathbb{C}$ denote the function

$$
g(z)=\lambda(f(z))
$$

It follows that $g$ is bounded and continuous. Further, for $z \neq w \in \mathbb{C}$

$$
\frac{f(w)-f(z)}{w-z}=\frac{(a-w)^{-1}-(a-z)^{-1}}{w-z}=(a-w)^{-1}(a-z)^{-1}=
$$

Hence, for $w \neq z$,

$$
\frac{g(w)-g(z)}{w-z}=\lambda\left((a-w)^{-1}(a-z)^{-1}\right) .
$$

Fixing $z$ and using continuity of the maps $w \mapsto a-w$ and of $b \mapsto b^{-1}$ as well as that of $\lambda$, it follows that $g$ is differentiable at $z$ and

$$
g^{\prime}(z)=\lambda\left((a-z)^{-2}\right)
$$

Thus $g$ is entire and bounded and therefore constant by Liouville's Theorem. Now $g$ vanishes at $\infty$ and hence is identically zero. Choosing, by Corollary $\frac{\text { cor: hb-cor }}{21.10, \lambda \text { such }}$ that $\lambda\left(a^{-1}\right) \neq 0$ gives a contradiction since $g(0)=\lambda(f(0))=\lambda\left(a^{-1}\right)$. Thus $\sigma(a)$ is not empty.

For $|z|>\|a\|,(a-z)$ is invertible. Thus $\sigma(a)$ is a bounded set. (In fact a subset of $\{z:|z| \leq\|a\|\}$.$) The set \mathscr{G}$ of invertible elements is open and the mapping $h: \mathbb{C} \rightarrow \mathscr{A}$ defined by $h(z)=a-z$ is continuous. Hence $O=h^{-1}(\mathscr{G})$ is open and therefore $\sigma(a)=O^{c}$ is closed.

### 26.3. Commutative Banach algebras.

prop:max Proposition 26.4. Suppose $\mathscr{A}$ is a unital commutative Banach algebra and $\mathscr{M} \subset \mathscr{A}$. If $\mathscr{M}$ is a maximal ideal, then $\mathscr{Q}=\mathscr{A} / \mathscr{M}$ has no proper (non-zero) ideals and thus each non-zero element of $\mathscr{Q}$ is invertible.

Proof. Let $\pi: \mathscr{A} \rightarrow \mathscr{A} / \mathscr{M}$ denote the quotient map. If $\mathscr{I} \subset \mathscr{A} / \mathscr{M}$ is a proper ideal, then $\pi^{-1}(\mathscr{I}) \subsetneq \mathscr{A}$ is an ideal containing $\mathscr{M}$. It is also proper as $\pi$ is onto. By maximality, $\pi^{-1}(\mathscr{I})=\mathscr{M}$ and hence, as $\pi$ is onto, $I=\pi\left(\pi^{-1}(I)\right)=\pi(\mathscr{M})=(0) \subset \mathscr{Q}$. Thus $\mathscr{Q}$ has no non-trivial proper ideals. If $q \in \mathscr{Q}$, then $q \mathscr{Q}$ is an ideal that is not proper; i.e., $q \mathscr{Q}=\mathscr{Q}$ and hence $q$ is invertible.
prop:BM Proposition 26.5 (Banach-Mazur). Suppose $\mathscr{A}$ is a unital Banach algebra. If each non-zero element of $\mathscr{A}$ is invertible, then $\mathscr{A}$ is $\mathbb{C}$; i.e., letting 1 denote the unit, $\mathscr{A}=$ $\{z 1: z \in \mathbb{C}\}$ isometrically isomorphically. In particular, if $\mathscr{M}$ is a maximal ideal, then $\mathscr{A} / \mathscr{M}=\mathbb{C}$.

Proof. Let $a \in \mathscr{A}$ be given. Since $\sigma(a) \neq \emptyset$, there is a $\lambda$ such that $a-\lambda$ is not invertible. Hence $a-\lambda=0$; that is, $a=\lambda 1$.

Proposition 26.6. Suppose $\mathscr{A}$ is a commutative Banach algebra with unit. If $a \in \mathscr{A}$ is not invertible, then there is a unital homomorphism $\varphi: \mathscr{A} \rightarrow \mathbb{C}$ such that $\varphi(a)=0$. In particular, $a \in \mathscr{A}$ is invertible if and only if $\varphi(a) \neq 0$ for every unital homomorphism $\varphi: \mathscr{A} \rightarrow \mathbb{C}$.

Proof. If $a$ is not invertible, then $\mathscr{I}=a \mathscr{A}$ is a proper ideal (it contains no invertible elements). Thus $a$ is contained in a maximal ideal $\mathscr{M}$. Let $\varphi: \mathscr{A} \rightarrow \mathscr{A} / \mathscr{M}=\mathbb{C}$ denote the quotient map. It is a unital homomorphism and $\varphi(a)=0$.
remaxideal Proposition 26.7. Suppose $\mathscr{A}$ is a commutative Banach algebra with unit. If $\mathscr{M}$ is a maximal ideal, then there is a unique unital homomorphism $\varphi: \mathscr{A} \rightarrow \mathbb{C}$ such that $\operatorname{ker}(\varphi)=\mathscr{M}$. Conversely, if $\varphi: \mathscr{A} \rightarrow \mathbb{C}$ is a unital homomorphism, then $\operatorname{ker}(\varphi)$ is a maximal ideal. In particular unital homomorphisms $\varphi: \mathscr{A} \rightarrow \mathbb{C}$ are continuous with $\|\varphi\|=1$ and there is a bijective correspondence between unital homomorphisms and maximal ideals.

Proof. Suppose $\mathscr{M}$ is a maximal ideal. By Propositions $\frac{p r o p: m a x}{26.4 \text { and } \frac{p r o p: B M}{26.5, ~ t h e r e ~ e x i s t s ~ a ~}}$ unital homomorphism $\varphi: \mathscr{A} \rightarrow \mathbb{C}$ with kernel $\mathscr{M}$. Suppose now that Suppose $\psi$ : $\mathscr{A} \rightarrow \mathbb{C}$ is a unital homomorphisms with the same kernel $\mathscr{M}$. If $\varphi(a)=0$, then $a \in \mathscr{M}$ and hence $\psi(a)=0$ too. Given $a \in \mathscr{A}$, since $\varphi(\varphi(a) 1-a)=0$, it follows that $0=\psi(\varphi(a) 1-a)=\varphi(a)-\psi(a)$. Hence $\psi=\varphi$ and $\varphi$ is unique.

Now suppose $\varphi: \mathscr{A} \rightarrow \mathbb{C}$ is a unital homomorphism and let $\mathscr{M}$ denote its kernel. Thus $\mathscr{M}$ is an ideal and there is a unital homomorphism $\rho: \mathscr{Q}=\mathscr{A} / \mathscr{M} \rightarrow \mathbb{C}$ such that $\varphi(a)=\rho(\pi(a))$ where $\pi$ is the quotient map. It follows that $\mathscr{Q}$ has no ideals and therefore $\mathscr{M}$ is a maximal ideal. Finally $\varphi=\tau \circ \pi$ where $\tau$ is the isometric isomorphism identifying $\mathscr{A} / \mathscr{M}$ with $\mathbb{C}$ since both are unital homomorphisms with kernel $\mathscr{M}$. Hence $\|\varphi\|=1$.
26.4. The $\frac{1}{f}$ theorem. The Wiener algebra $\mathscr{W}(\mathbb{T})$ consists of those functions $f \in L^{1}(\mathbb{T})$ such that $\hat{f} \in \ell^{1}(\mathbb{Z})$.

Proposition 26.8. $\mathscr{W}(\mathbb{T})$ is a subalgebra of $C(\mathbb{T})$ and, for $f, g \in \mathscr{W}(\mathbb{T})$,

$$
\widehat{f g}=\hat{f} * \hat{g}
$$

In particular, $\mathscr{W}(\mathbb{T})$ is a Banach algebra when given the norm $\|f\|=\|\hat{f}\|_{1}$ and this Banach algebra is isometrically isomorphic to $\ell^{1}(\mathbb{Z})$ under the mapping $f \mapsto \hat{f}$.

Proof. Suppose $f, g \in \mathscr{W}(\mathbb{T})$. In particular, $\hat{f}, \hat{g} \in \ell^{1}(\mathbb{Z})$ and recalling Example $\frac{\text { ex: } 26 \text { banachalgs }}{26.1, \text { for }}$ each $n$ the sum
converges and

$$
\hat{f} * \hat{g}(n)=\sum_{j=-\infty}^{\infty} \hat{f}(j) \hat{g}(n-j)
$$

$$
\sum_{n} \sum_{j=-\infty}^{\infty}|\hat{f}(j) \hat{g}(n-j)|=\|\hat{f}\|_{1}\|\hat{g}\|_{1}
$$

so that $\hat{f} * \hat{g} \in \ell^{1}(\mathbb{Z})$.

By $L^{1}-\ell^{1}$ inversion,

$$
f=\sum_{n=-\infty}^{\infty} \hat{f}(n) e^{i n t}, \quad g=\sum_{n=-\infty}^{\infty} \hat{g}(n) e^{i n t}
$$

with the series converging pointwise absolutely and uniformly by $L^{1}-\ell^{1}$ inversion. In particular, both $f$ and $g$ are continuous and thus $f g$ is continuous and in $L^{1}$. The sequence

$$
s_{k}=\sum_{j=-k}^{k} \hat{f}(j) e^{i j t}
$$

converges uniformly to $f$ and

$$
\begin{aligned}
s_{k} g & =\sum_{j=-k}^{k} \hat{f}(j) e^{i j t} \sum_{n=-\infty}^{\infty} \hat{g}(n) e^{i n t} \\
& =\sum_{m=-\infty}^{\infty}\left[\sum_{j=-k}^{k} \hat{f}(j) \hat{g}(m-j)\right] e^{i m t}
\end{aligned}
$$

converges uniformly to $f g$. Hence,

$$
\widehat{f g}(n)=\lim _{k \rightarrow \infty}\left[\sum_{j=-k}^{k} \hat{f}(j) \hat{g}(n-j)\right]=\hat{f} * \hat{g}(n)
$$

Thus $\widehat{f g}=\hat{f} * \hat{g} \in \ell^{1}(\mathbb{Z})$. Thus $f g \in \mathscr{W}$ and the mapping $f \rightarrow \hat{f}$ is a ring homomorphism that, as an exercise shows, is bijective.

