OPERATOR ALGEBRAS AND NON-COMMUTATIVE ANALYSIS:
An introductory course with application in quantum mechanics

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Abstract
We give an elementary introduction to the subject of operator algebras an non-commutative analysis with the emphasis on material related to a number of open problems arising from quantum mechanics.

1 Introduction
1.1 Basic definitions and notation

1.1 DEFINITION (Banach algebra). A Banach algebra is an algebra $\mathcal{A}$ over the complex numbers equipped with a norm $\| \cdot \|$ under which it is complete as a metric space such that

$$\|ab\| \leq \|a\|\|b\| \quad \text{for all} \quad a, b \in \mathcal{A}. \quad (1.1)$$

1.2 EXAMPLE. Let $X$ be a locally compact Hausdorff space, and let $\mathcal{C}_0(X)$ denote the set of continuous complex valued functions on $X$ that vanish at infinity, and equip it with the supremum norm. Then with the usual algebraic structure of pointwise addition and multiplication, $\mathcal{A} = \mathcal{C}_0(X)$ is a Banach algebra. This is the canonical example of a commutative Banach algebra. There is a multiplicative identity if and only if $X$ is compact.

1.3 EXAMPLE. Let $\mathcal{A} = L^1(\mathbb{R}^n)$ equipped with the convolution product

$$f * g(x) = \int_{\mathbb{R}^n} f(x - y)g(y)dy.$$

Let the norm be the $L^1$ norm. Then $\mathcal{A}$ is a commutative Banach algebra that does not have an identity.
1.4 EXAMPLE. Let \( \mathcal{H} \) be a Hilbert space, and let \( A = \mathcal{B}(\mathcal{H}) \), the set of all continuous linear mappings from \( \mathcal{H} \) to \( \mathcal{H} \), equipped with the composition product and the operator norm

\[
\|a\| = \sup\{ \|a\psi\|_{\mathcal{H}} : \psi \in \mathcal{H}, \|\psi\|_{\mathcal{H}} = 1 \} = \sup\{ \Re(<\varphi, a\psi>_{\mathcal{H}}) : \varphi, \psi \in \mathcal{H}, \|\varphi\|_{\mathcal{H}}, \|\psi\|_{\mathcal{H}} = 1 \},
\]

where \( \|\cdot\|_{\mathcal{H}} \) is the norm on \( \mathcal{H} \), and \( <\cdot, \cdot>_{\mathcal{H}} \) is the inner product in \( \mathcal{H} \). This is the canonical example of a non-commutative Banach algebra.

1.5 EXAMPLE. Let \( A \) be the algebra of \( n \times n \) matrices. The Frobenius, or Hilbert-Schmidt norm on \( A \) is the norm \( \|\cdot\|_2 \) given by

\[
\|a\|_2 = \left( \sum_{i,j=1}^{n} |a_{i,j}|^2 \right)^{1/2}
\]

where \( a_{i,j} \) denotes the \( i, j \)th entry of \( a \). By the Cauchy-Schwarz inequality, for all \( a, b \in A \),

\[
\|ab\|_2 = \left( \sum_{i,j=1}^{n} \sum_{k=1}^{n} a_{i,k}b_{k,j} \right)^{1/2} \leq \left( \sum_{i,j=1}^{n} \sum_{k=1}^{n} |a_{i,k}|^2 \right)^{1/2} \left( \sum_{k=1}^{n} |b_{k,j}|^2 \right)^{1/2} = \|a\|_2 \|b\|_2,
\]

and thus (1.1) is satisfied. Note that the algebra of \( n \times n \) matrices with the operator norm is the special case of Example 1.4 in which \( \mathcal{H} = \mathbb{C}^n \).

1.6 DEFINITION (\( C^* \)-algebra). A \( C^* \) algebra is a Banach algebra equipped with a conjugate linear map \( * : \mathcal{A} \rightarrow \mathcal{A} \), the action of which is written as \( a \mapsto a^* \), and which satisfies the properties

(i) The \( * \) map is an involution; for all \( a \in \mathcal{A} \), \( a^{**} = a \).

(ii) For all \( a, b \in \mathcal{A} \), \( (ab)^* = b^*a^* \).

(iii) For all \( a \in \mathcal{A} \),

\[
\|aa^*\| = \|a\|^2. \tag{1.3}
\]

When discussing a \( C^* \) algebra it is convenient and standard to refer to the map \( a \mapsto a^* \) as the involution in \( \mathcal{A} \).

In a \( C^* \) algebra, the involution is always an isometry. This is because

\[
\|a\|^2 = \|aa^*\| \leq \|a\|\|a^*\|,
\]

where we used (1.3) and (1.1) in succession. Then for \( a \neq 0 \), we have \( \|a\| \leq \|a^*\| \), and then \( \|a^*\| \leq \|a^{**}\| = \|a\| \), so that \( \|a^*\| = \|a\| \) for all \( a \in \mathcal{A} \). The condition (1.3) is much stronger than the condition that \( a \mapsto a^* \) be an isometry.

1.7 EXAMPLE. In Example 1.2, take the involution to be pointwise complex conjugation of the functions that constitute the algebra. The conditions (i), (ii) and (iii) are all clearly satisfied in this case. Thus, \( \mathcal{C}_0(X) \) equipped with this structure is a commutative \( C^* \)-algebra.

Similarly, in Example 1.4, take define the involution by taking \( a^* \) to be the Hermitian conjugate of \( a \). That is, for all \( \varphi, \psi \in \mathcal{H} \),

\[
\langle a^* \varphi, \psi \rangle_{\mathcal{H}} = \langle \varphi, a\psi \rangle_{\mathcal{H}}.
\]
It is immediate from this that \( a \mapsto a^* \) is conjugate linear and an involution satisfying (ii). Also, it is immediate from this and (1.2) that \( \|a^*\| = \|a\| \) for all \( a \). Moreover, for all \( \psi \in \mathcal{H} \),

\[
\langle \psi, aa^* \psi \rangle_{\mathcal{H}} = \langle a^* \psi, a^* \psi \rangle_{\mathcal{H}} = \|a^* \psi\|^2,
\]

and hence

\[
\|aa^*\| = \sup \{ \mathcal{R}(\langle \varphi, aa^* \psi \rangle_{\mathcal{H}}) : \varphi, \psi \in \mathcal{H}, \|\varphi\|_{\mathcal{H}}, \|\psi\|_{\mathcal{H}} = 1 \} \geq (\sup \{ \|a^* \psi\|_{\mathcal{H}} : \psi \in \mathcal{H}, \|\psi\|_{\mathcal{H}} = 1 \})^2 = \|a^*\|^2 = \|a\|^2.
\]

1.8 EXAMPLE. If we equip the convolution algebra of Example 1.3 with the convolution given by pointwise complex conjugation, this involution is an isometry since for any \( f \in L^1(\mathbb{R}) \), \( \|f\|_{L^1(\mathbb{R})} = \|f^*\|_{L^1(\mathbb{R})} \). However, the stronger property (1.3) does not hold in general in this algebra. To see this, let \( \rho \) and \( \sigma \) be two non-negative functions in \( L^1(\mathbb{R}^n) \). Fix \( \lambda \in \mathbb{R} \) and define functions \( f \) and \( g \) in \( L^1(\mathbb{R}^n) \) by

\[
f(x) = \rho(x)e^{i\lambda x} \quad \text{and} \quad g(x) = \sigma(x)e^{i\lambda x}.
\]

Then

\[
f * g(x) = e^{i\lambda x} \rho * \sigma(x) \quad \text{so that} \quad \|f * g\|_{L^1(\mathbb{R})} = \|\rho * \sigma\|_{L^1(\mathbb{R})}.
\]

However,

\[
f * g^*(x) = e^{i\lambda x} \int_{\mathbb{R}} \rho(x)\sigma(x-y)e^{-2i\lambda y} dy,
\]

and now a simple argument using the Riemann-Lebesgue Lemma and the Dominated Convergence Theorem shows that \( \|f * g^*\|_{L^1(\mathbb{R})} \) converges to zero as \( \lambda \) is taken to infinity. Hence (1.3) fails in this Banach algebra.

Now consider Example 1.5, and again define \( a^* \) to be the Hermitian conjugate of \( a \). This is a conjugate linear involution, and as above, \( (ab)^* = b^*a^* \) for all \( a, b \). This involution is even an isometry in the Frobenius norm since

\[
\|a\|_2^2 = \sum_{i,j=1}^n |a_{i,j}|^2 = \sum_{i,j=1}^n |a_{j,i}^*|^2 = \|a^*\|_2^2.
\]

However, the property (1.3) fails. Recall that this algebra has a multiplicative identity \( e \), the \( n \times n \) identity matrix. Evidently, \( e^* = e \), and so were (1.3) to hold, we would have \( \|e\|_2 = \|ee^*\|_2 = \|e\|_2^2 \), but for \( n > 1 \) this is false since \( \|e\|_2 = \sqrt{n} \).

The condition (1.3) is very strong. As we shall soon see, given an algebra \( \mathcal{A} \) with a conjugate linear involution \( * \) satisfying \( (ab)^* = b^*a^* \) for all \( a, b \in \mathcal{A} \), there is at most one norm on \( \mathcal{A} \) that makes it a \( C^* \) algebra. We shall also see that our two examples of \( C^* \)-algebras, given in Example 1.7 are universal. In particular, a theorem of Gelfand and Naimark says that every \( C^* \) algebra is isomorphic to a sub-algebra of \( \mathcal{B}(\mathcal{H}) \) for some Hilbert space \( \mathcal{H} \) that one constructs from the algebra itself using the Gelfand-Naimark-Segal construction, to which we shall come.

An operator \( C^* \) algebra is an operator norm closed subalgebra of \( \mathcal{B}(\mathcal{H}) \) that is closed under Hermitian conjugation and closed in the operator norm. The term \( C^* \)-algebra was first applied in this context by Irving Segal in 1947. The theorem of Gelfand and Naimark shows that “abstract” \( C^* \)-algebras, as defined here, are essentially the same thing.
There are other important topologies in $\mathcal{B}(\mathcal{H})$ that are weaker than the topology given by the operator norm. In particular, there is the weak operator topology on $\mathcal{B}(\mathcal{H})$ which is the weakest topology under which the maps

$$a \mapsto \langle \varphi, a \psi \rangle_{\mathcal{H}}$$

are continuous for all $\varphi, \psi \in \mathcal{H}$. Every subset (and hence subalgebra) of $\mathcal{B}(\mathcal{H})$ that is weakly closed is norm closed, but not vice-versa in the case that $\mathcal{H}$ is infinite dimensional. A von Neumann algebra is a subalgebra of $\mathcal{B}(\mathcal{H})$ that is closed under Hermitian conjugation and closed in the weak operator topology. Thus, von Neumann algebras are a special type of $C^*$ operator algebras, and these shall be important to us also.

We shall be especially interested in von Neumann algebras, but every von Neumann algebra is a $C^*$ algebra, and every $C^*$ algebra is a Banach algebra, and it is natural to begin developing the theory at this general level, which we do in the next subsection.

1.2 The spectrum and the resolvent set

Let $X$ be a locally compact Hausdorff space that is not compact. Then $\mathcal{A} = \mathcal{C}_0(X)$ equipped with the usual structures is a Banach algebra without an identity. Let $\tilde{\mathcal{A}}$ be the larger algebra obtained by adjoining to $\mathcal{A}$ the constant functions, $\lambda 1, \lambda \in \mathbb{C}$. Then every $\tilde{a} \in \tilde{\mathcal{A}}$ has the form $\tilde{a}(x) = \lambda + a(x)$ where $a \in \mathcal{C}_0(X)$. Then for $\lambda + a$ and $\mu + b$ in $\tilde{\mathcal{A}}$,

$$(\lambda + a)(\mu + b) = \lambda \mu + (\lambda b + \mu a + ab) .$$

The constant function $1$ is the multiplicative identity in $\tilde{\mathcal{A}}$.

The procedure can be done in general. Let $\mathcal{A}$ be any Banach algebra, with or without a unit. Define $\tilde{\mathcal{A}}$ to be $\mathbb{C} \oplus \mathcal{A}$ with the multiplication

$$(\lambda, a)(\mu, b) = (\lambda \mu, \lambda b + \mu a + ab) ,$$

and the norm

$$\|(\lambda a)\| = |\lambda| + \|a\| .$$

By the definitions,

$$\|(\lambda, a)(\mu, b)\| = \|(\lambda \mu, \lambda b + \mu a + ab)\| = |\lambda| \mu + \|\lambda b + \mu a + ab\|$$

$$\leq |\lambda| \mu + |\lambda||b| + |\mu||a| + \|a\||b|$$

$$= (|\lambda| + \|a\|)(|\mu| + \|b\|) = \|(\lambda, a)\||(\mu, b)\| .$$

This shows that (1.1) is satisfied, and hence that $\tilde{\mathcal{A}}$ is a Banach algebra. Now define $e = (1, 0) \in \tilde{\mathcal{A}}$. Then $(1,0)(\lambda, a) = (\lambda, a)(1,0) = (\lambda, a)$ so that $e$ is the identity in $\tilde{\mathcal{A}}$.

The original algebra $\mathcal{A}$ is embedded in $\tilde{\mathcal{A}}$ as the subalgebra consisting of elements of the form $(0, a)$. None of these elements are invertible even when $\mathcal{A}$ itself has an identity. Indeed, if $(\lambda, a)$ has an inverse $(\mu, b)$, then

$$(1,0) = (\lambda, a)(\mu, b) = (\lambda \mu, \lambda b + \mu a + ab) ,$$

and this is impossible if $\lambda = 0$. However, it will be important in what follows that if $\mathcal{A}$ has a unit 1, then $1 - a$ is invertible in $\mathcal{A}$ if and only if $(1,-a)$ is invertible in $\tilde{\mathcal{A}}$. 

1.9 PROPOSITION. Let $\mathcal{A}$ be a Banach algebra with unit 1. Then $1 - a$ is invertible if and only if there exists $b \in \mathcal{A}$ such that
\[ ab = ba = b - a \quad . \quad (1.6) \]
consequently, $1 - a$ is invertible in $\mathcal{A}$ if and only if $e - a$ is invertible in $\widetilde{\mathcal{A}}$.

Proof. Suppose that $1 - a$ is invertible. Define $b = (1 - a)^{-1} - 1$. Then $(1 - a)b = 1 - (1 - a) = a$, and hence $ab = b - a$. The proof of $ba = b - a$ is similar.

Now suppose that there exists $b \in \mathcal{A}$ such that $(1.6)$ is true. Then
\[ (1 + b)(1 - a) = 1 + b - a - ab = 1 \quad \text{and} \quad (1 - a)(1 + b) = 1 - b + a - ba = 1 \quad . \]
This proves the first part.

For the second part, suppose that $1 - a$ is invertible in $\mathcal{A}$. Then there exists $b \in \mathcal{A}$ such that $(1.6)$ is satisfied. Regarding $a$ and $b$ as elements of $\widetilde{\mathcal{A}}$, $(1.6)$ is satisfied also in $\widetilde{\mathcal{A}}$, and hence $(1 - a)$ is invertible in $\widetilde{\mathcal{A}}$.

Finally, suppose that $(1, -a)$ is invertible in $\widetilde{\mathcal{A}}$, and let $(\lambda, b)$ be the inverse. Then
\[ (1, 0) = (1, -a)(\lambda, b) = (\lambda, b - \lambda a - ab) \quad . \]
Evidently $\lambda = 1$, and then $b - a - ab = 0$. A similar argument shows that $b - a - ba = 0$, and now the first part implies that $1 - a$ is invertible in $\mathcal{A}$. \qed

1.10 DEFINITION ( Spectrum and resolvent set). Let $\mathcal{A}$ be a Banach algebra, and let $a \in \mathcal{A}$. If $\mathcal{A}$ has a unit, the spectrum of $a$ in $\mathcal{A}$, $\sigma_{\mathcal{A}}(a)$ is defined to be the set of all $\lambda \in \mathbb{C}$ such that $\lambda 1 - a$ is not invertible. If $\mathcal{A}$ does not have a unit, then $\sigma_{\mathcal{A}}(a)$ is defined to be the spectrum of $(0, a) \in \widetilde{\mathcal{A}}$. The resolvent set of $a$ in $\mathcal{A}$, $\rho_{\mathcal{A}}(a)$ is defined to be the complement of $\sigma_{\mathcal{A}}(a)$.

Let $\mathcal{A}$ be a Banach algebra with a identity 1. Then we can still carry out the process of adjoining an identity to form $\widetilde{\mathcal{A}}$, and can regard each $a \in \mathcal{A}$ also as an element of $\widetilde{\mathcal{A}}$. Since no element of $\mathcal{A}$ is invertible in $\widetilde{\mathcal{A}}$, $0 \in \sigma_{\widetilde{\mathcal{A}}}(a)$ for all $a \in \mathcal{A}$. However, for $\lambda \neq 0$, $\lambda 1 - a$ is invertible if and only if $1 - a/\lambda$ is invertible. Likewise, $(\lambda - a)$ is invertible if and only if $(1, -a/\lambda)$ is invertible. Then by Proposition 1.9, $\lambda 1 - a$ is invertible in $\mathcal{A}$ if and only if $(1, 0) - (0, a/\lambda)$ is invertible in $\widetilde{\mathcal{A}}$. This shows that for $\lambda \neq 0$, $\lambda \in \sigma_{\mathcal{A}}(a) \iff \lambda \in \sigma_{\widetilde{\mathcal{A}}}(0, a)$. We summarize:
\[ \{0\} \cup \sigma_{\mathcal{A}}(a) = \sigma_{\widetilde{\mathcal{A}}}(0, a) \quad . \quad (1.7) \]

1.11 LEMMA ( Spectral Mapping Lemma). Let $\mathcal{A}$ be a Banach algebra, and let $p$ be a polynomial. In case $\mathcal{A}$ has no identity, we suppose that $p$ has no constant term. Then
\[ p(\sigma_{\mathcal{A}}(a)) = \sigma_{\mathcal{A}}(p(a)) \quad . \]

Proof. We may suppose that $p$ is not identically constant. We first suppose that $\mathcal{A}$ has an identity. Fix $\lambda \in \sigma_{\mathcal{A}}(a)$. We shall show that $p(\lambda) 1 - p(a)$ is not invertible. The polynomial $p(\lambda) - p(z)$ has a root at $z = \lambda$, and hence
\[ p(\lambda) - p(z) = (\lambda - z)q(z) \]
for some polynomial $q(z)$. Replacing $z$ by $a$,
\[ p(\lambda) 1 - p(a) = (\lambda - a)q(a) \quad . \]
Were \( p(\lambda)1 - p(a) \) invertible, we would have \( 1 = (\lambda - a)[q(a)(p(\lambda) - p(a))^{-1}] \), and then since polynomials in \( a \) commute, \( 1 = [q(a)(p(\lambda) - p(a))^{-1}](\lambda - a) \). This would mean that \( \lambda 1 - a \) is invertible, with contradicts our hypothesis that \( \lambda \in \sigma_{\mathcal{A}}(a) \). Hence \( p(\lambda) - p(a) \) is not invertible, and hence \( p(\lambda) \in \sigma_{\mathcal{A}}(p(a)) \). This shows that \( p(\sigma_{\mathcal{A}}(a)) \subset \sigma_{\mathcal{A}}(p(a)) \).

Next, fix \( \mu \in \sigma_{\mathcal{A}}(p(a)) \), and factor
\[
\mu - p(z) = \alpha(\lambda_1 - a) \cdots (\lambda_n - z)
\]
where \( \alpha \neq 0 \) and \( n \geq 1 \). For each \( j \), \( \mu = p(\lambda_j) \). We have
\[
\mu 1 - p(a) = \alpha(\lambda_1 1 - a) \cdots (\lambda_n 1 - a)
\]
and if each \( \lambda_j 1 - a \) were invertible, then \( \mu 1 - p(a) \) would be invertible, but this is not the case. Hence Hence for some \( j \), \( \lambda_j \in \sigma_{\mathcal{A}}(a) \), and \( \mu = p(\lambda_j) \in \sigma_{\mathcal{A}}(p(a)) \). This shows that \( \sigma_{\mathcal{A}}(p(a)) \subset p(\sigma_{\mathcal{A}}(a)) \), and completes th proof when \( \mathcal{A} \) has an identity. The general case now follows by adjoining an identity and then appealing to (1.7).

\[\square\]

1.3 Properties of the inverse function

Now let \( \mathcal{A} \) be a Banach algebra with an identity 1. Let \( a \in \mathcal{A} \) be such that \( \|1 - a\| = r < 1 \). Then by the defining property (1.1), \( \|(1 - a)^n\| \leq r^n \) for all \( n \in \mathbb{N} \). For all \( n \in \mathbb{N} \), define
\[
s_n = \sum_{j=1}^{n} (1-a)^j
\]
where, as usual, we interpret \( (1-a)^0 = 1 \). Then for all \( n > m \), by the triangle inequality and (1.1),
\[
\|s_n - s_m\| \leq \sum_{j=m+1}^{n} \|(1-a)^j\| \leq \sum_{j=m+1}^{n} r^j = \frac{r^m - r^n}{r - 1}.
\]
Hence \( \{s_n\}_{n \in \mathbb{N}} \) is a Cauchy sequence in \( \mathcal{A} \). Now for the first time we use the metric completeness of \( \mathcal{A} \): There exists \( b \in \mathcal{A} \) such that \( \lim_{n \to \infty} \|b - s_n\| = 0 \). But then
\[
ba = \lim_{n \to \infty} s_n a = \lim_{n \to \infty} s_n (1 - (1-a)) = \lim_{n \to \infty} (1 - (1-a)^{n+1}) = 1.
\]
The same reasoning shows that \( ab = 1 \), and so \( a \) is invertible. Let \( \Omega \) denote the set of invertible elements in \( \mathcal{A} \). This brings us to:

1.12 LEMMA. Let \( \mathcal{A} \) be a Banach algebra with a unit. Let \( \Omega \) be the set of invertible elements of \( \mathcal{A} \). Then \( \Omega \) contains every \( a \in \mathcal{A} \) such that \( \|1-a\| < 1 \), and in this case \( a^{-1} \) is given by the convergent series
\[
a^{-1} = \sum_{j=0}^{\infty} (1-a)^j.
\]
Moreover, if \( |\lambda| > \|a\| \), then \( \lambda 1 - a \) is invertible, with
\[
\|(\lambda 1 - a)^{-1}\| \leq \frac{1}{|\lambda| - \|a\|}.
\]
In particular, \( \sigma_{\mathcal{A}}(a) \) is contained in the closed disk of radius \( \|a\| \in \mathbb{C} \).
Proof. It remains to prove the final part. If $|\lambda| > \|a\|$, the $\lambda 1 - a = \lambda(1 - \lambda^{-1} a)$ and $\|1 - (1 - \lambda^{-1} a)\| = |\lambda|^{-1} \|a\| < 1$, so that $(1 - \lambda^{-1} a)$ is invertible.

At this point, we do not know in general that $\sigma_{\mathcal{A}}(a)$ is not empty, but we do know this of $\rho_{\mathcal{A}}(a)$. We now claim that $\Omega$ is open. This has the immediate consequence that $\rho_{\mathcal{A}}(a)$ is open, and hence that $\sigma_{\mathcal{A}}(a)$ is closed, though at this point the possibility that $\sigma_{\mathcal{A}}(a) = \emptyset$ has not yet been eliminated.

Let $a_0 \in \Omega$ and $a \in \mathcal{A}$. Then $\|1 - aa_0^{-1}\| = \|(a_0 - a)a_0^{-1}\| \leq \|a - a_0\||a_0^{-1}\|$. Therefore, for any $r \in (0, 1)$,

$$\|a - a_0\| \leq r\|a_0^{-1}\|^{-1} \Rightarrow \|1 - aa_0^{-1}\| \leq r \Rightarrow aa_0^{-1} \in \Omega.$$ 

Since $\Omega$ is closed under multiplication, $a = (aa_0^{-1})a_0 \in \Omega$. This shows that for all $a_0 \in \Omega$, the open ball of radius $\|a_0^{-1}\|^{-1}$ an center $a_0$ is contained in $\Omega$. In particular, $\Omega$ is open.

Now recall that a function $F$ from a Banach space $X$ to itself is Frechet differentiable at $x_0 \in X$ in case there is a continuous linear transformation $L$ from $X$ to itself such that for all $x \in X$,

$$\|F(x_0 + x) - F(x_0) - Lx\| = o(\|x\|),$$

and in this case, $L$ is unique and is the Frechet derivative of $F$ at $x_0$. We now show that the inverse function $a \mapsto a^{-1}$ is Frechet differentiable at every $a_0 \in \mathcal{A}$, and that the derivative is the linear transformation

$$a \mapsto -a_0^{-1}aa_0^{-1}.$$

This is a simple consequence of an important identity that we record in a lemma:

**1.13 Lemma** (First resolvent identity). Let $\mathcal{A}$ be a Banach algebra with an identity $1$. Let $\Omega$ be the set of invertible elements. For all $a, b \in \Omega$,

$$a^{-1} - b^{-1} = a^{-1}(b - a)b^{-1}.$$  \hspace{1cm} (1.9)

*Proof.* Simply expand the right hand side. \hfill $\square$

Now suppose that $a_0, a_0 + a \in \Omega$. Then

$$(a_0 + a)^{-1} - a_0^{-1} = -(a_0 + a)^{-1}aa_0^{-1} = -a_0^{-1}aa_0^{-1} + [a_0^{-1} - (a_0 + a)^{-1}]aa_0^{-1}.$$ 

By (1.1) once more and the continuity proved above,

$$\|a_0^{-1} - (a_0 + a)^{-1}aa_0^{-1}\| \leq \|a_0^{-1}\|\|a_0^{-1} - (a_0 + a)^{-1}\||a\| = o(\|a\|).$$

We are now ready to show that for all $a$ in any Banach algebra, $\sigma_{\mathcal{A}}(a) \neq \emptyset$. Let $\varphi$ be any continuous linear functional on $\mathcal{A}$, regarded as a Banach space. Such functionals exist (and are plentiful) by the Hahn-Banach Theorem. Define a complex valued function $f$ in the resolvent set $\rho_{\mathcal{A}}(a)$ by

$$f(\zeta) = \varphi((\zeta 1 - a)^{-1}).$$

Note that the resolvent set includes $\{\zeta : |\zeta| > \|a\|\}$, and that by (1.8),

$$\lim_{\zeta \to \infty} f(\zeta) = 0.$$  \hspace{1cm} (1.10)
Next, by the identity (1.9),
\[ f(\zeta + \eta) - f(\zeta) = \eta \varphi[((\zeta + \eta)1 - a)^{-1})(\zeta1 - a)^{-1}] . \]
From this identity and the continuity of the inverse function, it follows that
\[ \lim_{\eta \to 0} \frac{f(\zeta + \eta) - f(\zeta)}{\eta} = \varphi[(\zeta1 - a)^{-2}] , \]
which shows that \( f \) is an analytic function on \( \rho_{\mathcal{A}}(a) \).

If the resolvent set \( \rho_{\mathcal{A}}(a) \) were all of \( \mathbb{C} \), \( f \) would be an entire analytic function, and on account of (1.10), \( f \) would also be bounded. By Liouville’s Theorem it would then be constant, and by (1.10), the constant would have to be zero. In particular, we would have \( f(0) = 0 \). Therefore, for every continuous linear functional \( \varphi \) on \( \mathcal{A} \), it would be the case that \( \varphi(a^{-1}) = 0 \). This contradicts the Hahn-Banach Theorem. We summarize:

1.14 THEOREM. Let \( \mathcal{A} \) be any Banach algebra with an identity \( 1 \). Then for all \( a \in \mathcal{A} \), \( \sigma_{\mathcal{A}}(a) \) is a nonempty closed set contained in the closed disc of radius \( \|a\| \) centered at 0 in \( \mathbb{C} \).

It is now a simple matter to prove:

1.15 THEOREM (Gelfand-Mazur Theorem). Let \( \mathcal{A} \) be a Banach algebra with identity \( 1 \). If \( \mathcal{A} \) is a division algebra, then \( \mathcal{A} \) is isomorphic to \( \mathbb{C} \). More specifically, each element \( a \) of \( \mathcal{A} \) satisfies \( a = \lambda 1 \) for some necessarily unique \( \lambda \in \mathbb{C} \), and \( a \mapsto \lambda \) is an isomorphism with \( \mathbb{C} \).

Proof. Suppose that \( \mathcal{A} \) is a division algebra. By Theorem 1.14, there exists \( \lambda \in \sigma_{\mathcal{A}}(a) \). Thus \( \lambda 1 - a \) is not invertible. Since the only non-invertible element in a division algebra is 0, \( a = \lambda 1 \).

1.16 DEFINITION (Spectral radius). The spectral radius of an element \( a \) of a Banach algebra \( \mathcal{A} \) is
\[ \nu(a) = \max \{ |\lambda| : \lambda \in \sigma_{\mathcal{A}}(a) \} . \]

1.17 THEOREM (Gelfand’s formula). The spectral radius of an element \( a \) of a Banach algebra \( \mathcal{A} \) is given by
\[ \nu(a) = \lim_{n \to \infty} \|a^n\|^{1/n} . \]
In particular, the limit exists.

Proof. Suppose that \( \lambda \in \sigma_{\mathcal{A}}(a) \). Then by the Spectral Mapping Lemma, \( \lambda^n \in \sigma_{\mathcal{A}}(a^n) \), and then by Theorem 1.14 and (1.7) in case \( \mathcal{A} \) lacks an identity, \( |\lambda|^n \leq \|a^n\| \). Taking the \( n \)th root, we obtain \( \nu(a) \leq \|a^n\|^{1/n} \) for all \( n \in \mathbb{N} \).

It remains to show that
\[ \limsup_{n \to \infty} \|a^n\|^{1/n} \leq \nu(a) . \]
To this end, pick \( \lambda \) with \( |\lambda| > \nu(a) \) so that \( \lambda \in \rho_{\mathcal{A}}(a) \). Let \( \varphi \) be any continuous linear functional on \( \mathcal{A} \). Then, as in the proof of Theorem 1.14, the function \( f \) defined by \( f(\lambda) = \varphi((\lambda 1 - a)^{-1}) \) is analytic on \( \rho_{\mathcal{A}}(a) \). Define \( \zeta = 1/\lambda \), \( g(\zeta) = f(1/\lambda) = \zeta \varphi((1 - \zeta a)^{-1}) \), which is analytic on the open disc about 0 with radius \( 1/\nu(a) \).
For $\zeta$ with $|\zeta| < \|a\|^{-1}$, $(1 - \zeta a)^{-1}$ has the convergent power series $(1 - \zeta a)^{-1} = \sum_{n=0}^{\infty} \zeta^n a^n$. Therefore, by the uniqueness of the power series representation, $g(z) = \sum_{n=0}^{\infty} \zeta^{n+1} \varphi(a^n)$ is a convergent power series for all $\zeta$ with $|\zeta| \leq 1/\nu(a)$. It follows that for all such $\zeta$, $\lim_{n \to \infty} \zeta^{n+1} \varphi(a^n) = 0$. In particular, there exists a finite constant $C_\varphi$ such that

$$|\zeta^{n+1} \varphi(a^n)| \leq C_\varphi \quad \text{for all } n \in \mathbb{N}. \quad (1.14)$$

Now, for each $n \in \mathbb{N}$ define a linear functional $\Lambda_n$ on $\mathcal{A}^*$, the Banach space dual of $\mathcal{A}$, by

$$\Lambda_n(\varphi) = \zeta^{n+1} \varphi(a^n).$$

Then (1.15) says that

$$\sup_{n \in \mathbb{N}} \{ |\Lambda_n(\varphi)| \} \leq C_\varphi. \quad (1.15)$$

The Uniform Boundedness Principle then implies that there exists a finite constant $M$ such that

$$|\zeta|^{n+1} |\varphi(a^n)| \leq M \quad \text{for all } n \in \mathbb{N}.$$

The Hahn-Banach Theorem provides $\varphi \in \mathcal{A}^*$ with $\|\varphi\| = 1$ such that $\varphi(a^n) = \|a^n\|$. Hence we have $|\zeta|^{n+1} \|a^n\| \leq M$. Taming the $n$th root of both sides,

$$|\zeta| \|a^n\|^{1/n} \leq \left( \frac{M}{|\zeta|} \right)^{1/n}.$$

This proves that $|\zeta| \limsup_{n \to \infty} \|a^n\|^{1/n} \leq 1$. However, $\zeta$ was any complex number with $|\zeta| < 1/\nu(a)$, this proves (1.13). \hfill \Box

We close this section with the following results that is trivial for commutative Banach algebras, and familiar for the algebra of $n \times n$ matrices.