Proposition 26.9. If $\varphi: \mathscr{W}(\mathbb{T}) \rightarrow \mathbb{C}$ is a homomorphism, then there is a $\theta$ such that

$$
\varphi(f)=f\left(e^{i \theta}\right)
$$

Proof. Let $\iota: \mathbb{T} \rightarrow \mathbb{T}$ denote the identity function, $\iota\left(e^{i x}\right)=e^{i x}$. Let $w=\varphi(\iota)$. For $|z|<1$, the function $(1-z \iota)$ is invertible and hence so is $\varphi=1-z w$. It follows that $|w| \leq 1$. The same argument with $\iota^{-1}$ in place of $\iota$ shows $|w|=1$. Hence $w=e^{i \theta}$ for some $\theta$. For a trigonometric polynomial $p\left(e^{i x}\right)=\sum_{-N}^{N} p_{j} e^{i j x}$, we have $\varphi(p)=p\left(e^{i \theta}\right)$ and since polynomials are dense in $\mathscr{W}(\mathbb{T})$, the proof is complete.

Theorem 26.10. If $f \in \mathscr{W}(\mathbb{T})$ and if $f$ is never 0 , then $\frac{1}{f} \in \mathscr{W}(\mathbb{T})$; i.e., an element $g \in \ell^{1}(\mathbb{Z})$ is invertible if and only if the function $f=\sum g_{j} e^{i j t}$ has no zeros.

Proof. An element $f \in \mathscr{W}(\mathbb{T})$ is invertible if and only if $\varphi(f) \neq 0$ for every homomorphism $\varphi: \mathscr{W}(\mathbb{T}) \rightarrow \mathbb{C}$. Each such homomorphism is evaluation is of the form $e^{i \theta}$ (for
some real $\theta$ ); i.e., $\varphi(f)=f\left(e^{i \theta}\right)$ and since, by assumption, this value is not 0 , the proof is complete.
26.5. The Weak-star topology and Gelfand Transform. Let $\mathcal{X}$ denote a Banach space with its dual $\mathcal{X}^{*}$. Given $x \in \mathcal{X}$, let $p_{x}: \mathcal{X}^{*} \rightarrow \mathbb{C}$ denote the (linear) mapping $p_{x}(\lambda)=\lambda(x)$. The weak-star topology on $\mathcal{X}^{*}$ is the topology generated by the family of functions $\left\{p_{x}: x \in \mathcal{X}\right\}$. The (closed) unit ball $\mathcal{X}_{1}^{*}$ of $\mathcal{X}^{*}$ is the set

$$
\mathcal{X}_{1}^{*}=\left\{\lambda \in \mathcal{X}^{*}:\|\lambda\| \leq 1\right\} .
$$

Theorem 26.11 (Banach-Alaoglu). The unit ball $\mathcal{X}_{1}^{*}$ is compact in the weak-star topology.

Proof sketch. For $x \in \mathcal{X}$, let $D_{x}=\{z \in \mathbb{C}:|z| \leq\|x\|\}$ and let $P=\prod_{x \in \mathcal{X}} D_{x}$ in the product topology (the topology generated by the canonical projections). The Tychonoff's Theorem says $P$ is compact. The space $P$ can be identified with functions $g: X \rightarrow \mathbb{C}$ satisfying

$$
|g(x)| \leq\|x\|
$$

Define $\tau: \mathcal{X}_{1}^{*} \rightarrow P$ by

$$
\tau(\lambda)(x)=\lambda(x)
$$

(Thus $\tau$ is simply identifies, in the canonical fashion, $\mathcal{X}_{1}^{*}$ as a subset of $P$.)

Evidently $\tau$ is injective.

Suppose $f$ is in the (product topology) closure of $\mathcal{X}_{1}^{*} \subset P$. Given $\epsilon>0$ and $x, y \in \mathcal{X}$ the set

$$
\{g \in P:|g(x)-f(x)|,|g(y)-f(y)||g(x+y)-f(x+y)|<\epsilon\}
$$

is a neighborhood of $f$ in the product topology. Hence there is a $\lambda \in \mathcal{X}_{1}^{*}$ in this neighborhood. Since $\lambda$ is linear, it follows that

$$
|f(x+y)-f(x)-f(y)|<3 \epsilon
$$

Thus $f$ preserves sums. Similarly $f$ preserves scalar multiplication. Hence $f$ is linear. A similar argument shows $\|f\| \leq 1$. Hence $f \in \mathcal{X}_{1}^{*}$. Thus $R=\tau\left(\mathcal{X}_{1}^{*}\right)$ is closed in the product topology and therefore compact. Now view $\tau: \mathcal{X}_{1}^{*} \rightarrow R$. A subbasic open set in $\mathcal{X}_{1}^{*}$ takes the form $p_{x}^{-1}\left(\{z \in \mathbb{C}:|z-w|<\delta\}=\left\{\lambda \in \mathcal{X}_{1}^{*}:|\lambda(x)-w|<\delta\right\}\right.$ for some $x \in \mathcal{X}, w \in \mathbb{C}$ and $\delta>0$. Its preimage under $\tau^{-1}$ (so image under $\tau$ ) is the (relatively) open set

$$
\tau\left(\mathcal{X}_{1}^{*}\right) \cap\{f \in P:|f(x)-w|<\delta\}
$$

Hence $\tau^{-1}$ is continuous and since the continuous image of a compact set is compact, $\tau^{-1}(R)=\mathcal{X}_{1}^{*}$ is compact.

Let $\mathscr{A}$ denote a commutative Banach algebra with unit. The maximal ideal space $\Sigma=\Sigma_{\mathscr{A}}$ of $\mathscr{A}$ is

$$
\Sigma=\{\varphi: \mathscr{A} \rightarrow \mathbb{C}: \varphi \text { is a homomorphism }\} \subset \mathscr{A}_{1}^{*}
$$

given the relative weak-star topology. (We have used Proposition $\frac{\text { prop: premaxideal } 26.7 \text { that each }\|\varphi\|=}{\varphi}$ 1.)

Proposition 26.12. $\Sigma$ is a compact Hausdorff space.

Proof. It suffices to show that $\Sigma$ is closed (in the weak-star topology); that is, if $f$ is in the closure of $\Sigma$, then $f$ is multiplicative and unital. Given $a, b \in \mathscr{A}$ and $\epsilon>0$, the set

$$
\left\{h \in \mathscr{A}^{*}:|h(a)-f(a)|,|h(b)-f(b)|,|h(a b)-f(a b)|<\epsilon\right\}
$$

is a neighborhood of $f$. Hence it contains some $h \in \Sigma$. It follows that

$$
\begin{aligned}
|f(a b)-f(a) f(b)| & \leq|f(a b)-h(a b)|+|h(a)||h(b)-f(b)|+|f(b)||h(a)-f(a)| \\
& <(1+|h(a)|\|a\|+|f(b)|\|b\|) \epsilon
\end{aligned}
$$

Since $a, b$ are fixed and $\epsilon>0$ is arbitrary, $f(a b)-f(a) f(b)=0$. Thus $f$ is multiplicative.

A proof that $\Sigma$ is Hausdorff is left to the gentle reader.

Proposition 26.13. Given $a \in \mathscr{A}$ the function $\hat{a}: \Sigma \rightarrow \mathbb{C}$ defined by $\hat{a}(\varphi)=\varphi(a)$ is continuous. Moreover, the mapping $\mathscr{G}: \mathscr{A} \rightarrow C(\Sigma)$ defined by $\mathscr{G}(a)=\hat{a}$ is a continuous unital homomorphism of norm one. In particular,

$$
\widehat{p(a)}(\varphi)=\varphi(p(a))=p(\varphi(a))
$$

for $\varphi \in \Sigma$.

The mapping $\mathscr{G}$ is the Gelfand Transform.

Proof. The weak-star topology is defined exactly to make each $\mathscr{G}(a)$ continuous. Since $\varphi \in \Sigma$ has norm one, $\|\mathscr{G}(a)\| \leq\|a\|$. On the other hand, $\mathscr{G}(1)$ is the function that is identically equal to 1 and thus $\|\mathscr{G}(1)\|=1$. Thus $\|\mathscr{G}\|=1$. An easy exercise shows $\mathscr{G}$ is multiplicative and hence a unital homomorphism.
:sigma-gel Proposition 26.14. For $a \in \mathscr{A}$,

$$
\sigma(a)=\{\varphi(a): \varphi \in \Sigma\} .
$$

Moreover, if $\mathscr{A}$ is generated by a, then $\mathscr{G}(a): \Sigma \rightarrow \sigma(a)$ is a homeomorphism.

Proof. The main part of the proposition is a restatement of Proposition $\frac{\text { prop:premaxideal }}{26.7 . \text { For the }}$ moreover, the function $\mathscr{G}(a)$ is continuous with compact domain. It is onto, since $\sigma(a)=$ $\{\varphi(a): \varphi \in \Sigma\}$. Thus $\mathscr{G}(a)$ is a homeomorphism.
26.6. The spectral radius formula. Given $\mathscr{A}$ a unital Banach algebra and $a \in \mathscr{A}$, the spectral radius of $a$ is

$$
r(a)=\max \{|z|: z \in \sigma(a)\} .
$$

Proposition 26.15. $r(a)=\lim _{n \rightarrow \infty}\left\|a^{n}\right\|^{\frac{1}{n}}$.