1.18 THEOREM (Spectrum of $ab$ and $ba$). If $\mathcal{A}$ is a Banach algebra, then for all $a, b \in \mathcal{A}$,

$$\{0\} \cup \sigma_{\mathcal{A}}(ab) = \{0\} \cup \sigma_{\mathcal{A}}(ba). \quad (1.16)$$

Proof. By passing to $\widetilde{\mathcal{A}}$, we may suppose that $\mathcal{A}$ has an identity. For each $\lambda \neq 0$, we must show that $(\lambda 1 - ab)$ is invertible if and only if $(\lambda 1 - ba)$ is invertible. Dividing through by $\lambda$, we may take $\lambda = 1$. Therefore, suppose that $(1 - ab)$ is invertible, and let $z = (1 - ab)^{-1}$. Then

$$(1 - ba)(1 + bza) = 1 - ba + bza - babza = 1 - ba + b(1 - ab)za = 1 - ba + ba = 1.$$

Likewise, $(1 + bza)(1 - ba) = 1$, and so $(1 - ba)$ is invertible with inverse $(1 + bza)$. \hfill \Box

1.19 THEOREM (Spectral Contraction Theorem). Let $\mathcal{A}$ and $\mathcal{B}$ be Banach algebras, and let $\pi : \mathcal{A} \to \mathcal{B}$ be a homomorphism. Then for all $a \in \mathcal{A}$,

$$\sigma_{\mathcal{B}}(\pi(a)) \subset \{0\} \cup \sigma_{\mathcal{A}}(a). \quad (1.17)$$
Proof. Adjoin identities to $\mathcal{A}$ and $\mathcal{B}$, and define $\tilde{\pi} : \mathcal{A} \to \mathcal{B}$ by $\tilde{\pi}(1, a) = (1, \pi(a))$. This is a homomorphism, and takes the identity in $\mathcal{A}$ to the identity in $\mathcal{B}$. Since adjoining an identity had no effect on non-zero spectrum, we may assume that $\mathcal{A}$ and $\mathcal{B}$ have identities 1$_{\mathcal{A}}$ and 1$_{\mathcal{B}}$ respectively, and that $\pi(1_{\mathcal{A}}) = 1_{\mathcal{B}}$.

Now suppose that $\lambda \in \rho_{\mathcal{B}}(a)$. Then $1_{\mathcal{B}} = (\lambda 1_{\mathcal{A}} - a)(\lambda 1_{\mathcal{A}} - a)^{-1}$. Since $\pi$ is a homomorphism,

$$1_{\mathcal{B}} = \pi(1_{\mathcal{A}}) = (\lambda 1_{\mathcal{B}} - \pi(a))\pi((\lambda 1_{\mathcal{A}} - a)^{-1}) .$$

Thus $\pi((\lambda 1_{\mathcal{A}} - a)^{-1})$ is a right inverse of $\lambda 1_{\mathcal{B}} - \pi(a)$, and the same reasoning shows it is also a left inverse. Hence $\lambda \in \rho_{\mathcal{B}}(\pi(a))$. This shows that $\rho_{\mathcal{B}}(a) \subset \rho_{\mathcal{B}}(\pi(a))$, which is equivalent to the statement $\sigma_{\mathcal{B}}(\pi(a)) \subset \sigma_{\mathcal{A}}(a)$, and even shows that when $\mathcal{A}$ and $\mathcal{B}$ have identities and $\pi$ takes the identity in $\mathcal{A}$ to that in $\mathcal{B}$, it is not necessary to adjoin $\{0\}$ on the right side in (1.17). \hfill $\square$

### 1.4 Characters and the Gelfand Transform

#### 1.20 Definition (Characters). A character of a Banach algebra $\mathcal{A}$ is a non-zero algebraic homomorphism from $\mathcal{A}$ to $\mathbb{C}$. The set of characters of $\mathcal{A}$ is denoted $\Delta(\mathcal{A})$, and the set $\{0\} \cup \Delta(\mathcal{A})$ is denoted $\Delta'(\mathcal{A})$.

Though characters are defined with respect to the algebraic structure alone, they are necessarily continuous:

#### 1.21 Lemma. If $\mathcal{A}$ is a Banach algebra and $\varphi$ is a character of $\mathcal{A}$, then $\varphi(a) \in \sigma_{\mathcal{A}}(a)$, and

$$|\varphi(a)| \leq \|a\| \tag{1.18}$$

for all $a \in \mathcal{A}$. That is $\varphi$ is a contraction from $\mathcal{A}$ to $\mathbb{C}$. Moreover, if $\mathcal{A}$ has an identity 1, then $\varphi(1) = 1$.

Proof. Suppose first that $\mathcal{A}$ contains an identity 1. We first prove the final statement. Since $\varphi(1) = \varphi(1^2) = (\varphi(1))^2$, $\varphi(1)$ solves $\zeta - \zeta^2 = 0$, so either $\varphi(1) = 0$ or $\varphi(1) = 1$. But if $\varphi(1) = 0$, then for all $a \in \mathcal{A}$, $\varphi(a) = \varphi(1a) = \varphi(1)\varphi(a) = 0$, and this is excluded by the definition. Hence $\varphi(1) = 1$.

Next, for any $a \in \mathcal{A}$, $\varphi(a)1 - a$ is not invertible, and hence $\varphi(a) \in \sigma_{\mathcal{A}}(a)$. To see this, note that $\varphi(\varphi(a)1 - a) = 0$, but if $\varphi(a)1 - a$ had even a right inverse $b$, we would have

$$1 = \varphi(1) = \varphi((\varphi(a)1 - a)b) = 0\varphi(b) = 0 .$$

Then since $\sigma_{\mathcal{A}}(a)$ is contained in the closed centered disc of radius $\|a\|$, (1.18) is proved.

Now suppose that $\mathcal{A}$ lacks a unit. Let $\mathcal{A}'$ be the algebra obtained by adjoining an identity, and let $\tilde{\varphi}$ be the character on $\mathcal{A}'$ given by

$$\tilde{\varphi}(\lambda, a) = \lambda + \varphi(a) ,$$

which is easily seen to be a character. Since $\sigma_{\mathcal{A}}(a) = \sigma_{\mathcal{A}'}((0, a))$ be definition, and $\tilde{\varphi}((0, a)) = \varphi(a)$, it follows from the above that $\varphi(a) \in \sigma(a)$, and then that $\|\varphi(a)\| \leq \|(0, a)\| = \|a\|$. \hfill $\square$
Note that if $\varphi \in \Delta(\mathcal{A})$, then for all $a, b \in \mathcal{A}$,
$$\varphi(ab) = \varphi(a)\varphi(b) = \varphi(b)\varphi(a) = \varphi(ba).$$
Consequently, $\varphi(ab - ba) = 0$ for all $a, b$. When the algebra $\mathcal{A}$ is not commutative, this can be a stringent constraint, and there may not exist any characters at all.

**1.22 Example.** Let $\mathcal{A}$ be the algebra of $2 \times 2$ matrices. The Pauli matrices are
\[\sigma_1 = \begin{bmatrix} 0 & 1 \\ 1 & 0 \end{bmatrix}, \quad \sigma_2 = \begin{bmatrix} 0 & -i \\ i & 0 \end{bmatrix} \text{ and } \sigma_3 = \begin{bmatrix} 1 & 0 \\ 0 & -1 \end{bmatrix}.\]
Then with $[a, b]$ denoting the commutator $ab - ba$,
\[[\sigma_1, \sigma_2] = i2\sigma_3, \quad [\sigma_2, \sigma_3] = i2\sigma_1 \text{ and } [\sigma_3, \sigma_1] = i2\sigma_2.\]
It follows that for any homomorphism $\varphi$ of $\mathcal{A}$ into $\mathbb{C}$, $\varphi(\sigma_j) = 0$ for $j = 1, 2, 3$. Next, the identity matrix $I$ satisfies $I = \sigma_1^2$, and so $\varphi(I) = (\varphi(\sigma_1))^2 = 0$. Thus, for all $(z_0, z_1, z_2, z_3) \in \mathbb{C}^4$,
$$\varphi(z_0I + z_1\sigma_1 + z_2\sigma_2 + z_3\sigma_3) = 0.$$
Since every evidently $\{I, \sigma_1, \sigma_2, \sigma_3\}$ is linearly independent and $\mathcal{A}$ is 4 dimensional, $\mathcal{A}$ is the span of $\{I, \sigma_1, \sigma_2, \sigma_3\}$, and hence $\varphi$ vanishes identically on $\mathcal{A}$. Thus, if $\mathcal{A}$ is the algebra of $2 \times 2$ matrices, $\Delta(\mathcal{A}) = \emptyset$ and $\Delta'(\mathcal{A})$ is the one-point space $\{0\}$.

Even when $\mathcal{A}$ is commutative, there may be no non-trivial characters. However, as we shall see in the next chapter, when $\mathcal{A}$ is a commutative $C^*$ algebra, characters are plentiful enough to justify our present considerations. In the rest of this chapter, commutativity of the algebras will not play any role in the proofs, and so we shall state the results without making any reference to commutativity. However, one should keep in mind that without commutativity, and even with commutativity alone, $\Delta(\mathcal{A})$ may be empty and $\Delta'(\mathcal{A})$ may be a one-point space, as in the previous example.

**1.23 Definition (Gelfand topology).** For a Banach algebra $\mathcal{A}$, the *Gelfand topology* on $\Delta'(\mathcal{A})$ is the relative weak-$*$ topology on $\Delta'(\mathcal{A})$ considered as a subset of $\mathcal{A}^*$, the Banach space dual to $\mathcal{A}$. That is, the Gelfand topology is the weakest topology on $\Delta'(\mathcal{A})$ that makes the functions $\varphi \mapsto \varphi(a)$ continuous for all $a \in \mathcal{A}$.

**1.24 Lemma.** Let $\mathcal{A}$ be a Banach algebra. Then $\Delta'(\mathcal{A})$, equipped with the Gelfand topology is a compact Hausdorff space. If $\mathcal{A}$ does not have an identity, then with the Gelfand topology, $\Delta(\mathcal{A})$ is a locally compact Hausdorff space, and $\Delta'(\mathcal{A})$ is its one-point compactification. If $\mathcal{A}$ has an identity, $\Delta(\mathcal{A})$ itself is compact and 0 is an isolated point in $\Delta'(\mathcal{A})$.

*Proof.* Equip $\mathcal{A}^*$ with the weak-$*$ topology; i.e., the weakest topology making all of functions $\varphi \mapsto \varphi(a)$ continuous for all $a \in \mathcal{A}$. The Banach-Alaoglu Theorem asserts that the unit ball in $\mathcal{A}^*$ is compact in the weak-$*$ topology. For each $a, b \in \mathcal{A}$, define a function $f_{a,b}$ on $\mathcal{A}^*$ by
$$f_{a,b}(\varphi) = \varphi(ab) - \varphi(a)\varphi(b).$$
This is evidently continuous for the weak-* topology. Now note that
\[ \Delta'(\mathcal{A}) = \bigcap_{a,b \in \mathcal{A}} \{ \varphi \in \mathcal{A}^* : f_{a,b}(\varphi) = 0 \} . \]
This displays \( \Delta'(\mathcal{A}) \) as an intersection of closed sets. Hence \( \Delta'(\mathcal{A}) \) is a closed subset of the unit ball in \( \mathcal{A}^* \), and hence is compact.

For \( \varphi_1, \varphi_2 \in \Delta'(\mathcal{A}) \) with \( \varphi_1 \neq \varphi_2 \), there exists \( a \in \mathcal{A} \) such that \( \varphi_1(a) \neq \varphi_2(a) \). Let \( U_1 \) and \( U_2 \) be disjoint open sets in \( C \) that contain \( \varphi_1(a) \) and \( \varphi_2(a) \) respectively. Then \( \{ \psi \in \Delta'(\mathcal{A}) : \psi(a) \in U_1 \} \) and \( \{ \psi \in \Delta'(\mathcal{A}) : \psi(a) \in U_2 \} \) are disjoint open sets in \( \Delta'(\mathcal{A}) \) that contain \( \varphi_1 \) and \( \varphi_2 \) respectively.

In particular, for each \( \varphi \in \Delta(\mathcal{A}) \), there disjoint open neighborhoods \( V_1 \) of \( \varphi \) and \( V_2 \) of \( 0 \), and then since \( V_1 \subset V_2^c \), \( V_2^c \) is a compact neighborhood of \( \varphi \). Thus, \( \Delta(\mathcal{A}) \) is locally compact. If \( \mathcal{A} \) has an identity \( 1 \), \( \varphi(1) = 1 \) for all \( \varphi \in \Delta(\mathcal{A}) \), while \( 0(1) = 0 \). Consequently, the zero homomorphism is an isolated point of \( \Delta'(\mathcal{A}) \) in this case.

**1.25 DEFINITION** (Gelfand transform). Let \( \mathcal{A} \) be a Banach algebra. The **Gelfand transform** is the map \( \gamma \) from \( \mathcal{A} \) to \( C(\Delta'(\mathcal{A})) \) given by
\[
(\gamma(a))[\varphi] = \varphi(a) .
\]
That is, \( \gamma(a) \) is the function of evaluation at \( a \), and it is continuous by the definition of the Gelfand topology.

**1.26 THEOREM.** Let \( \mathcal{A} \) be a Banach algebra. The Gelfand transform is a norm reducing homomorphism from \( \mathcal{A} \) to \( C(\Delta'(\mathcal{A})) \). That is, the Gelfand transform is a homomorphism of algebras and for all \( a \in \mathcal{A} \),
\[
\|\gamma(a)\|_{C(\Delta'(\mathcal{A}))} \leq \|a\| .
\]

*Proof.* The homomorphism property is evident since for all \( a, b \in \mathcal{A} \) and all \( \varphi \in \Delta'(\mathcal{A}) \),
\[
(\gamma(ab))[\varphi] = \varphi(ab) = \varphi(a)\varphi(b) = (\gamma(a))[\varphi](\gamma(b))[\varphi] .
\]

Next, suppose that \( \mathcal{A} \) has an identity \( 1 \). If \( \varphi \in \Delta(\mathcal{A}) \) and \( a \in \mathcal{A} \), then
\[
\varphi(\varphi(a)1 - a) = \varphi(a) - \varphi(a) = 0 ,
\]
and so \( \varphi(a)1 - a \) is not invertible. This means that \( \varphi(a) \in \sigma_\mathcal{A}(a) \), and this is contained in the closed centered disc of radius \( \nu(a) \leq \|a\| \).

If \( \mathcal{A} \) lacks an identity, adjoin an identity to form \( \tilde{\mathcal{A}} \). For \( \varphi \in \Delta(\mathcal{A}) \), define \( \tilde{\varphi} \) on \( \tilde{\mathcal{A}} \) by
\[
\tilde{\varphi}(\lambda, a) = \lambda + \varphi(a) .
\]
It is easy to check that \( \tilde{\varphi} \in \Delta(\tilde{\mathcal{A}}) \). Let \( e = (1, 0) \) denote the identity in \( \tilde{\mathcal{A}} \). Then for all \( a \in \mathcal{A} \),
\[
\tilde{\varphi}(\varphi(a)e - (0, a)) = \varphi(a) - \varphi(a) = 0 ,
\]
so that once again, we have that \( \varphi(a) \in \sigma_\mathcal{A}(a) \).
This result, as it stands, does not take us far at all. The problem is that at this level of generality, there may be no characters at all, and the transform may be a trivial homomorphism into a trivial algebra. As indicated above, characters can only be expected to be plentiful for commutative algebras. Even then, there may be too few characters for the Gelfand transform to be of interest. However, a fundamental theorem of Gelfand and Naimark says that for commutative $C^*$-algebras, the Gelfand transform is an isometric isomorphism. This is explained in the next chapter. We close this chapter with some examples, and then an important theorem on characters in a commutative Banach algebra.

1.27 EXAMPLE. Let $a_0$ be the $n \times n$ matrix, $n > 1$, with

$$a_{i,j} = \begin{cases} 1 & j = i + 1 \\ 0 & j \neq i + 1 \end{cases}.$$  

That is $a_0$ is the $n \times n$ matrix with 1 in every entry just above the diagonal, and zero elsewhere. It is easy to see that $a_0^2 = 0$.

Let $\mathcal{A}$ denote that subalgebra of the $n \times n$ matrices that are polynomials in $a_0$. That is, every $a \in \mathcal{A}$ has the form

$$a = \sum_{j=0}^{n-1} p_j a_0^j,$$  \hspace{1cm} (1.20)

where higher order terms are zero. This is a commutative algebra with an identity. Let $\varphi \in \Delta'(\mathcal{A})$. Then $0 = \varphi(a_0^n) = (\varphi(a_0))^n$ so that $\varphi(a_0) = n$. Then for $a$ given by (1.20), $\varphi(a) = p_0 \varphi(I) = p_0$. Thus, the only candidate for a character on $\mathcal{A}$ is the map $\varphi_0$ given by

$$\varphi_0 \left( \sum_{j=0}^{n-1} p_j a_0^j \right) = p_0 ,$$

It is readily checked that this is indeed a homomorphism and it is non-zero. Hence $\Delta(\mathcal{A}) = \{ \varphi_0 \}$ and $\Delta'(\mathcal{A}) = \{ \varphi_0 \} \cup \{ 0 \}$. Since $\Delta'(\mathcal{A})$ consists of two isolated points, we may identify $\mathcal{C}(\Delta'(\mathcal{A}))$ with $\mathbb{C}^2$ in the usual way, and then we may write the Gelfand transform as

$$\gamma \left( \sum_{j=0}^{n-1} p_j a_0^j \right) = (p_0, 0) ,$$

which is indeed a norm reducing homomorphism, but not very interesting.

Before leaving this example, we note that for elements of $\mathcal{A}$, the spectrum is as trivial as Theorem 1.14 allows: For all $a \in \mathcal{A}$, $\sigma_{\mathcal{A}}(a)$ consists of a single point: $\sigma_{\mathcal{A}}(a) = \{ \varphi_0(a) \}$. This is true since when $a$ is given by (1.20), then $a$ is invertible if and only if $p_0 \neq 0$.

Finally, note that while the algebra of all $n \times n$ matrices equipped with the usual norm is a $C^*$ algebra, this subalgebra is not closed under the Hermitian adjoint, and hence is not a $C^*$ algebra.

1.28 EXAMPLE. This example illustrates not what can go wrong when $\mathcal{A}$ is not a commutative $C^*$ algebra, but what the utility of these considerations might be when $\mathcal{A}$ is a commutative $C^*$ algebra.
Let $a_0$ be any $n \times n$ normal matrix; i.e., $a_0a_0^* = a_0^*a_0$. Let $\mathcal{A}$ be the algebra of all polynomials in $a_0$ and $a_0^*$. (We may unambiguously evaluate a polynomial $p(\eta, \zeta)$ in two the variables $\eta, \zeta$ at $\eta = a_0$ and $\zeta = a_0^*$ precisely because $a_0$ and $a_0^*$ commute.) This is a commutative algebra with an identity.

The Spectral Theorem for $n \times n$ matrices says that there exists a orthonormal basis $\{\phi_1, \ldots, \phi_n\}$ of $\mathbb{C}^n$ such that each $\phi_j$ is an eigenvector of $a_0$. Let $\lambda_j$ be the corresponding eigenvalue. That is, $a_0\phi_j = \lambda_j \phi_j$. Then $a_0^*\phi_j = \lambda_j^* \phi_j$, and especially, for any polynomial $p$,

$$p(a_0, a_0^*)\phi_j = p(\lambda_j, \lambda_j^*)\phi_j.$$ 

For each $j = 1, \ldots, n$ define a linear functional $\varphi_j$ on $\mathcal{A}$ by

$$\varphi_j(a) = \langle \phi_j, a\phi_j \rangle.$$ 

By what we have noted above, for any polynomial

$$\varphi_j(p(a_0, a_0^*)) = p(\lambda_j, \lambda_j^*).$$ 

It is evident that each $\varphi_j$ is a character, and that if $\lambda_j \neq \lambda_k$ than $\varphi_j \neq \varphi_k$.

In this case we have plenty of characters. We shall see in the next chapter that there are no other characters besides these. Granted that, $\Delta(\mathcal{A})$ can be identified with the set $\{\mu_1, \ldots, \mu_m\}$ of distinct eigenvalues of $a_0$, and the Gelfand transform identities $p(a_0, a_0^*)$ with the function on $\{\mu_1, \ldots, \mu_m\}$ given by $\mu \mapsto p(\mu, \mu^*)$. Since the operator norm of a normal matrix is the maximum of the absolute values of its eigenvalues, it is evident that the Gelfand transform is an isometry in this case.

### 1.5 Characters and spectrum in commutative Banach algebras

The Hahn-Banach Theorem, which provides the existence of continuous linear functionals on a Banach space, may be viewed as a theorem asserting the existence of maximal closed subspaces containing a given subspace. In the Banach algebra setting, the kernel of a homomorphism of a Banach algebra $\mathcal{A}$ to $\mathbb{C}$ is not only a closed subspace, it is a closed ideal, as we now explain, and consideration of maximal ideals leads to a Banach algebra version of the Hahn-Banach Theorem for commutative Banach algebras. Much of what is introduced here is also useful without assuming the $\mathcal{A}$ is commutative. We therefore start in general, and shall be clear about the key point when commutativity enters.

#### 1.29 DEFINITION. Let $\mathcal{A}$ be a Banach algebra. An ideal of $\mathcal{A}$ is a subspace of $\mathcal{A}$ such that for all $b \in \mathcal{J}$ and $a \in \mathcal{A}$, $ba \in \mathcal{J}$ and $ab \in \mathcal{J}$. An ideal of $\mathcal{A}$ is proper in case it is not equal to $\mathcal{A}$ itself. An ideal of $\mathcal{A}$ is a closed in case it is topologically closed as a subset of $\mathcal{A}$. If $\mathcal{J}$ is an ideal, an element $u$ of $\mathcal{A}$ is called a unit mod $\mathcal{J}$ in case

$$au - a \in \mathcal{J} \quad \text{and} \quad ua - a \in \mathcal{J} \quad \text{for all } a \in \mathcal{A}. \quad (1.21)$$

An ideal $\mathcal{J}$ is called a modular ideal in case there exists a unit mod $\mathcal{J}$.
Evidently if \( \mathcal{J} \) is an ideal in \( \mathcal{A} \), and \( \overline{\mathcal{J}} \) is the norm closure of \( \mathcal{J} \), then \( \overline{\mathcal{J}} \) is also an ideal in \( \mathcal{A} \).

Given a Banach algebra \( \mathcal{A} \) and an ideal \( \mathcal{J} \), there is a natural equivalence relation \( \sim \) on \( \mathcal{A} \) given by
\[
a \sim b \iff a - b \in \mathcal{J}.
\]
Let \( \{a\} \) and \( \{b\} \) denote the equivalence classes of \( a \) and \( b \) respectively. Let \( \tilde{a} \) and \( \tilde{b} \) be any other representative of \( \{a\} \) and \( \{b\} \) respectively. Then for some \( x, y \in \mathcal{J} \), \( \tilde{a} = a + x \) and \( \tilde{b} = b + y \). Then
\[
\tilde{a}\tilde{b} = (a + x)(b + y) = ab + (ay + xb + xy) \sim ab.
\]
Even more simply one sees that \( \tilde{a} + \tilde{b} \sim a + b \) and for all \( \lambda \in \mathbb{C} \), \( \lambda\tilde{a} \sim \lambda a \). Hence \( \mathcal{A}/\mathcal{J} \), the set of equivalence classes in \( \mathcal{A} \), equipped with the operations
\[
\{a\}\{b\} = \{ab\} \quad \text{and} \quad \{a\} + \{b\} = \{a + b\} \quad \text{and} \quad \lambda\{a\} = \{\lambda a\}
\]
is an algebra, and \( a \mapsto \{a\} \) is a homomorphism of \( \mathcal{A} \) onto \( \mathcal{A}/\mathcal{J} \).

Now introduce a norm on \( \mathcal{A}/\mathcal{J} \) by
\[
\|\{a\}\| = \inf \{ \|\tilde{a}\| : \tilde{a} \sim a \} = \inf \{ \|a - b\| : b \in \mathcal{J} \}.
\]
Note that \( \|\{a\}\| \leq \|a\| \). To see that
\[
\|\{a\}\{b\}\| \leq \|\{a\}\|\|\{b\}\| \quad (1.22)
\]
for all \( \{a\}, \{b\} \in \mathcal{A}/\mathcal{J} \), let \( 0 < \epsilon < \min\{\|\{a\}\|, \|\{b\}\|\} \), and pick \( \tilde{a} \in \{a\} \) and \( \tilde{b} \in \{b\} \) so that \( \{a\} > \|\tilde{a}\| - \epsilon \) and \( \{b\} > \|\tilde{b}\| - \epsilon \). Then
\[
\|\{a\}\{b\}\| = \|\tilde{a}\tilde{b}\| \leq \|\tilde{a}\|\|\tilde{b}\| \leq (\|\{a\}\| + \epsilon)(\|\{b\}\| + \epsilon).
\]
Since \( \epsilon \) can be taken arbitrarily small, (1.22) is proved.

Therefore, \( \mathcal{A}/\mathcal{J} \) will be a Banach algebra with this norm provided it is complete in this norm. Consider a Cauchy sequence \( \{\{a\}_n\}_{n \in \mathbb{N}} \) in \( \mathcal{A}/\mathcal{J} \). A standard argument shows that this sequence always has a limit if \( \mathcal{J} \) is closed. Thus, when \( \mathcal{J} \) is a closed ideal, \( \mathcal{A}/\mathcal{J} \) is a Banach algebra, and the map \( a \mapsto \{a\} \) is a contractive homomorphism of \( \mathcal{A} \) onto \( \mathcal{A}/\mathcal{J} \). This homomorphism is called the natural homomorphism of \( \mathcal{A} \) onto \( \mathcal{A}/\mathcal{J} \).

It is possible for \( \mathcal{A}/\mathcal{J} \) to have an identity even when \( \mathcal{A} \) does not. Suppose that \( \mathcal{J} \) is modular, and that \( u \) is a unit mod \( \mathcal{J} \). Then for all \( a \in \mathcal{A} \), \( \{u\}\{a\} = \{ua\} = \{a\} \) and \( \{a\}\{u\} = \{au\} = \{a\} \). Thus, \( \{u\} \) is a multiplicative identity in \( \mathcal{A}/\mathcal{J} \). Clearly if \( \mathcal{A} \) has an identity \( 1 \), \( 1 \) is a unit mod \( \mathcal{J} \).

There is a close connection between closed ideals and kernels of continuous homomorphisms of Banach algebras.

1.30 PROPOSITION. Let \( \mathcal{A} \) and \( \mathcal{B} \) be Banach algebras, and let \( \pi : \mathcal{A} \to \mathcal{B} \) be a continuous homomorphism. Then \( \mathcal{J} = \ker(\pi) \) is a closed ideal in \( \mathcal{A} \). Conversely, if \( \mathcal{J} \) is a closed ideal in \( \mathcal{A} \), then the map \( a \mapsto \{a\} \mathcal{J} \), sending \( a \) to its equivalence class mod \( \mathcal{J} \), is a homomorphism of \( \mathcal{A} \) onto \( \mathcal{A}/\mathcal{J} \).
Proof. Suppose that $\pi : \mathcal{A} \to \mathcal{B}$ be a continuous homomorphism. Then evidently $\mathcal{J} = \ker(\mathcal{A})$ is a closed by the continuity of $\pi$, and it is a subspace by the linearity of $\pi$. Next, for all $x \in \mathcal{J}$ and $a, b \in \mathcal{A}$, $\pi(AXB) = \pi(a)\pi(x)\pi(b) = \pi(a)0\pi(b) = 0$. Hence $AXB \in \ker(\pi)$, and so $\mathcal{J}$ is an ideal. The converse is clear from the construction of $\mathcal{A} / \mathcal{J}$ described above.

Now consider a commutative Banach algebra $\mathcal{A}$. For any $x_0 \in \mathcal{A}$, we can define $\mathcal{J}(x_0)$ to be the subset of $\mathcal{A}$ given by

$$\mathcal{J}(x_0) = \{ \, yx_0 : y \in \mathcal{A} \, \}.$$  \hfill (1.23)

Then for all $yx_0 \in \mathcal{J}(x_0)$ and all $a, b \in \mathcal{A}$, $ax_0b = (ay)x_0 \in \mathcal{J}(x_0)$, and evidently $\mathcal{A}$ is a subspace of $\mathcal{J}$. Hence $\mathcal{J}(x_0)$ is an ideal, and it is called the ideal generated by $x_0$.

In the non-commutative setting, one could consider the set $\{ yx_0 : y, z \in \mathcal{A} \}$ which is closed under left and right multiplication by elements of $\mathcal{A}$. However, without some additional hypothesis on $x_0$, such as that $x_0$ commutes with all elements of $\mathcal{A}$, it need not be a subspace, and the closure of its span might be all of $\mathcal{A}$.

Let $\mathcal{A}$ be a commutative Banach algebra with identity $1$. Let $x_0$ be a non-invertible element of $\mathcal{A}$. Let $\mathcal{J}(x_0)$ be the ideal generated by $x_0$. Then no element of $\mathcal{J}(x_0)$ is invertible. Indeed, if $x_0y$ were invertible, there would exist $z \in \mathcal{A}$ such that $(x_0y)z = x_0(zy) = 1$, and then (since $\mathcal{A}$ is commutative), $zy$ would be an inverse of $x_0$, which is not possible. Hence, for all non-invertible $x_0$, $\mathcal{J}(x_0)$ consists entirely of non-invertible elements. Since the open unit ball about the identity consists of invertible elements, $\mathcal{J}(x_0)$ does not intersect the open unit ball about the identity $1$.

In particular, $1$ does not belong to $\overline{\mathcal{J}(x_0)}$, the closure of $\mathcal{J}(x_0)$.

Now we come to a crucial construction of characters in a commutative Banach algebra:

1.31 THEOREM. Let $\mathcal{A}$ be a commutative Banach algebra with identity $1$. Then for all non-invertible $x_0 \in \mathcal{A}$, there exists a character $\varphi \in \Delta(\mathcal{A})$ such that $\varphi(x_0) = 0$.

Proof. Since $x_0$ is not invertible, $\mathcal{J}(x_0)$ is a proper ideal in $\mathcal{A}$, and in fact, as explained above, the open unit ball about $1$ does not intersect $\mathcal{J}(x_0)$. Now consider any chain of proper ideals in $\mathcal{A}$, ordered by inclusion. Since no proper ideal contains the identity, the union of this chain is again a proper ideal. Hence by Zorn’s Lemma, there exists a maximal proper ideal $\mathcal{M}$ containing $\mathcal{J}(x_0)$. Since no proper ideal can contain any invertible elements, this ideal does not intersect the open unit ball about $1$. Hence its closure $\overline{\mathcal{M}}$ also contains $\mathcal{J}(x_0)$ and is proper. Since $\mathcal{M}$ is maximal among such ideals, $\mathcal{M} = \overline{\mathcal{M}}$. Hence in a commutative Banach algebra $\mathcal{A}$ with identity $1$, for each non-invertible $x_0 \in \mathcal{A}$, there exists a closed proper ideal $\mathcal{M}$ that contains any ideal in $\mathcal{A}$ that contains $\mathcal{J}(x_0)$.

We now claim that the Banach algebra $\mathcal{B} = \mathcal{A} / \mathcal{M}$ is a division algebra. Suppose not. Then it contains a non-zero, non-invertible element $\{y_0\}_{\mathcal{M}}$. Let $\mathcal{N}$ be the closure of the ideal in $\mathcal{B}$ generated by $\{y_0\}_{\mathcal{M}}$. Let $\pi_1$ be the natural homomorphism of $\mathcal{A}$ onto $\mathcal{B}$, and let $\pi_2$ be the natural homomorphism of $\mathcal{B}$ onto $\mathcal{B} / \mathcal{N}$. Then $\pi_2 \circ \pi_1$ is a homomorphism of $\mathcal{A}$ onto $\mathcal{B} / \mathcal{N}$. By Proposition 1.30, $\ker(\pi_2 \circ \pi_1)$ is a closed ideal that contains $\mathcal{M} = \ker(\pi_1)$. The containment is proper since $\pi_2 \circ \pi_1(y_0) = 0$, but $y_0 \notin \mathcal{M}$. Finally, $1 \notin \ker(\pi_1 \circ \pi_2)$ since $\{1\}_{\mathcal{M}}$ is a unit $\mathcal{B} = \mathcal{A} / \mathcal{M}$, and $\mathcal{N}$ does not contain any invertible elements, so $\pi_2(\{1\}_{\mathcal{M}}) = \pi_2(\pi_1(1)) \neq 0$. Thus, $\ker(\pi_2 \circ \pi_1)$ is a closed proper ideal that strictly contains $\mathcal{M}$, which is impossible. Hence the hypothesis that $\mathcal{B} = \mathcal{A} / \mathcal{M}$ contains a non-zero, non-invertible element is false. This shows that
$\mathcal{B} = \mathcal{A} / \mathcal{M}$ is a division algebra, and then the Gelfand-Mazur Theorem tells us that $\mathcal{B} = \mathcal{A} / \mathcal{M}$ is canonically isomorphic to $\mathbb{C}$. Hence $\pi_1$ may be regarded as a character of $\mathcal{A}$, and by construction $x_0 \in \mathcal{J}(x_0) \subset \mathcal{M} = \ker(\pi_1)$.

This theorem has the following important consequence:

1.32 COROLLARY. Let $a \in \mathcal{A}$, where $\mathcal{A}$ is a commutative Banach algebra. Let $\lambda \in \sigma_{\mathcal{A}}(a)$. Then there exists $\varphi \in \Delta(\mathcal{A})$ such that $\varphi(a) = \lambda$. In particular, the spectral radius $\nu(a)$ of $a$ is given by

$$\nu(a) = \sup \{ |\varphi(a)| : \varphi \in \Delta(\mathcal{A}) \}.$$  \hfill (1.24)

Proof. Adjoining an identity if need be has no effect on the spectral radius, so we may assume that $\mathcal{A}$ has an identity $1$. We have already seen that for all $\varphi \in \Delta(\mathcal{A})$, $\varphi(a) \in \sigma_{\mathcal{A}}(a)$. We now show that for every $\lambda \in \sigma_{\mathcal{A}}(a)$, there exists $\varphi \in \Delta(\mathcal{A})$ with $\varphi(a) = \lambda$.

Since $\lambda 1 - a$ is not invertible, by Theorem 1.31, there exists $\varphi \in \Delta(\mathcal{A})$ such that $\varphi(\lambda 1 - a) = 0$. But $\varphi(\lambda 1 - a) = \lambda \varphi(1) - \varphi(a) = \lambda - \varphi(a)$.

2 The Spectral Theorem for $C^*$ Algebras

Let $\mathcal{A}$ be a $C^*$ algebra. The involution $*$ allows us to define certain classes of elements in $\mathcal{A}$:

2.1 DEFINITION (Self-adjoint, normal and unitary elements of a $C^*$ algebra). Let $\mathcal{A}$ be a $C^*$ algebra. Then:

(i) $a \in \mathcal{A}$ is self-adjoint in case $a = a^*$.

(ii) $a \in \mathcal{A}$ is normal in case $aa^* = a^*a$.

(iii) In case $\mathcal{A}$ has an identity $a$, $a \in \mathcal{A}$ is unitary in case $a^*a = a^*a = 1$.

This definition generalizes these notions from the basic example in which $\mathcal{A}$ is the algebra of $n \times n$ matrices or the bounded operators on some Hilbert space $\mathcal{H}$.

2.2 LEMMA. Let $\mathcal{A}$ be a $C^*$ algebra with an identity $1$. Then $1$ is self adjoint and $\|1\| = 1$. Moreover, for any unitary $u \in \mathcal{A}$, $\|u\| = 1$.

Proof. $1^* = 1^*1$. Applying the involution $1^* = 1^*1$, showing that $1$ is self adjoint. The next two parts use the strong condition on the norm in a $C^*$ algebra, which is that for all $a \in \mathcal{A}$, $\|a^*a\| = \|a\|^2$. We use this first in

$$\|1\| = \|1^2\| = \|1^*1\| = \|1\|^2,$$

where the second equality is true since $1 = 1^*$. Thus $\|1\| = 1$ or $\|1\| = 0$. The latter is impossible. Finally, if $u$ is unitary, $1 = \|1\| = \|u^*u\| = \|u\|^2$, so that $\|u\| = 1$.

2.3 THEOREM (In a $C^*$ algebra, self-adjointness implies real spectrum). Let $\mathcal{A}$ be a $C^*$ algebra, and let $a \in \mathcal{A}$ then if $a = a^*$, $\sigma_{\mathcal{A}}(a) \subset (\mathbb{R})$.  \hfill $\square$
Proof. It is no loss of generality to assume that \( \mathcal{A} \) has an identity since we may adjoin one if need be without any effect on the spectrum apart from possibly adjoining 0 to it. Therefore, suppose that \( \mathcal{A} \) has an identity but contains some self-adjoint element \( a \) with some \( \lambda \in \sigma_{\mathcal{A}}(a) \) such that \( \lambda \notin \mathbb{R} \). Then taking an appropriate real multiple of \( a \) (so that the multiple is still self-adjoint), we may suppose that \( e^{i\lambda} = 2 \) for some \( \lambda \in \sigma_{\mathcal{A}}(a) \).

For each \( n \in \mathbb{N} \), define the polynomial \( p_n(\zeta) = \sum_{j=0}^{n} (i\zeta)^j/j! \). By the Spectral Mapping Lemma, for each \( n \), \( p_n(\lambda) \in p_n(a) \). For \( n > m \),

\[
\|p_n(a) - p_m(a)\| \leq \sum_{j=m+1}^{n} \|a\|^j/j!,
\]

and hence by standard estimates on the exponential power series for numbers, \( \{p_n(a)\}_{n \in \mathbb{N}} \) is a Cauchy sequence in \( \mathcal{A} \). Therefore, there exist \( u \in \mathcal{A} \) with \( u = \lim_{n \to \infty} p_n(a) \). Evidently, for all \( n \), \( (p_n(a))^* = p_n(-a) \), so that once again, by standard estimates for the exponential power series, \( u^*u = uu^* = 1 \); that is, \( u \) is unitary, and by Lemma 2.2, \( \|u\| = 1 \). Therefore, for all \( \mu \) with \( |\mu| > 1 \), \( \mu 1 - u \) is invertible. But

\[
\lim_{n \to \infty} (p_n(\lambda)1 - p_n(a)) = e^{i\lambda}1 - u.
\]

Since \( |e^{i\lambda}| = 2 \), \( e^{i\lambda}1 - u \) is invertible. However, \( p_n(\lambda)1 - p_n(a) \) is non-invertible for each \( n \). Since the set of invertible elements is open, it cannot be that a sequence of non-invertible elements converges to an invertible element. Thus, the hypothesis that \( \mathcal{A} \) contains some self-adjoint element \( a \) with some \( \lambda \in \sigma_{\mathcal{A}}(a) \) such that \( \lambda \notin \mathbb{R} \) leads to contradiction. \( \square \)

2.4 DEFINITION (Hermitian character). Let \( \mathcal{A} \) be a \( C^* \) algebra. A character \( \varphi \) of \( \mathcal{A} \) is Hermitian in case for all \( a \in \mathcal{A} \),

\[
\varphi(a^*) = (\varphi(a))^*.
\]

2.5 L EMMA. All characters of a \( C^* \) algebra are Hermitian.

Proof. For any \( a \in \mathcal{A} \) define \( x = \frac{1}{2}(a + a^*) \) and \( y = \frac{1}{2i}(a - a^*) \). Then \( x \) and \( y \) are self-adjoint, and \( a = x + iy \). For any character \( \varphi \) of \( \mathcal{A} \),

\[
\varphi(a) = \varphi(x + iy) = \varphi(x) + i\varphi(y) \quad \text{and} \quad \varphi(a^*) = \varphi(x - iy) = \varphi(x) - i\varphi(y).
\]

By Theorem 2.3, \( \varphi(x) \) and \( \varphi(y) \) are real, and hence \( \varphi(a^*) = (\varphi(a))^* \). \( \square \)

The next theorem again makes use of the strong condition on the norm in a \( C^* \) algebra, which is that for all \( a \in \mathcal{A} \), \( \|a^*a\| = \|a\|^2 \).

2.6 THEOREM (Norm and spectral radius in a \( C^* \) algebra). Let \( \mathcal{A} \) be a \( C^* \) algebra. Then for all \( a \in \mathcal{A} \),

\[
\|a\|^2 = \nu(a^*a)
\]

and if \( a \) is normal,

\[
\|a\| = \nu(a) .
\]
Proof. Suppose first that $a$ is normal. Then $(a^*a) = (a^*)^*(a^2)$. Then by the $C^*$-algebra identity $\|b^*b\| = \|b\|^2$ applied twice, and the isometry property of the involution,

$$\|a\|^2 = \|a\|^2\|a\|^2 = \|a^*a\|^2 = \|(a^*a)\| = \|(a^*)^*(a^2)\| = \|a^2\|^2.$$ 

Therefore, $\|a^2\| = \|a\|^2$, and by an obvious induction, for all $m \in \mathbb{N}$, $\|a^{2m}\| = \|a\|^{2m}$. Then by Gelfand’s Formula,

$$\nu(a) = \lim_{m \to \infty} (\|a^{2m}\|)^{1/2m} = \|a\|.$$

This proves (2.1). Next, for any $a \in \mathcal{A}$, $a^*a$ is self-adjoint and so $\nu(a^*a) = \|a^*a\|$. Then since $\mathcal{A}$ is a $C^*$ algebra, $\|a^*a\| = \|a\|^2$, and this proves (2.2).

2.7 THEOREM (Commutative Gelfand-Naimark Theorem). Let $\mathcal{A}$ be a commutative $C^*$-algebra. Then the Gelfand transform is an isometric isomorphism of $\mathcal{A}$ onto $\mathcal{C}_0(\Delta(\mathcal{A}))$.

Proof. By Lemma 2.5, for all $a \in \mathcal{A}$ and all $\varphi \in \Delta(\mathcal{A})$,

$$\gamma(a^*)[\varphi] = \varphi(a^*) = (\varphi(a))^* = \gamma(a)[\varphi]$$

since the involution in $\mathcal{C}_0(\Delta(\mathcal{A}))$ is pointwise complex conjugation.

Next, $|\gamma(a)[\varphi]| = |\varphi(a)|$. By the easy Lemma 1.21, $\sup_{\varphi \in \Delta(\mathcal{A})} \{ |\varphi(a)| \} \leq \nu(a)$. By the deeper Corollary 1.32 of Theorem 1.31, $\sup_{\varphi \in \Delta(\mathcal{A})} \{ |\varphi(a)| \} \geq \nu(a)$. Combining these two inequalities with Theorem 2.6, and noting that in a commutative $C^*$ algebra, every element is normal,

$$\sup_{\varphi \in \Delta(\mathcal{A})} \{ |\gamma(a)[\varphi]| \} = \nu(a) = \|a\|,$$

which proves that the Gelfand transform is an isometry, and hence is injective onto a subalgebra of $\gamma(\mathcal{A})$ of $\mathcal{C}_0(\Delta(\mathcal{A}))$. However, $\gamma(\mathcal{A})$ separates points, and does not vanish at any $\varphi \in \Delta(\mathcal{A})$, and is closed under complex conjugation. Hence by the Stone-Wierstrass Theorem, and the closure of $\gamma(\mathcal{A})$, $\gamma(\mathcal{A}) = \mathcal{C}_0(\Delta(\mathcal{A}))$. $\blacksquare$

2.1 Spectral invariance and the Abstract Spectral Theorem

Let $\mathcal{A}$ be a Banach algebra with identity, and let $\mathcal{B}$ be a Banach subalgebra. It can happen that some $b \in \mathcal{B}$ is not invertible in $\mathcal{B}$, but is invertible in $\mathcal{A}$.

2.8 EXAMPLE. Let $D$ denote the closed unit disc in $\mathbb{C}$, and let $C$ denote its boundary, the unit circle. Let $\mathcal{A} = \mathcal{C}(C)$, the algebra of continuous functions on $C$. Let $\mathcal{B}$ denote the algebra of continuous functions on $D$ that are holomorphic in the interior of $D$. These functions are determined by their values on $C$, and their maximum absolute value is attained on $C$. Therefore, restriction to $C$ is an isometric embedding of $\mathcal{B}$ in $\mathcal{A}$, so we may regard $\mathcal{B}$ as a subalgebra of $\mathcal{A}$.

Let $b$ denote the function $f(e^{i\theta}) = e^{i\theta}$, the identity function on $C$, which evidently belongs to $\mathcal{B}$. Then $1 - b$ is invertible in $\mathcal{A}$ if and only if $\lambda \not\in \mathcal{C}$, in which case the inverse is the function $g(e^{i\theta}) = (\lambda - e^{i\theta})^{-1}$. However, for $\lambda$ in the interior of $D$, $\zeta \mapsto (\lambda - \zeta)^{-1}$ is not holomorphic in the interior of $D$, and so the inverse of $b$ in $\mathcal{A}$ does not belong to $\mathcal{B}$. That is $\sigma_{\mathcal{A}}(b) = C$, but $\sigma_{\mathcal{B}}(b) = D$. 

Now we specialize to $C^*$ algebras, first introducing certain minimal subalgebras:

### 2.9 Definition

Let $\mathcal{A}$ be a $C^*$ algebra with unit $1$. For all $b \in \mathcal{A}$, $C(b)$ is the smallest $C^*$ subalgebra of $\mathcal{A}$ that contains $b$ and $1$.

### 2.10 Theorem

Let $\mathcal{A}$ be a $C^*$ algebra with unit $1$, and let $b \in \mathcal{A}$. If $b$ is invertible in $\mathcal{A}$, $b^{-1} \in C(b)$, and hence $b$ is invertible in every $C^*$ subalgebra of $\mathcal{A}$ that contains $1$ and $b$. In particular, if $\mathcal{B}$ is a $C^*$ subalgebra of $\mathcal{A}$ that contains $1$ and $b$, then

$$\sigma_\mathcal{B}(b) = \sigma_\mathcal{A}(b).$$

**Proof.** Suppose first that $b$ is self adjoint and invertible in $\mathcal{A}$. By Theorem 2.3, $\sigma_{C(b)}(b) \subset \mathbb{R}$, and consequently, for all $n \in \mathbb{N}$, $b - (i/n)1$ is invertible in $C(b)$. Since $\lim_{n \to \infty}(b - (i/n)1) = b$ in $\mathcal{A}$, and since the inverse is continuous, $\lim_{n \to \infty}(b - (i/n)1)^{-1} = b^{-1}$ in $\mathcal{A}$. But since $(b - (i/n)1)^{-1} \in C(b)$ for all $n$, and since $C(b)$ is closed,

$$b^{-1} = \lim_{n \to \infty}(b - (i/n)1)^{-1} \in C(b).$$

Hence, $b$ is invertible within $C(b)$.

Now let $b$ be any invertible element of $\mathcal{A}$. Then $b^*$ and $b^*b$ are invertible in $\mathcal{A}$, and also belong to $C(b)$. Since $b^*b$ is self adjoint, what we have proved above says that $(b^*b)^{-1} \in C(b)$. Define $x = (b^*b)^{-1}b^* \in C(b)$. Evidently $xb = 1$. Thus, $b$ has a left inverse in $C(b)$. The same argument shows that $y = b^*(bb^*)^{-1}$ is a well defined right inverse of $b$ in $C(b)$, and then $x = x(by) = (xb)y = y$ so $x = y$ is the inverse of $b$ in $C(b)$. In particular, for all $\lambda \in C$, $\lambda 1 - b$ is invertible in $C(b)$ if and only if it is invertible in $\mathcal{A}$. Thus, $\lambda 1 - b$ is invertible in $\mathcal{A}$ if and only if it is invertible in $C(b)$, and this proves the final statement. \(\square\)

### 2.11 Lemma

Let $\mathcal{A}$ be a $C^*$ algebra with identity $1$, and let $a \in \mathcal{A}$ be normal. Then the map $\varphi \mapsto \varphi(a)$ is a homeomorphism of $\Delta(C(a))$ onto $\sigma_\mathcal{A}(a)$.

**Proof.** Since $a$ and $a^*$ commute, the closure of the linear span of $\{a^m(a^*)^n : m, n \geq 0\}$ is a $C^*$ algebra that contains $1$ and $a$. Evidently, it is $C(a)$. Hence if $\varphi \in \Delta(C(a))$, $\varphi$ is determined by its values on $a$ and $a^*$. In fact, since $\varphi$ is necessarily Hermitian, $\varphi$ is determined by its value on $a$. That is, for any $\varphi, \psi \in \Delta(C(a))$,

$$\varphi = \psi \iff \varphi(a) = \psi(a).$$

We have also seen that for all $\varphi \in \Delta(C(a))$, $\varphi(a) \in \sigma_{C(a)}(a) = \sigma_\mathcal{A}(a)$, and for all $\lambda \in \sigma_\mathcal{A}(a) = \sigma_{C(a)}(a)$, there is a $\varphi_\lambda \in \Delta(C(a))$ such that $\varphi_\lambda(a) = \lambda$. This shows that the map $\varphi \mapsto \varphi(a)$ is a one-to-one map of $\Delta(C(a))$ onto $\sigma_\mathcal{A}(a)$. This map is also continuous by the definition of the Gelfand topology, and continuous bijections between compact spaces are homeomorphisms. \(\square\)

We now come to the Abstract Spectral Theorem:

### 2.12 Theorem (Abstract Spectral Theorem)

Let $\mathcal{A}$ be a $C^*$ algebra with identity $1$, and let $a \in \mathcal{A}$ be normal. Then identifying $\mathcal{C}_{\Delta(C(a))}$ and $\mathcal{C}(\sigma_\mathcal{A}(a))$ through the homeomorphism provided by Lemma 2.11, we may regard the Gelfand transform as a homomorphism of $C(a)$ into $\mathcal{C}(\sigma_\mathcal{A}(a))$. Then the Gelfand transform $\gamma$ is an isometric isomorphism of $C(a)$ onto $\mathcal{C}(\sigma_\mathcal{A}(a))$. For all non-negative integers $m, n$, $\gamma(a^n(a^*)^m)$ is the function on $\sigma_\mathcal{A}(a)$ given by

$$\lambda \mapsto \lambda^m(\lambda^*)^n.$$  \(2.3\)
Proof. The Commutative Gelfand-Naimark Theorem says that $\gamma$ is an isometric isomorphism, and if $\varphi \in \Delta(C(a))$,  
$$
\gamma(a^m(a^*)^m)[\varphi] = \varphi(a)^m((\varphi(a))^*)^n = \lambda^m(\lambda^n)^n
$$
for $\lambda = \varphi(a)$ so that under the identification provided by Lemma 2.11, $\gamma(a^m(a^*)^m)$ is indeed given by (2.3).

2.13 DEFINITION. For $\mathcal{A}$ a $C^*$ algebra with identity 1, a a normal element of $\mathcal{A}$, and $f \in \mathcal{C}(\sigma_{\mathcal{A}}(a))$, $f(a)$ is defined by $\gamma^{-1}(f)$; i.e., $f(a)$ is the inverse image of $f$ under the isometric isomorphism of $C(a)$ onto $\mathcal{C}(\sigma_{\mathcal{A}}(a))$ that is provided by the Commutative Gelfand Naimark Theorem.

2.14 THEOREM (Spectral Mapping Theorem). Let $\mathcal{A}$ be a $C^*$ algebra with identity 1, $a$ a normal element of $\mathcal{A}$, and $f \in \mathcal{C}(\sigma_{\mathcal{A}}(a))$. Then  
$$
\sigma_{\mathcal{A}}(f(a)) = f(\sigma_{\mathcal{A}}(a)) .
$$

Proof. For all $\mu \in \mathbb{C}$, the function $\lambda \mapsto \mu - f(\lambda)$ is invertible in $\mathcal{C}(\sigma_{\mathcal{A}}(a))$ if and only if $\mu$ does not belong to the range of $f$, which is $f(\sigma_{\mathcal{A}}(a))$. Then, using the isomorphism provided by the Commutative Gelfand Naimark Theorem, we see that $\mu 1 - f(a)$ is invertible in $C(a)$ if and only if $\mu \notin f(\sigma_{\mathcal{A}}(a))$, and hence $\sigma_{C(a)}(f(a)) = \sigma_{\mathcal{A}}(a)$. Finally, by Theorem 2.10, the spectrum of $f(a)$ is the same in all $C^*$ subalgebras of $\mathcal{A}$ that contain $f(a)$ and 1. In particular, $\sigma_{C(a)}(f(a)) = \sigma_{cA}(f(a))$.

2.2 Continuity of the spectrum and the functional calculus

2.15 THEOREM (Newburgh’s Theorem). Let $\mathcal{A}$ be a Banach algebra and $a \in \mathcal{A}$. Let $U$ be an open subset of $\mathbb{C}$ with $\sigma_{\mathcal{A}}(a) \subset U$. Then there exists a $\delta > 0$ such that if $\|b - a\| \leq \delta$,  
$$
\sigma_{\mathcal{A}}(b) \subset U .
$$

Proof. First note that for all $b \in \mathcal{A}$ with $\|b - a\| < 1$, $\|b\| < \|a\| + 1$, and hence for all $\lambda \in \mathbb{C}$ with $\lambda \geq \|a\| + 1$, $\lambda \in \rho_{\mathcal{A}}(b)$. Hence when $\|b - a\| \leq 1$, $\sigma_{\mathcal{A}}(b)$ is contained in the closed centered disc of radius $\|a\| + 1$.

Let $K = U^c \cap \{ \lambda : |\lambda| \leq \|a\| + 1 \}$ which is a compact subset of $\rho_{\mathcal{A}}(a)$. It suffices to show that there is an $r > 0$ so that for all $\mu \in K$, $(\mu 1 - b)$ is invertible whenever $\|b - a\| < r$.

Let $\lambda \in K$. Then $\lambda \in \rho_{\mathcal{A}}(a)$, and for all $b \in \mathcal{A}$ and $\mu \in \mathbb{C}$ with  
$$
|\mu - \lambda| + \|b - a\| < \|\lambda 1 - a\|^{-1} \Rightarrow \|\mu 1 - b - (\lambda 1 - a)\| \leq \|\lambda 1 - a\|^{-1} ,
$$
and hence $\mu \in \rho_{\mathcal{A}}(b)$. For each $\lambda \in K$, define $U_\lambda = \{ \mu : |\mu - \lambda| < \frac{1}{2}\|\lambda 1 - a\|^{-1} \}$. Since $K$ is compact, there exists a finite subcover $\{U_{\lambda_1}, \ldots, U_{\lambda_n}\}$. Define  
$$
r = \min\{\frac{1}{2}\|\lambda_1 1 - a\|^{-1}, \ldots, \frac{1}{2}\|\lambda_n 1 - a\|^{-1} \} .
$$

Then for any $b$ with $\|b - a\| < r$ and any $\mu \in K$, $\mu \in U_{\lambda_j}$ for some $j = 1, \ldots, n$, and then  
$$
\|\mu 1 - b - (\lambda_j 1 - a)\| \leq |\mu - \lambda_j| + \|b - a\| < \|\lambda_j 1 - a\|^{-1} .
$$

Therefore, $(1\mu - b)$ is invertible. Thus, for all $\mu \in K$, whenever $\|b - a\| < r$, $\mu \in \rho_{\mathcal{A}}(b)$.  
\qed
Now let $\mathcal{A}$ be an Banach algebra, and let $x, y \in \mathcal{A}$, and $t \in \mathbb{R}$. Then for all $n \in \mathbb{N}$, by the telescoping sum identity
\[
(x + y)^n - x^n = \sum_{j=0}^{n-1} ((x + y)^{n-j}x^j - (x + y)^{n-j-1}x^{j+1}) = \sum_{j=0}^{n-1} (x + y)^{n-j-1}yx^j
\]
Therefore,
\[
\|(x + y)^n - x^n\| \leq \left( \sum_{j=0}^{n-1} \|(x + y)^{n-j-1}\|\|x^j\| \right) \|y\| \leq n(\|x\| + \|y\|)^n \|y\|.
\]
Therefore, when $\|y\| < \delta$, $\|(x + y)^n - x^n\| < n(\|x\| + \delta)^n \delta$, and this proves that $x \mapsto x^n$ is continuous.

In a $C^*$ algebra we can say more.

2.16 THEOREM. Let $\mathcal{A}$ be a $C^*$ algebra with identity $1$. Let $U \subset \mathbb{C}$ be open with $\overline{U}$ compact. Let $\mathcal{N}_U$ be given by
\[
\mathcal{N}_U = \{ a \in \mathcal{A} : aa^* = a^*a \text{ and } \sigma_{\mathcal{A}}(a) \subset U \}.
\]

Then $\mathcal{N}_U$ is an open subset of the normal elements of $\mathcal{A}$. Moreover, let $f$ be a continuous complex valued function on $\overline{U}$, and for all $a \in \mathcal{N}_U$, define $f(a) \in \mathcal{A}$ using the Gelfand-Naimark isomorphism. Then the map $a \mapsto f(a)$ is continuous on $\mathcal{N}_U$.

Proof. The first assertion is an immediate consequence of Newburg’s Theorem. For the second, consider any sequence $\{p_n\}$ of polynomials converging uniformly to $f$ on $\overline{U}$. Then for all $a \in \mathcal{N}_U$,
\[
\|p_n(a) - f(a)\| \leq \sup_{\lambda \in U} \{ |p_n(\lambda) - f(\lambda)| \}.
\]
That is,
\[
\lim_{n \to \infty} \left( \sup_{a \in \mathcal{N}_U} \{ \|p_n(a) - f(a)\| \} \right) = 0.
\]
Thus, the function $a \mapsto f(a)$ is the uniform limit of the continuous functions $a \mapsto p_n(a)$.

For normal elements of a $C^*$ algebra, there is a quantitative version of Newburgh’s Theorem.

2.17 THEOREM. Let $\mathcal{A}$ be a $C^*$ algebra, and let $a, x \in \mathcal{A}$ be normal. Then
\[
\sigma_{\mathcal{A}}(a + x) \subset \{ \lambda : \text{dist}(\lambda, \sigma_{\mathcal{A}}(a)) \leq \|x\| \}.
\]

Proof. Let $\lambda \in \rho_{\mathcal{A}}(a)$. By the Gelfand-Naimark isomorphism,
\[
\| (\lambda 1 - a)^{-1} \| = \sup \{ |\lambda - \mu|^{-1} : \mu \in \sigma_{\mathcal{A}}(a) \} = \frac{1}{\text{dist}(\lambda, \sigma_{\mathcal{A}}(a))}.
\]
Let $x \in \mathcal{A}$ be normal, Then $[\lambda 1 - (a+x)] - [\lambda 1 - a] = -x$. Therefore, as long as $\|x\| < \text{dist}(\lambda, \sigma_{\mathcal{A}}(a))$, $[\lambda 1 - (a+x)]$ is invertible.
2.3 Positivity in $C^*$ algebras

2.18 DEFINITION. Let $\mathscr{A}$ be a $C^*$ algebra. Then a self adjoint element $a$ in $\mathscr{A}$ is positive in case $\sigma_\mathscr{A}(a) \subset [0, \infty)$. The set of all positive elements of $\mathscr{A}$ is denoted $\mathscr{A}^+$. If $a \in \mathscr{A}^+$, we may use the Abstract Spectral Theorem to define $\sqrt{a}$, and then $a = (\sqrt{a})^2 = (\sqrt{a})^*(\sqrt{a})$. It is also true that in any $C^*$ algebra, every element of the form $b^*b$ is positive. This was not known to Gelfand and Naimark when they wrote their 1943 paper, in which they raised the question as to whether it was true or not. They included an extra hypothesis in their paper, namely that for all $b$ in $\mathscr{A}$, $1 \notin \sigma_\mathscr{A}(b^*b)$.

The fact that for all $b$ in a $C^*$ algebra $\mathscr{A}$, $b^*b \in \mathscr{A}^+$ was finally proved in 1952 and 1953 through the contributions of Fukamiya and Kaplansky. The history is interesting: Kaplansky had managed to prove that if the sum of two positive elements is necessarily positive, then $b^*b$ is necessarily positive. However, he was unable to show that $\mathscr{A}^+$ was closed under sum. He published nothing, but showed his proof to many people. When Fukamiya proved the closure of $\mathscr{A}^+$ in 1952, Kaplansky communicated his proof to the reviewer of Fukamiya’s paper for Math Reviews, and the proof was published there.

2.19 THEOREM (Fukamiya’s Theorem). Let $\mathscr{A}$ be a $C^*$ algebra. Then $\mathscr{A}^+$ is a pointed convex cone. That is:

(1) For all $\lambda \in \mathbb{R}^+$, and all $a \in \mathscr{A}^+$, $\lambda a \in \mathscr{A}^+$, and for all $a, b \in \mathscr{A}^+$, $a + b \in \mathscr{A}^+$.

(2) $-\mathscr{A}^+ \cap \mathscr{A}^+ = \{0\}$.

(The first part says that $\mathscr{A}^+$ is a convex cone; the second part says that this cone is pointed.)

Proof. We may suppose that $\mathscr{A}$ has an identity 1 since otherwise we may adjoin an identity without affecting positivity.

Let $B_\mathscr{A}$ denote the closed unit ball in $\mathscr{A}$. Fukamiya observed that that $\mathscr{A}^+ \cap B_\mathscr{A}$ consists precisely of the self-adjoint elements $a$ with both $a$ and $1 - a$ are in $B_\mathscr{A}$. To see this suppose that $a \in \mathscr{A}^+ \cap B_\mathscr{A}$. Then since $a$ is self adjoint, $\nu(a) = \|a\| \leq 1$, and so $\sigma_\mathscr{A}(a) \subset [0, 1]$. By (an easy case of) the Spectral Mapping Lemma, $\sigma_\mathscr{A}(1 - a) \subset [0, 1]$, and hence $\|1 - a\| = \nu(1 - a) \leq 1$.

Conversely, suppose $a$ is self-adjoint and both $a$ and $1 - a$ are in $B_\mathscr{A}$.

Now let $a, b \in B_\mathscr{A}$. Then by Minkowski's inequality $\|(a + b)/2\| \in B_\mathscr{A}$ and

$$\left\| 1 - \frac{a + b}{2} \right\| \leq \frac{1}{2}(\|1 - a\| + \|1 - b\|).$$

If furthermore, $a, b \in \mathscr{A}^+$, then we also have that $\|1 - a\| \leq 1$ and $\|1 - b\| \leq 1$, and then from (2.6), $\|1 - (a + b)/2\| \leq 1$. Thus, $(a + b)/2$ is self adjoint and belongs to both $B_\mathscr{A}$ and $1 - B_\mathscr{A}$, and hence $(a + b)/2 \in \mathscr{A}^+$. 

Since the closure of $\mathcal{A}^+$ under positive multiples is clear, it then clear that $\mathcal{A}^+$ is closed under sums. Finally, if $a \in \mathcal{A}^+$ and $-a \in \mathcal{A}^+$, then $\sigma_{\mathcal{A}}(a) \subset (-\infty,0] \cap [0,\infty) = \{0\}$, so that $\|a\| = \nu(a) = 0$, and hence $a = 0$. □

2.20 THEOREM (Fukamiya-Kaplansky Theorem). For all $a \in \mathcal{A}$, a $C^*$ algebra, $a^*a \in \mathcal{A}^+$.

Proof. We first show that if $a^*a \in -\mathcal{A}^+$, then $a^*a = 0$. Since $a^*a$ and $aa^*$ have the same spectrum, Fukamiya’s Theorem says that $a^*a + aa^* \in -\mathcal{A}^+$. However, writing $a = x + iy$ with $x$ and $y$ self adjoint,

$$a^*a + aa^* = 2(x^2 + y^2) \in \mathcal{A}^+$$

where once again we have used Fukamiya’s Theorem, and the Spectral Mapping Lemma. Since $\mathcal{A}^+$ is a pointed cone, this means that $a^*a + aa^* = 0$. But then $a^*a = (a^*a + aa^*) - aa^* = -aa^* \in \mathcal{A}^+$. Again since $\mathcal{A}^+$ is pointed, this means that $a^*a = 0$, as claimed.