Proof. First observe, for $m \in \mathbb{N}^{+}$and $z \in \mathbb{C}$,

$$
z^{m+1}-a^{m+1}=(z-a) \sum_{j=0}^{m} z^{j} a^{m-j}
$$

Thus, if $z-a$ is not invertible, then neither is $z^{m+1}-a^{m+1}$. Consequently, $\left|z^{m+1}\right| \leq$ $\left\|a^{m+1}\right\|$ and therefore $|z| \leq \lim \inf \left\|a^{n}\right\|^{\frac{1}{n}}$. Hence, if $z \in \sigma(a)$, then $|z| \leq \lim \inf \left\|a^{n}\right\|^{\frac{1}{n}}$.

Given $\rho \in \mathscr{A}^{*}$, consider the function

$$
g(z)=\rho\left((1-z a)^{-1}\right) .
$$

It is analytic on the set $\left\{z \in \mathbb{C}:|z|<\frac{1}{r(a)}\right\}$.

For $|z|<\frac{1}{\|a\|}$, the function $g$ has the power series expansion,

$$
g(z)=\sum \rho\left(a^{n}\right) z^{n}
$$

Hence this series has radius of convergence at least $\frac{1}{r(a)}$ and hence, if $|z|<\frac{1}{r(a)}$, then the set $\left\{\rho\left(a^{n}\right) z^{n}\right\}$ is bounded. Fix $|z|<\frac{1}{r(a)}$, view $f_{n}=z^{n} a^{n} \in \mathscr{A}$ as an element of $\mathscr{A}^{* *}$ so that $f_{n}(\rho)=\rho\left(f_{n}\right)=z^{n} \rho\left(a^{n}\right)$ for $\rho \in \mathscr{A}^{*}$. For each $\rho \in \mathscr{A}^{*}$ there is a constant $C_{\rho}$ such that

$$
\left|f_{n}(\rho)\right|=\left|\rho\left(f_{n}\right)\right| \leq C_{\rho}
$$

Thus $\left(f_{n}\right)$ is a pointwise bounded set of linear functionals on the Banach space $\mathscr{A}^{*}$. By the principle of uniform boundedness, there is a constant $C$ such that $\left\|f_{n}\right\| \leq C$ independent of $n$. Thus $\left\|z^{n} a^{n}\right\| \leq C$ and therefore $\lim \sup \left\|a^{n}\right\|^{\frac{1}{n}} \leq \frac{1}{|z|}$. Hence, by choosing $z$ such that $|z|<\frac{1}{r(a)}$ is close to $\frac{1}{r(a)}$,

$$
\limsup \left\|a^{n}\right\|^{\frac{1}{n}} \leq r(a)
$$

Corollary 26.16. Suppose $\mathscr{A}$ is a commutative Banach algebra with unit and maximal
ideal space $\Sigma$. Given $a \in \mathscr{A}$,

$$
\|\mathscr{G}(a)\|=\lim \left\|a_{n}\right\|^{\frac{1}{n}}=r(a) .
$$

Proof. Recall $\mathscr{G}(a)$ is the function on $\Sigma$ defined by $\mathscr{G}(a)(h)=h(a)$. Hence, in view of Proposition $\frac{\text { prop:sigma-gel }}{26.14, \mid \nmid(a) \|}=\max \{|h(a)|: h \in \Sigma\}=r(a)$.

## 27. C-star algebras and the functional calculus

A C-star algebra $\mathscr{A}$ is Banach algebra with norm $\|\cdot\|$ and an involution $\mathscr{A} \rightarrow \mathscr{A}$, denoted $a \rightarrow a^{*}$, such that for $a, b \in \mathscr{A}$ and $c \in \mathbb{C}$,
(i) $\left(a^{*}\right)^{*}=a$ (involutive);
(ii) $(a b)^{*}=b^{*} a^{*}$ (anti-multiplicative);
(iii) $(c a+b)^{*}=\bar{c} a^{*}+b^{*}=c^{*} a^{*}+b^{*}$ (anti-linear); and
(iv)

$$
\left\|a^{*} a\right\|=\|a\|^{2}
$$

In the case $\mathscr{A}$ has a unit 1 , as with Banach algebras, we assume $\|1\|=1$. Note $1^{*}=1^{*} 1=\left(1^{*} 1\right)^{* *}=\left(1^{*} 1\right)^{*}=\left(1^{*}\right)^{*}=1=1$.

For $X$ a compact metric space (or just a compact Hausdorff space), $C(X)$ is a C-star algebra given the usual norm and involution $f^{*}(z)=\overline{f(z)}=f(z)^{*}$. For $H$ a (complex) Hilbert space, $B(H)$, the bounded operators on $H$, is a C-star algebra given the usual norm and involution $T \mapsto T^{*}$ given by the map taking an operator $T$ to its adjoint,

$$
\langle T x, y\rangle=\left\langle x, T^{*} y\right\rangle, \quad x, y \in H
$$

An element $a \in \mathscr{A}$ is hermitian if $a^{*}=a$.

Proof. Since $a=a^{*}$, we have $\left\|a^{2}\right\|=\left\|a^{*} a\right\|=\|a\|^{2}$. By induction $\left\|a^{2^{n}}\right\|=\|a\|^{2^{n}}$. Hence, by the spectral radius formula,

$$
\|a\|=\lim \left\|a^{2^{n}}\right\| \|^{\frac{1}{2^{n}}}=r(a) .
$$

An element $u \in \mathscr{A}$ is an isometry if $u^{*} u=1$.

Proposition 27.2. Suppose $\mathscr{A}$ is a $C$-star algebra with unit and $h: \mathscr{A} \rightarrow \mathbb{C}$ is a unital homomorphism and $a, u \in \mathscr{A}$.
(1) If $a=a^{*}$, then $h(a) \in \mathbb{R}$;
(2) $h\left(a^{*}\right)=h(a)^{*}(=\overline{h(a)})$;
(3) $h\left(a^{*} a\right) \geq 0$;
(4) If $u$ is an isometry, then $|h(u)|=1$.

In particular, if a is hermitian, then $\sigma(a) \subset \mathbb{R}$; and any homomorphism $h: \mathscr{A} \rightarrow \mathbb{C}$ is actually $a$ *-homomorphism.

Proof. Recall the norm of a unital homomorphism of a Banach algebra is 1. (Identify it with the mapping $\pi: \mathscr{A} \rightarrow \mathscr{A} / \operatorname{ker}(h)=\mathbb{C}$.) Assuming $a=a^{*}$, for $t \in \mathbb{R}$,

$$
|h(a+i t)|^{2} \leq\|a+i t\|^{2}=\left\|(a+i t)^{*}(a+i t)\right\|=\left\|a^{2}+t^{2}\right\| \leq\|a\|^{2}+t^{2} .
$$

Writing $h(a)=\alpha+i \beta$,

$$
\alpha^{2}+(\beta+t)^{2} \leq\|a\|^{2}+t^{2} .
$$

Hence $\beta=0$ and thus $h(a) \in \mathbb{R}$, proving item (i).

To prove item (ii), write $a=x+i y$ where $x, y$ are hermitian (the cartesian decomposition) and apply the first part of the proposition.

For the third part, $h\left(a^{*} a\right)=h\left(a^{*}\right) h(a)=h(a)^{*} h(a)=|h(a)|^{2} \geq 0$.

Finally $1=h(1)=h\left(u^{*} u\right)=h\left(u^{*}\right) h(u)=h(u)^{*} h(u)=1$. Thus $|h(u)|=1$.

Theorem 27.3. If $\mathscr{A}$ is a commutative $C$-star algebra with unit and maximal ideal space $\Sigma$, then the gelfand transform $\mathscr{G}: \mathscr{A} \rightarrow C(\Sigma)$ is an isometric $*$-isomorphism; i.e., $\mathscr{A}$ and $C(\Sigma)$ are equal as $C$-star algebras.

Proof. If $a \in \mathscr{A}$ is hermitian, then $\|a\|=r(a)=\|\mathscr{G}(a)\|$ by Proposition $\frac{\text { prop:normherm }}{27.1 \text { and }}$ Corollary cor: $\frac{\text { chelra }}{26}$
Corollary 26.16.

For $a \in \mathscr{A}$ and $\varphi \in \Sigma$,

$$
\mathscr{G}\left(a^{*}\right)(\varphi)=\varphi\left(a^{*}\right)=\varphi(a)^{*}=\mathscr{G}(a)^{*}(\varphi) .
$$

Thus $\mathscr{G}\left(a^{*}\right)=\mathscr{G}(a)^{*}$ and $\mathscr{G}$ is a $*$-homomorphism. Further, given $a \in \mathscr{A}$, since $a^{*} a$ is hermitian and $\mathscr{G}$ is a $*$-homomorphism,

$$
\|a\|^{2}=\left\|a^{*} a\right\|=\left\|\mathscr{G}\left(a^{*} a\right)\right\|=\left\|\left.\mathscr{G}(a)\right|^{2}\right\|=\|\mathscr{G}(a)\|^{2} .
$$

Hence $\|a\|=\|\mathscr{G}(a)\|$ and $\mathscr{G}$ is an isometry and thus has closed range. Therefore to show $\mathscr{G}$ is onto, it suffices to show it has dense range. Now $\mathscr{G}(\mathscr{A})$ is a subalgebra of $C(\Sigma)$ that is closed under complex conjugation (the involution), contains the constants and
separates points. Hence by Stone-Weierstrass it is dense. Thus $\mathscr{G}$ is a bijective isometric *-isomorphism.
27.1. Normal elements and the functional calculus. Given an element $a$ in a Cstar algebra $\mathscr{A}$, the C-star algebra generated by $a$, denoted $C^{*}(a)$, is the smallest C-star subalgebra of $\mathscr{A}$ containing $a$.

Proposition 27.4 (Spectral Permanence). Let $\mathscr{A}$ be a unital C-star algebra. An $a \in \mathscr{A}$ is invertible in $\mathscr{A}$ if and only if it is invertible in $C^{*}(a)$. In particular, if $\mathscr{B} \subset \mathscr{A}$ is a unital $C$-star algebra and $a \in \mathscr{B}$, then

$$
\sigma_{\mathscr{B}}(a)=\sigma_{\mathscr{A}}(a)
$$

Thus the spectrum of a does not depend upon the C-star algebra.