Now suppose that $x = b^*b$ for some $b \in \mathcal{A}$, Define continuous functions $f,g : \mathbb{R} \to \mathbb{R}$ by $f(t) = \max\{t,0\}$ and $g(t) = t - f(t)$. Note that $f(t)g(t) = 0$ for all $t$. By the Abstract Spectral Theorem, if we define $y = f(x)$ and $z = g(x)$, then $yz = 0$, and $y + z = x = b^*b$. Now define $w = bz$, Then

$$w^*w = bz^*bz = z(y + z)z = z^3.$$

Since $\sigma_{\mathcal{A}}(z) \subset (-\infty,0]$, $z^3 \in -\mathcal{A}^+$, and the first part of the proof says that $w^*w = 0$. Therefore, $z = 0$, and so $b^*b = f(b^*b) \in \mathcal{A}^+$. □

2.4 Homeomorphisms of $C^*$ algebras

2.21 THEOREM. Let $\mathcal{A}$ be a $C^*$ algebra, and let $\mathcal{I}$ be a norm-closed ideal in $\mathcal{A}$. Then $\mathcal{I}$ is closed under the involution.

The heart of the proof is the following approximation lemma:

2.22 LEMMA. Let $\mathcal{A}$ be a $C^*$ algebra, and let $\mathcal{I}$ be a norm-closed ideal in $\mathcal{A}$. Then for every $a \in \mathcal{I}$, there is a sequence $\{u_n\}_{n \in \mathbb{N}}$ of positive elements of $\mathcal{I}$ with $\|u_n\| \leq 1$ for all $n$ such that

$$\lim_{n \to \infty} \|au_n - a\| = 0.$$

Proof. Consider the sequence of continuous functions $f_n : \mathbb{R}_+ \to \mathbb{R}_+$ given by $f_n(t) = \min\{nt,1\}$. Note that

$$t(1 - f_n(t))^2 = \begin{cases} t(1 - nt)^2 & t \leq 1/n \\ 0 & t > 1/n. \end{cases}$$

Evidently $\sup_{t \geq 0} |\{t(1 - f_n(t))^2\}| \leq 1/n$. Consequently, by the Abstract Spectral Theorem applied to $a^*a$ for any $a \in \mathcal{A}$, $|(f_n(a^*a) - 1)a^*(f_n(a^*a) - 1)| \leq 1/n$. Note that

$$\|(f_n(a^*a) - 1)a^*(f_n(a^*a) - 1)\| = \|af_n(a^*a) - a\|^2.$$

Thus, $\lim_{n \to \infty} \|af_n(a^*a) - a\| = 0$. By the Abstract Spectral Theorem, $\|f_n(a^*a)\| \leq 1$ and $f_n(a^*a) \in \mathcal{A}^+$ for all $n$. It remains to show that when $a \in \mathcal{I}$, then $f_n(a^*a) \in \mathcal{I}$ for all $n$. Clearly when $a \in \mathcal{I}$, $a^*a$ and all polynomials in $a^*a$ belong to $\mathcal{I}$. But then since $f_n$ may be uniformly approximated by polynomials, and since $\mathcal{I}$ is norm closed, $f_n(a^*a) \in \mathcal{I}$. □
Proof of Theorem 2.21. Let \( a \in \mathcal{J} \), and let \( \{u_n\}_{n \in \mathbb{N}} \) be a sequence of positive elements of \( \mathcal{J} \) such that \( \lim_{n \to \infty} \|au_n - a\| = 0 \). Since \( \|u_n a^* - a^*\| = \|au_n - a\| \), and \( u_n a^* \in \mathcal{J} \), \( \lim_{n \to \infty} \|u_n a^* - a^*\| = 0 \). Then since \( \mathcal{J} \) is closed, \( a^* \in \mathcal{J} \).

Now let \( \mathcal{A} \) be a C* algebra, and let \( \mathcal{J} \) be a closed ideal in \( \mathcal{J} \). As usual, let \( \{a\} \) denote equivalence class of \( a \) mod \( \mathcal{J} \), and let \( \|\{a\}\| \) denote the quotient norm of \( \{a\} \); that is,

\[
\|\{a\}\| = \inf \{ \|a - b\| : b \in \mathcal{J} \}.
\]

Then \( \mathcal{A}/\mathcal{J} \) is a Banach algebra with the quotient space norm. By Theorem 2.21, \( a - b \in \mathcal{J} \iff a^* - b^* \in \mathcal{J} \), and therefore we may define an involution on \( \mathcal{J} / \mathcal{A} \) by

\[
\{a\}^* = \{a^*\}.
\]

Evidently this involution is an isometry, and so for all \( a \in \mathcal{A} \), \( \|\{a\}^*\{a\}\| \leq \|\{a\}\|^2 \). To show that \( \mathcal{A}/\mathcal{J} \) is a C* algebra with this involution, we need only show that for all \( a \in \mathcal{A} \),

\[
\|\{a\}\|^2 \leq \|\{a\}^*\{a\}\|.
\]

We shall use the following lemma:

**2.23 LEMMA.** Let \( \mathcal{A} \) be a C* algebra, and let \( \mathcal{J} \) be a closed ideal in \( \mathcal{A} \). For all \( a \in \mathcal{A} \), the quotient norm of \( \{a\} \) is given by

\[
\|\{a\}\| = \inf \{ \|a - au\| : u^* = u \in \mathcal{J} \text{ and } u \in \mathcal{A}^+ \cap B_{\mathcal{A}} \}.
\]

*Proof.* Whenever \( u \in \mathcal{J} \), \( a \sim (a - au) \) so that \( \|\{a\}\| \) is no greater than the right hand side of (2.8). To prove the equality, pick \( \epsilon > 0 \) and \( b \in \mathcal{J} \) so that \( \|\{a\}\| \geq \|a - b\| - \epsilon \). Then by (2.5), \( \|1 - u\| \leq 1 \), and so

\[
\|a - b\| \geq \|a - b\| \|1 - u\| \geq \|(a - b)(1 - u)\| = \|(a - au) - (b - bu)\| \geq \|(a - au)\| - \|(b - bu)\|.
\]

By Lemma 2.22 we can choose \( u \) so that \( \|b - bu\| < \epsilon \). We then have \( \|\{a\}\| \geq \|a - au\| - 2\epsilon \), and since \( \epsilon > 0 \) is arbitrary, (2.8) is proved.

Now to prove (2.7), pick \( \epsilon > 0 \) and \( u = u^* \in \mathcal{J} \) with \( u \in \mathcal{A}^+ \cap B_{\mathcal{A}} \) so that

\[
\|a^*a(1 - u)\| \leq \|\{a\}^*\{a\}\| + \epsilon = \|\{a\}\|^2 + \epsilon.
\]

Then

\[
\|\{a\}\|^2 \leq \|a(1 - u)\|^2 = \|(1 - u)a^*a(1 - u)\| \leq (1 - u)\|a^*a(1 - u)\| \leq \|a^*a(1 - u)\|.
\]

where the last in equality is valid since by (2.5), \( \|1 - u\| \leq 1 \). Altogether, \( \|\{a\}\|^2 \leq \|\{a\}^*\{a\}\| + \epsilon \), and since \( \epsilon > 0 \) is arbitrary, (2.7) is proved. We have shown:

**2.24 THEOREM.** Let \( \mathcal{A} \) be a C* algebra, and let \( \mathcal{J} \) be a norm closed ideal in \( \mathcal{A} \), then \( \mathcal{J} \) is closed under the involution, and the definition \( \{a\}^* = \{a^*\} \) defines an involution on \( \mathcal{A}/\mathcal{J} \) so that, equipped with the quotient norm, \( \mathcal{A}/\mathcal{J} \) is a C* algebra.
2.25 **Lemma.** Let \( \mathcal{A} \) and \( \mathcal{B} \) be \( C^* \) algebras, and let \( \pi : \mathcal{A} \to \mathcal{B} \) be a \(*\)-homomorphism. Then \( \pi \) is a contraction; i.e., \( \|\pi(a)\| \leq \|a\| \) for all \( a \in \mathcal{A} \). If moreover \( \pi \) is one-to-one, \( \pi \) is an isometry.

**Proof.** For all \( a \in \mathcal{A} \), by the Spectral Contraction Theorem, \( \nu(\pi(a)^*\pi(a)) = \nu(a^*a) \leq \nu(a^*a) \). Then since for self adjoint elements of a \( C^* \) algebra, the norm is the spectral radius, \( \|\pi(a)^*\pi(a)\| \leq \|a^*a\| \). Then by the crucial defining property of a \( C^* \) algebra relating the norm and the involution, \( \|\pi(a)\|^2 \leq \|a\|^2 \), and this proves that \( \pi \) is a contraction.

Notice that if \( \nu(\pi(a)^*a) = \nu(a^*a) \), the argument gives \( \|\pi(a)\| = \|a\| \). Hence it remains to show that if \( \pi \) is one-to-one, \( \pi \) cannot decrease the spectral radius of any self adjoint element of \( \mathcal{A} \).

Indeed, let \( a = a^* \in \mathcal{A} \), and suppose that \( \nu(\pi(a)) < \nu(a) \). Then there is a non-zero continuous bounded function \( f \) supported on \( [-\nu(a), \nu(a)] \) that vanishes identically on \( \nu(\pi(a)), \nu(a)) \].

since \( f \) may be approximated by polynomials, \( \pi(f(a)) = f(\pi(a)) \). However, since \( f \) vanishes identically on the spectrum of \( \pi(a) \), \( f(\pi(a)) = 0 \). Thus, \( f(a) \) is in the kernel of \( \pi \), which is a contradiction. Hence, when \( \pi \) is one-to-one, it preserves the spectral radius of self adjoint elements.

We summarize with the following theorem:

2.26 **Theorem** (Homomorphisms of \( C^* \) algebras). Let \( \mathcal{A} \) and \( \mathcal{B} \) be \( C^* \) algebras, and let \( \pi : \mathcal{A} \to \mathcal{B} \) be a \(*\)-homomorphism. Then \( \pi \) is a contraction, \( \pi(\mathcal{A}) \) is a \( C^* \)-subalgebra of \( \mathcal{B} \), and \( \pi \) induces an isometric isomorphism of \( \mathcal{A}/\ker(\pi) \) onto \( \pi(\mathcal{A}) \).

2.5 **Projections in \( C^* \) algebras**

2.27 **Definition.** Let \( \mathcal{A} \) be a \( C^* \) algebra. A self adjoint element \( e \) of \( \mathcal{A} \) is a projection in case \( e^2 = e \). A projection \( e \) is a central projection in case \( e \) commutes with every element of \( \mathcal{A} \).

Note that 0 is a projection, as is 1 when \( \mathcal{A} \) has an identity. Any other projections, should they exist, are non-trivial projections. Suppose that \( e \) is a non-trivial projection in \( \mathcal{A} \). Then \( 1 - e \) is also a non-trivial projection in \( \mathcal{A} \).

Associated to \( e \) are the two subalgebras, namely \( e\mathcal{A}e \) and \( (1 - e)\mathcal{A}(1 - e) \), where \( e\mathcal{A}e \) consists of all elements of \( \mathcal{A} \) of the form \( eae \), \( a \in \mathcal{A} \), and likewise \( (1 - e)\mathcal{A}(1 - e) \) consists of all elements of \( \mathcal{A} \) of the form \( (1 - e)a(1 - e) \), \( a \in \mathcal{A} \). Evidently, these are both \( C^* \) subalgebras of \( \mathcal{A} \). Note that \( e \) is the identity in \( e\mathcal{A}e \), and \( (1 - e) \) is the identity in \( (1 - e)\mathcal{A}(1 - e) \).

The name “corner algebra” come from the case in which \( \mathcal{A} = M_n(\mathbb{C}) \), the algebra of \( n \times n \) complex matrices, and for some \( 1 \leq m \leq m - 1 \), \( e \) is the orthogonal projection onto the subspace of \( \mathbb{C}^n \) consisting of vectors \( (\eta_1, \ldots, \eta_n) \) such that \( \eta_j = 0 \) for \( j \geq m + 1 \). The general element of \( a \in \mathcal{A} \) can then be written in “block form” as \( a = \begin{bmatrix} x & y \\ z & w \end{bmatrix} \) where \( x \in M_m(\mathbb{C}), \, w \in M_{n-m}(\mathbb{C}) \) and \( y \) and \( z^T \) are \( m \times (n - m) \) matrices. Then

\[
eae = \begin{bmatrix} x & 0 \\ 0 & 0 \end{bmatrix} \quad \text{and} \quad (1 - e)a(1 - e) = \begin{bmatrix} 0 & 0 \\ 0 & y \end{bmatrix}.
\]
Continuing with this example, note that if \( a = eae + (1 - e)a(1 - e) \) so that \( a = \begin{bmatrix} x & 0 \\ 0 & w \end{bmatrix} \), then \( a \) is invertible if and only if both \( x \) and \( z \) are invertible, and so for such matrices \( a \),
\[
\sigma_{M_n(\mathbb{C})}(a) = \sigma_{M_m(\mathbb{C})}(x) \cup \sigma_{M_{n-m}(\mathbb{C})}(w) .
\]

We shall be especially interested in the case in which \( e \) is central. Note that this is not the case in the example we just considered.

If \( e \) commutes with every element of \( eA \), then \( eAe = \{ ea : a \in A \} = \{ ae : a \in A \} \), and \( eAe \) is then evidently an ideal in \( A \), as is \( (1 - e)A(1 - e) \). In this case,
\[
A = eAe \oplus (1 - e)A(1 - e)
\]

since for all \( a \in A \), \( a = ae + a(1 - e) \) and if \( a \in eAe \cap (1 - e)A(1 - e) \), \( a = ea \) and \( a = (1 - e)a \) so that \( a = e(1 - e)a = 0 \).

For all \( a \in A \), \( a \) is invertible if and only if \( eae \) and \( (1 - e)a \) is invertible in \( (1 - e)A(1 - e) \). To see this, suppose that \( a \) is invertible in \( A \). Then \( e = eaa^{-1} = e^2aa^{-1} = (ea)(ea^{-1}) = (ea^{-1})(ea) \), and so the inverse of \( ea \) in \( eAe \) is \( ea^{-1} \). The same reasoning shows that the inverse of \( (1 - e)a \) in \( (1 - e)A(1 - e) \) is \( (1 - e)a^{-1} \). For the converse, suppose that \( ea \) has the inverse \( ex \) in \( eAe \) and that \( (1 - e)a \) has the inverse \( (1 - e)y \) in \( (1 - e)A(1 - e) \). Then
\[
\begin{align*}
a(ex + (1 - e)y) &= (ae + a(1 - e))(ex + (1 - e)y) = ae^2x + a(1 - e)^2y \\
&= (ea)(ex) + (1 - e)a(1 - e)y = e + (1 - e) = 1,
\end{align*}
\]

thus showing that \( (ex + (1 - e)y) \) is a right inverse of \( a \). A similar computation shows that it is a left inverse. This leads to the following result:

2.28 THEOREM. Let \( A \) be a C* algebra with identity 1, and let \( e \) be a central projection in \( A \). For any \( a \in A \), let \( a = ea + (1 - e)a \) be the unique decomposition of \( a \) corresponding to \( A = eAe \oplus (1 - e)A(1 - e) \). Then
\[
\sigma_A(a) = \sigma_{eAe}(ea) \cup \sigma_{(1 - e)A(1 - e)}(a).
\]

Proof. For all \( \lambda \in \mathbb{C} \), \( \lambda 1 - a = (\lambda e - ea) + (\lambda (1 - e) - (-1)e)a \). By the remarks above the theorem, \( \lambda \in \rho_A(a) \) if and only if \( \lambda \in \rho_{eAe}(ea) \cap \rho_{(1 - e)A(1 - e)}((1 - e)a) \).

3 Lin’s Theorem

3.1 Almost commuting and nearly commuting

In 1995, Huaxin Lin proved a theorem that settled an old conjecture arising from the work of John von Neumann on quantum mechanics. His theorem concerns the C* algebra \( M_n(\mathbb{C}) \) of complex \( n \times n \) matrices:

3.1 THEOREM (Lin’s Theorem). For every \( \epsilon > 0 \), there is a \( \delta > 0 \) such that for any \( n \in \mathbb{N} \) and any pair of self-adjoint \( a, b \in M_n(\mathbb{C}) \) with \( \|a\|, \|b\| \leq 1 \) and
\[
\|ab - ba\| \leq \delta ,
\]
there exists a commuting pair of self adjoint \( \tilde{a}, \tilde{b} \in M_n(\mathbb{C}) \) such that
\[
\|a - \tilde{a}\| + \|b - \tilde{b}\| \leq \epsilon . \tag{3.2}
\]

When a pair \( a, b \in M_n(\mathbb{C}) \) satisfies (3.1) for small \( \delta \), we may say that they almost commute. When a pair \( a, b \in M_n(\mathbb{C}) \) is such that there exists a commuting pair \( \tilde{a}, \tilde{b} \in M_n(\mathbb{C}) \) such that (3.2) is satisfied for small \( \epsilon \), we may say that \( a \) and \( b \) are nearly commuting – they are near to matrices that exactly commute.

The theorem may be rephrased as a theorem about “almost normal” and “nearly normal” matrices. Let \( a, b \in M_n(\mathbb{C}) \) be self adjoint \( n \times n \) matrices. Let \( x = a + ib \). Then \( x^*x - xx^* = 2i(ab - ba) \) so that
\[
\|x^*x - xx^*\| = 2\|ab - ba\| \tag{3.3}
\]
so that \( \|x^*x - xx^*\| \leq 2\delta \).

Let \( \tilde{a} \) and \( \tilde{b} \) be another pair of self adjoint \( n \times n \) matrices, and define \( \tilde{x} = \tilde{a} - \tilde{b} \). Note that \( x - \tilde{x} = (a - \tilde{a}) + i(b - \tilde{b}) \), so that \( \|x - \tilde{x}\| \leq \|a - \tilde{a}\| + \|b - \tilde{b}\| \). However,
\[
a - \tilde{a} = \frac{(x - \tilde{x}) + (x^* - \tilde{x}^*)}{2}
\]
so that \( \|a - \tilde{a}\| \leq \|x - \tilde{x}\| \). Likewise, we have \( \|b - \tilde{b}\| \leq \|x - \tilde{x}\| \). Altogether,
\[
\|x - \tilde{x}\| \leq \|a - \tilde{a}\| + \|b - \tilde{b}\| \leq 2\|x - \tilde{x}\| . \tag{3.4}
\]

Combining (3.3) and (3.4), we arrive at an alternate formulation of Lin’s Theorem:

**3.2 THEOREM** (Lin’s Theorem, Alternate Formulation). For every \( \epsilon > 0 \), there is a \( \delta > 0 \) such that for any \( n \in \mathbb{N} \) and every \( x \in M_n(\mathbb{C}) \) with \( \|x\| \leq 1 \), and
\[
\|x^*x - xx^*\| \leq \delta , \tag{3.5}
\]
there exists a normal \( \tilde{x} \in M_n(\mathbb{C}) \) such that
\[
\|x - \tilde{x}\| \leq \epsilon . \tag{3.6}
\]

The crucial feature of Lin’s Theorem is that \( \delta \) depends only on \( \epsilon \) and not on \( n \). Without the requirement that \( \delta \) be independent of \( n \), the result is trivial. It then suffices to show that for fixed \( n \), and fixed \( \epsilon > 0 \), there does not exist a sequence \( \{x_j\}_{j \in \mathbb{N}} \) of \( n \times n \) matrices with \( \|x_j\| \leq 1 \) for all \( j \) and such that \( \lim_{j \to \infty} \|x_j^*x_j - x_jx_j^*\| = 0 \) but \( \|x_j - x\| > \epsilon \) for all normal \( x \) and all \( j \).

Suppose such a sequence exists. By the compactness of the unit ball in the space of \( n \times n \) matrices, there is a subsequence \( \{x_{j_k}\}_{k \in \mathbb{N}} \) and an \( x \) with \( \|x\| \leq 1 \) such that \( \lim_{k \to \infty} \|x_{j_k} - x\| = 0 \). Evidently
\[
x^*x - xx^* = \lim_{k \to \infty} (x_{j_k}^*x_{j_k} - x_{j_k}x_{j_k}^*) = 0 .
\]
Therefore, \( x \) is normal but \( \|x_{j_k} - x\| < \epsilon \) for all sufficiently large \( k \).
3.2 The finite spectrum problem

Lin’s proof of his theorem turns on the analysis of two $C^*$ algebras that we now define: For any sequence $\{n_j\}_{j \in \mathbb{N}}$ of natural numbers, define two sets of sequences of matrices as follows:

$$\mathcal{A} = \{ \{a_j\}_{j \in \mathbb{N}} : a_j \in M_{n_j}(\mathbb{C}) \text{ and } \sup_{j \in \mathbb{N}} \|a_j\| < \infty \}$$ (3.7)

and

$$\mathcal{J} = \{ \{b_j\}_{j \in \mathbb{N}} : b_j \in M_{n_j}(\mathbb{C}) \text{ and } \lim_{j \to \infty} \|b_j\| = 0 \}$$ (3.8)

Obviously $\mathcal{J} \subset \mathcal{A}$. Equip $\mathcal{A}$ with the operations of term by term addition and multiplication and the norm

$$\|\{a_j\}_{j \in \mathbb{N}}\| = \sup_{j \in \mathbb{N}} \|a_j\|.$$ This makes $\mathcal{A}$ a Banach algebra. Equip $\mathcal{A}$ with the involution consisting of term by term Hermitian conjugation. This makes $\mathcal{A}$ a $C^*$ algebra, and $\mathcal{J}$ a closed ideal in $\mathcal{A}$. Let $B$ denote the quotient algebra $\mathcal{A}/\mathcal{J}$, and let $\pi$ denote the natural homomorphism of $\mathcal{A}$ onto $B$. This notation will be used throughout this section.

The relevance of this construction is as follows: If Theorem 3.2 were false, there would exist a sequence $\{n_j\}_{j \in \mathbb{N}}$ of natural numbers, and a sequence $\{x_j\}_{j \in \mathbb{N}}$ with each $x_j \in M_{n_j}(\mathbb{C})$ such that

$$\|x_j - \tilde{x}_j\| \geq \epsilon \text{ for all } j \in \mathbb{N} \text{ and all normal } \tilde{x}_j \in M_{n_j}(\mathbb{C}),$$ (3.9)

and

$$\lim_{j \to \infty} \|x_j^*x_j - x_jx_j^*\| = 0.$$ (3.10)

Let us write $x$ to denote $\{x_j\}_{j \in \mathbb{N}}$ considered as an element of $\mathcal{A}$, and let us write $y$ to denote $\pi(x) \in B$. By (3.10), which says that $\{x_j^*x_j - x_jx_j^*\}_{j \in \mathbb{N}} \in \mathcal{J}$,

$$y^*y - yy^* = \pi(x^*)\pi(x) - \pi(x)\pi(x^*) = \pi(x_j^*x_j - x_jx_j^*) = 0.$$ Thus, for any $x = \{x_j\}_{j \in \mathbb{N}}$ satisfying (3.10), $y = \pi(x)$ is normal in $B$.

We say that an element of a Banach algebra has finite spectrum if its spectrum is a finite subset of $\mathbb{C}$. There are two parts to Lin’s proof. One is to show that every normal $y \in B$ can be approximated arbitrarily well in norm by a normal element $\tilde{y}$ that has finite spectrum. The other is to show that for any $x = \{x_j\}_{j \in \mathbb{N}} \in \mathcal{A}$, if $\pi(x)$ is normal with finite spectrum, then (3.9) is impossible. We begin with the latter point.

3.3 Lemma. Let $x = \{x_j\}_{j \in \mathbb{N}} \in \mathcal{A}$ and suppose that $y = \pi(x)$ is normal and has finite spectrum. Then there exists a normal $\tilde{x} = \{\tilde{x}_j\}_{j \in \mathbb{N}} \in \mathcal{A}$ such that $\pi(\tilde{x}) = y$, and, consequently, such that

$$\lim_{j \to \infty} \|x_j - \tilde{x}_j\| = 0.$$ (3.11)

In other words, normal equivalence classes in $\mathcal{A}/\mathcal{J}$ that have finite spectrum have a normal representative.
Proof. Let \( \{\lambda_1, \ldots, \lambda_m\} \) be the points in the spectrum of \( y \). Let \( p \) and \( q \) be complex polynomials with
\[
p(\lambda_j) = j \quad \text{and} \quad q(j) = \lambda_j \quad \text{for} \quad j = 1, \ldots, n.
\]
Notice that \( q \circ p(\lambda) = \lambda \) on \( \sigma_\mathcal{A}(y) \) so that \( q(p(y)) = y \). Since \( p(\lambda) \in \mathbb{R} \) for all \( \lambda \in \sigma_\mathcal{A}(y) \), \( p(y) \) is self adjoint.

Let \( z \) be any element of \( \mathcal{A} \) with \( \pi(z) = p(y) \). Then \( \pi(z) \) is self adjoint, and so \( \pi((z^* + z)/2) = p(y) \). Then
\[
q(\pi((z^* + z)/2)) = q(q((z^* + z)/2)) = q(p(y)) = y,
\]
Then since \((z^* + z)/2\) is self adjoint, \( q((z + z^*)/2) \) is normal, and thus the equivalence class of \( y \) contains a normal representative, namely \( q((z^* + z)/2) \), that we denote by \( \hat{x} \).

By the definition of the norm in the quotient algebra, for all \( \epsilon > 0 \), there exists \( b = \{b_j\}_{j \in \mathbb{N}} \in \mathcal{A} \) such that \( \|x - \hat{x} - b\| \leq \epsilon \). This means that
\[
\|x_j - \tilde{x}_j\| \leq \epsilon + \|b_j\|.
\]
Then since \( \epsilon > 0 \) is arbitrary and \( \lim_{j \to \infty} \|b_j\| = 0 \), (3.11) is proved. \( \square \)

### 3.3 Approximation of normals by normal elements with finite spectrum

It remains to show that every normal element of \( \mathcal{B} \) can be well-approximated by normal elements with finite spectrum. To prepare for this, we make several observations about the algebra \( \mathcal{A} \).

Consider \( x \in M_n(\mathbb{C}) \) for some \( n \in \mathbb{N} \). Then \( x \) has a singular value decomposition
\[
x = \tilde{u} s \tilde{v}^* 
\]
where \( s \) is a diagonal matrix with non-negative entries, and \( u \) and \( v \) are unitary matrices. Since \( \tilde{u}^* \tilde{u} = 1 \), \( x = \tilde{v} u^* (\tilde{u} s \tilde{v}^*) \). We define \( u = \tilde{v} u^* \) and \( |x| = \tilde{u} s \tilde{v}^* \). Then we have
\[
x = u |x| \quad \text{and} \quad |x| = \sqrt{x^* x},
\]
where the square root is defined by the functional calculus. Because \( u \) is unitary, this is called a unitary polar decomposition of \( x \).

Now we observe that every element \( x = \{x_j\}_{j \in \mathbb{N}} \in \mathcal{A} \) has a unitary polar decomposition \( x = u |x| \): Simply choose such a decomposition \( x_j = u_j |x_j| \) for each \( j \), and then \( u = \{u_j\}_{j \in \mathbb{N}} \) and \( |x| = \{|x_j|\}_{j \in \mathbb{N}} \).

Next consider any \( y \in \mathcal{B} \), and any \( x \in \mathcal{A} \) such that \( \pi(x) = y \). Let \( u |x| \) be a unitary polar decomposition of \( x \). Then
\[
y = \pi(u) \pi(|x|) = \pi(u) \pi(\sqrt{x^* x}) = \pi(u) \sqrt{y^* y} = \pi(u) |y|,
\]
and \( \pi(u) \) is unitary. Therefore, each element of \( \mathcal{B} \) has a unitary polar decomposition.

Essentially the same argument shows that every unitary \( v \in \mathcal{B} \), has a unitary representative in \( \mathcal{A} \); i.e, there exists a unitary \( u \in \mathcal{A} \) such that \( \pi(u) = v \). To see this, consider any \( x \in \mathcal{A} \) such that \( \pi(x) = y \), and and let \( x = u |x| \) be a unitary polar decomposition of \( x \). Then \( y = \pi(u) \pi(|x|) = \pi(u) |\pi(x)| = \pi(u) |y| = \pi(u) \). While we have to do significant work to obtain even an approximate normal representative for normal \( y \in \mathcal{B} \), for unitary \( v \in \mathcal{B} \), things are much simpler: There is always an exact unitary representative in \( \mathcal{A} \). This will be used below.
3.4 LEMMA. Let any \( \epsilon > 0 \) and any countable subset \( F \) of \( \mathbb{C} \) be given. Then for all normal \( y \in \mathcal{B} \), there exists a normal \( \tilde{y} \in \mathcal{B} \) such that \( \|y - \tilde{y}\| \leq \epsilon \) and \( F \cap \sigma_{\mathcal{B}}(\tilde{y}) = \emptyset \).

Proof. The set of invertible normal elements is dense in the set of normal elements of \( \mathcal{B} \). To see this let, \( y \in \mathcal{B} \) be normal and let \( y = v|y| \) be a unitary polar decomposition. Then \( |y|^2 = y^*y = yy^* = v|y|^2v^* \), which mean that \( |y|^2v = v|y|^2 \) so that \( v \) computes with \( |y|^2 \), and hence any polynomial in \( |y|^2 \), and hence any continuous function of \( |y|^2 \). In particular, \( v \) commutes with \( |y|^2 \).

Evidently, \( v(|y| + 1) \) is invertible and normal since \( v \) is unitary and commutes with \( |y| \). Clearly \( \|y - v(|y| + 1)\| \leq \epsilon \) and this justifies the claim that the set of invertible normal elements is dense in the set of normal elements of \( \mathcal{B} \).

It follows that for each \( \lambda \in F \), the set of normal elements \( z \in \mathcal{B} \) such that \( \lambda 1 - z \) is invertible is dense and open in the relative topology. By Baire’s Theorem, the intersection of these sets over all \( \lambda \in F \) is dense in the normal elements of \( \mathcal{B} \).

This lemma shall be applied to approximate an arbitrary normal \( y \in \mathcal{B} \) by another normal \( \tilde{y} \in \mathcal{B} \) where \( \sigma_{\mathcal{B}}(\tilde{y}) \) lies on the \( \epsilon \) grid \( \Gamma_\epsilon \subset \mathbb{C} \), where for \( \epsilon > 0 \),

\[
\Gamma_\epsilon = \{ s + it \in \mathbb{C} : s \in \epsilon \mathbb{Z} \quad \text{or} \quad t \in \epsilon \mathbb{Z} \}.
\]

(3.12)

To do this, fix \( \epsilon > 0 \), and let \( F \) be the set of the centers of the squares in \( \Gamma_\epsilon \). That is, the set

\[
F_\epsilon := \{ s + it \in \mathbb{C} : s+ \in \epsilon(\mathbb{Z} + 1/2) \quad \text{and} \quad t \in \epsilon(\mathbb{Z} + 1/2) \}.
\]

Let \( f \) be the obvious continuous contraction from \( \mathbb{C} \setminus F_\epsilon \) onto \( \Gamma_\epsilon \), such that for all \( \lambda \in \mathbb{C} \setminus F_\epsilon \),

\[
|f(\lambda) - \lambda| \leq \epsilon/\sqrt{2}.
\]

(3.13)

Define \( \tilde{y} = f(y) \). Then \( \tilde{y} \) is normal and by the Spectral Mapping Theorem, \( \sigma_{\mathcal{B}}(\tilde{y}) \in \Gamma_\epsilon \). By (3.13), \( \|\tilde{y} - y\| < \epsilon/\sqrt{2} \). This proves:

3.5 LEMMA. For all normal \( y \in \mathcal{B} \) there and all \( \epsilon > 0 \), there exists a normal \( \tilde{y} \in \mathcal{B} \) such that \( \sigma_{\mathcal{B}}(\tilde{y}) \subset \Gamma_\epsilon \) and \( \|y - \tilde{y}\| < \epsilon \).

Now fix \( \epsilon > 0 \) and consider any normal \( y \in \mathcal{B} \) such that \( \sigma_{\mathcal{B}}(\tilde{y}) \subset \Gamma_\epsilon \). Then since \( \sigma_{\mathcal{B}}(\tilde{y}) \) is a closed subset of \( \mathbb{C} \) contained in \( D_{||y||} \), the closed centered disc of radius \( ||y|| \) in \( \mathbb{C} \),

\[
\sigma_{\mathcal{B}}(\tilde{y}) \subset \Gamma_\epsilon \cap D_{||y||}.
\]

At this point we have that the spectrum of \( y \) lies in a subset of \( \mathbb{C} \) that looks something like the following:
Now consider two sets

\[
\Lambda_\epsilon = \left\{ s + it \in \mathbb{C} : s \in \epsilon \mathbb{Z} \text{ and } t \in \epsilon(\mathbb{Z} + \frac{1}{2}) \text{ or } s \in \epsilon \mathbb{Z} \text{ and } t \in \epsilon(\mathbb{Z} + \frac{1}{2}) \right\}
\]

(3.14)

and

\[
\tilde{\Lambda}_\epsilon = \Lambda_\epsilon = \left\{ s + it \in \mathbb{C} : s \in \epsilon \mathbb{Z} \text{ and } t \in \epsilon(\mathbb{Z} + \frac{1}{2}) \right\}.
\]

(3.15)

Note that \(\Lambda_\epsilon\) is the set of midpoints on the elementary segments of the grid \(\Gamma_\epsilon\), and \(\tilde{\Lambda}_\epsilon\) is the set of intersection points of the grid \(\Gamma_\epsilon\).

There is an obvious continuous \(g\) retraction of \(\Gamma_\epsilon \setminus \Lambda_\epsilon\) onto \(\tilde{\Lambda}_\epsilon\) such that for all \(\lambda \in \Gamma_\epsilon\),

\[|g(\lambda) - \lambda| \leq \epsilon/2.\]

Therefore, if \(y\) is any normal element of \(\mathcal{B}\) with spectrum in \(\Gamma_\epsilon \setminus \Lambda_\epsilon\), \(g(y)\) is a normal element with \(\|y - g(y)\| \leq \epsilon/2\), and \(\sigma_{\mathcal{B}}(g(y)) \subset \tilde{\Lambda}_\epsilon \cap \mathcal{D}_{\|y\|}\), a finite set whose cardinality depends only on \(\epsilon\) and \(\|y\|\).

We now turn to the lemma that will enable us to remove, one at a time, the finitely many points of \(\Lambda_\epsilon \cap \mathcal{D}_{\|y\|}\) from the spectrum of our normal element \(y\), (whose spectrum lies in \(\Gamma_\epsilon \cap \mathcal{D}_{\|y\|}\)). This will give us the approximation by elements of finite spectrum that we seek.

**3.6 Lemma.** Let \(y \in \mathcal{B}\) be normal. Let \(V\) be an open subset in \(\mathbb{C}\) such that \(V \cap \sigma_{\mathcal{B}}(y)\) is contained in a subset \(X\) of \(\mathbb{C}\) that is homeomorphic to the open unit interval. Let \(y_0 \in V \cap \sigma_{\mathcal{B}}(y)\) and suppose that \(\lambda_0\) is not an isolated point of \(\sigma_{\mathcal{B}}(y)\). Then for each \(\epsilon > 0\), there exists a normal \(\tilde{y} \in \mathcal{B}\) such that

\[
\sigma_{\mathcal{B}}(\tilde{y}) \subset \sigma_{\mathcal{B}}(y) \setminus \{\lambda_0\} \quad \text{and} \quad \|y - \tilde{y}\| < \epsilon.
\]

We preface the proof with remarks on the strategy. Suppose we can find a commutative subalgebra \(\mathcal{C}\) of \(\mathcal{B}\) that contains \(C(y)\) and a projection \(e\), necessarily central, such that for some small neighborhood \(U\) of \(\lambda_0\)

\[
\sigma_{\mathcal{C}}(ey) \subset \mathcal{U} \quad \text{and} \quad \sigma_{(1-e)\mathcal{C}}((1-e)y) \subset \sigma_{\mathcal{B}}(y) \setminus U.
\]

(3.16)

Pick any \(\lambda_1 \neq \lambda_0 \in \sigma_{\mathcal{B}}(y) \cap U\). The function \(\lambda \mapsto |\lambda_1 - \lambda|\) is bounded by \(\text{diam}(U)\) on \(\sigma_{\mathcal{C}}(ey)\). Therefore, by the Gelfand-Naimark Theorem, \(\|\lambda_1 e - ey\| \leq \text{diam}(U)\).
Define \( \tilde{y} = \lambda_1 e + (1 - e)y \) Then \( \tilde{y}_1 \) is normal, and \( \|\tilde{y} - y\| = \|\lambda_1 e - ey\| \leq \text{diam}(U) \). Finally, by Theorem 2.28 and (3.16), \( \sigma_{\mathcal{A}}(\tilde{y}) = \{\lambda_1\} \cup \sigma_{(1-e)\mathcal{B}(1-e)}((1-e)y) \subset \{\lambda_1\} \cup \sigma_{\mathcal{B}}(y) \setminus U \). Then by Theorem 2.10,

\[
\sigma_{\mathcal{B}}(\tilde{y}) \subset \{\lambda_1\} \cup \sigma_{\mathcal{B}}(y) \setminus U \subset \{\lambda_0\} \cup \sigma_{\mathcal{B}}(y) \setminus \{\lambda_0\} .
\]

The construction of \( e \) requires some ingenuity: If we could apply the characteristic function \( 1_U \) to \( y \), we would readily obtain \( e \). However, \( 1_U \) need not be continuous on \( \sigma_{\mathcal{B}}(y) \), and so the Abstract Spectral Theorem is not available. If \( y \) had finite spectrum, then of course there would be a continuous function \( f \) agreeing with \( 1_U \) on \( \sigma_{\mathcal{B}}(y) \), and then we could define \( e = f(y) \). However, \( \sigma_{\mathcal{B}}(y) \) need not have any isolated points, and then there will be no such continuous function.

We will use the fact that \( y \) is an equivalence class of sequences of matrices, represented by some \( x = \{x_j\}_{j \in \mathbb{N}} \). If each \( x_j \) were normal, we could apply the spectral theorem to define \( 1_u(x_j) \) for each \( j \), and this would provide us with a projection of the sort we seek. However, we do not know that in general that normal \( y \in \mathcal{B} \) have normal representatives – except in the special case that \( y \) is not only normal, but unitary. Therefore, we use the continuous functional calculus to convert \( y \) into a unitary, and then we work with a unitary representative of this in \( \mathcal{A} \) to produce our desired projection.

Proof. Choose a relatively open set \( U \subset X \) with

\[
\lambda_0 \in U \subset \overline{U} \subset X \quad \text{and} \quad \text{diam}(U) < \varepsilon .
\]

Let \( f_0 \) be a homeomorphism of \( X \) onto \( \mathbb{T} \setminus \{-1\} \) where \( \mathbb{T} \) is the unit circle in \( \mathbb{C} \). Extend \( f_0 \) to a continuous function \( f : \sigma_{\mathcal{B}}(y) \to \mathbb{C} \) by

\[
f(\lambda) = \begin{cases} 
 f_0(\lambda) & \lambda \in X \\
 -1 & \lambda \in \sigma_{\mathcal{B}}(y) \cap X^c
\end{cases} .
\]

Set \( v = f(y) \). Observe that \( v \) is unitary. Let \( u \) be any unitary in \( \mathcal{A} \) with \( \pi(u) = v \). Since \( f_0 \) is a homeomorphism of \( X \) onto \( \mathbb{T} \setminus \{-1\} \), and since \( U \) is open in \( X \), \( W = f_0(U) \) is open in \( \mathbb{T} \). Let \( 1_W \) denote the characteristic function of \( W \).

We now use the Spectral Theorem for \( n \times n \) matrices to define \( 1_W(u_j) \) for each \( j \in \mathbb{N} \). For each \( j \in \mathbb{N} \), \( 1_W(u_j) \) is a projection in \( M_{n_j}(\mathbb{C}) \). Therefore, \( e = \pi(\{1_W(u_j)\}_{j \in \mathbb{N}}) \) is a projection in \( \mathcal{B} \). For each \( j \in \mathbb{N} \), \( u_j 1_W(u_j) = 1_W(u_j) u_j \), and hence, \( u e = e u \).

Let \( \hat{\varphi} \) be any continuous function on \( \mathbb{T} \). Then \( \hat{\varphi}(u) \in C(u) \), and since \( e \) commutes with \( u \), \( e \) commutes with \( \hat{\varphi}(u) \). Now let \( \varphi : \sigma_{\mathcal{B}}(y) \to \mathbb{C} \) be any continuous function with \( \varphi(\lambda) = 0 \) on \( \sigma_{\mathcal{B}}(y) \setminus V \). Define a function \( \hat{\varphi} : \mathbb{T} \to \mathbb{C} \) by

\[
\hat{\varphi}(\lambda) = \begin{cases} 
 \varphi(f_0^{-1}(\lambda)) & \lambda \in \mathbb{T} \setminus \{-1\} \\
 0 & \lambda = -1
\end{cases} .
\]

Then \( \varphi(y) = \hat{\varphi}(u) \) so that \( e \) commutes with \( \varphi(y) \).

Suppose that \( \varphi = 1 \) everywhere on \( U \). Then \( \hat{\varphi} = 1 \) everywhere on \( W \). Hence, for each \( j \in \mathbb{N} \),

\[
1_W(u_j) \hat{\varphi}(u_j) = \hat{\varphi}(u_j) 1_W(u_j) = 1_W(u_j) , \quad (3.17)
\]
and hence \( e\varphi(y) = \varphi(y)e = e \). Finally, if \( \varphi|_{X\setminus U} = 0 \), then \( \tilde{\varphi}|_{U \setminus W} = 0 \), and then

\[
1_W(u_j)\tilde{\varphi}(u_j) = \tilde{\varphi}(u_j)1_W(u_j) = \tilde{\varphi}(u_j),
\]

with the consequence that \( \varphi(y)e = e\varphi(y) = \varphi(y) \).

Summarizing the last two paragraphs, when \( \varphi \) is continuous on \( \sigma_B(y) \), then:

\[
\varphi|_{\sigma_B(y) \setminus V} = 0 \Rightarrow e\varphi(y) = \varphi(y)e,
\]

\[
\varphi|_{\sigma_B(y) \setminus V} = 0 \quad \text{and} \quad \varphi|_{\Gamma} = 1 \Rightarrow e\varphi(y) = \varphi(y)e = e,
\]

and

\[
\varphi|_{\sigma_B(y) \setminus U} = 0 \Rightarrow e\varphi(y) = \varphi(y)e = \varphi(y).
\]

Now let \( h : X \rightarrow [0, 1] \) be a continuous function such that \( h|_{\Gamma} = 1 \) and \( h|_{X \setminus V} = 0 \). Then using (3.20), (3.19), the commutativity of \( C(y) \), and then (3.20) once more,

\[
ye = yh(y)e = eyh(y) = eyh = ey.
\]

Thus, \( e \) commutes with \( y \).

Let \( C \) be the smallest \( C^* \) algebra containing \( y, e \) and \( 1 \). Evidently \( C \) is commutative, and \( e \) is a central projection. We now claim that \( \sigma_{\epsilon C}(ey) \subset U \). To see this, it suffices to show that whenever \( \psi \) is continuous on \( \sigma_B(y) \) with \( \psi|_{\Gamma} = 0 \), then \( \psi(ey) = 0 \). In fact, it suffices to do this for continuous \( \psi \) such that \( \psi|_{\sigma_B(y) \setminus V} = 1 \), since by the Gelfand-Naimark Theorem, \( \psi(ey) \neq 0 \) whenever \( \psi \) takes on any non-zero value anywhere on the spectrum of \( ey \). Then since \( y \) and \( e \) commute and \( e \) is a projection, \( \psi(ey) = e\psi(y) \). But \((1 - \psi)|_{\sigma_B(y) \setminus V} = 0 \) and \((1 - \psi)|_{\Gamma} = 1 \), and so by (3.19), \( e = e(1 - \psi(y)) \). Altogether,

\[
\psi(ey) = e\psi(y) = e(1 - (1 - \psi(y))) = e - e = 0.
\]

More simply, let \( \psi \) be continuous on \( \sigma_B(y) \) with \( \psi|_{\sigma_B(y) \setminus U} = 0 \). Then by (3.21), \( (1 - e)\psi(y) = 0 \), but as above \( \psi((1 - e)y) = (1 - e)\psi(y) \). Hence there is no spectrum of \( \psi((1 - e)y) \) outside \( \sigma_B(y) \setminus U \).

By Theorem 2.10 and the Spectral Invariance Theorem, putting \( \hat{y} = \lambda_1e + (1 - e)y \), we have \( \sigma_B(\hat{y}) \subset \{\lambda_1\} \cup (\sigma_B(y) \setminus U) \). Finally, \( \|\hat{y} - y\| = \|\lambda_1e - ey\| \leq \epsilon \).

**Proof of Lin’s Theorem.** Let \( \epsilon > 0 \), and let \( y \) be normal in \( B \), and suppose that \( \sigma_B(y) \subset \Gamma_\epsilon \). Since \( \sigma_B(y) \) is contained in the disc of radius \( \|y\| \), there are at most \( 2\|y\|(2\|y\| + 1)/\epsilon^2 \) edges of the elementary squares in \( \Gamma_\epsilon \) whose midpoints intersect \( \sigma_B(y) \). That is, with \( \Lambda_\epsilon \) defined as in (3.14), there are at most \( 2\|y\|(2\|y\| + 1)/\epsilon^2 \) points of \( \Lambda_\epsilon \) within \( \sigma_B(y) \).

We claim that there is a normal \( \tilde{y} \) such that \( \|\tilde{y}\| \leq \|y\| \), \( \|\tilde{y} - y\| \leq \epsilon \) and

\[
\sigma_B(\tilde{y}) \subset \Gamma_\epsilon \setminus \Lambda_\epsilon.
\]

This is true because if \( \lambda_0 \in \sigma_B(y) \cap \Lambda_\epsilon \), we have the following alternative: Either \( \lambda_0 \) is an isolated point of \( \sigma_B(y) \), or it is not.

If \( \lambda_0 \) is an isolated point of \( \sigma_B(y) \), then we can find a continuous function \( f : \Gamma_\epsilon \rightarrow \Gamma_\epsilon \) such that \( f(\lambda) = \lambda \) except on a small neighborhood of \( \lambda_0 \), and such that \( \lambda_0 \) is not in the range of \( f \). We may choose the neighborhood small enough that

\[
\sup_{\lambda \in \Gamma_\epsilon} \{|f(\lambda) - \lambda|\} \leq \epsilon^3/(2\|y\|(2\|y\| + 1)) \,.
\]
Moreover, we can always arrange that applying $f$ does not increase the spectral radius. Then $\|f(y)\| \leq \|y\|$, Then $\|f(y) - y\| \leq \epsilon^3/(2\|y\|(2\|y\| + 1))$, $f(y)$ is normal, $\sigma_B(f(y)) \subset \Gamma_{\epsilon} \setminus \{\lambda_0\}$. In this way, we remove all points in $\lambda_\epsilon$ that are isolated points of the spectrum.

Now we apply Lemma 3.6 to remove all points in $\lambda_\epsilon$ that are isolated points of the spectrum, noting that we only affect the spectrum near the each such point of $\lambda_\epsilon$ at each step. We may arrange that the shift in $y$ at each step has norm no more than $\epsilon^3/(2\|y\|(2\|y\| + 1))$.

At the end of at most $2\|y\|(2\|y\| + 1)/\epsilon^2$ operations, each of which shifted $y$ by at most $\epsilon^3/(2\|y\|(2\|y\| + 1))$ in norm, we arrive at $\tilde{y}$ which is normal and has $\|\tilde{y} - y\| \leq \epsilon$ and

$$
\sigma_B(\tilde{y}) \subset \Gamma_{\epsilon} \setminus \Lambda_\epsilon.
$$

The set $\Gamma_{\epsilon} \setminus \Lambda_\epsilon$ is a disjoint union of open crosses, and there is a continuous function $f$ on $\Gamma_{\epsilon} \setminus \Lambda_\epsilon$ that retracts each cross onto its center. That is, there is a continuous function $f : \Gamma_{\epsilon} \setminus \Lambda_\epsilon \to \tilde{\Lambda}_\epsilon$ with

$$
\sup_{\lambda \in \Gamma_{\epsilon} \setminus \Lambda_\epsilon} \{ |f(\lambda) - \lambda| \} \leq \epsilon/2.
$$

Then $f(\tilde{y})$ is normal with finite spectrum and $\|f(\tilde{y}) - \tilde{y}\| \leq \epsilon/2$. Consequently, $\|f(\tilde{y}) - y\| \leq 3\epsilon/2$.

Combining this with Lemma 3.5, we see that for every normal element of $B$, there is a normal element with finite spectrum arbitrarily close in norm, which is what we had to show.

\[\square\]

### 3.4 The Bott invariant and obstructions to commutativity

One might hope that one could extend Lin’s Theorem to three or more matrices. That is, one might conjecture that for all $\epsilon > 0$, there is a $\delta > 0$ such that if $\{h_1, h_2, h_3\}$ is a set of $n \times n$ self adjoint matrices such that

$$
||[h_1, h_2]| + |[h_2, h_3]| + |[h_3, h_1]| | \leq \delta,
$$

then there exists a set of three self adjoint commuting matrices $\{k_1, k_2, k_3\}$ such that

$$
||h_1 - k_1|| + ||h_2 - k_2|| + ||h_3 - k_3|| \leq \epsilon.
$$

This is false. Hastings and Loring, building on previous work, have shown the following:

#### 3.7 THEOREM. For all $j \in \frac{1}{2} \mathbb{N}$ there exists a set $\{h_1, h_2, h_3\}$ a set of self adjoint $(2j+1) \times (2j+1)$ matrices such that

$$
||[h_1, h_2]| + |[h_2, h_3]| + |[h_3, h_1]| | \leq \frac{1}{\sqrt{j(j+1)}},
$$

and such that if $\{k_1, k_2, k_3\}$ is any set of commuting self adjoint $(2j + 1) \times (2j + 1)$ matrices, then

$$
||h_1 - k_1|| + ||h_2 - k_2|| + ||h_3 - k_3|| \geq \sqrt{1 - 4/\sqrt{j(j+1)}}.
$$

#### 3.8 DEFINITION. Let $\delta > 0$ be given. A $\delta$ representation of the sphere in $M_n(\mathbb{C}$) is a set $\{h_1, h_2, h_3\}$ of self adjoint $n \times n$ matrices such that

$$
||[h_1, h_2]| \leq \delta,\quad |[h_2, h_3]| \leq \delta\quad \text{and}\quad |[h_3, h_1]| \leq \delta,
$$

and

$$
|1 - (h_1^2 + h_2^2 + h_3^2)| \leq \delta.
$$

(3.22)
3.9 EXAMPLE. A set of three self-adjoint $n \times n$ matrices $\{s_1, s_2, s_3\}$ that satisfy

$$[s_1, s_2] = is_3 \quad \text{,} \quad [s_2, s_3] = is_1 \quad \text{and} \quad [s_3, s_1] = is_2 .$$  \tag{3.24}$$
is an $n$ dimensional representation of the Lie algebra $\mathfrak{su}(2)$. The representation is irreducible in case there is no subspace of $\mathbb{C}^n$ that is invariant under each of $s_1$, $s_2$ and $s_3$. The simplest example is provided by the Pauli matrices, multiplied by $1/2$:

$$s_1 = \frac{1}{2} \sigma_1 = \frac{1}{2} \begin{bmatrix} 0 & 1 \\ 1 & 0 \end{bmatrix}, \quad s_2 = \frac{1}{2} \sigma_2 = \frac{1}{2} \begin{bmatrix} 0 & -i \\ i & 0 \end{bmatrix} \quad \text{and} \quad s_3 = \frac{1}{2} \sigma_3 = \frac{1}{2} \begin{bmatrix} 1 & 0 \\ 0 & -1 \end{bmatrix} .$$

For each $j \in \frac{1}{2} \mathbb{N}$ there is an irreducible representation by $(2j+1) \times (2j+1)$ matrices $\{s_1, s_2, s_3\}$ and these matrices satisfy $s_1^2 + s_2^2 + s_3^2 = j(j+1)1$, and they each have the same spectrum consisting of $\{-j, -j+1, \ldots, j-1, j\}$. It is then easy to see that defining $h_j = (j(j+1))^{-1/2}s_j$, $j = 1, 2, 3$, we obtain a $(j(j+1))^{-1/2}$ representation of the sphere.

We now explain the representation theory on which this construction depends, partly for completeness, and partly because it will be useful for some calculations that follow.

Notice that for any representation, $s_3s_1^2 = s_1s_3s_1 + is_3s_1s_1 = s_1^2s_3 + is_3s_1 + is_1s_3$, Using (3.24), this reduces to $[s_3, s_1^2] = -s_2s_1 - s_1s_2$. A similar calculation shows that $[s_3, s_2^2] = s_1s_2 + s_2s_1$. Altogether, $[s_3, (s_1^2 + s_2^2 + s_3^2)] = 0$, and by symmetry, $s_1^2 + s_2^2 + s_3^2$ commutes with $s_1$ and $s_2$ as well. In summary, defining the positive matrix $s$ by

$$s^2 := s_1^2 + s_2^2 + s_3^2 , \tag{3.25}$$

$$[s_1, s^2] = [s_2, s^2] = [s_3, s^2] = 0 . \tag{3.26}$$

Suppose that $\{s_1, s_2, s_3\}$ is an irreducible $n$-dimensional representation of $\mathfrak{su}(2)$. By (3.25), the eigenspaces of $s^2$ are invariant under each of $s_1$, $s_2$ and $s_3$. Since the representation is irreducible, it must be that $s^2$ is a multiple of the identity. Let $\mu$ temporarily denote this multiple, so that $s^2 = \mu 1$.

Define operators $s_+$ and $s_-$ by

$$s_+ = s_1 + is_2 \quad \text{and} \quad s_- = s_1 - is_2 . \tag{3.27}$$

We compute $s_3(s_1 + is_2) = (s_1 + is_2)s_3 + ([s_3, s_1] + i[s_3, s_2]) = (s_1 + is_2)s_3 + (is_2 + s_1)$. That is $[s_3, s_+] = s_+$. Taking the adjoint, $[s_3, s_-] = -s_-$, and we have

$$[s_3, s_+] = s_+ \quad \text{and} \quad [s_3, s_-] = -s_- . \tag{3.28}$$

Therefore, if $\zeta$ is an eigenvector of $s_3$ with $s_3\zeta = \lambda s_3$,

$$s_3(s_+\zeta) = s_+(s_3\zeta) + s_+\zeta = (\lambda + 1)s_+\zeta .$$

That is, either $\lambda + 1$ is an eigenvalue of $s_3$, or $s_+\zeta = 0$. In the same way we see that either $\lambda - 1$ is an eigenvalue of $s_3$ or else $s_-\zeta = 0$.

Now let $\zeta_1$ be an eigenvector of $s_3$ with minimal eigenvalue. (This is a least weight vector in the language of representation theory.) Then $s_-\zeta_1 = 0$. Define vectors $\zeta_k = (s_+)^{k-1}\zeta$. Suppose that for some $m \in \mathbb{N}$, no vector in $\{\zeta_1, \ldots, \zeta_m\}$ is zero. By what we have noted above, each is
an eigenvector of $s_3$, and the successive eigenvalues are all different, so that this set is orthogonal. Evidently, there is some least $m \in \mathbb{N}$ for which $(s_+)^{m+1} = 0$. Let $m$ be this integer.

By construction, $s_+ \zeta_k = \zeta_{k+1}$ for $k < m$, and $s_+ \zeta_m = 0$. Next, we compute that $s_+ s_- = s_1^2 + s_2^2 + s_3$ and $s_- s_+ = s_1^2 + s_2^2 - s_3$. Adding and subtracting $s_3^2$,

$$s_+ s_- = s^2 - s_3^2 + s_3 \quad \text{and} \quad s_- s_+ = s^2 - s_3^2 - s_3. \quad (3.29)$$

Since each vector in $\{\zeta_1, \ldots, \zeta_m\}$ is an eigenvector of $s^2 - s_3^2 + s_3$, it is an eigenvector of $s_- s_+$ and of $s_+ s_-$. For each $k = 2, \ldots, m$, $\zeta_k$ is a multiple of $s_+ \zeta_{k-1}$. Hence $s_- \zeta_k$ is a multiple of $s_- s_+ \zeta_{k-1}$ which, by the above, is a multiple $\zeta_{k-1}$. For $k = 1$, $s_- \zeta_k = 0$ since otherwise it would be an eigenvector of $s_3$ with an eigenvalue lower by one than the least eigenvalue. Hence the span of $\{\zeta_1, \ldots, \zeta_m\}$ is invariant under $s_-$ as well as $s_+$ and $s_3$. Hence it is invariant under each of $s_1$, $s_2$ and $s_3$. Since the representation is irreducible, $m = n$ and the span is all of $\mathbb{C}^n$.

Now let $\lambda$ be the least eigenvalue of $s_3$, and recall that $\mu$ denotes the single eigenvalue of $s^2$. By construction, $\zeta_n$ is an eigenvector of $s_3$ with eigenvalue $\lambda + n - 1$. Since $s_- \zeta_1 = 0$ and $s_+ \zeta_n = 0$, (3.29) gives us

$$0 = \mu - \lambda^2 + \lambda \quad \text{and} \quad 0 = \mu - (\lambda + n - 1)^2 - (\lambda + n - 1).$$

Thus, $\lambda^2 + 2\lambda(n - 1) + (n - 1)^2 + \lambda + (n - 1) = \lambda^2 - \lambda$ so that

$$\lambda 2n = -(n - 1) - (n - 1)^2 = -n(n - 1).$$

We obtain

$$\lambda = -\frac{n - 1}{2} \quad \text{and} \quad \mu = \frac{n^2 - 1}{2}. \quad (3.30)$$

At this point it is traditional to introduce $j \in \frac{1}{2} \mathbb{N}$ by $j = \frac{n - 1}{2}$, so that $n = 2j + 1$, and then $s^2 = j(j+1)1$. The eigenvalues of $s_3$ are then given, in increasing order, by $\{-j - (j - 1), \ldots, j - 1, j\}$, and by symmetry, $s_1$ and $s_2$ have the same spectrum.

So far we have seen that if for some $j \in \frac{1}{2} \mathbb{N}$ there is a $2j+1$ dimensional irreducible representation of $\mathfrak{su}(2)$, then there is an orthonormal basis $\{\text{eta}_j, \ldots, \text{eta}_1\}$ of $\mathbb{C}^{2j+1}$ such that $s_3 \text{eta}_k = k \text{eta}_k$ for each $k = -2j - 1, \ldots, 2j + 1$. This gives us the (diagonal) form of the matrix for $s_3$ in this basis.

Moreover, we have seen that for all $k = -2j - 1, \ldots, 2j$, $a_+ \text{eta}_k = t_k \text{eta}_k$ for some positive multiple $t$, while $a_+ \text{eta}_{2j+1} = 0$. We compute

$$t_k^2 = \langle \text{eta}_k, s_- s_+ \text{eta}_k \rangle = \langle \text{eta}_k, (s^2 - s_3^2 - s_3) \text{eta}_k \rangle = j(j+1) - k(k+1).$$

That is,

$$s_+ \text{eta}_k = \sqrt{j(j+1) - k(k+1)} \eta_{j+1} \quad \text{for all} \quad k = -2j - 1, \ldots, 2j + 1. \quad (3.31)$$

This gives us the form of the matrix representing $s_+$ in this basis, and taking the hermitian conjugate we get the matrix that represents $s_-$. Finally, it is easy to check that the matrices determine a triple $\{s_1, s_2, s_3\}$ of self adjoint $(2j + 1) \times (2j + 1)$ matrices that satisfy (3.24). Hence, for each $j$, there is a representation of $\mathfrak{su}(2)$ by $(2j + 1) \times (2j + 1)$ matrices, and any two such representations are unitarily equivalent. Any such representation is called a spin $j$ representation.

Each such representation gives rise to a natural example of a $(j(j+1))^{-1/2}$ representation of the sphere: Since

$$s_1^2 + s_2^2 + s_3^2 = j(j+1)1,$$
we define \( h_j = (j(j+1))^{-1/2}s_j \), \( j = 1, 2, 3, \{h_1, h_2, h_3\} \), and this provides a \((j(j+1))^{-1/2}\) representation of the sphere.

We will show, following Hastings and Loring, that if \( \{k_1, k_2, k_3\} \) is a set of three commuting self adjoint \((2j+1) \times (2j+1)\) matrices, then for the spin \( j \) representation of \( su(2) \),

\[
\|h_1 - k_1\| + \|h_2 - k_2\| + \|h_3 - k_3\| \geq \sqrt{1 - 2/j}.
\]

The method involves a topological invariant, the Bott invariant that we now define.

**3.10 DEFINITION** (Bott invariant). For any set of three \( n \times n \) hermitian matrices \( h_1, h_2 \) and \( h_3 \), define the \( 2n \times 2n \) matrix \( b(h_1, h_2, h_3) \) by

\[
b(h_1, h_2, h_3) = \sum_{j=1}^{3} \sigma_j \otimes h_j = \begin{bmatrix} h_3 & h_1 - ih_2 \\ h_1 + ih_2 & -h_3 \end{bmatrix}.
\]

Note that \( b(h_1, h_2, h_3) \) is self adjoint so that all of its eigenvalues are real. Let \( N_+(h_1, h_2, h_3) \) be the number of strictly positive eigenvalues of \( b(h_1, h_2, h_3) \), and let \( N_-(h_1, h_2, h_3) \) be the number of strictly negative eigenvalues of \( b(h_1, h_2, h_3) \). Suppose that 0 is not an eigenvalue of \( b(h_1, h_2, h_3) \), so that \( N_+(h_1, h_2, h_3) + N_-(h_1, h_2, h_3) = 2n \). Then \( N_+(h_1, h_2, h_3) - N_-(h_1, h_2, h_3) = 2n - 2N_-(h_1, h_2, h_3) \) is an even integer, so that

\[
bott(h_1, h_2, h_3) := \frac{1}{2}[N_+(h_1, h_2, h_3) - N_-(h_1, h_2, h_3)]
\]

is an integer. This integer is the *Bott invariant* of \( \{h_1, h_2, h_3\} \). Note that \( bott(h_1, h_2, h_3) \) is only defined when 0 is not an eigenvalue of \( b(h_1, h_2, h_3) \).