Proof. First suppose $a$ is hermitian and invertible in $\mathscr{A}$. In this case $a$ is invertible in the commutative C-star algebra $\mathscr{D}$ generated by $a$ and $a^{-1}$, namely the closure of expressions of the form $r(a) \neq \sum_{j=-N}^{N} r_{j} a^{-j}$. The Gelfand transform $\mathscr{G}_{\mathscr{D}}: \mathscr{D} \rightarrow \Sigma_{\mathscr{D}}$ is an isometric $*$-isomorphism. Suppose $\varphi, \psi \in \Sigma_{\mathscr{D}}$ and $\varphi(a)=\psi(a)$. Since $a$ is invertible this common value is not zero and moreover $\varphi\left(a^{-1}\right)=\varphi(a)^{-1}=\psi(a)^{-1}=\psi\left(a^{-1}\right)$. Hence $\varphi=\psi$. Therefore the subalgebra $\mathscr{G}_{\mathscr{D}}\left(C^{*}(a)\right) \subset C\left(\Sigma_{\mathscr{D}}\right)$ separates points. It is also contains the constants, is closed under pointwise conjugation and closed. Hence, by the Stone-Weierstrass Theorem, $\mathscr{G}_{\mathscr{D}}\left(C^{*}(a)\right)=C\left(\Sigma_{\mathscr{D}}\right)$ and therefore $C^{*}(a)=\mathscr{D}$. Hence $a$ is invertible in $C^{*}(a)$.

For the general case, if $a$ is invertible in $\mathscr{A}$ with inverse $b \in \mathscr{A}$, then $a^{*} a$ is invertible in $\mathscr{A}$ with inverse $b b^{*}$. Thus $a^{*} a$ is invertible in $C^{*}\left(a^{*} a\right)$; i.e., $b b^{*} \in C^{*}\left(a^{*} a\right)$. Finally, $b=b 1=b\left(b^{*} a^{*}\right)=\left(b b^{*}\right) a \in C^{*}(a)$.

An element $a$ in a C-star algebra $\mathscr{A}$ is normal if $a^{*} a=a a^{*}$. Assuming $a$ is normal, given a polynomial $p(z, \bar{z})=\sum p_{j, k} z^{j} \bar{z}^{k}$, let

$$
p\left(a, a^{*}\right)=\sum p_{j, k} a^{j} a^{* k}
$$

(since $a$ and $a^{*}$ commute the order of the products are not important). In this case, the C-star algebra generated by $a$ is the closure of $\left\{p\left(a, a^{*}\right): p=p(z, \bar{z})\right.$ a polynomial $\}$. It is commutative.

Proposition 27.5. Let $\mathscr{A}=C^{*}(a)$ be a C-star algebra generated by a normal element a and let $\Sigma$ denote its maximal ideal space. The mapping $\mathscr{G}(a): \Sigma \rightarrow \sigma(a)$ is a homeomorphism. Moreover, if $p=p(z, \bar{z})$ is a polynomial, then $\mathscr{G}\left(p\left(a, a^{*}\right)\right)=p\left(\mathscr{G}(a), \mathscr{G}(a)^{*}\right)$.

Sketch of proof. The moreover part of the Proposition is routine. $\mathscr{G}(a)$ is a continuous function on $\Sigma$ and since $\sigma(a)=\{h(a): h \in \Sigma\}$ its codomain can be taken to be $\sigma(a)$. If $h, g \in \Sigma$ and $h(a)=g(a)$, then, for all polynomials $p\left(z, z^{*}\right)$, we have $h\left(p\left(a, a^{*}\right)\right)=$ $g\left(p\left(a, a^{*}\right)\right)$ by the moreover portion of proposition. Hence $h=g$ and $\mathscr{G}(a)$ is one-one. Since $\mathscr{G}(a)$ is a continuous bijection with compact domain, it is a homeomorphism.

The proof of the following lemma is left as an exercise.

Lemma 27.6. If $X$ and $Y$ are compact Hausdorff spaces and $\tau: X \rightarrow Y$ is a homeomorphism, then $\tau^{*}: C(Y) \rightarrow C(X)$ defined by $\tau^{*}(f)(x)=f(\tau(x))$ is an isometric *-isomorphism.

Given $a \in \mathscr{A}$ normal, let $\mathscr{G}$ denote the Gelfand transform $\mathscr{G}: C^{*}(a) \rightarrow C\left(\Sigma_{C^{*}(a)}\right)$ (where $\Sigma=\Sigma_{C^{*}(a)}$ is the maximal ideal space of $\left.C^{*}(a) \subset \mathscr{A}\right)$. The lemma applies to $\tau=\mathscr{G}(a): \Sigma_{C^{*}(a)} \rightarrow \sigma(a)$. In this case, for $f \in C(\sigma(a))$, we have $\tau^{*}(f)(\varphi)=f(\tau(\varphi))=$ $f(\varphi(a))$. Hence we obtain an isometric $*$-isomorphism $\rho=\mathscr{G}^{-1} \circ \tau^{*}: C(\sigma(a)) \rightarrow$ $C^{*}(a) \subset \mathscr{A}$. This mapping is the functional calculus for $a$. Note that, for a polynomial $p=p(z, \bar{z})$ viewed as an element of $C(\sigma(a))$, that $\rho(p)=p\left(a, a^{*}\right)$ and moreover $\rho$ is the unique isometric $*$-homomorphism from $C(\sigma(a)) \rightarrow \mathscr{A}$ sending $p$ to $p\left(a, a^{*}\right)$ (since polynomials in $z$ and $z^{*}=\bar{z}$ are dense in $C(\sigma(a))$ by Stone-Weierstrass). It is customary to write $f(a)=\rho(f)$.

Proposition 27.7 (Spectral mapping). Suppose $a \in \mathscr{A}$ is normal. For $f \in C(\sigma(a))$,

$$
\sigma(f(a))=f(\sigma(a))
$$

and $\|f(a)\|=\max \{|f(z)|: z \in \sigma(a)\}$.

Proof. Since $\rho$ is a $*$-isomorphism, $f-\lambda$ is invertible if and only if $\rho(f-\lambda)=\rho(f)-\lambda$ is invertible. Hence $f(\sigma(a))=\sigma(f)=\sigma(f(a))$.

An element $u \in \mathscr{A}$ is unitary if $u^{*} u=u u^{*}=1$. In particular $u$ is invertible and normal.
ary-fn-cal Proposition 27.8. If $u$ is unitary, then $\sigma(u) \subset \mathbb{T}=\{|z|=1\}$. In particular, if $p$ is an analytic polynomial $p(z)=\sum p_{j} z^{j}$, then

$$
\|p(u)\|=\max (\{|p(z)|: z \in \sigma(u)\} \leq \max (\{|p(z)|: z \in \mathbb{T}\})=\sup \{|p(z)|:|z|<1\}
$$

Proof. Let $\Sigma$ denote the maximal ideal space of $C^{*}(u)$. In particular each $\varphi \in \Sigma$ is a *-homomorphism. Thus $1=\varphi(1)=\varphi\left(u^{*} u\right)=|\varphi(u)|^{2}$. Hence $|\varphi(u)|=1$ and the proof of the first part is complete. (Or appeal to the earlier result about isometric elements in a C-star algebra.)

The second part (the equality) is simply an instance of the maximum principle.
27.2. Positive and contractive elements in a C-star algebra. An element $a$ in a C-star algebra with unit $\mathscr{A}$ is positive (semidefinite), denoted $a \succeq 0$ if $a=a^{*}$ and $\sigma(a) \subset[0, \infty)$. For instance, if $X$ is a compact Hausdorff space, then an element $f \in$ $C(X)$ is positive if and only if the range of $f$ lies in $[0, \infty)$. Moreover, if $\rho: C(X) \rightarrow \mathscr{A}$ is a $*$-homomorphism and $f \succeq 0$, then $\rho(f) \succeq 0$; and if $\rho$ is an isometric $*$-isomorphism, then $\rho(f) \succeq 0$ if and only if $f \succeq 0$.

Lemma 27.9. Suppose $\mathscr{A}$ is a unital C-star algebra and $a \in \mathscr{A}$ is hermitian.
(i) If $\sigma(a)=\{0\}$, then $a=0$;
(ii) If $\pm a \succeq 0$, then $a=0$;
(iii) If $a^{2}=0$, then $a=0$;
(iv) $a^{2}$ is positive;
(v) There exists positive elements $p, q \in C^{*}(a)$ such that $p q=0$ and $q=p^{2}-q^{2}$. In particular, if $a$ is positive, then $q=0$;
(vi) If $a$ is postive, then each $n \in \mathbb{N}^{+}$there exists a positive $b$ such that $b^{n}=a$;
(vii) If $\|1-a\| \leq 1$, then $a$ is positive;
(viii) If $a$ is positive, then $\|t-a\| \leq t$ for all $t \geq\|a\|$.

Proof. Since $a$ is hermitian, $\|a\|=r(a)$, the spectral radius. Hence, if $\sigma(a)=\{0\}$, then $a=0$. If $\pm a \succeq 0$, then $\pm \sigma(a) \subset[0, \infty)$ in which case $\sigma(a)=\{0\}$ and hence $a=0$. If $a^{2}=0$, then $\sigma\left(a^{2}\right)=\{0\}$ and hence, using spectral mapping $\sigma(a)=\sigma(a)^{2}=\{0\}$. Thus $a=0$. Since $a$ is hermitian, $\sigma(a) \subset \mathbb{R}$. Thus using the functional calculus, $\sigma\left(a^{2}\right)=\sigma(a)^{2} \subset[0, \infty)$. Hence $a$ is positive.