Note that the three matrices \( \sigma_1, \sigma_2 \) and \( \sigma_3 \) are have the same spectrum, namely \( \{-1, 1\} \). Then for any self adjoint \( h \in M_n(C) \), \( \sigma_1 \otimes h, \sigma_2 \otimes h \) and \( \sigma_3 \otimes h \) all have the same eigenvalues, namely \( \{\pm \lambda_1, \ldots, \pm \lambda_n\} \) where \( \{\lambda_1, \ldots, \lambda_n\} \) is the set of eigenvalues of \( h \). Hence

\[
\|\sigma_1 \otimes h\| = \|\sigma_2 \otimes h\| = \|\sigma_3 \otimes h\| = \|h\|.
\]

It follows that for any self adjoint triple \( \{h_1, h_2, h_3\} \),

\[
\|b(h_1, h_2, h_3)\| \leq \|h_1\| + \|h_2\| + \|h_3\|.
\]  \hspace{1cm} (3.32)

Suppose that \( h_1, h_2 \) and \( h_3 \) are three commuting Hermitian \( n \times n \) matrices. Then there is a unitary \( n \times n \) matrix \( u \) such that \( k_j := u^*h_ju \) is diagonal for each \( j = 1, 2, 3 \). Evidently

\[
\begin{bmatrix} u^* & 0 \\ 0 & u^* \end{bmatrix} b(h_1, h_2, h_3) \begin{bmatrix} u & 0 \\ 0 & u \end{bmatrix} = b(k_1, k_2, k_3).
\]

Therefore,

\[
bott(h_1, h_2, h_3) = bott(k_1, k_2, k_3).
\]

Let \( \alpha_1, \ldots, \alpha_n, \beta_1, \ldots, \beta_n, \) and \( \gamma_1, \ldots, \gamma_n \) be the diagonal entries of \( k_1, k_2 \) and \( k_3 \) respectively. Then \( b(k_1, k_2, k_3) \) is unitarily equivalent to

\[
\bigoplus_{\ell=1}^{n} \begin{bmatrix} \gamma_\ell & \alpha_\ell - i\beta_\ell \\ \alpha_\ell + i\beta_\ell & -\gamma_\ell \end{bmatrix}.
\]  \hspace{1cm} (3.33)
A simple computation shows that \[
\begin{bmatrix}
\gamma \ell & \alpha \ell - i\beta \ell \\
\alpha \ell + i\beta \ell & -\gamma \ell
\end{bmatrix}^2 = (\alpha^2 + \beta^2 + \gamma^2) \begin{bmatrix} 1 & 0 \\ 0 & 1 \end{bmatrix}.
\]
Since the trace of \[
\begin{bmatrix}
\gamma \ell & \alpha \ell - i\beta \ell \\
\alpha \ell + i\beta \ell & -\gamma \ell
\end{bmatrix}
\] is zero, it then follows that the eigenvalues are \(\pm \sqrt{\alpha^2 + \beta^2 + \gamma^2}\).

Thus, as long no non-zero vector is in the null space of each of \(h_1, h_2\) and \(h_3\), each of the blocks in (3.33) has one strictly positive eigenvalue and one strictly negative eigenvalue. In particular, this is the case if each of \(h_1, h_2\) and \(h_3\) are invertible. It follows that in this case, \(\text{bott}(h_1, h_2, h_3) = 0\).

We have proved:

3.11 LEMMA. Let \(h_1, h_2\) and \(h_3\) be three commuting invertible Hermitian \(n \times n\) matrices. Then \(\text{bott}(h_1, h_2, h_3) = 0\).

3.12 EXAMPLE. Let \(\{h_1, h_2, h_3\}\) be the \((j(j + 1))^{-1/2}\) representation of the sphere provided by a spin \(j\) representation of \(\mathfrak{su}(2)\). Let \(\{\eta_{-2j-1}, \ldots, \eta_{2j+1}\}\) be the corresponding sequence of eigenvectors of \(s_3\) and hence \(h_3\). Since, using the notation of the previous example,

\[
b(h_1, h_2, h_3) = \frac{1}{\sqrt{j(j + 1)}} \begin{bmatrix} s_3 & s_- \\ s_+ & -s_3 \end{bmatrix},
\]

is easy to see that the vectors of the form \((\eta_k, \pm \eta_{k+1}), -2j - 1, \ldots, 2j,\) together with \((0, \eta_{-2j-1})\) and \((\eta_{2j+1}, 0)\), are a set of \(2n\) orthonormal eigenvalues of \(b(h_1, h_2, h_3)\). A simple calculation show that the pairs \((\eta_k, \pm \eta_{k+1})\) contribute a positive and a negative eigenvalue each, while the two special case eigenvectors have positive eigenvalues. Hence

\[
\text{bott}(h_1, h_2, h_3) = 1
\]

for all \(j\).

Next, we show that the Bott invariant is defined for all \(\delta\) representations of the sphere with \(\delta < 1/4\).

3.13 LEMMA. Let \(\{h_1, h_2, h_3\}\) be a \(\delta\)-representation of the sphere in \(M_n(C)\) with \(\delta < 1/4\). Then

\[
\sigma(b(h_1, h_2, h_3)) \subset [-\sqrt{1 + 4\delta}, -\sqrt{1 - 4\delta}] \cup [\sqrt{1 - 4\delta}, \sqrt{1 + 4\delta}].
\]  

Moreover, if \(\{k_1, k_2, k_3\}\) is any triple of self adjoint operators with

\[
\gamma = \|h_1 - k_1\| + \|h_2 - k_2\| + \|h_3 - k_3\| < \sqrt{1 - 4\delta},
\]

then for all \(t \in [0, 1]\),

\[
\sigma(s((1 - t)h_1 + tk_1, (1 - t)h_2 + tk_2, (1 - t)h_3 + tk_3)) \subset \\
[-\gamma - \sqrt{1 + 4\delta}, \gamma - \sqrt{1 - 4\delta}] \cup [\sqrt{1 - 4\delta} - \gamma, \sqrt{1 + 4\delta} + \gamma].
\]  

and consequently, \(\text{bott}((1 - t)h_1 + tk_1, (1 - t)h_2 + tk_2, (1 - t)h_3 + tk_3)\) is well-defined for all \(t \in [0, 1]\).
Proof. We compute that
\[ (s(h_1, h_2, h_3))^2 = 1 \otimes (h_1^2 + h_2^2 + h_3^2) + \sigma_3 \otimes i [h_1, h_2] + \sigma_1 \otimes i [h_2, h_3] + \sigma_2 \otimes i [h_3, h_1] . \]
Therefore, \( \|s(h_1, h_2, h_3)\|^2 - 1 \leq 4\delta \), and then by the Spectral Mapping Lemma,
\[ \sigma(s(h_1, h_2, h_3)) \subset \{ t \in \mathbb{R} : |t^2 - 1| < 4\delta \} . \]
Next, note that by (3.32),
\[ ||b((1 - t)h_1 + tk_1, (1 - t)h_2 + tk_2, (1 - t)h_3 + tk_3)) - b(h_1, h_2, h_3)|| = t\|b(k_1 - h_1, k_2 - h_2, k_3 - h_3))\| \leq t\gamma . \] (3.37)
Then (3.34) and Theorem 2.17 yield (3.39), and \( \text{bott}((1 - t)h_1 + tk_1, (1 - t)h_2 + tk_2, (1 - t)h_3 + tk_3)) \) is well-defined for all \( t \in [0, 1] \).

3.5 The Bott invariant as a trace function

Recall that for an \( n \times n \) matrix \( a \), the trace of \( a \), \( \text{Tr}[a] \), is defined by
\[ \text{Tr}[a] = \sum_{j=1}^{n} a_{i,j} \] (3.38)
where \( a_{i,j} \) denotes the \( i, j \) entry of \( a \).

A simple computation shows that for any \( a \in M_n(\mathbb{C}) \), and any invertible \( b \in M_n(\mathbb{C}) \), \( \text{Tr}[b^{-1}ab] = \text{Tr}[a] \). Let \( \{\eta_1, \ldots, \eta_n\} \) be any orthonormal basis of \( \mathbb{C}^n \), and let \( \{\chi_1, \ldots, \chi_n\} \) be the standard basis. Let \( u \) be the unitary matrix with \( u\chi_j = \eta_j \) for \( j = 1, \ldots, n \). Then
\[ \text{Tr}[a] = \text{Tr}[u^*au] = \sum_{j=1}^{n} \langle \chi_i, u^*au\chi_i \rangle = \sum_{j=1}^{n} \langle \eta_j, b^{-1}ab\eta_j \rangle , \]
showing that the trace may be computed at the sum of the diagonal elements in any orthonormal basis. If \( a \) is self adjoint, there is an orthonormal basis consisting of eigenvectors of \( a \); \( a\eta_j = \lambda_j\eta_j \) for each \( j = 1, \ldots, n \). Then evidently \( \text{Tr}[a] - \sum_{j=1}^{n} \lambda_j \). The function \( a \mapsto \text{Tr}[a] \) is evidently continuous.

Now let \( \epsilon > 0 \) be given, and let \( f_\epsilon \) be any continuous function form \( \mathbb{R} \) to \([-1, 1]\) such that \( f(t) = -1 \) for \( t \leq -\epsilon \) and \( f(t) = 1 \) for \( t \geq \epsilon \).

Let \( a \in M_n(\mathbb{C}) \) be self adjoint and such that \( (-\epsilon, \epsilon) \cap \sigma_{\text{off}}(a) = \emptyset \). Let \( \{\eta_1, \ldots, \eta_n\} \) be an orthonormal basis of \( \mathbb{C}^n \) consisting of eigenvectors of \( a \) with \( a\eta_j = \lambda_j\eta_j \) for each \( j = 1, \ldots, n \). Define
\[ N_+(a) = \sum_{j=1}^{n} n_1(0, \infty)(\lambda_j) \quad \text{and} \quad N_-(a) = \sum_{j=1}^{n} n_1(-\infty, 0)(\lambda_j) . \]
Then by considering a sequence of polynomial approximations of \( f_\epsilon \) on \([-\|a\|, \|a\|]\), we have that
\[ \text{Tr}[f(a)] = \sum_{j=1}^{n} \langle \eta_j, f(a)\eta_j \rangle = \sum_{j=1}^{n} f_\epsilon(\lambda_j) = N_+(a) - N_-(a) . \]
It follows that when \( \{h_1, h_2, h_3\} \) is a \( \delta \) representation of the sphere in \( M_n(\mathbb{C}) \) with \( \delta < 1/4 \),
\[ \text{bott}(\{h_1, h_2, h_3\}) = \text{Tr}[f_{1-4\delta}(s(h_1, h_2, h_3))] . \]
Proof of Theorem 3.7. Consider the $1/\sqrt{j(j+1)}$ representation of the sphere associated to the spin $j$ representation of $\mathfrak{su}(2)$. By Lemma 3.13, if $\{k_1, k_2, k_3\}$ is any triple of self adjoint $(2j+1) \times (2j+1)$ matrices with

$$\gamma = \|h_1 - k_1\| + \|h_2 - k_2\| + \|h_3 - k_3\| < \sqrt{1 - 4/\sqrt{j(j+1)}} ,$$

then for all $t \in [0,1]$,

$$\sigma(s((1-t)h_1 + tk_1, (1-t)h_2 + tk_2, (1-t)h_3 + tk_3)) \subset [-\gamma - \sqrt{1 + 4\sqrt{j(j+1)}}, \gamma - \sqrt{1 - 4\sqrt{j(j+1)}}] \cup [\sqrt{1 - 4\sqrt{j(j+1)}} - \gamma, \sqrt{1 + 4\sqrt{j(j+1)}} + \gamma].$$

and consequently, for any $\epsilon < \sqrt{1 - 4\sqrt{j(j+1)}} - \gamma$,

$$\text{bott}((1-t)h_1 + tk_1, (1-t)h_2 + tk_2, (1-t)h_3 + tk_3) - \text{Tr}[f_s((1-t)h_1 + tk_1, (1-t)h_2 + tk_2, (1-t)h_3 + tk_3)],$$

with $f_s$ defined as in the paragraphs above. Then the right hand side is a continuous integer valued function of $t$, and so

$$\text{bott}(k_1, k_2, k_3) = \text{bott}(h_1, h_2, h_3) = 1 .$$

Therefore, $\{k_1, k_2, k_3\}$ cannot be a commuting triple.

\hfill \Box

4 Operators on Hilbert space

4.1 Topologies on $\mathcal{B(H)}$

Let $\mathcal{H}$ be a separable Hilbert space with inner product $\langle \cdot, \cdot \rangle_{\mathcal{H}}$ and norm $\| \cdot \|_{\mathcal{H}}$, and let $\mathcal{B(\mathcal{H})}$, as usual, denote the $C^*$-algebra of bounded linear operators on $\mathcal{H}$. There are two important non-metric topologies in $\mathcal{B(\mathcal{H})}$, waker than the norm topology, that are essential to what follows.

4.1 DEFINITION (Strong and weak operator topologies). The strong operator topology on $\mathcal{B(\mathcal{H})}$ is the weakest topology such that for each $\xi \in \mathcal{H}$, the function $a \mapsto a\xi$ from $\mathcal{B(\mathcal{H})}$ to $\mathcal{H}$ is continuous with the usual norm topology on $\mathcal{H}$. The weak operator topology on $\mathcal{B(\mathcal{H})}$ is the weakest topology such that for each $\xi, \zeta \in \mathcal{H}$, that function $a \mapsto \langle \zeta, a\xi \rangle_{\mathcal{H}}$ is continuous from $\mathcal{B(\mathcal{H})}$ to $\mathbb{C}$.

It follows from the definitions that a basic set of neighborhoods of 0 for the strong operator topology is given by the sets

$$U_{\epsilon, \xi_1, \ldots, \xi_n} = \{ a \in \mathcal{B(\mathcal{H})} : \|a\xi_j\|_{\mathcal{H}} < \epsilon \text{ for } j = 1, \ldots, n \}$$

(4.1)

where $\epsilon > 0$ and $\xi_1, \ldots, \xi_n \in \mathcal{H}$. Likewise, it follows that a basic set of neighborhoods of 0 for the weak operator topology is given by the sets

$$V_{\epsilon, \zeta_1, \ldots, \zeta_n} = \{ a \in \mathcal{B(\mathcal{H})} : |\langle \zeta_j, a\xi_j \rangle_{\mathcal{H}}| < \epsilon \text{ for } j = 1, \ldots, n \}$$

(4.2)
\( \epsilon > 0 \) and \( \zeta_1, \ldots, \zeta_n, \xi_1, \ldots, x_n \in H \). Note that both topologies are evidently Hausdorff.

It is clear that for each \( \xi \in H \), \( a \mapsto a\xi \) is continuous in the norm topology on \( B(H) \), so that the norm topology is stronger than the strong operator topology. Furthermore, since for all \( \zeta \in H \), \( \zeta \mapsto \langle \zeta, \xi \rangle \) is continuous on \( H \), the function \( a \mapsto \langle \zeta, a\xi \rangle \) is continuous in the strong operator topology on \( B(H) \), being the composition of continuous functions, and hence the strong operator topology is stronger than the weak operator topology.

The following proposition shows that the norm topology is strictly stronger than the strong operator topology, which is in turn strictly stronger than the weak operator topology.

**4.2 PROPOSITION** (Continuity of the norm and adjoint). Let \( H \) be an infinite dimensional Hilbert space. Then:

(1) The function \( a \mapsto \|a\| \) from \( B(H) \) to \( \mathbb{R}_+ \) is continuous in the norm topology, but is only lower semicontinuous in the strong and weak operator topologies.

(2) The function \( a \mapsto a^* \) is continuous from \( B(H) \) to \( B(H) \) in the norm and the weak operator topologies, but not in the strong operator topology.

**Proof.** Let \( \{\zeta_j\} \) be an orthonormal sequence in \( H \). For each \( n \in \mathbb{N} \), let \( p_n \) denote the orthogonal projection onto the span of \( \{\zeta_1, \ldots, \zeta_n\} \). Then for all \( \xi \in H \), \( \lim_{n \to \infty} \|p_n\xi\| = 0 \) by Bessel’s inequality, so that \( \lim_{n \to \infty} p_n = 0 \) in the strong operator topology. However, for \( n \neq m \), \( \|p_n - p_m\| = 1 \), so that the sequence \( \{p_n\} \) is not even Cauchy in the norm topology. Hence the norm is discontinuous in the strong operator topology, and hence also in the weak operator topology.

To see that the norm is lower semicontinuous in these topologies, it suffices to show that the sub-level sets \( \{ a \in B(H) : \|a\| \leq t \} \) are closed for each \( t > 0 \). Fix \( t > 0 \) and \( b \) in the closure of \( \{ a \in B(H) : \|a\| \leq t \} \). Then for each unit vector \( \xi \in H \), and each \( n \in \mathbb{N} \) there is an \( a_n \in \{ a \in B(H) : \|a\| \leq t \} \) such that \( b - a_n \in U_{1/n,\xi} \), which means that \( \|(b - a_n)\xi\| < 1/n \). This means that \( \|b\xi\| \leq \|a_n\| + 1/n \leq t + 1/n \). Since \( n \) is arbitrary, \( \|b\xi\| \leq t \). Then since \( \xi \) is an arbitrary unit vector in \( H \), \( \|b\| \leq t \). This proves the closure in the strong operator topology, and a very similar argument proves the closure for the weak operator topology.

For the second part, since every infinite dimensional Hilbert space contains a copy of \( \ell_2 \), the Hilbert space of all square summable functions from \( \mathbb{N} \) to \( \mathbb{C} \), we may suppose without loss of generality that \( H = \ell_2 \). Define the shift operator \( a \in B(H) \) by

\[
(a\zeta)_j = \begin{cases} 
\zeta_{j-1} & j \geq 2 \\
0 & j = 1 
\end{cases}
\]

Evidently, for all \( \zeta \), \( \|a\zeta\|_\mathcal{F} = \|\zeta\|_\mathcal{F} \). The adjoint is given by \( (a^*\zeta)_j = \zeta_{j+1} \) for all \( j \in \mathbb{N} \). Therefore, \( \|a^*\zeta\|_\mathcal{F}^2 = \sum_{j=2}^{\infty} |\zeta_j|^2 = \|\zeta\|^2 - |\zeta_1|^2 \). It follows that for all \( \zeta \in H \),

\[
\lim_{n \to \infty} \|(a^n)^*\zeta\|_\mathcal{F} = 0 \quad \text{while} \quad \|a^n\zeta\|_\mathcal{F} = \|\zeta\|_\mathcal{F}.
\]

Hence the sequence \( \{(a^n)^*\} \) converges to zero in the strong operator topology, but the sequence \( \{a^n\} \) does not. Since \( \{a^n\} = \{(a^n)^{**}\} \) this shows that the involution is not continuous in the strong operator topology.

The continuity of the involution is obvious in the norm topology since the involution is an isometry, and in the weak operator topology it follows from the fact that

\[
(V_{\varepsilon,\xi_1,\ldots,\xi_n,\xi_1,\ldots,\xi_n})^* = V_{\varepsilon,\xi_1,\ldots,\xi_n,\xi_1,\ldots,\xi_n}.
\]
As far as sequences are concerned, a sequence \( \{a_n\} \) in \( \mathcal{B}(\mathcal{H}) \) converges to \( a \in \mathcal{B}(\mathcal{H}) \) in the strong operator topology if and only if for all \( \zeta, \xi \in \mathcal{H} \), \( \lim_{n \to \infty} a_n \zeta = a \zeta \), and likewise, converges to \( a \in \mathcal{B}(\mathcal{H}) \) in the weak operator topology if and only if for all \( \zeta, \xi \in \mathcal{H} \), \( \lim_{n \to \infty} \langle \zeta, a_n \xi \rangle = \langle a \zeta, \xi \rangle \). A sequence \( \{a_n\} \) in \( \mathcal{B}(\mathcal{H}) \) is a Cauchy sequence for the weak operator topology in case for every basic open neighborhood \( V_{\epsilon, \zeta_1, \ldots, \zeta_n} \) of 0, \( a_m - a_n \in V_{\epsilon, \zeta_1, \ldots, \zeta_n} \) for all but finitely many \( m, n \). Cauchy sequences for the strong operator topology are defined analogously.

**4.3 THEOREM.** Let \( \{a_n\} \) be a Cauchy sequence for the weak operator topology. Then \( \{\|a_n\|\} \) is a bounded sequence, and there exists an \( a \in \mathcal{B}(\mathcal{H}) \) with \( \|a\| \leq \sup_{n \in \mathbb{N}} \{\|a_n\|\} \) and such that \( \lim_{n \to \infty} a_n = a \) in the weak operator topology. Moreover, the analogous statement for the strong operator topology is also true.

**Proof.** Let \( \{a_n\} \) be a Cauchy sequence for the weak operator topology. We first show that \( \{\|a_n\|\} \) is a bounded sequence. To see this, note that for each \( \zeta, \xi \in \mathcal{H} \), \( \{\langle \zeta, a_n \xi \rangle \} \) is a Cauchy sequence in \( \mathbb{C} \), and hence convergent and bounded. Thus, if we define the sets \( C_m \subset \mathcal{H} \times \mathcal{H} \) by

\[
C_m = \{ (\zeta, \xi) \in \mathcal{H} \times \mathcal{H} : \sup_{n \in \mathbb{N}} |\langle \zeta, a_n \xi \rangle| \leq m \}
\]

we have that \( \cup_{m \in \mathbb{N}} C_m = \mathcal{H} \times \mathcal{H} \). If \( \{\zeta_k, \xi_k\} \) is a convergent sequence in \( C_m \) with limit \( (\zeta, \xi) \), then for all \( n \),

\[
|\langle \zeta, a_n \xi \rangle| = \lim_{k \to \infty} |\langle \zeta_k, a_n \xi_k \rangle| \leq m
\]

so that \( C_m \) is closed. Since \( \mathcal{H} \times \mathcal{H} \) with the product metric is a complete metric space, by Baire’s Theorem, for at least one \( m \in \mathbb{N} \), \( C_m \) contains an open set, and then it is clear that \( \{\|a_n\|\} \) is a bounded sequence.

Now let \( L = \sup_{n \in \mathbb{N}} \{\|a_n\|\} \), and for all \( \zeta, \xi \in \mathcal{H} \), define \( q(\zeta, \xi) = \lim_{n \to \infty} \langle \zeta, a_n \xi \rangle = \langle a \zeta, \xi \rangle \), which exists since the sequence on the right is Cauchy in \( \mathbb{C} \). It is easy to see that \( \zeta, \xi \mapsto q(\zeta, \xi) \) is a sesquilinear form on \( \mathcal{H} \), with

\[
|q(\zeta, \xi)| \leq L \|\zeta\| \|\xi\|.
\]

For each \( \zeta \in \mathcal{H} \), the map \( \zeta \mapsto q(\zeta, \xi) \) is a conjugate linear functional on \( \mathcal{H} \), and hence by the Riesz Representation Theorem, there is a uniquely determined vector \( \eta_\zeta \in \mathcal{H} \) such that \( q(\zeta, \xi) = \langle \zeta, \eta_\zeta \rangle \), for all \( \zeta, \xi \in \mathcal{H} \), and \( \|\eta_\zeta\| \|\xi\| \leq r \|\xi\| \|\xi\| \). Since \( q \) is sesquilinear, the map \( \xi \mapsto \eta_\zeta \) is linear, and thus there exists \( a \in \mathcal{B}(\mathcal{H}) \) such that \( \|a\| \leq L \) and \( \eta_\zeta = a \xi \) for all \( \xi \in \mathcal{H} \). It now follows that for each \( \zeta, \xi \in \mathcal{H} \), \( \lim_{n \to \infty} \langle \zeta, a_n \xi \rangle = \langle \zeta, a \xi \rangle \), and hence that \( \lim_{n \to \infty} a_n = a \) in the weak operator topology. The corresponding proof for the strong operator topology is easier, and is left as an exercise.

We next claim that the strong operator topology is not metrizable when \( \mathcal{H} \) is infinite dimensional. The basic open set \( U_{\epsilon, \zeta_1, \ldots, \zeta_n} \) contains all \( a \in \mathcal{B}(\mathcal{H}) \) with \( a \xi_j = 0 \) for each \( j = 1, \ldots, n \). If \( \mathcal{H} \) is infinite dimensional, then there is a non-trivial subspace of \( \mathcal{B}(\mathcal{H}) \) contained in every \( U_{\epsilon, \zeta_1, \ldots, \zeta_n} \), and hence in every open set about the origin.

For each \( n \in \mathbb{N} \), the set \( C_n := \{ a \in \mathcal{B}(\mathcal{H}) : \|a\| \leq n \} \) is closed in the strong operator topology by the lower semicontinuity of the norm. By what we have just said, each \( C_n \) is nowhere dense,
since no ball can contain a non-trivial subspace. Since evidently $\mathcal{B}(\mathcal{H}) = \bigcup_{n=1}^{\infty} C_n$, it follows that $\mathcal{B}(\mathcal{H})$ is a countable union of closed, nowhere dense sets in the strong operator topology. Suppose the strong operator topology were metrizable. Then by Theorem refwscom, $\mathcal{B}(\mathcal{H})$ equipped with this topology would be a complete metric space. Baire’s Theorem says that a complete metric space is never the countable union of closed nowhere dense sets, so the strong topology on $\mathcal{B}(\mathcal{H})$ cannot be metrized. A similar argument applies to the weak operator topology. However, as we show next, The relative weak and strong operator topologies on bounded subsets of $\mathcal{B}(\mathcal{H})$ are metrizable when $\mathcal{H}$ is separable.

4.4 THEOREM. For $r > 0$, let $\overline{B}_r$ denote the closed unit ball of radius $r$ in $\mathcal{B}(\mathcal{H})$. That is, $\overline{B}_r = \{ a \in \mathcal{B}(\mathcal{H}) : \|a\| \leq r \}$. Then there are metrics $\rho_w$ and $\rho_s$ on $\overline{B}_r$ such that the metric topologies are equivalent to the relative weak and strong operator topologies respectively, and such that $(\overline{B}_r, \rho_w)$ and $(\overline{B}_r, \rho_s)$ are complete metric spaces.

Proof. Let $\{\eta_j\}$ be any sequence of unit vectors that is dense in the unit sphere of $\mathcal{H}$. For all $a, b \in \mathcal{B}(\mathcal{H})$, define

$$\rho_s(a, b) = \sum_{j=1}^{\infty} 2^{-j} \| (a - b)\eta_j \|_\mathcal{H} \quad \text{and} \quad \rho_w(a, b) = \sum_{j,k=1}^{\infty} 2^{-j-k} | \langle \eta_k (a - b)\eta_j \rangle_\mathcal{H} |. \quad (4.3)$$

It is easy to verify that these are indeed metrics.

We now show that the relative strong operator topology on $\overline{B}_r$ coincides with the metric topology on $\overline{B}_r$ induced by the metric $\rho_s$. First, we first show that for every $t > 0$, $\{ a : \rho_s(a, 0) < t \}$ contains a neighborhood of 0 in the relative strong operator topology. Choose $n$ so that $r 2^{-n} < t/2$. Then for $b \in U_{t/2, \eta_1, \ldots, \eta_n} \cap \overline{B}_r$,

$$\rho_s(b, 0) = \sum_{j=1}^{\infty} 2^{-j} \| b\eta_j \| \leq \sum_{j=1}^{n} 2^{-j} \frac{t}{2} + \sum_{j=n+1}^{\infty} r \leq t$$

and consequently, $U_{t/2, \eta_1, \ldots, \eta_n} \cap \overline{B}_r \subset \{ a : \rho_s(a, 0) < t \}$.

We next show that every basic strong operator topology neighborhood $U_{r, \xi_1, \ldots, \xi_m}$ contains an open ball about 0 in the relative metric topology. By decreasing epsilon as necessary, we may suppose that $\xi_j$ is a unit vector for each $j$. Choose $\{\eta_{j_1}, \ldots, \eta_{j_m}\}$ such that $\|\eta_{j_k} - \xi_k\| < \epsilon/2$ for $k = 1, \ldots, m$. Let $M = \max\{j_1, \ldots, j_M\}$. Then for $b \in \overline{B}_r \cap \{ a : \rho_s(a, 0) < 2^{-M}\epsilon \}$, $\| b\eta_j \| \leq \epsilon$ for each $j = 1, \ldots, m$, and consequently $b \in U_{r, \xi_1, \ldots, \xi_m}$. A similar argument shows that on each $\overline{B}_r$, the relative weak operator topology is metrizable.

We shall be especially concerned with bounded subsets of the self adjoint elements of $\mathcal{B}(\mathcal{H})$, for which there is an even simpler description of the relative weak operator topology, for which there is an even simpler criterion for weak convergence:

$$\lim_{n \to \infty} \langle \eta_j a_n \eta_j \rangle_\mathcal{H} = \langle \eta_j a \eta_j \rangle_\mathcal{H} \quad \text{for all} \quad j \in \mathbb{N}.$$  

4.5 LEMMA (Polarization identify). Let $a \in \mathcal{B}(\mathcal{H})$ be self adjoint. Then for all $\zeta$ and $\xi$ in $\mathcal{H}$,

$$\langle \zeta, a\xi \rangle_\mathcal{H} = \frac{1}{4} \left[ \langle \langle \zeta + \xi, a(\zeta + \xi) \rangle_\mathcal{H} \rangle_\mathcal{H} - \langle \langle \zeta - \xi, a(\zeta - \xi) \rangle_\mathcal{H} \rangle_\mathcal{H} \right] - \frac{i}{4} \left[ \langle \langle \zeta + i\xi, a(\zeta + i\xi) \rangle_\mathcal{H} \rangle_\mathcal{H} - \langle \langle \zeta - i\xi, a(\zeta - i\xi) \rangle_\mathcal{H} \rangle_\mathcal{H} \right].$$
4.6 REMARK. It follows that for the relative weak operator topology on the self-adjoint elements of $\mathcal{B}(\mathcal{H})$, a basic set of neighborhoods at the origin is given by the sets

$$V_{\epsilon,\xi_1,\ldots,\xi_n} = \{ a \in \mathcal{B}(\mathcal{H}) : |\langle \xi_j, a\xi_j \rangle_{\mathcal{H}} | < \epsilon \quad \text{for} \quad j = 1, \ldots, n \}$$

(4.4)

$\epsilon > 0$ and $\xi_1, \ldots, \xi_n \in \mathcal{H}$.

4.7 THEOREM (Continuous linear functions for the strong operator topology). Let $\mathcal{H}$ be a Hilbert space, and let $\varphi$ be a linear functional on $\mathcal{B}(\mathcal{H})$ that is continuous in the strong operator topology. Then there exists $n \in \mathbb{N}$ and two sets of vectors $\{\zeta_1, \ldots, \zeta_n\}$ and $\{\xi_1, \ldots, \xi_n\}$ such that for all $a \in \mathcal{B}(\mathcal{H})$,

$$\varphi(a) = \sum_{j=1}^{n} \langle \zeta_j, a\xi_j \rangle_{\mathcal{H}}.$$  

(4.5)

Evidently, every such linear functional is weakly continuous, and hence every strongly continuous linear functional is weakly continuous. Consequently, a convex subset of $\mathcal{B}(\mathcal{H})$ is strongly closed if and only if it is weakly closed.

Proof. If $\varphi$ is strongly continuous, then $\varphi^{-1}(\{\lambda : |\lambda| < 1\})$ contain a neighborhood of 0 in $\mathcal{B}(\mathcal{H})$. Thus, there exists an $\epsilon > 0$ and a set of $n$ vectors $\xi_1, \ldots, \xi_n$, which without loss of generality we may assume to be orthonormal, such that if $\|a\xi_j\| < \epsilon$ for $j = 1, \ldots, n$, $|\varphi(a)| < 1$. Note that if $a\xi_j = 0$ for $j = 1, \ldots, n$, then $t > 0$, $\|ta\xi_j\| < \epsilon$ for $j = 1, \ldots, n$, and consequently $t|\varphi(a)| < 1$. It follows that

$$a\xi_j = 0 \quad \text{for} \quad j = 1, \ldots, n \quad \Rightarrow \quad \varphi(a) = 0.$$  

(4.6)

For any $a \in \mathcal{B}(\mathcal{H})$, define $\hat{a}$ by $\hat{a} = \sum_{j=1}^{n} a\xi_j \langle \xi_j, \cdot \rangle_{\mathcal{H}}$. Evidently $(a - \hat{a})\xi_j = 0$ for $j = 1, \ldots, n$, and hence by (4.6),

$$\varphi(a) = \varphi(\hat{a}) = \sum_{j=1}^{n} \varphi(a\xi_j) \langle \xi_j, \cdot \rangle_{\mathcal{H}}.$$  

(4.7)

For each fixed $j$, and any $\eta \in \mathcal{H}$, consider the rank-one operator $\eta \langle \xi_j, \cdot \rangle_{\mathcal{H}}$. Then $\eta \mapsto \varphi(\eta \langle \xi_j, \cdot \rangle_{\mathcal{H}})$ is a bounded linear functional on $\mathcal{H}$, and therefore by the Riesz Representation Theorem, there is a vector $\zeta_j \in \mathcal{H}$ such that $\langle \zeta_j, \eta \rangle_{\mathcal{H}} = \varphi(\eta \langle \xi_j, \cdot \rangle_{\mathcal{H}})$ for all $\eta \in \mathcal{H}$. Combining this with (4.7) yields (4.5). The final statement is a standard application of the Hahn-Banach Theorem.

4.2 The measurable functional calculus

Let $a \in \mathcal{B}(\mathcal{H})$ be self-adjoint, and for brevity let $\sigma(a)$ denote $\sigma_{\mathcal{B}(\mathcal{H})}(a)$. Let $\eta \in \mathcal{H}$, and define a linear functional $\mu_\eta$ on $C(\sigma(a))$ through

$$\mu_\eta(f) = \langle \eta, f(a)\eta \rangle_{\mathcal{H}}.$$  

(4.8)
Then $\mu$ is evidently a positive linear functional with $\mu_\eta(1) = \|\eta\|_{\mathscr{H}}^2$. By the Reisz-Markoff Theorem, there is a positive Borel measure of total mass $\|\eta\|_{\mathscr{H}}^2$, also denoted by $\mu_\eta$, so that for all $f \in C(\sigma(a))$,

$$\mu_\eta(f) = \int_{\sigma(a)} f \, d\mu_\eta. \quad (4.9)$$

Combining (4.8) and (4.9), we conclude that for all $f \in C(\sigma(a))$, $\langle \eta, f(a) \eta \rangle_{\mathscr{H}} = \int_{\sigma(a)} f \, d\mu_\eta$.

Now let $\{\eta_j\}$ be any dense sequence in the unit sphere of $\mathscr{H}$, and define the probability measure $\nu$ on $\sigma(a)$ by

$$\nu = \sum_{j=1}^{\infty} 2^{-j} \mu_{\eta_j}. \quad (4.10)$$

By (4.8) and (4.10), for all $f, g \in C(\sigma(a))$, and all $k$,

$$2^{-k} \| (f(a) - g(a)) \eta_k \|_{\mathscr{H}}^2 \leq \sum_{j=1}^{\infty} 2^{-j} \langle \eta_j, |f(a) - g(a)|^2 \eta_j \rangle_{\mathscr{H}} \leq \int_{\sigma(a)} |f - g|^2 \, d\nu. \quad (4.11)$$

Recall that the continuous functions on $\sigma(a)$ are dense in $L^1(\sigma(a), \mu)$ for any Borel measure $\mu$, and that from any sequence that converges in $L^1(\sigma(a), \mu)$, one can extract a subsequence that converges a.e. to $f$. It follows that if $f$ is any bounded Borel function of $\sigma(a)$, there exists a sequence $\{f_n\}$ of uniformly bounded continuous functions on $\sigma(a)$ with $\lim_{n \to \infty} f_n(\lambda) = f(\lambda)$ for $\nu$ a.e. $\lambda$, then by the Lebesgue Dominated Convergence Theorem, $\lim_{n \to \infty} \int_{\sigma(a)} |f_n - f|^2 \, d\nu = 0$. Consequently for all $\epsilon > 0$, $\int_{\sigma(a)} |f_n - f|^2 \, d\nu < \epsilon$ for all but finitely many $m$ and $n$. Then by (4.10), $\{f_n(a)\}$ is a Cauchy sequence for the strong operator topology, and hence $\lim_{n \to \infty} f_n(a) = b$ exists for this topology. In particular, for all $\xi \in \mathscr{H}$,

$$\lim_{n \to \infty} f_n(a) \xi = b \xi. \quad (4.12)$$

We would like to define $f(a) = b$, but at this point, one might suppose that the definition depends on the approximating sequence of continuous functions, or on the choice of the dense sequence $\{\eta_j\}$ in the unit sphere of $\mathscr{H}$. In fact, it does not.

First, let $f$ be a bounded Borel function on $\sigma(a)$, and let $\{f_n\}$ and $\{\tilde{f}_n\}$ be two sequences of continuous functions that converge $\nu$ a.e. to $f$, where $\nu$ is defined by (4.10) for some choice of a dense sequence $\{\eta_j\}$ in the unit sphere of $\mathscr{H}$. Define the “interlaced” sequence $\{g_n\}$ by $g_{2n-1} = f_n$ and $g_{2n} = \tilde{f}_n$. Then evidently $\{g_n\}$ converges $\nu$ a.e. to $f$, and so $b = \lim_{n \to \infty} g_n(a)$ exists in the weak operator topology. Since subsequences of convergent sequences converge to the same limit, we have

$$b = \lim_{n \to \infty} f_n(a) = \lim_{n \to \infty} \tilde{f}_n(a). \quad (4.13)$$

Next, let $\{\eta_j\}$ and $\{\tilde{\eta}_j\}$ be two dense sequences in the unit sphere of $\mathscr{H}$, and define $\nu$ and $\tilde{\nu}$ in terms of them as in (4.10). Then $\nu$ and $\tilde{\nu}$ are equivalent measures, meaning that a Borel set $E \subset \sigma(a)$ is a null set for one if and only if it is also a null set for the other. To see this, suppose on the contrary that $E$ is a Borel subset of $\sigma(a)$ with $\nu(E) > 0$ but $\tilde{\nu}(E) = 0$. Let $f$ be the indicator function of $E$. Let $\mu = \nu + \tilde{\nu}$, and let $\{f_n\}$ be a sequence of continuous non-negative functions that
converges \( \mu \) a.e. to \( f \), and hence converges both \( \nu \) and \( \tilde{\nu} \) a.e. Since each \( f_n(a) \) is self adjoint and nonnegative, so it the weak limit \( b \). Therefore

\[
\lim_{n \to \infty} \int_{\sigma(a)} f_n(\lambda) d\nu = \nu(E) > 0 \quad \text{while} \quad \lim_{n \to \infty} \int_{\sigma(a)} f_n(\lambda) d\tilde{\nu} = \tilde{\nu}(E) = 0 .
\]

This would imply that

\[
\sum_{j=1}^{\infty} 2^{-j} \langle \eta_j b \eta_j \rangle_{\mathcal{H}} > 0 \quad \text{while} \quad \sum_{j=1}^{\infty} 2^{-j} \langle \tilde{\eta}_j b \tilde{\eta}_j \rangle_{\mathcal{H}} = 0 ,
\]

and then since each term in the sum on the right is non-negative, \( \langle \tilde{\eta}_j b \tilde{\eta}_j \rangle_{\mathcal{H}} = 0 \) for all \( j \), while for at least one \( j_0 \), \( \langle \eta_{j_0} b \eta_{j_0} \rangle_{\mathcal{H}} > 0 \). But since \( \{ \tilde{\eta}_j \} \) is dense in the unit sphere, there is a subsequence \( \{ \tilde{\eta}_{j_k} \} \) with \( \lim_{k \to \infty} \tilde{\eta}_{j_k} = \eta_{j_0} \), and this then forces \( \langle \eta_{j_0} b \eta_{j_0} \rangle_{\mathcal{H}} = 0 \). The contradiction shows there is no such Borel set \( E \) and hence the two measures are equivalent.

In other words, the self adjoint operator \( a \) determines a class of mutually equivalent Borel measures, and if \( E \) is a null set for this class then \( \mu_\eta(E) = 0 \) for all \( \eta \in \mathcal{H} \) since for \( \eta \neq 0 \), we may include \( \| \eta \|_{\mathcal{H}}^{-1} \eta \) is any dense sequence in the unit sphere. Thus, when discussing a.e. convergence of functions on the spectrum of \( a \), we shall always mean almost everywhere with respect to any one of these equivalent measures, and then for any Bounded Borel function \( f \) on \( \sigma(a) \), we define

\[
f(a) = \lim_{n \to \infty} f_n(a) \quad (4.13)
\]

where \( \{ f_n \} \) is any sequence of continuous functions on \( \sigma(a) \) that converges almost everywhere to \( f \) in this sense. By what we have noted above, such sequences always exist, the limit always exists, and the limit is independent of the approximating sequence \( \{ f_n \} \) and of the particular reference measure used in the construction. We have prepared the way for an easy proof of the following theorem:

**4.8 THEOREM** (Functional Calculus For Bounded Self-Adjoint Operators). Let \( \mathcal{H} \) be a separable Hilbert space, and let \( a \) be a self-adjoint element of \( \mathcal{B}(\mathcal{H}) \). Let \( \mathcal{B}(\sigma(a)) \) denote the bounded Borel functions on \( \sigma(a) \). Then for \( f \in \mathcal{B}(\sigma(a)) \), \( f(a) \) is defined through (4.13) as described above.

The function \( f \mapsto f(a) \) is a norm-reducing \( * \)-isomorphism from \( \mathcal{B}(\sigma(a)) \) into \( \mathcal{B}(\mathcal{H}) \). Moreover:

1. Its restriction to the continuous functions \( \mathcal{B}(\sigma(a)) \) agrees with the function \( f \mapsto f(a) \) given by that Abstract Spectral Theorem. In particular, \( a \) is the image of \( \lambda \mapsto \lambda \) and the identity is the image of \( \lambda \mapsto 1 \).
2. Let \( \{ f_n \} \) be a sequence in \( \mathcal{B}(\sigma(a)) \) sup_{n \in \mathbb{N}} \| f_n \|_\mathcal{H} < \infty \), and such that \( \lim_{n \to \infty} f_n(\lambda) = f(\lambda) \) for all \( \lambda \in \sigma(a) \). Then \( f(a) = \lim_{n \to \infty} f_n(a) \) in the strong operator topology.
3. The function \( f \mapsto f(a) \) preserves order: If \( g \geq f \) in \( \mathcal{B}(\sigma(a)) \), then \( g(a) - f(a) \) is non-negative.

*Proof.* Let \( f \in \mathcal{B}(\sigma(a)) \) and let \( \{ f_n \} \) be a bounded sequence in \( \mathcal{C}(\sigma(a)) \) converging a.e. to \( f \). Then for all \( \eta, \xi \in \mathcal{H} \),

\[
\langle \zeta, (f(a))^* \zeta \rangle_{\mathcal{H}} = \langle f(a) \zeta, \xi \rangle_{\mathcal{H}} = \lim_{n \to \infty} \langle f_n(a) \zeta, \xi \rangle_{\mathcal{H}} = \lim_{n \to \infty} \langle \zeta, f_n^*(a) \xi \rangle_{\mathcal{H}} = \langle \zeta, f^*(a) \xi \rangle_{\mathcal{H}} .
\]
The map \( f \mapsto f(a) \) is evidently linear, and we conclude the proof that it is a 8-homomorphism by showing that for \( f, g \in \mathcal{B}(\sigma(a)) \), \( fg(a) = f(a)g(a) \). Let \( \{f_n\} \) and \( \{g_n\} \) be bounded sequences in \( \mathcal{C}(\sigma(a)) \) converging a.e. to \( f \) and \( g \) respectively. For any \( \zeta, \xi \in \mathcal{H} \),

\[
\langle \zeta, f(a)g(a)\xi \rangle_{\mathcal{H}} = \langle f^*(a)\zeta, g(a)\xi \rangle_{\mathcal{H}} = \lim_{n \to \infty} \langle f_n^*(a)\zeta, g_n(a)\xi \rangle_{\mathcal{H}} = \lim_{n \to \infty} \langle \zeta, f_ng_n(a)\xi \rangle_{\mathcal{H}} = \langle \zeta, fg(a)\xi \rangle_{\mathcal{H}}.
\]

Next, for all \( f \in \mathcal{B}(\sigma(a)) \), and all \( \xi \in \mathcal{H} \) with \( \|\xi\|_{\mathcal{H}} = 1 \)

\[
\|f(a)\xi\|^2_{\mathcal{H}} = \int_{\sigma(a)} |f(\lambda)|^2 d\mu_\xi \leq \|f\|^2_{L^2}
\]

since \( \mu_\xi \) is a probability measure. This completes the proof that \( f \mapsto f(a) \) is a 8-homomorphism from \( \mathcal{B}(\sigma(a)) \) to \( \mathcal{B}(\mathcal{H}) \).

Properties (1) and (2) have been proved above, and to prove (3) write \( g - f = h^2 \) and use the \( * \)-homomorphism property.

The \( * \)-homomorphism provided by Theorem 4.8 need not be an isomorphism. The following example is useful elsewhere: Let \( \lambda_0 \in \sigma(a)\mathbb{N} \) and consider the function \( 1_{\lambda_0} \) given by \( 1_{\lambda_0}(\lambda) = 1 \) for \( \lambda = \lambda_0 \) and zero otherwise. Then for all \( \lambda, \lambda_01_{\lambda_0}(\lambda) = \lambda1_{\lambda_0}(\lambda) \). By the \( * \)-homomorphism property,

\[
\lambda_01_{\lambda_0}(a) = a1_{\lambda_0}(a).
\]

It follows that any non-zero vector in the range of \( 1_{\lambda_0}(a) \) is an eigenvector of \( a \) with eigenvalue \( \lambda_0 \), and conversely any such eigenvector \( \xi \) is in the range of \( 1_{\lambda_0}(a) \) as one sees by considering a continuous approximation \( \{f_n\} \) to \( 1_{\lambda_0} \): We suppose that \( \|\xi\| = 1 \), and note that

\[
\langle \xi, 1_{\lambda_0}(a)\xi \rangle_{\mathcal{H}} = \lim_{n \to \infty} \langle \xi, f_n(a)\xi \rangle_{\mathcal{H}} = \lim_{n \to \infty} f_n(\lambda_0) = 1.
\]

By the conditions for equality in the Cauchy-Schwarz inequality, and the nor reduction property,

\( 1_{\lambda_0}(a)\xi = \xi \).

In fact, for any Borel set \( E \subset \mathbb{R} \), the indicator function \( 1_E \) is a projector in \( \mathcal{B}(\sigma(a)) \), and hence, by the \( * \)-homomorphism property, \( 1_E(a) \) is a projector in \( \mathcal{B}(\mathcal{H}) \). Taking \( E = (s,t) \) for \( s < t \) yields a useful family of projectors that we shall encounter later on.

### 4.3 The polar decomposition

#### 4.9 DEFINITION (Operator absolute value).

Let \( \mathcal{H} \) be a Hilbert space and let \( a \in \mathcal{B}(\mathcal{H}) \). Then the operator absolute value of \( a \) is the operator \( |a| \) defined by

\[
|a| = \sqrt{a^*a},
\]

where the square root is taken using the Abstract Spectral Theorem.

#### 4.10 REMARK. One should not be misled by the notation: It is not in general true that \( |ab| = |a||b| \), or that \( |a^*| = |a| \) or even that \( |a+b| \leq |a| + |b| \).
Next, or each \( t > 0 \), define the operator
\[
u_t := a(t + |a|)^{-1}.
\]
This does not require the Abstract Spectral Theorem; since \(|a| \geq 0\), \((t + |a|)\) is invertible. Now note that for \( s, t > 0 \), by the Resolvent Identity
\[
u_t - \nu_s = (s - t)a[(t + |a|)^{-1}(s + |a|)^{-1}] .
\]
Hence for any \( \xi \in \mathcal{H} \), and \( 0 < s < t \),
\[
\| (\nu_t - \nu_s)\xi \|^2 = (s - t)^2 \langle \xi, |a|^2(t + |a|)^{-2}(s + |a|)^{-2}\rangle \mathcal{H}
\]
\[
= (s - t)^2 \int_{\sigma(|a|)} \frac{\lambda^2}{(t + \lambda)^2(s + \lambda)^2}d\mu_\xi
\]
\[
\leq \int_{\sigma(|a|) \setminus \{0\}} \frac{t^2}{(t + \lambda)^2}d\mu_\xi
\]
Since \( 0 \leq t^2/(t + \lambda)^2 \leq 1 \) for all \( \lambda > 0 \), and since \( \lim_{t \to 0} t^2/(t + \lambda)^2 = 0 \) for all \( \lambda > 0 \), the Lebesgue Dominated Convergence Theorem yields
\[
\lim_{t \to 0} \left( \sup_{s < t} \{ \| (\nu_t - \nu_s)\xi \|^2 \} \right) = 0 .
\]
Thus, the strong limit \( u = \lim_{t \to 0} \nu_t \) exists. Note that \( u|a| = \lim_{t \to 0} \nu_t|a| = x \lim_{t \to 0} f_t(|a|) \) where \( f_t(\lambda) = \lambda/(t + \lambda) \) since \( \lim_{t \to 0} f_t(\lambda) = 1_{(0, \infty)}(\lambda) \) for all \( \lambda \geq 0 \), it follows from Theorem 4.8 that \( \lim_{t \to 0} f_t(|a|) = 1_{(0, \infty)}(|a|) = 1 - 1_{\{0\}}(|a|) \). Since \( 1_{\{0\}}(|a|) \) is the projector onto the null space of \(|a|\), which is the null space of \( a, a1_{\{0\}}(|a|) = 0 \), and hence
\[
u|a| = a . \quad (4.15)
\]
Next note that \( u^*u = \lim_{t \to \infty} f_t^2(|a|) \) with \( f_t(\lambda) = \lambda/(t + \lambda) \) once more. It follows that
\[
u^*u = 1_{(0, \infty)}(|a|) \quad (4.16)
\]
which is the projector onto \( \ker(a)^\perp \). It follows from (4.15) that \( \text{ran}(u) = \text{ran}(a) \), and hence \( u \) is a partial isometry from \( \ker(a)^\perp \) onto \( \text{ran}(a) \).

Taking the adjoint of (4.15), we obtain \( a^* = |a|u^* \) and hence \( aa^* = ua^*au^* \). Squaring both sides and observing that \( au^*u = a \) follows from (4.16), we obtain \( (aa^*)^2 = u(a^*a)^2u^* \). An induction now yields \( (aa^*)^n = u(a^*a)^nu^* \) for all \( n \), and then taking a polynomial approximation to the square root, we conclude that
\[
u|a|u^* = |a^*|
\]

5 Representations of \( C^* \) algebras

5.1 Irreducible representations

5.1 DEFINITION. A representation of a \( C^* \)-algebra \( \mathcal{A} \) is a \(*\)-homomorphism \( \pi \) from \( \mathcal{A} \) into \( B(\mathcal{H}) \) for some Hilbert space \( \mathcal{H} \). For any subspace \( \mathcal{K} \) of \( \mathcal{H} \), we define
\[
\pi(\mathcal{A})\mathcal{K} = \{ \pi(a)\eta : a \in \mathcal{A}, \eta \in \mathcal{K} \} .
\]
A subspace \( \mathcal{K} \) of \( \mathcal{H} \) is invariant under \( \pi \) in case \( \pi(\mathcal{A})\mathcal{K} \subset \mathcal{K} \). The representation \( \pi \) is irreducible in case no non-trivial subspace \( \mathcal{K} \) of \( \mathcal{H} \) is invariant under \( \pi \). The representation \( \pi \) is non-degenerate in case \( \pi(\mathcal{A})\mathcal{H} = \mathcal{H} \). Let \( \pi_1 \) and \( \pi_2 \) be two representations of \( \mathcal{A} \) on Hilbert spaces \( \mathcal{H}_1 \) and \( \mathcal{H}_2 \), respectively. Then \( \pi_1 \) and \( \pi_2 \) are equivalent representations of \( \mathcal{A} \) in case there exists a unitary transformation from \( u \) from \( \mathcal{H}_1 \) onto \( \mathcal{H}_2 \) such that for all \( a \in \mathcal{A} \),

\[
\pi_2(a)u = u\pi_1(a).
\]

The notion of the commutant of a subset \( S \subset \mathcal{B}(\mathcal{H}) \) plays a crucial role in the study of irreducibility.

5.2 DEFINITION. Let \( \mathcal{H} \) be a Hilbert space, and \( S \subset \mathcal{B}(\mathcal{H}) \). The commutant \( S' \) of \( S \) is the subset of \( \mathcal{B}(\mathcal{H}) \) given by

\[
S' = \{ a \in \mathcal{B}(\mathcal{H}) : ab - ba = 0 \quad \text{for all} \quad b \in S \}.
\]

5.3 LEMMA. Let \( \mathcal{H} \) be a Hilbert space, and \( S \subset \mathcal{B}(\mathcal{H}) \). The commutant \( S' \) of \( S \) has the following properties:

(1) \( S' \) is a closed in the weak operator topology on \( \mathcal{B}(\mathcal{H}) \), and contains the identity 1.

(2) \( S' \) is a subalgebra of \( \mathcal{B}(\mathcal{H}) \).

(3) If \( S \) is closed under the involution, then so is \( S' \), so that \( S' \) is a weakly closed \( * \)-subalgebra of \( \mathcal{B}(\mathcal{H}) \) that contains the identity.

Proof. It is evident that for all \( S, 1 \in S' \). Moreover, for any \( \zeta, \xi \in \mathcal{H} \) and \( b \in S \), define the linear functional \( \varphi_{\zeta,\xi,b} \) on \( \mathcal{B}(\mathcal{H}) \) by

\[
\varphi_{\zeta,\xi,b}(a) = \langle \zeta, (ab - ba)\xi \rangle_\mathcal{H} = \langle \zeta, a(b\xi) \rangle_\mathcal{H} - \langle (b^*\zeta), a\xi \rangle_\mathcal{H}.
\]

Since \( \varphi_{\zeta,\xi,b} \) is weakly continuous, \( \varphi_{\zeta,\xi,b}^{-1}(\{0\}) \) is weakly closed. Then since

\[
S' = \bigcap \{ \varphi_{\zeta,\xi,b}^{-1}(\{0\}) : \zeta, \xi \in \mathcal{H}, b \in S \},
\]

(1) is proved. (2) is evident, and the (3) follows from the fact that \( (ab - ba)^* = (b^*a^* - a^*b^*) \) together with (1) and (2).

5.4 DEFINITION. A von Neumann algebra is a \( * \)-subalgebra \( \mathcal{M} \) of \( \mathcal{B}(\mathcal{H}) \) for some Hilbert space \( \mathcal{H} \) such that \( 1 \in \mathcal{M} \) and such that \( \mathcal{M} \) is a weakly closed subset of \( \mathcal{B}(\mathcal{H}) \).

By Lemma 5.2, the commutant of any \( * \)-subalgebra of \( \mathcal{B}(\mathcal{H}) \) is a von Neumann algebra. Note that every von Neumann algebra \( \mathcal{M} \) is generated by the projections it contains. Indeed, \( \mathcal{M} \) is generated by is self adjoint elements, and by the Spectral Theorem, each self adjoint \( a \in \mathcal{M} \) is the strong limit os a sequence of finite linear combinations of the spectral projections of \( a \), which themselves belong to \( \mathcal{M} \), being strong limits of polynomials in \( a \).

5.5 LEMMA. Let \( \mathcal{A} \) be a \( C^* \) algebra, and let \( \pi \) be a non-zero representation of it as an algebra of operators on some Hilbert space \( \mathcal{H} \). Then a closed subspace \( \mathcal{K} \) of \( \mathcal{H} \) is invariant under \( \pi(\mathcal{A}) \) if and only if the orthogonal projection of \( \mathcal{H} \) onto \( \mathcal{K} \) belongs to \( (\pi(\mathcal{A}))' \).
Proof. First, \( \mathcal{K} \) is invariant under \( \pi(\mathcal{A}) \) if and only if \( \mathcal{K}^\perp \) is invariant under \( \pi(\mathcal{A}) \). To see this, let \( \zeta \in \mathcal{K}^\perp \) and \( \xi \in \mathcal{K} \), and \( a \in \mathcal{A} \). If \( \mathcal{K} \) is invariant, \( \pi(a^*)\xi \in \mathcal{K} \), and hence

\[
(\pi(a)\zeta, \xi)_{\mathcal{K}} = (\zeta, \pi(a^*)\xi)_{\mathcal{K}} = 0.
\]

Thus the invariance of \( \mathcal{K} \) implies the invariance of \( \mathcal{K}^\perp \), and then by symmetry, the reverse implication is valid as well.

Now let \( p \) be the orthogonal projection onto \( \mathcal{K} \). Then when \( \mathcal{K} \) is invariant, for all \( a \in \mathcal{A} \),

\[
0 = p\pi(a)(1-p) = p\pi(a) - p\pi(a)p = p\pi(a) - \pi(a)p
\]

where the last equality is true since the range of \( \pi(a)p \) lies in \( \mathcal{K} \). Therefore, \( p \in (\pi(\mathcal{A}))' \). Conversely, if \( p \in (\pi(\mathcal{A}))' \) and \( \xi \in \mathcal{K} \), then for all \( a \in \mathcal{A} \),

\[
\pi(a)\xi = \pi(a)p\xi = p\pi(a)\xi \in \mathcal{K},
\]

which shows the invariance of \( \mathcal{K} \).

\[\square\]

Lemma 5.5 permits us to make the following definition:

5.6 DEFINITION. For a representation \( \pi \) of a \( \mathbb{C}^* \) algebra \( \mathcal{A} \) on a Hilbert space \( \mathcal{H} \), and a non-zero projector \( p \in (\pi(\mathcal{A}))' \), \( \pi_p \) is the subrepresentation obtained by restricting \( \pi \) to \( \text{ran}(p) \).

5.7 THEOREM. Let \( \mathcal{A} \) be a \( \mathbb{C}^* \) algebra, and let \( \pi \) be a non-zero representation of it as an algebra of operators on some Hilbert space \( \mathcal{H} \). Then \( \pi \) is irreducible if and only if \( (\pi(\mathcal{A}))' \) consists of scalar multiples of the identity.

Proof. If \( (\pi(\mathcal{A}))' \) consists of scalar multiples of the identity, then \( (\pi(\mathcal{A}))' \) contains no non-trivial orthogonal projections, and hence by Lemma 5.5, \( \pi \) is irreducible. On the other hand, if \( (\pi(\mathcal{A}))' \) contains some operator that is not a multiple of the identity, then it contains a self-adjoint operator \( a \) that is not a multiple of the identity. Any such \( a \in (\pi(\mathcal{A}))' \) has a non-trivial spectral projection that is also in \( (\pi(\mathcal{A}))' \) since \( (\pi(\mathcal{A}))' \) is a von Neumann algebra containing \( a \).

\[\square\]

5.8 THEOREM (von Neumann Double Commutant Theorem). Let \( \mathcal{A} \) be a \( \ast \)-subalgebra of \( \mathcal{B}(\mathcal{H}) \) that contains the identity. Then \( \mathcal{A}'' \) is the weak operator topology closure of \( \mathcal{A} \).

Proof. Since \( \mathcal{A} \) is convex, the weak and strong operator topology closes of \( \mathcal{A} \) coincide. Hence it suffices to show that for all \( a \in \mathcal{A}'' \), every strong neighborhood of \( a \) contains some \( b \in \mathcal{A} \). That is, it suffices to show that for all \( n \in \mathbb{N} \) and all \( \{\eta_1, \ldots, \eta_n\} \subset \mathcal{H} \), and all \( \epsilon > 0 \), there is some \( b \in \mathcal{A} \) such that \( \| (b - a)\eta_j \| < \epsilon \) for all \( j = 1, \ldots, n \).

Let \( \mathcal{H} = \mathcal{H} \oplus \cdots \oplus \mathcal{H} \), the direct sum of \( n \) copies of \( \mathcal{H} \). The elements of \( \mathcal{B}(\mathcal{H}) \) are \( n \times n \) matrices \( [b_{i,j}] \) with entries in \( \mathcal{B}(\mathcal{H}) \).

Let \( \mathcal{A} \) be the algebra of all operators on \( \mathcal{H} \) of the form \( [a\delta_{i,j}] \) with \( a \in \mathcal{A} \). Evidently, its commutator \( \mathcal{A}' \) consists of all \( [b_{i,j}] \) with each \( b_{i,j} \in \mathcal{A}' \). Thus, for all \( a \in \mathcal{A}' \), \( [a\delta_{i,j}] \in \mathcal{A}'' \).

Let \( \eta = \eta_1 \oplus \cdots \oplus \eta_n \), and define \( \mathcal{K} = \mathcal{A}\eta \) which is a closed subspace of \( \mathcal{H} \) that is invariant under \( \mathcal{A} \). By Lemma 5.5, the orthogonal projection \( p \) of \( \mathcal{H} \) onto \( \mathcal{K} \) belongs to \( \mathcal{A}' \), and hence to \( \mathcal{A}'' \). Then by Lemma 5.5 again, for \( a \in \mathcal{A}' \), \( \mathcal{K} \) is invariant under \( \mathcal{A}' \). In particular, for all \( a \in \mathcal{A}' \), \( \mathcal{K} \) is invariant under \( [a\delta_{i,j}] \).
Since \( \mathcal{A} \) contains the identity, \( \eta \in \mathcal{H} \), so that \( a\eta_1 \oplus \cdots \oplus a\eta_n \in \mathcal{H} \). Therefore, for all \( \epsilon > 0 \), there exists \( b \in \mathcal{A} \) such that
\[
\|b\eta_1 \oplus \cdots \oplus b\eta_n - a\eta_1 \oplus \cdots \oplus a\eta_n\|_{\mathcal{H}}^2 \leq \epsilon^2.
\]
\qed

In particular, the weak operator topological closure of any \(*\)-subalgebra of \( \mathcal{B}(\mathcal{H}) \) containing the identity is again a \(*\)-algebra containing the identity, and hence is a von Neumann algebra, though this can be seen directly.

For any self-adjoint operator \( a \) in \( \mathcal{B}(\mathcal{H}) \), note that \( \{a\}' = (C(a))' \) where \( C(a) \) is the \( C^* \) algebra generated by \( a \). Hence \( \{a\}'' \) is the smallest von Neumann algebra that contains \( a \). That is, \( \{a\}'' \) is the von Neumann algebra generated by \( a \).

**5.9 Theorem.** On a separable Hilbert space \( \mathcal{H} \), every abelian von Neumann algebra \( \mathcal{Z} \) is generated by a single self adjoint operator; i.e, for some self adjoint \( a \in \mathcal{Z} \), \( \mathcal{Z} = \{a\}'' \).

**Proof.** Recall that subsets of separable spaces are separable. Let \( \{p_n\} \) be a sequence of projections in \( \mathcal{Z} \) that is dense for the strong operator topology in the set of all projections in \( \mathcal{Z} \). Define
\[
a = \sum_{j=1}^{\infty} 3^{-j}p_j.
\]
The sum converges in operator norm, and hence belongs to \( \mathcal{Z} \). Note also that \( \|\sum_{j=2}^{\infty} 3^{-j}p_j\| \leq 2/9 \), and hence \( \|a - 3^{-1}p_1\| \leq 2/9 \).

Pick \( \lambda_0 \in (2/9, 1/3) \) and let \( q = 1_{(\lambda_0, 1)}(a) \). Then \( q \) and \( p_1 \) are commuting projections, and hence \( q^\perp p_1 \) and \( p_1 q^\perp \) are projections.

If \( qp_1^\perp \neq 0 \), there is a unit vector \( \eta \) with \( q\eta = \eta \) and \( p_1^\perp \eta = \eta \). Then since \( qaq \geq \lambda_0 q \),
\[
\lambda_0 \leq \langle \eta a q \eta \rangle_{\mathcal{H}} = \langle \eta a \rangle_{\mathcal{H}} = \langle \eta p_1^\perp a p_1^\perp \eta \rangle_{\mathcal{H}} \leq \langle \eta (a - 3^{-1}p_1) \eta \rangle_{\mathcal{H}} \leq 2/9.
\]
This is a contradiction, and so \( qp_1^\perp = 0 \).

If \( q^\perp p_1 \neq 0 \), there is a unit vector \( \eta \) with \( q^\perp \eta = \eta \) and \( p_1 \eta = \eta \). Then since \( \lambda_0 q^\perp \geq q^\perp a q^\perp \),
\[
\lambda_0 \geq \langle \eta q^\perp a q^\perp \eta \rangle_{\mathcal{H}} = \langle \eta a \rangle_{\mathcal{H}} = \langle \eta p_1 a p_1 \eta \rangle_{\mathcal{H}} \geq \langle \eta (3^{-1}p_1) \eta \rangle_{\mathcal{H}} \geq 1/3.
\]
This is a contradiction, and so \( q^\perp p_1 = 0 \).