Let $\rho: C(\sigma(a)) \mapsto C^{*}(a)$ denote the functional calculus. Consider the functions $u, v: \mathbb{R} \rightarrow \mathbb{R}$ defined by $u(t)=\max \{0, t\}$ and $v=-\min \{0, t\}$. Thus $u-v=\iota$ is the identity function $\iota(t)=t$ and both $u, v$ are positive functions. Hence $a=\rho(\iota)=$ $\rho(u)-\rho(v)$. Letting $u^{\frac{1}{2}}$ and $v^{\frac{1}{2}}$ denote the (pointwise) positive square roots and $p=\rho\left(u^{\frac{1}{2}}\right)$ and $q=\rho\left(v^{\frac{1}{2}}\right)$, it follows that $a=p^{2}-q^{2}$ and $p q=0$. If $a$ is positive, then $v=0$ and hence $q=0$. In this case, for $n \in \mathbb{N}^{+}$, we have $a=\rho\left(\text { iota }^{\frac{1}{n}}\right)^{n}$.

Using again the functional calculus, $\rho(\iota)=a$, observe that $\|1-\iota\| \leq 1$ implies $|1-\iota| \leq 1$ pointwise. Hence the range $(\iota) \subset[0,2]$ and therefore $\iota$, and hence $a$, is positive. Finally, if $a$ is positive, then, for $t \geq\|a\|$, pointwise $|\iota| \leq\|a\| \leq t$ and therefore $0 \leq t-\iota \leq t$. Hence $\|t-\iota\| \leq t$.

Let $\mathscr{A}_{+}$denote the positive elements in the unital C-star algebra $\mathscr{A}$.

Proposition 27.10. The set $\mathscr{A}_{+}$is a closed cone; i.e., it is a closed set that is closed under addition and multiplication by nonnegative scalars.

Proof. That $\mathscr{A}_{+}$is closed under multiplication by a $\lambda \geq 0$ is evident. Now suppose $a, b \in \mathscr{A}_{+}$. By multiplying by a $\lambda>0$, we may assume that $\|a\|,\|b\| \leq 1$. Thus,

$$
\|2-(a+b)\| \leq[\|1-a\|+\|1-b\|] \leq 2
$$

by the last part of the lemma. Hence, by the next to last part of the lemma, $\frac{a+b}{2}$, and hence, $a$ is positive.

To prove $\mathscr{A}_{+}$is closed, use, $a=a^{*}$, the fact that $\|\|a\|-a\| \leq\|a\|$ if and only if $a \succeq 0$.

Proposition 27.11. For $a \in \mathscr{A}$, a unital C-star algebra, the following are equivalent.
(i) $a$ is positive;
(ii) there is a positive $b \in \mathscr{A}$ such that $a=b^{2}$;
(iii) there is a a hermitian $b \in \mathscr{A}$ such that $a=b^{2}$;
(iv) there is an $x \in \mathscr{A}$ such that $a=x^{*} x$;

Proof. The previous lemma shows (i) and (ii) are equivalent. Evidently (ii) implies (iii) implies (iv).

If (iv) holds with $a=x^{*} x$, then $a=a^{*}$ and, from the lemma, there exists $p, q \in C^{*}(a)$ such that $p, q \succeq 0, p q=0$ and $a=p^{2}-q^{2}$. Let $b+i c$ denote the cartesian decomposition
of $x q$. Thus,

$$
b^{2}+c^{2}+i(b c-c b)=(b-i c)(b+i c)=q x^{*} x q=q a q=q\left(p^{2}-q^{2}\right) q=-q^{4} .
$$

It follows that $\pm q^{4} \succ 0$ and therefore $q^{4}=0$. Hence $q=0$.

Proposition 27.12. Let $H$ be a Hilbert space. An element $P \in B(H)$ is positive if and only if $\langle P x, x\rangle \geq 0$ for all $x \in H$.
(Caution: the condition $P=P^{*}$ needs to be added in the case of real Hilbert space.)

Proof. If $P$ is positive, then $P=X^{*} X$ for some $X \in B(H)$. Hence, for $x \in H$,

$$
\langle P x, x\rangle=\left\langle X^{*} X x, x\right\rangle=\langle X x, X x\rangle \geq 0 .
$$

Conversely, suppose $\langle P x, x\rangle \geq 0$ for all $x \in H$. It follow that

$$
\left\langle\left(P-P^{*}\right) x, x\right\rangle=\langle P x, x\rangle-\langle x, P x\rangle=0
$$

since both terms are real. Hence $P=P^{*}$ and in particular $\sigma(P) \subset \mathbb{R}$. Now suppose $\lambda>0$. For $x \in X$,

$$
\|(P+\lambda) x\|^{2}=\|P x\|^{2}+2\langle P x, x\rangle+\lambda^{2}\|x\|^{2} \geq \lambda^{2}\|x\|^{2}
$$

It follows that $P+\lambda$ has no kernel and the range of $P+\lambda$. Thus range $(P+\lambda)=$ $\operatorname{ker}\left((P+\lambda)^{*}\right)^{\perp}=\operatorname{ker}(P+\lambda)^{\perp}=H$. By the Banach isomorphism theorem, $P+\lambda$ is invertible. Hence $\sigma(P) \subset[0, \infty)$.
27.3. Contractions on Hilbert space and the de Branges-Rovnyak construc-
tion. Let $H$ denote a Hilbert space (over $\mathbb{C}$ ). An operator $C \in B(H)$ is a contraction if $\|C\| \leq 1$.

Proposition 27.13. $C$ is a contraction if and only if $I-C^{*} C \succeq 0$.

Proof. We have $I-C^{*} C \succeq 0$ if and only if for all $x \in H$,

$$
0 \leq\left\langle\left(I-C^{*} C\right) x, x\right\rangle=\langle x, x\rangle-\langle C x, C x\rangle=\|x\|^{2}-\|C x\|^{2}
$$

if and only if $\|C\| \leq 1$.

The operator $I-C^{*} C$ being positive (semidefinite) has a positive square root ( $1-$ $\left.C^{*} C\right)^{\frac{1}{2}}$ denoted $D_{C}$ and referred to as the defect of $T$.

An operator $T$ on a Hilbert space $H$ lifts to an operator $J$ on a Hilbert space $K$ if there exist an isometry $V: H \rightarrow K$ such that

$$
V T=J V
$$

Note, in this case, if $p \in \mathbb{C}[z]$ is a polynomial, then

$$
V p(T)=p(J) V
$$

Hence sometimes a lift is referred to as a $\mathbb{C}[z]$-lift. An operator $T$ on a Hilbert space $H$ is a $C_{0}$-contraction if $\|T\| \leq 1$ and $\lim T^{n} h=0$ for each $h \in H$. The adjoint $S^{*}$ of the shift operator $S$ is an example of a $C_{0}$-contraction.

Given a Hilbert space $E$, let $K_{+}(E)=\oplus_{j=0}^{\infty} E=\ell_{E}^{2}(\mathbb{N})$ consisting of those sequences $\left\{\left(e_{j}\right)=\oplus_{j=0}^{\infty} e_{j}: \sum\left\|e_{j}\right\|^{2}<\infty\right\}$ with the inner product,

$$
\langle e, f\rangle=\left\langle\left(e_{j}\right),\left(f_{k}\right)\right\rangle=\sum\left\langle e_{j}, f_{j}\right\rangle
$$

It is an exercise to verify that $K_{+}$is a Hilbert space. The corresponding shift operator $S_{E}$ is defined by $\left(S_{E} e\right)_{j}=0$ if $j=0$ and $(S e)_{j}=e_{j-1}$ for $j>0$. The multiplicity of $S_{E}$ is the dimension of $E$ and any two shifts of the same multiplicity are unitarily equivalent.

Proposition 27.14. The operator $S_{E}$ is an isometry with $\sigma\left(S_{E}\right)=\overline{\mathbb{D}}$. Its adjoint is given by

$$
S_{E}^{*} \oplus e_{j}=\oplus e_{j+1}
$$

In particular, $S_{E}^{*}$ has kernel $\left\{e: e_{j}=0, j>0\right\}$ (which we identify with $E$ ) and is a $C_{0}$-contraction.

Proof. We will prove the proposition for the case of the shift $S$ of multiplicity one where $E=\mathbb{C}$. The general case is both similar and easily derived from this case. We already know $S$ is an isometry and in particular, $\|S\|=1$. Hence $\sigma(S) \subset \mathbb{D}$. Likewise we have already seen that $S^{*}$ is the backward shift. Given $\lambda \in \mathbb{D}$, let $s_{\lambda}=\left(\bar{\lambda}^{j}\right)_{j=0}^{\infty} \in \ell^{2}(\mathbb{N})$ and observe that $S^{*} s_{\lambda}=\left(\bar{\lambda}^{j+1}\right)_{j=0}^{\infty}=\bar{\lambda} s_{\lambda}$. Hence $S^{*}-\bar{\lambda}$ has a non-trivial kernel and is thus not invertible. Hence $\lambda \in \sigma(S)$ and therefore $\mathbb{D} \subset \sigma(S) \subset \overline{\mathbb{D}}$. Since the spectrum is closed, the result follows.

An operator $T$ is a shift operator if it is unitarily equivalent to $S_{E}$ for some $E$.

Theorem 27.15. An operator $T$ is a $C_{0}$-contraction if and only if $T$ lifts to the adjoint of a shift operator.

Proof. It is straightforward to verify if $T$ lifts to the adjoint of a shift operator, then $T$ is a $C_{0}$-contraction.

The proof of the converse is constructive - the de Branges-Rovnyak construction (a variation of a construction of Rota). Suppose $T$ is a $C_{0}$-contraction. Let $E$ denote the closure of the range of $D_{T}$, the defect of $T$. Define $V: H \rightarrow \ell_{E}^{2}(\mathbb{N})$ by

$$
V h=\oplus D_{T} T^{j} h
$$

A computation shows $V$ is an isometry. Moreover,

$$
S_{E}^{*} V h=S_{E}^{*} \oplus_{j=0}^{\infty} D_{T} T^{j} h=\nsubseteq D_{T} T^{j+1} h=V T h
$$

Hence $S_{E}^{*} \operatorname{lifts} T$.

Given a Hilbert space $E$, let $K=\ell_{E}^{2}(\mathbb{Z})=\oplus_{j=-\infty}^{\infty} E$ and define $U_{E}: K \rightarrow K$ by

$$
U \oplus e_{j}=e_{j+1}
$$

A computation shows $U^{*}=U^{-1}$ is given by

$$
U^{*} \oplus e_{j}=e_{j-1}
$$

In particular $U$ is a unitary operator. An operator $T$ is a bilateral shift if there is a Hilbert space $E$ such that $T$ is unitarily equivalent to $U_{E}$.

Proposition 27.16. The shift operator $S_{E}$ lifts to the bilateral shift operator $U_{E}$. In particular, a shift operator lifts to a bilateral shift (unitary operator).

An operator $T \in B(H)$ dilates to an operator $J \in B(K)$ or has a $\mathbb{C}[z]$-dilation to $J$ if there is an isometry $V: H \rightarrow K$ such that

$$
p(T)=V^{*} p(J) V
$$

for all $p \in \mathbb{C}[z]$. Note that if $T$ lifts to $J$, then $T$ dilates to $J$ and $T^{*}$ dilates to $J^{*}$.
gy-dil-dBR Proposition 27.17. If $C \in B(H)$ is a $C_{0}$-contraction, then $C$ dilates to a bilateral shift.

Proof. There is a Hilbert space $E$ and isometry $V: H \rightarrow \ell_{E}^{2}(\mathbb{N})$ such that $V p(C)=$ $p\left(S_{E}^{*}\right) V$ for $p \in \mathbb{C}[z]$. Hence $q\left(C^{*}\right)=V^{*} q\left(S_{E}\right) V$ for all $q \in \mathbb{C}[z]$. The inclusion $\iota$ : $\ell_{E}^{2}(\mathbb{N}) \rightarrow \ell_{E}^{2}(\mathbb{Z})$ is an isometry and $q\left(S_{E}\right)=\iota^{*} q\left(U_{E}\right) \iota$ for all polynomials $q \in \mathbb{C}[z]$. Hence $q\left(C^{*}\right)=W^{*} q\left(U_{E}\right) W$ where $W=\iota V$. It now follows that $p(C)=W^{*} p\left(U_{E}^{*}\right) W$ for all polynomials $p \in \mathbb{C}[z]$.

Remark 27.18. The Sz.-Nagy dilation theorem is the assertion that a contraction operator dilates to a unitary operator. In particular, the hypothesis and conclusion of the Nagy dilation theorem are both slightly different than that of Proposition $\frac{\text { prop:Nagy-dil-dBR }}{27.17 . \mathrm{ft}}$ has a relatively simple geometric proof.
inequality Theorem 27.19 (The von Neumann inequality). If $C \in B(H)$ is a contraction and $p(z)=\sum_{j=0}^{N} p_{j} z^{j}$ is a polynomial, then $\|p(C)\| \leq\|p\|_{\infty, \mathbb{D}}$.

Proof. By the Sz.-Nagy dilation theorem there is a Hilbert space $K$ and isometry $V$ such that for all polynomials $q \in \mathbb{C}[z]$,

$$
q(C)=V^{*} q(U) V
$$

By Proposition $\frac{\text { prop:unitary-fn-cal }}{27.8}\|p(U)\| \leq\|p\|_{\infty, \mathbb{D}}$. Thus,

$$
\|p(C)\| \leq\left\|V^{*}\right\|\|p(U)\|\|V\| \leq\|p\|_{\infty, \mathbb{D}} .
$$

27.4. Spectral measures. Let $H$ denote a Hilbert space (over $\mathbb{C}$ ). For $h \in H$, let $r_{h}$ : $B(H) \rightarrow[0, \infty)$ denote the function $r_{h}(T)=\|T h\|$. The strong operator topology (SOT) on $H$ is the topology generated by the functions $\left\{r_{h}: h \in H\right\}$. In particular, a sequence $\left(T_{n}\right)$ from $B(H)$ SOT converges to $T \in B(H)$ if and only if $\lim _{n \rightarrow \infty}\left\|\left(T-T_{n}\right) h\right\|=0$ for all $h \in H$. In particular, $T$ is a $C_{0}$-contraction if and only if $\|T\| \leq 1$ and $\left(T^{n}\right)$ SOT-converges to 0.

An $B(H)$-valued spectral measure on a measurable space $(X, \Sigma)$ is a mapping $E$ : $\Sigma \rightarrow B(H)$ such that
(i) $E(X)=I$, the identity mapping;
(ii) $E(\emptyset)=0$;
(iii) $E$ is projection valued meaning if $\sigma \in \Sigma$, then $E(\sigma)$ is a projection;
(iv) if $\left(\omega_{j}\right)_{j=1}^{\infty}$ is a disjoint sequence of sets from $\Sigma$, then the series

$$
\sum_{j=1}^{\infty} E\left(\omega_{j}\right)
$$

SOT-converges to $E\left(\cup \omega_{j}\right)$.

Given self-adjoint (hermitian) operators $P, R$ on a Hilbert space $H$ the notation $P \preceq R$ means $R-P \succeq 0$. This partial order on the self-adjoint elements of $B(H)$ is the Loewner ordering.

Lemma 27.20. Suppose $P, Q \in B(H)$. If $P, Q$ and $P+Q$ are all projections, then $P Q=Q P=0$.

Proof. The equality $P+Q=(P+Q)^{2}$ implies $P Q+Q P=0$. Multiplying this last equation on both sides by $Q$ gives, $0=2 Q P Q=2 Q P P Q=2 Y^{*} Y$, where $Y=P Q$. Thus $P Q=Y=0$.

Proposition 27.21. Suppose $E$ is an $B(H)$-valued spectral measure on $(X, \Sigma)$ and $\omega, \rho \in \Sigma$.
(i) if $\omega \cap \rho=\emptyset$, then $E(\omega) E(\rho)=0$;
(ii) if $\omega \subset \rho$, then

$$
\begin{equation*}
E(\rho)=E(\rho \backslash \omega)+E(\omega) \tag{89}
\end{equation*}
$$

and therefore
(a) $E(\omega) \preceq E(\rho)$; and
(b) $E(\omega) E(\rho)=E(\omega)$;
(iii) $E(\omega \cap \rho)=E(\omega) E(\rho)=E(\rho) E(\omega)$.

Proof. Letting $P=E(\omega)$ and $Q=E(\rho)$ by additivity of $E$, all of $P, Q$ and $P+Q$ are projections. Hence $P Q=0$. Item (ii)(a) follows immediately from (i). For (ii)(b), multiply equation (eq9) Eadd (18) on right by $E(\omega)$ and apply (i).

To prove (iii) observe

$$
\begin{equation*}
E(\omega \cup \rho)+E(\omega \cap \rho)=E(\omega)+E(\rho), \tag{90}
\end{equation*}
$$

eq:EdeM
an identity that follows from

$$
\begin{aligned}
A \cup B & =(A \backslash B) \cup(B \backslash A) \cup(A \cap B) \\
A & =(A \backslash B) \cup(A \cap B) \\
B & =(B \backslash A) \cup(A \cap B) .
\end{aligned}
$$

Multiplying equation equation ( $(\mathrm{eg} ; \mathrm{Edem})$ on the right by $E(\omega)$ gives

$$
E(\omega \cup \rho) E(\omega)+E(\omega \cap \rho) E(\omega)=E(\omega)^{2}+E(\rho) E(\omega)
$$

Using (ii)(b) on each term on the left hand side, gives

$$
E(\omega)+E(\omega \cap \rho)=E(\omega)+E(\rho) E(\omega)
$$

from which the result follows.

Example 27.22. Let $N$ be a normal matrix of size $n$. From undergraduate linear algebra, $N$ is unitarily equivalent to a diagonal matrix; i.e., there exists a unitary $U$ and a diagonal matrix $D=\operatorname{diag}\left(\lambda_{1}, \ldots, \lambda_{n}\right)$ such that $N=U D U^{*}$. Letting $u_{j}$ denote the columns of $u$

$$
N=\sum \lambda_{j} u_{j} u_{j}^{*}
$$

and of course $\sigma(N)=\left\{\lambda_{1}, \ldots, \lambda_{n}\right\}$. Letting $P_{j}$ denote the rank one matrix $u_{j} u_{j}^{*}$. Since $U$ is unitary, each $u_{j} u_{j}^{*}$ is a projection and moreover, $P_{j} P_{k}=0$ for $j \neq k$ Letting $E$ denote the mapping on the Borel sets of $\mathbb{C}$ defined by

$$
E(\omega)=\sum_{\lambda_{j} \in \omega} u_{j} u_{j}^{*}=\sum_{\lambda_{j} \in \omega} P_{j}
$$

gives a spectral measure and

$$
N=\sum \lambda_{k} E\left(\left\{\lambda_{k}\right\}\right) .
$$

Moreover, if $f$ is a polynomial, then

$$
f(N)=\sum f\left(\lambda_{j}\right) P_{j} .
$$

Example 27.23. Let $(X, \mathscr{M}, \mu)$ denote a measure space. Define a $L^{2}(\mu)$-valued spectral measure $E$ on $\mathscr{M}$ as follows. Given $A \in \mathscr{M}$, let $E(A): L^{2}(\mu) \rightarrow L^{2}(\mu)$ denote the mapping defined by $E(A) f=\mathbf{1}_{A} f$. Given $\left(A_{j}\right)$ a disjoint sequence from $\mathscr{M}$ and $f \in$ $L^{2}(\mu)$, observe that

$$
\sum_{j=1}^{N} E\left(A_{j}\right) f=E\left(\cup_{j=1}^{N} A_{j}\right) f=\sum \mathbf{1}_{\cup^{N} A_{j}} f
$$

converges to $\mathbf{1}_{\cup A_{j}} f$ by the dominated convergence theorem. Hence $\sum E\left(A_{j}\right)$ converges SOT to $E\left(\cup A_{j}\right)$.
prop:Egh Proposition 27.24. If $E$ is an $B(H)$-valued spectral measure on $(X, \Sigma)$ and if $h, k \in H$, then $\mu_{h, k}: \Sigma \rightarrow \mathbb{C}$ defined by $\mu_{h, k}(\omega)=\langle E(\omega) h, k\rangle$ is a measure with total variation $\left|\mu_{h, k}\right|(X):=\left\|\mu_{h, k}\right\| \leq\|h\|\|k\|$.

Part of proof. Let $\Pi=\left\{\omega_{1}, \ldots, \omega_{n}\right\}$ be a given measurable partition of $X$. For each $j$ there is a unimodular $\gamma_{j} \in \mathbb{C}$ such that

$$
\gamma_{j}\left\langle E\left(\omega_{j}\right) h, k\right\rangle \geq 0
$$

Hence,

$$
\begin{aligned}
\sum\left|\left\langle E\left(\omega_{j}\right) h, k\right\rangle\right| & =\sum \gamma_{j}\left\langle E\left(\omega_{j}\right) h, k\right\rangle \\
& =\left\langle\sum E\left(\omega_{j}\right) \gamma_{j} h, k\right\rangle \\
& =\left|\left\langle\sum E\left(\omega_{j}\right) \gamma_{j} h, k\right\rangle\right| \\
& \leq\left\|\sum E\left(\omega_{j}\right) \gamma_{j} h\right\|\|k\| .
\end{aligned}
$$

Now the vectors $h_{j}^{\prime}=E\left(\omega_{j}\right) \gamma_{j} h$ and $h_{j}^{\prime \prime}=E\left(\omega_{j}\right) h$ are pairwise orthogonal (since the $\omega_{j}$ are disjoint). Hence,

$$
\left\|\sum E\left(\omega_{j}\right) \gamma_{j} h\right\|^{2}=\sum\left\|E\left(\omega_{j}\right) \gamma_{j} h\right\|^{2}=\sum\left\|E\left(\omega_{j}\right) h\right\|^{2}=\left\|\sum E\left(\omega_{j}\right) h\right\|^{2}=\|h\|^{2} .
$$

Putting things together gives,

$$
\sum\left|\mu_{h, k}\left(\omega_{j}\right)\right|=\sum\left|\left\langle E\left(\omega_{j}\right) h, k\right\rangle\right| \leq\|h\|\|k\|
$$

and the result follows by the definition of total variation norm.

Spectral measures share many properties of measures and we will use these properties without (or with little) comment in what follows. As an example, we can assume if $E(\rho)=0$, then $\Sigma$ contains all subsets of $\rho$. A measurable function $f: X \rightarrow \mathbb{C}$ is essentially bounded (with respect to $E$ ) if there is a $C$ such that $E(\{|f(x)|>C\})=0$ and in this case $\|f\|_{\infty}$ is the infimum of such $C$. The collection of such function, $L^{\infty}(E)$, is a $C$-star algebra. Given a measurable space $(X, \mathscr{M})$, let $\mathbb{B}(X)=\mathbb{B}(X, \mathscr{M})$ denote the
bounded ( $\mathbb{C}$-valued) functions on $X$. It is a C -star algebra (under the usual pointwise operations and involution given by pointwise conjugation) and the supremum norm.
p:specint0
Proposition 27.25. If $E$ is an $B(H)$-valued spectral measure on $(X, \mathscr{M})$ and $f \in$ $\mathbb{B}(X)$, then there is a unique operator $I(f) \in B(H)$ such that for every $\epsilon>0$, each measurable partition $\left\{A_{1}, \ldots, A_{n}\right\}$ of $X$ with $\sup \left\{|f(x)-f(y)|: x, y \in A_{j}\right\}<\epsilon$ for each $j$, and each choice of $x_{j} \in A_{j}$,

$$
\left\|I(f)-\sum f\left(x_{j}\right) E\left(A_{j}\right)\right\|<\epsilon
$$

Moreover, if $A \in \mathscr{M}$, then $I\left(\mathbf{1}_{A}\right)=E(A)$.

A mapping $J: H \times H \rightarrow \mathbb{C}$ is sesquilinear if, for all $g, h, k \in H$ and $c \in \mathbb{C}, g, h, k \in H$ and $c \in \mathbb{C}$,

$$
\begin{aligned}
J(g, h+c k) & =J(g, h)+\bar{c} J(g, k) \\
J(g+c h, k) & =J(g, k)+c J(h, k) \\
J(g, h) & =\overline{J(h, g)}
\end{aligned}
$$

The form is bounded if there is a (real) constant $C$ such that

$$
|J(h, k)| \leq C\|h\|\|k\|
$$

In this case the infimum over all such $C$ is the norm of $J$. If $T \in B(H)$, then $J(h, k)=$ $\langle T h, k\rangle$ is a bounded sesquilinear form with $\|J\|=\|T\|$.

Lemma 27.26. If $J$ is a bounded sesquilinear form on the Hilbert space $H$, then there is a uniquely determined bounded operator $T$ on $H$ such that

$$
\langle T h, k\rangle=J(h, k) .
$$

Moreover, $\|T\|$ is the norm of $J$.

Sketch of proof. Fix $k \in H$. The mapping $\lambda_{k}: H \rightarrow \mathbb{C}$ defined by $\lambda_{k}(h)=J(h, k)$ is linear and bounded with norm at most $C\|k\|$. Hence, by the Riesz representation theorem, there is a vector $k_{*} \in H$ such that $\left\|k_{*}\right\| \leq C\|k\|$ and

$$
J(h, k)=\left\langle h, k_{*}\right\rangle
$$

Since $J$ is conjugate linear in the second variable, the mapping $k \rightarrow k_{*}$ is linear. Let $A$ denote this mapping. In particular, $\|A k\|=\left\|k_{*}\right\| \leq C\|k\|$. Hence $A$ is bounded and $\|A\| \leq C$ and for $h, k \in H$,

$$
J(h, k)=\langle h, A k\rangle .
$$

Choose $T=A^{*}$.

Sketch of proof of Proposition prop:specinto $\in \mathbb{B}(X)$. Define $J: H \times H \rightarrow \mathbb{C}$ by $J(h, k)=\int f d \mu_{h, k}$. Verify that $J$ is sesquilinear and $|J(h, k)| \leq\|f\|_{\infty}\|h\|\|k\|$. Thus $J$ is a bounded sesquilinear form. Thus, there is a (uniquely determined) bounded linear operator $I(f)$ such that

$$
\langle I(f) h, k\rangle=J(h, k)=\int f d \mu_{h, k}
$$

Now suppose $\epsilon>0$, a partition and points $x_{j}$ as specified in the statement of the proposition are given and estimate, for $h, k \in H$,

$$
\begin{aligned}
\left|\left\langle\left[I(f)-\sum f\left(x_{j}\right) E\left(A_{j}\right)\right] h, k\right\rangle\right| & =\left|\sum \int_{A_{j}}\left(f-f\left(x_{j}\right)\right) d \mu_{h, k}\right| \\
& \leq \sum \epsilon\left|\mu_{h, k}\right|\left(A_{j}\right) \\
& =\epsilon\left|\mu_{h, k}\right|(X) \leq \epsilon\|h\|\|k\| .
\end{aligned}
$$

We write $\int f d E$ for $I(f)$ and call it the spectral integral of $f$.

Proposition 27.27. Suppose $(X, \mathscr{M})$ is a measure space and $E$ is an $B(H)$-valued spectral measure on $X$. The mapping $I: \mathbb{B}(X) \rightarrow B(H)$ is a contractive $*$-homomorphism.

Sketch of proof. While not particularly difficult, the most challenging part is to show $I$ is multiplicative. The remainder of the proof is left to the gentle reader.

Fix $f, g \in \mathbb{B}(X)$ and let $\epsilon>0$ be given. Choose a measurable partition $\left\{A_{1}, \ldots, A_{n}\right\}$ of $X$ such that $\sup \left\{|F(x)-F(y)|: x, y \in A_{j}\right\}<\epsilon$ for each $j$ and $F \in\{f, g, f g\}$. Given $x_{j} \in A_{j}$, note that

$$
\sum_{j} f\left(x_{j}\right) g\left(x_{j}\right) E\left(A_{j}\right)=\sum_{j, k} f\left(x_{j}\right) g\left(x_{k}\right) E\left(A_{j}\right) E\left(A_{k}\right)
$$

since $E\left(A_{j}\right) E\left(A_{k}\right)=E\left(A_{j} \cap A_{k}\right)=0$ for $j \neq k$. From here the proof is messy, but standard.

Proposition 27.28. If $X$ is a compact Hausorff space and $\rho: C(X) \rightarrow B(H)$ is a unital isometric *-homomorphism, then there is a uniquely determined spectral measure
on the Borel sets of $X$ such that, for $f \in C(X)$,

$$
\rho(f)=\int f d E
$$

Sketch of proof. The first step is to extend $\rho$ to an contractive $*$-homomorphism $\tau$ : $\mathbb{B}(X) \rightarrow B(H)$. Given $h, k \in H$, let $\rho_{h, k}: C(X) \rightarrow \mathbb{C}$ denote the linear functional

$$
\rho_{h, k}(f)=\langle\rho(f) h, k\rangle .
$$

By the Riesz-Markov Theorem, for each $h, k \in H$ there is a unique Borel measure $\mu_{h, k}$ such that

$$
\rho_{h, k}(f)=\int f d \mu_{h, k}
$$

and moreover, $\|h\|\|k\| \geq\left\|\rho_{h, k}\right\|=\left|\mu_{h, k}\right|(X)$. One checks that the mapping $H \times H \rightarrow$ $M(X)$ (the space of measures) given by $(h, k) \rightarrow \mu_{h, k}$ is sesquilinear.

Given $g \in \mathbb{B}(X)$ we obtain the sesquilinear form $J: H \times H \rightarrow \mathbb{C}$ given by

$$
J(h, k)=\int g d \mu_{h, k} .
$$

Evidently $|J(h, k)| \leq\|g\|_{\infty}\left|\mu_{h, k}\right|(X) \leq\|g\|_{\infty}\|h\|\|k\|$. Hence there is a uniquely determined bounded linear operator $\tau(g)$ such that

$$
\langle\tau(g) h, k\rangle=J(h, k)=\int g d \mu_{h, k} .
$$

In particular, if $g \in C(X)$, then $\tau(g)=\rho(g)$ and $\|\tau(g)\| \leq\|g\|_{\infty}$.
We will verify that $\tau$ is multiplicative, leaving the rest of the verification that $\tau$ is a $*$-homomorphism (and hence contractive) to the gentle reader. If $f, g \in C(X)$, then
$\tau(f g)=\rho(f g)=\rho(f) \rho(g)$. Hence,

$$
\begin{aligned}
\langle\tau(f g) h, k\rangle & =\int f g d \mu_{h, k} \\
& =\langle\rho(f) \rho(g) h, k\rangle \\
& =\int f d \mu_{\rho(g) h, k}
\end{aligned}
$$

In particular, $g d \mu_{h, k}=d \mu_{\rho(g) h, k}$ as measures. Thus, for $f \in \mathbb{B}(X)$,

$$
\begin{aligned}
\langle\tau(f g) h, k]\rangle & =\int f g d \mu_{h, k} \\
& =\int f d \mu_{\rho(g) h, k} \\
& =\langle\tau(f) \rho(g) h, k\rangle \\
& =\left\langle\rho(g) h, \tau(f)^{*} k\right\rangle \\
& =\int g d \mu_{h, \tau(f) * k}
\end{aligned}
$$

Hence $\tau(f g)=\tau(f) \tau(g)$ for $f \in \mathbb{B}(X)$ and $g \in C(X)$. Moreover, $f d \mu_{h, k}=d \mu_{h, \tau(f)^{*} k}$. Thus, for $g \in \mathbb{B}(X)$,

$$
\begin{aligned}
\langle\tau(f g) h, k\rangle & =\int f g d \mu_{h, k} \\
& =\int g d \mu_{h, \tau(f)^{*} k} \\
& =\left\langle\tau(g) h, \tau(f)^{*} k\right\rangle \\
& =\langle\tau(f) \tau(g) h, k\rangle
\end{aligned}
$$

and hence $\tau(f g)=\tau(f) \tau(g)$ for all $f, g \in \mathbb{B}(X)$. That $\tau$ is contractive, additive, respects scalar multiplication and satisfies $\tau\left(f^{*}\right)=\tau(f)^{*}$ are left as exercises.

Given a Borel set $\omega$, let $E(\omega)=\tau\left(\mathbf{1}_{\omega}\right)$. Thus $E(\emptyset)=0, E(X)=\tau(1)=I$ and $E$ is finitely additive. Now suppose $\left(f_{n}\right)$ is a bounded sequence from $\mathbb{B}(X)$ that converges
pointwise to $f$ (automatically in $\mathbb{B}(X)$ ). Given $h \in H$,

$$
\begin{aligned}
\left\|\tau\left(f_{n}-f\right) h\right\|^{2} & =\left\langle\left(\tau\left(f_{n}-f\right)^{*} \tau\left(f_{n}-f\right)\right) h, h\right\rangle \\
& =\left\langle\tau\left(\left|f_{n}-f\right|^{2}\right) h, k\right\rangle \\
& =\int\left|f_{n}-f\right|^{2} d \mu_{h, k}
\end{aligned}
$$

Now $\left|f_{n}-f\right|^{2}$ is a (uniformly) bounded sequence converging to 0 pointwise. Thus, since $\left|\mu_{h, k}\right|$ is a finite measure, the sequence of integrals above converges to 0 by dominated convergence. Thus $\left(\tau\left(f_{n}\right)\right)$ SOT-converges to $\tau(f)$. Now let $\left(\omega_{j}\right)_{j=1}^{\infty}$ be a sequence of pairwise disjoint measurable sets and let $f_{n}=\sum_{j=1}^{n} \boldsymbol{1}_{\omega_{j}}$. Hence $\left(f_{n}\right)$ is a bounded sequence of measurable functions that converges pointwise to $f$, the characteristic function of $\omega=\cup \omega_{j}$. Hence $\tau\left(f_{n}\right)$ SOT-converges to $f$; i.e.,

$$
\operatorname{SOT} \lim \tau\left(f_{n}\right)=\sum_{j=1}^{\infty} E(\omega)=E(\omega)=\tau(f)
$$

Thus $E$ is a spectral measure on the Borel sets of $X$.
If $s=\sum c_{j} \mathbf{1}_{\omega_{j}}$ is a simple measurable function, then

$$
\tau(s)=\sum c_{j} \tau\left(\mathbf{1}_{\omega_{j}}\right)=\sum c_{j} E\left(\omega_{j}\right)=\int s d E
$$

Since an $f \in \mathbb{B}(X)$ is a uniform limit of simple functions $\left(s_{n}\right)$ and $\tau$ is contractive, $\left(\tau\left(s_{n}\right)\right)$ converges in norm to $\tau(f)$. Likewise, since the mapping $I: \mathbb{B}(X) \rightarrow B(H)$ given by $I(h)=\int h d E$ is contractive, $\left(I\left(s_{n}\right)\right)$ converges to $I(f)$. Using $I\left(s_{n}\right)=\int s_{n} d E$, it follows that

$$
\tau(f)=\int f d E
$$

for $f \in \mathbb{B}(X)$.

Theorem 27.29 (Spectral Theorem). Suppose $\mathscr{A} \subset B(H)$ is a commutative sub-C-star algebra with maximal ideal space $\Sigma$ and let $\mathscr{G}: \mathscr{A} \rightarrow C(\Sigma)$ denote the Gelfand transform. There is an $B(H)$-valued spectral measure $E$ such that
(1) for $T \in \mathscr{A}$,

$$
T=\int_{\Sigma} \mathscr{G}(T) d E
$$

(2) if $\emptyset \neq O \subset \Sigma$ is open, then $E(O) \neq 0$;
(3) $R \in B(H)$ commutes with $\mathscr{A}$ if and only if $R E(\omega)=E(\omega) R$ for all $\omega \in \Sigma$.

Sketch of proof. The inverse of the Gelfand transform $\mathscr{G}^{-1}: C(\Sigma) \rightarrow \mathscr{A} \subset B(H)$ is an isometric $*$-isomorphism. Hence there is an $B(H)$-valued spectral measure on the Borel sets of $\Sigma$ such that

$$
\mathscr{G}^{-1}(f)=\int f d E
$$

for $f \in C(\Sigma)$. Choose $f=\mathscr{G}(T)$ to obtain the first item.

If $\emptyset O \subset \Sigma$ is open, then there exists a non-zero continuous function $f$ that is 0 on $O^{C}$. Hence, $\left\|\mathscr{G}^{-1}(f)\right\|>0$ and hence $E(O) \neq 0$.

Suppose $R$ commutes with each $E(\omega)$. Given $h, k \in H$ and $\omega$ a Borel set,

$$
\langle E(\omega) R h, k\rangle=\left\langle E(\omega) h, R^{*} k\right\rangle .
$$

Thus, $d \mu_{R h, k}=d \mu_{h, R^{*} k}$. Hence, for $f \in C(\Sigma)$,

$$
\begin{aligned}
\left\langle\left[\int f d E\right] R h, k\right\rangle & =\int f d \mu_{R h, k} \\
& =\int f d \mu_{h, R^{*} k} \\
& =\left\langle\int f d E h, R^{*} k\right\rangle \\
& =\left\langle R \int f d E h, k\right\rangle
\end{aligned}
$$

Conversely, if $R$ commutes with $\mathscr{A}$, then for each $f \in C(\Sigma)$ and pair of vectors $h, k$,

$$
\begin{aligned}
\int f d \mu_{R h, k} & =\left\langle\left[\int f d E\right] R h, k\right\rangle \\
& =\left\langle\left[\int f d E\right] h, R^{*} k\right\rangle \\
& =\int f d \mu_{h, R^{*} k}
\end{aligned}
$$

Hence $\mu_{R h, k}=\mu_{h, R^{*} k}$ and consequently $\langle E(\omega) R h, k\rangle=\langle R E(\omega) h, k\rangle$ for all measurable sets $\omega$ and vectors $h, k \in H$. The result follows.

Corollary 27.30 (Spectral Theorem). If $T \in B(H)$ is normal, then there exists an $B(H)$-valued spectral measure $E$ on the Borel sets of $\sigma(T)$ such that

$$
f(T)=\int f d E
$$

i.e., the isometric $*$-homomorphism $I: L^{\infty}(E) \rightarrow B(H)$ given by $I(f)=\int f d E$ extends the functional calculus.

Proof. Let $C^{*}(T)$ denote the $C$-star algebra generated by $T$. Thus $C^{*}(T)$ is a commutative C-star algebra and its maximal ideal space $\Sigma$ is identified with $\sigma(T) \subset \mathbb{C}$ via the homeomorphism $\psi: \Sigma \rightarrow \sigma(T)$ given by $\psi(h)=\mathscr{G}(T)(h)$. Hence there is a uniquely
determined spectral measure $E$ on $\sigma(T)$ such that, for polynomials $p\left(z, z^{*}\right)$ (viewed as elements of $C(\sigma(T))$,

$$
p\left(T, T^{*}\right)=\int_{\sigma(T)} \mathscr{G}\left(p\left(T, T^{*}\right) \circ \psi^{-1} d E\right.
$$

Unraveling the definitions, $\mathscr{G}\left(p\left(T, T^{*}\right) \circ \psi^{-1}\left(z, z^{*}\right)=p\left(z, z^{*}\right)\right.$; i.e.,

$$
\rho(p)=\int p d E
$$

where $\rho$ is the functional calculus. By continuity, $\rho(f)=\int f d E$ for $f \in C(\sigma(T))$.