Then since \( q^\perp p_1 = qp_1^\perp = 0 \), \( q = qp_1 + qp_1^\perp = qp_1 = qp_1 + q^\perp p_1 = p_1 \). This shows that the spectral projection of \( a \) for the interval \( (\lambda_0, 1) \) is \( p_1 \). Inductively, one finds that each \( p_j \) is a spectral projection for \( a \), and hence belongs to \( \{a\}''' \).
\qed

**5.2 Central covers**

Let \( \pi \) be a non-degenerate representation of a \( C^* \) algebra \( \mathcal{A} \) on a Hilbert space \( \mathcal{H} \). If \( q \) is any projection in the center of \( (\pi(\mathcal{A}))' \), which is \( (\pi(\mathcal{A}))' \cap (\pi(\mathcal{A}))'' \), the range of \( q \) is invariant under both \( \pi(\mathcal{A}) \) and \( (\pi(\mathcal{A}))' \) and thus the restriction of \( \pi_q \) is a subrepresentation of \( \pi \).
5.10 DEFINITION. (Central projection for a representation) Let $\pi$ be a non-degenerate representation of a $C^*$ algebra $\mathcal{A}$ on a Hilbert space $\mathcal{H}$. A central projection for $\pi$ is a projection in the center of $(\pi(\mathcal{A}))'$.

5.11 LEMMA. Let $\pi$ be a non-degenerate representation of a $C^*$ algebra $\mathcal{A}$ on a Hilbert space $\mathcal{H}$, and $p$ be a projection in $(\pi(\mathcal{A}))'$. There exists a central projection $\overline{p}$ that is dominated by every central projection that dominates $p$. Moreover, $\overline{p}$ is the projection onto $(\pi(\mathcal{A}))'\text{ran}(p)$.

Finally, if $\pi_p$ is irreducible, $\overline{p}$ is the central cover of every projection $q \in (\pi(\mathcal{A}))'$ dominated by $\overline{p}$.

Proof. If $q$ is a projection in the center of $(\pi(\mathcal{A}))$ that dominates $p$, then the range of $q$ contains $\mathcal{H}$, and since $q$ commutes with $(\pi(\mathcal{A}))'$, $(\pi(\mathcal{A}))'\mathcal{H}$ is contained in the range of $q$.

For the final part, suppose that $q \in (\pi(\mathcal{A}))'$ is dominated by $\overline{p}$. Since $\overline{p}$ is a central projection that dominates $q$, $\overline{q} \leq \overline{p}$. It suffices to show that $p \leq \overline{q}$.

Note that $pq\overline{q}$ and $pq\overline{q}^\perp$ are projections in $(\pi(\mathcal{A}))'$ that are dominated by $p$. Since $\pi_p$ is irreducible, one must be zero, and the other must be $p$. If $pq\overline{q}^\perp = p$, then $pq\overline{q}$ is a central projection dominated $p$, and hence $pq\overline{q}^\perp = \overline{p}$. This is impossible since $q \leq \overline{p}$. Hence it must be the case that $pq\overline{q} = p$, which is what we needed to show. \qed

5.12 DEFINITION. (Central cover and central subrepresentations) Let $\pi$ be a non-degenerate representation of a $C^*$ algebra $\mathcal{A}$ on a Hilbert space $\mathcal{H}$, and let $\sigma$ be a subrepresentation of $\pi$ on a subspace $\mathcal{K}$ of $\mathcal{H}$. Let $p$ be the projector onto $\mathcal{K}$. Then the central cover of $\sigma$ is the representation $\overline{\sigma} = \pi_p$. A subrepresentation $\sigma$ of $\pi$ is a central subrepresentation in case $\sigma = \overline{\sigma}$.

Unless $\sigma$ is already a central representation, its central cover is a strictly larger subrepresentation of $\pi$. The precise sense in which it is larger makes notion of central representations fundamentally important in the study of the structure of representations.

5.13 LEMMA. Let $\pi$ be a non-degenerate representation of a $C^*$ algebra $\mathcal{A}$ on a Hilbert space $\mathcal{H}$. Let $p$ and $q$ be two projections in $(\pi(\mathcal{A}))'$ and suppose that they have the same central cover; i.e., $\overline{p} = \overline{q}$. Then there is a partial isometry $u \in (\pi(\mathcal{A}))'$ such that $uu^* \leq q$ and $u^*u \leq p$.

Proof. Let $\mathcal{K}_1$ and $\mathcal{K}_2$ denote the ranges of $p$ and $q$ respectively. Since $(\pi(\mathcal{A}))'\mathcal{K}_1 = (\pi(\mathcal{A}))'\mathcal{K}_2$, there is an $a \in (\pi(\mathcal{A}))'$ and vectors $\eta_1$ and $\eta_2$ in $\mathcal{K}_1$ and $\mathcal{K}_2$ respectively so that $\langle \eta_2, a\eta_2 \rangle_{\mathcal{H}} \neq 0$. This means that $z = qap$ is a non-zero element of $(\pi(\mathcal{A}))'$. Let $z = u|z|$ be its polar decomposition. Then $u$ is a partial isometry in $(\pi(\mathcal{A}))'$, such that $uu^* \leq q$ and $u^*u \leq p$. Thus, $u^*u$ is a subrepresentation of $\pi_p$ that is equivalent to $\pi_{uu^*}$, a subrepresentation of $q$. \qed

Lemma 5.13 has the following consequence: Since for any partial isometry $u$ $u^*u$ and $uu^*$ are, respectively, the projectors onto the final and initial spaces of $u$, $\pi_{uu^*}$ and $\pi_{u^*u}$ are non-zero equivalent subrepresentations of the representation $\pi_p$ and $\pi_q$ discussed in the lemma. In particular, if $\pi_p$ is irreducible, there is a partial isometry $u \in (\pi(\mathcal{A}))'$ such that $u^*u = p$ and $uu^* \leq q$, so that $\pi_p$ is equivalent to a subrepresentation of $\pi_q$. The fact that the equivalence is due to a partial isometry in $(\pi(\mathcal{A}))'$ is important in what follows.
5.3 The structure of type I factors of von Neumann algebras

5.14 DEFINITION. Let $H$ be a Hilbert space, and let $\mathcal{M}$ be a von Neumann algebra on $H$. $\mathcal{M}$ is a factor in case $\mathcal{M}$ has a trivial center.

Note that the center of $\mathcal{M}$ is $\mathcal{M} \cap \mathcal{M}'$, which by the Double Commutant Theorem, is the same as $\mathcal{M}' \cap \mathcal{M}''$, so that $\mathcal{M}$ and $\mathcal{M}'$ have the same center.

The center $Z$ of $\mathcal{M}$ is evidently an abelian von Neumann algebra, and therefore, by Theorem 5.9, when $\mathcal{M}$ is a von Neumann algebra on a separable Hilbert space $H$, its center $Z$ is generated by a single self-adjoint $a \in Z$, and every spectral projection of $a$ is a central projection for the identity representations of $\mathcal{M}$ and $\mathcal{M}'$.

Now suppose that this operator $a$ happens to have finite spectrum, as it must in case $H$ is finite dimensional. Then there is a finite set $\{p_1, \ldots, p_n\}$ of central projections. For each $j = 1, \ldots, n$, let $\mathcal{H}_j$ denote the range of $p_j$. Then $\mathcal{H} = \bigoplus_{j=1}^n \mathcal{H}_j$, and defining $\mathcal{M}_j = \mathcal{M} p_j$,

$$\mathcal{M} = \bigoplus_{j=1}^n \mathcal{M}_j \,.$$

(5.1)

Evidently, each $\mathcal{M}_j$ has a trivial center; its center is spanned by the identity on $\mathcal{H}_j$. The Spectral Theorem can be used to give a “direct integral” decomposition on $\mathcal{M}$ without making any assumption on the spectrum of $\mathcal{A}$, as was shown by von Neumann. This line of reasoning reduces the investigation of the structure of von Neumann algebras on a separable Hilbert space that that of von Neumann algebras with trivial center, which motivates the following definition:

5.15 DEFINITION (Factor). A factor $\mathcal{M}$ is a von Neumann algebra $\mathcal{M}$ with a trivial center. A factor is type I in case it contains a non-zero minimal projection; i.e., a non-zero projection $p$ such that the only projection in $\mathcal{M}$ that is dominated by $p$ is the zero projection.

Evidently every factor on a finite dimensional Hilbert space contains a minimal projection – any projection whose range has minimal dimension – and so every factor on a finite dimensional Hilbert space is type I. This is not true for infinite dimensional Hilbert spaces, and we shall return to a classification of types of factors and investigate their structure later. For the rest of this subsection, we focus on the structure of type I factors.

Looking at (5.1), one might think “summand” would be better terminology than “factor”, but the following theorem justifies the terminology:

5.16 THEOREM. Let $\mathcal{H}$ be a separable Hilbert space, and let $\mathcal{M}$ be a factor on $\mathcal{H}$. Then there exist Hilbert spaces $\mathcal{K}_1$ and $\mathcal{K}_2$ and a unitary $u : \mathcal{H} \to \mathcal{K}_1 \otimes \mathcal{K}_2$ such that

$$u^* \mathcal{M} u = \mathcal{B}(\mathcal{K}_1) \otimes 1_{\mathcal{K}_2} \,.$$

(5.2)

5.17 REMARK. The commutant of $\mathcal{M}'$ of $\mathcal{M} = \mathcal{B}(\mathcal{K}_1) \otimes 1_{\mathcal{K}_2}$ in $\mathcal{B}(\mathcal{K}_1 \otimes \mathcal{K}_2)$ is evidently $1_{\mathcal{K}_1} \otimes \mathcal{B}(\mathcal{K}_2)$, and hence $\mathcal{M} \cap \mathcal{M}'$ consists of multiples of the identity. Hence $\mathcal{B}(\mathcal{K}_1) \otimes 1_{\mathcal{K}_2}$ is a factor in $\mathcal{B}(\mathcal{K}_1 \otimes \mathcal{K}_2)$. The theorem says that all factors are of this type, and when $\mathcal{M}$ is a factor on a separable Hilbert space $\mathcal{H}$, $\mathcal{B}(\mathcal{H})$ is the closed span of elements of the form $ab$ where $a \in \mathcal{M}$ and $b \in \mathcal{M}'$, which may be viewed as a kind of factorization of $\mathcal{B}(\mathcal{H})$. 
Proof of Theorem 5.1. If \( \mathcal{M} \) consists of multiples of the identity, then we may take \( \mathcal{K}_1 = \mathbb{C} \), and the conclusion is obvious. Therefore, let us assume that \( \mathcal{M} \) does not consist of multiples of the identity, or, what is the same thing by Theorem 5.7, that the identity representation of \( \mathcal{M}' \) is reducible.

Let \( \pi \) denote the identity representation of \( \mathcal{M}' \), whose commutant is \( \mathcal{M} \). Let \( p_1 \) be a minimal projection in \( \mathcal{M} \). We now apply Lemma 5.13 in this setting. Since the center of \( \mathcal{M} \) is trivial, the central covers of both \( p_1 \) and \( p_1^\perp \) are the identity in \( \mathcal{M} \), and since \( p_1 \) is minimal, \( \pi_{p_1} \) is an irreducible representation of \( \mathcal{M}' \). Hence by Lemma 5.13, there is a partial isometry \( u \in \mathcal{M} \) such that \( u^*u = p_1 \) and \( uu^* \leq p_1^\perp \). Since \( p_1 = p_1^2 = u^*(uu^*)u \), \( uu^* \) is also minimal.

Define \( \mathcal{H}_j = p_j \mathcal{H} \) for \( j = 1, 2 \). If \( \mathcal{H} = \mathcal{K}_1 \oplus \mathcal{K}_2 \), we have decomposed \( \mathcal{H} \) as a direct sum of subspaces on which \( \mathcal{M}' \) acts irreducibly and equivalently. If not, repeat the argument made above with \( p_1^\perp \) replaced by \( 1 - p_1 - p_2 \), thus producing a minimal projection \( p_3 \) in \( \mathcal{M} \) with \( p_3p_j = 0 \) for \( j = 1, 2 \), and \( u_3 \), an isometry in \( \mathcal{M} \) that maps \( \mathcal{K}_1 \) onto the range of \( p_3 \). If after some finite number \( n \) of such steps, \( \mathcal{H} \) is exhausted, we have produced a set \( \{p_1, \ldots, p_n\} \) of minimal projections in \( \mathcal{M} \) with \( p_ip_j = 0 \) for \( i \neq j \), and a set \( \{u_1, \ldots, u_n\} \) of partial isometries in \( \mathcal{M} \) where \( u_j \) maps \( \mathcal{K}_1 \) onto \( \mathcal{H}_j \), the range of \( p_j \). (Note that \( u_1 \) is \( p_1 \) itself.)

Consider any \( a \in \mathcal{M} \). We claim that there is a matrix \( [a] \in M_n(\mathbb{C}) \) such that

\[
a = \sum_{i,j=1}^n [a]_{i,j} u_j u_i^* .
\]

(5.3)

To see this observe that

\[
a = \sum_{i,j=1}^n p_i ap_j = \sum_{i,j=1}^n u_i (u_i^* au_j) u_j^* .
\]

However, for each \( i, j \), \( u_i^* au_j \in \pi_{p_1} \mathcal{M} p_1 \), and since \( p_1 \) is minimal, \( \pi_{p_1} \mathcal{M} p_1 = \mathbb{C} p_1 \). Hence for some \( \lambda_{i,j} \in \mathbb{C} \), \( u_i^* au_j = \lambda_{i,j} p_1 \). Define \( [a]_{i,j} = \lambda_{i,j} \) to obtain (5.3).

Let \( \{\zeta_1, \ldots, \zeta_n\} \) denote the standard basis of \( \mathbb{C}^n \). Define a linear transformation \( u \) from \( \mathbb{C}^n \otimes \mathcal{K}_1 \) to \( \mathcal{H} \) by

\[
u \left( \sum_{j=1}^n \zeta_j \otimes \eta_j \right) = \sum_{j=1}^n u_j \eta_j .
\]

It is evident that this map is unitary. Moreover, for any \( a \in \mathcal{M} \), using (5.3), we have

\[
u u \left( \sum_{j=1}^n \zeta_j \otimes \eta_j \right) = \sum_{i,j=1}^n [a]_{i,j} u_j u_i^* u_i \eta_j = \sum_{i,j=1}^n [a]_{i,j} u_j u_i^* \eta_i = u \left( \sum_{j=1}^n \left( \sum_{i=1}^n [a]_{i,j} \zeta_i \right) \otimes \eta_j \right) = u \left( \sum_{j=1}^n [a] \zeta_j \otimes \eta_j \right)
\]

That is, with \( \mathcal{K}_1 = \mathbb{C}^n \) and \( \mathcal{K}_2 = \mathcal{K}_1 \),

\[
u u a u^* = [a] \otimes 1_{\mathcal{K}_2} .
\]

This proves (5.2) in case the procedure for producing a sequence of orthogonal minimal projections terminates in finitely many steps.
and moreover, there is a sequence \( \{p_n\} \) of minimal projections in \( \mathcal{M} \) such that \( p_mp_n = 0 \) for \( m \neq n \) and \( \mathcal{H} = \bigoplus_{n=1}^{\infty} p_n \mathcal{H} \), and moreover, there is a sequence \( \{u_n\} \) of partial isometries in \( \mathcal{M} \) such that \( u_n \) maps \( \mathcal{H}_1 \) onto \( \mathcal{H}_n \). Then the strong closure of the set of operators of the form (5.3) for some \( n \in \mathbb{N} \) is easily seen to be dense in \( \mathcal{M} \), and then with \( \mathcal{H}_1 = \ell_2 \), we obtain (5.2) in this case as well.

5.4 States on a \( C^* \) algebra

5.18 DEFINITION. Let \( \mathcal{A} \) be a \( C^* \) algebra. A linear functional \( \varphi \) on \( \mathcal{A}^* \), the Banach space dual to \( \mathcal{A} \) regarded as a Banach space, is positive in case \( \varphi(a) \geq 0 \) for all \( a \geq 0 \). If \( \mathcal{A} \) has an identity \( 1 \), a state on \( \mathcal{A} \) is a positive linear functional \( \varphi \) such that \( \varphi(1) = 1 \). We denote the set positive linear functionals by \( \mathcal{A}^*_+ \) and the states by \( \mathcal{A}^*_{+1} \). A state \( \varphi \in \mathcal{A}^*_{+1} \) is faithful in case

\[
\varphi(a^*a) = 0 \quad \Rightarrow \quad a = 0 .
\]

Evidently, for all \( \varphi \in \mathcal{A}^*_+ \), the map

\[
(a,b) \mapsto \varphi(a^*b) = \langle a,b \rangle_{\varphi}
\]

defines a (possibly degenerate) inner product on \( \mathcal{A} \); this inner product is non-degenerate if and only if \( \varphi \) is faithful. In any case, the fact that \( \langle a,a \rangle_{\varphi} \geq 0 \) for all \( a \in \mathcal{A} \) yields the Cauchy-Schwarz inequality:

\[
|\langle a,b \rangle_{\varphi}| \leq \langle a,a \rangle_{\varphi}^{1/2} \langle b,b \rangle_{\varphi}^{1/2} .
\]

5.19 THEOREM (Positivity and continuity). Let \( \mathcal{A} \) be a \( C^* \) algebra with identity \( 1 \). Then:

1. Every \( \varphi \in \mathcal{A}^*_{+1} \) is bounded, and \( \|\varphi\| = \varphi(1) \).
2. Every bounded linear functional \( \varphi \) such that \( \|\varphi\| = \varphi(1) \) is positive.

Proof. Let \( \varphi \in \mathcal{A}^*_+ \). For all \( a \in \mathcal{A} \),

\[
|\varphi(a)| = |\varphi(1a)| = |\langle 1,a \rangle_{\varphi}| \leq \langle 1,1 \rangle_{\varphi}^{1/2} \langle a,a \rangle_{\varphi}^{1/2} = \varphi(1)^{1/2} \varphi(a^*a)^{1/2} .
\]

Since \( \sigma_{\mathcal{A}}(a^*a) \subset [0,\|a\|^2], \|a\|^2 - a^*a \geq 0 \), and hence \( \varphi(a^*a)^{1/2} \leq \|a\|\varphi(1)^{1/2} \). Combining these inequalities, we have \( |\varphi(a)| \leq \varphi(1)\|a\| \) which proves (1).

For the second part, suppose that \( \varphi \in \mathcal{A}^*_{+1} \) and \( \varphi(1) = \|\varphi\| \). If \( \varphi = 0 \), it is positive. If \( \varphi \neq 0 \), we may divide by \( \|\varphi\| \) and thus may suppose that \( \|\varphi\| = \varphi(1) = 1 \).

We claim that for all \( \varphi \in \mathcal{A}^*_{+1} \) such that \( \varphi(1) = \|\varphi\| \), \( \varphi(a) \) belongs to the convex hull of \( \sigma_{\mathcal{A}}(a) \) for all \( a \geq 0 \) in \( \mathcal{A} \). To see this suppose that the closed disc of radius \( r \) centered on \( \lambda \) contains \( \sigma_{\mathcal{A}}(a) \). Then \( \lambda - a \) is normal, and its spectrum is continued in \( \{ \lambda - t : t \in \sigma_{\mathcal{A}}(a) \} \), and hence the spectral radius of \( \lambda - a \) is at most \( r \). Since \( \lambda - a \) is normal, \( \|\lambda - a\| \leq r \). Therefore,

\[
|\lambda - \varphi(a)| = |\varphi(\lambda - a)| \leq \|\lambda - a\| \leq r .
\]

Thus for all \( r > 0 \) and \( \lambda \in \mathbb{C} \), \( \varphi(a) \) is contained in the closed disc of radius \( r \) centered on \( \lambda \) contains \( \sigma_{\mathcal{A}}(a) \). The intersection over all such discs is the convex hull of \( \sigma_{\mathcal{A}}(a) \).

5.20 LEMMA. Let \( \mathcal{A} \) be a \( C^* \) algebra with identity \( 1 \). For all self adjoint \( a \in \mathcal{A} \), there exists a state \( \varphi \) such that \( |\varphi(a)| = \|a\| \).
Proof. Consider the C* algebra $C(a)$ generated by $a$ and 1. This is a commutative C* algebra, and so there is a character $\varphi_0$ of $C(a)$ such that $|\varphi_0(a)| = \|a\|$, and since $\varphi_0$ is a character $\varphi(1) = 1$. Then by Theorem 5.19, $\varphi_0 \in \mathcal{A}^*_+$, and so $\varphi$ is a state on $C(a)$.

By the Hahn-Banach Theorem, there is a norm preserving extension $\varphi$ of $\varphi_0$ (as a linear functional) to $\mathcal{A}$. Then $\varphi(1) = \varphi_0(1) = 1$, and hence by Theorem 5.19, $\varphi$ is a state, and since $\varphi$ extends $\varphi_0$, $\varphi(x) = \|x\|$.

Lemma 5.20 says, in particular, that $\mathcal{A}^*_+$ is not empty. It is evidently a closed subset of the unit ball in $\mathcal{A}^*$ in the weak-* topology, and hence is compact. $\mathcal{A}^*_+$ is also evidently convex. The Krein-Milman Theorem says that every non-empty convex set in $\mathcal{A}^*$ that is compact in the weak-* topology is the convex hull of its extreme points. Hence there exist extreme points in $\mathcal{A}^*_+$.

5.21 DEFINITION (Pure state). Let $\mathcal{A}$ be a C* algebra with identity 1. A pure state is an extreme point of $\mathcal{A}^*_+$.

5.22 THEOREM. Let $\mathcal{A}$ be a C* algebra with identity 1. For all self adjoint $a \in \mathcal{A}$, there exists a pure state $\varphi$ such that $|\varphi(a)| = \|a\|$.

Proof. By Lemma 5.20, the set $S$ of states $\varphi$ such that $\varphi(a) = \|a\|$ is non-empty, and evidently it is convex and closed in the weak-* topology. By the Krein-Milman, $S$ has at least one extreme point $\psi$. We now show that $\psi$ is extreme in $\mathcal{A}^*_+$ as well as in $S$.

Suppose that $\psi_1, \psi_2 \in \mathcal{A}^*_+$ and that $\psi = t\psi_1 + (1 - t)\psi_2$ for some $t \in (0, 1)$. Evaluating both sides at $a$,

$$\|a\| = \psi(a) = t\psi_1(a) + (1 - t)\psi_2(a) \leq t\|a\| + (1 - t)\|a\| = \|a\|.$$ 

Hence $\psi_1, \psi_2 \in S$, and so $\psi_1 = \psi_2 = \psi$. □

5.23 DEFINITION. Let $\pi$ be a representation of a C* algebra $\mathcal{A}$ on a Hilbert space $\mathcal{H}$. A vector $\eta \in \mathcal{H}$ is cyclic for $\pi$ in case $a \mapsto \pi(a)\eta$ has dense range, and is a separating vector for $\pi$ in case $a \mapsto \pi(a)\eta$ is injective. If a cyclic vector exists, then $\pi$ is a cyclic representation.

For any representation $\pi$ of $\mathcal{A}$ on $\mathcal{H}$, and any unit vector $\eta \in \mathcal{H}$, the functional $\eta_\eta \in \mathcal{A}^*$ defined by

$$\varphi_\eta(a) = \langle \eta, \pi(a)\eta \rangle_{\mathcal{H}}$$

is a state. Evidently, $\varphi_\eta(a^*a) = \langle \eta, \pi(a^*a)\eta \rangle_{\mathcal{H}} = \|\pi(a)\eta\|^2_{\mathcal{H}}$, and hence $\eta$ is separating for $\pi$ is and only if $\varphi_\eta$ is faithful. The next theorem links gives an cyclcity an irreducibility.

5.24 THEOREM. Let $\pi$ be a representation of a C* algebra $\mathcal{A}$ on a Hilbert space $\mathcal{H}$, and let $\eta$ be a cyclic unit vector for $\pi$. Then with $\varphi_\eta$ denoting the state defined in (5.6). Then $\pi$ is irreducibly iff and only if $\varphi_\eta$ is pure.

The heart of the matter is the following lemma:

5.25 LEMMA. Let $\pi$ be a representation of a C* algebra $\mathcal{A}$ on a Hilbert space $\mathcal{H}$, and let $\eta$ be a cyclic unit vector for $\pi$. Then with $\varphi_\eta$ denoting the state defined in (5.6). Suppose that $\psi \in \mathcal{A}^*_+$, and that for some $r \in (0, \infty)$,

$$\psi(a) \leq r\varphi_\eta(a) \quad \text{for all} \quad a \in \mathcal{A}.$$
Then there is a positive operator $x \in (\pi(\mathcal{A}))'$ such that $\|x\| \leq r$ and for all $a, b \in \mathcal{A}$,

$$\psi(a^* b) = \langle \eta, \pi(a), x \pi(b) \eta \rangle_{\mathcal{H}}.$$  \hspace{1cm} (5.7)

**Proof.** Define a sesquilinear form $q$ on $\pi(\mathcal{A}) \eta$ by $q(\pi(a) \eta, \pi(b) \eta) = \psi(a^* b)$. We have

$$|q(\pi(a) \eta, \pi(b) \eta)| \leq r |\pi(a) \eta, \pi(b) \eta\rangle_{\mathcal{H}}| \leq r \|\pi(a)\|_{\mathcal{H}} \|\pi(b)\|_{\mathcal{H}}.$$

Since $\eta$ is cyclic, $q$ is densely defined on $\mathcal{H}$ and extends to a sesquilinear form on all of $\mathcal{H}$, still denoted by $q$, that satisfies $|q(\zeta, \xi)| \leq r \|\zeta\|_{\mathcal{H}} \|\xi\|_{\mathcal{H}}$ for all $\zeta, \xi \in \mathcal{H}$. By Reisz's Lemma, there exists a self adjoint operator $x \in \mathcal{B}(\mathcal{H})$ such that $q(\zeta, \xi) = \langle \zeta, x \xi \rangle_{\mathcal{H}}$ for all $\zeta, \xi \in \mathcal{H}$, and $\|x\| \leq r$.

Since $q(\zeta, \zeta) \geq 0$ for $\zeta$ in the dense set $\pi(\mathcal{A}) \eta$, $x$ is positive.

Finally, note that for all $a, b, c \in \mathcal{A}$, $a^* (bc) = (b^* a)^* c$, and hence $\psi((bc)^*) = \psi((b^* a)^* c)$. This means that $q(\pi(a) \eta, \pi(b) \pi(c) \eta) = q(\pi(b^* \pi(a) \eta, \pi(c) \eta)$ which is the same as

$$\langle \pi(a) \eta, x \pi(b) \pi(c) \eta \rangle_{\mathcal{H}} = \langle \pi(a) \eta, \pi(b) x \pi(c) \eta \rangle_{\mathcal{H}}.$$

Thus for all $\zeta, \xi$ in a dense subset of $\mathcal{H}$, $\langle \zeta, x \pi(b) \xi \rangle_{\mathcal{H}} = \langle \zeta, \pi(b) x \xi \rangle_{\mathcal{H}}$ and this shows that $x$ commutes with $\pi(b)$ for arbitrary $b \in \mathcal{A}$. \hfill $\square$

**Proof of Theorem 5.24.** Suppose that $\pi$ is irreducible. Let $\psi_1, \psi_2$ be two states such that $\varphi_\eta = t \psi_1 + (1-t) \psi_2$ for some $t \in (0, 1)$. By Lemma 5.25, applied to $\psi_1$, which satisfies $\psi_1 \leq t^{-1} \varphi_\eta$, there is a positive $x \in (\pi(\mathcal{A}))'$

$$\psi_1(a^* b) = \langle \eta, \pi(a), x \pi(b) \eta \rangle_{\mathcal{H}} \quad \text{for all} \quad a, b \in \mathcal{A}. \hspace{1cm} (5.8)$$

Since $\pi$ is irreducible, $x$ must be a scalar multiple of the identity. Since $\psi_1$ is a state, taking $a = b = 1$ in (5.8), $1 = \psi_1(1) = \langle \eta, x \eta \rangle_{\mathcal{H}}$, which shows that $x = 1$. Then taking $a = 1$ in (5.8) shows that $\psi(b) = \varphi_\eta(b)$ so that $\psi_1 = \varphi_\eta$. By symmetry, $\psi_2 = \varphi_\eta$ as well, and this proves $\varphi_\eta$ is extreme.

For the converse, suppose that $\pi$ is not irreducible. Then there exists a projection $p \in (\pi(\mathcal{A}))'$ such that neither $p$ nor $p^*$ is zero. Suppose that $p \eta = 0$. Then for all $a \in \mathcal{A}$, $\pi(q)p \eta = p \pi(a) \eta = 0$ and this would mean that $p$ vanishes on a sense subspace, which is not the case. Hence $\|p \eta\|_{\mathcal{H}} > 0$, and the same reasoning shows that $\|p^* \eta\|_{\mathcal{H}} > 0$. Define $\eta_1 = \|p \eta\|_{\mathcal{H}}^{-1} p \eta$ and $\eta_2 = \|p^* \eta\|_{\mathcal{H}}^{-1} p^* \eta$. For all $a \in \mathcal{A}$,

$$\langle \eta_1, \pi(a) \eta_2 \rangle_{\mathcal{H}} = \langle p \eta_1, \pi(a) p^* \eta_2 \rangle_{\mathcal{H}} = \langle \eta_1, p p^* \pi(a) \eta_2 \rangle_{\mathcal{H}} = 0.$$

Define $t \in (0, 1)$ by $t = \|p \eta\|_{\mathcal{H}}^2 / \|p \eta\|_{\mathcal{H}}^2$. Since $\|p \eta\|_{\mathcal{H}}^2 + \|p^* \eta\|_{\mathcal{H}}^2 = 1$, $\|p^* \eta\|_{\mathcal{H}}^2 = 1 - t$. Then by the orthogonality proved just above, for all $a \in \mathcal{A}$,

$$\varphi_\eta(a) = \langle [\sqrt{t} \eta_1 + \sqrt{1-t} \eta_2], \pi(a) [\sqrt{t} \eta_1 + \sqrt{1-t} \eta_2] \rangle_{\mathcal{H}} = t \langle \eta_1, \pi(a) \eta_1 \rangle_{\mathcal{H}} + (1-t) \langle \eta_2, \pi(a) \eta_2 \rangle_{\mathcal{H}},$$

and this displays $\varphi_\eta$ as a non-trivial convex combination of states. Hence $\varphi_\eta$ is not extreme. \hfill $\square$
5.5 The GNS construction

A construction due to Gelfand, Naimark and Segal, known as the GNS construction, associates to every state \( \varphi \) on a \( C^* \) algebra \( \mathcal{A} \) a representation \( \pi \) of \( \mathcal{A} \) on a Hilbert space built out of \( \mathcal{A} \) itself and the state \( \varphi \).

5.26 THEOREM (The GNS construction). Let \( \mathcal{A} \) be a \( C^* \) algebra with identity 1, and let \( \varphi \) be a state on \( \mathcal{A} \). Then there exists a Hilbert space \( \mathcal{H} \) and a cyclic representation \( \pi \) of \( \mathcal{A} \) on \( \mathcal{H} \) with a distinguished cyclic unit vector \( \eta \) such that for all \( a \in \mathcal{A} \),

\[
\varphi(a) = \langle \eta, \pi(a)\eta \rangle_{\mathcal{H}}.
\] (5.9)

The representation \( \pi \) is irreducible if and only if \( \varphi \) is a pure state.

Proof. Let \( \langle a, b \rangle_\varphi \) be the possibly degenerate inner product on \( \mathcal{A} \) defined by \( \langle a, b \rangle_\varphi = \varphi(a^*b) \). Define

\[ \mathcal{N} := \{ a \in \mathcal{A} : \langle a, a \rangle_\varphi = 0 \} . \]

Since \( \varphi \) is continuous, \( \mathcal{N} \) is closed. In fact, \( \mathcal{N} \) is a closed left ideal. To see this, consider \( b \in \mathcal{A} \) and \( a \in \mathcal{N} \). Then

\[
\langle ba, ba \rangle_\varphi = \varphi(a^*b^*ba) = \langle a, b^*ba \rangle_\varphi \leq \langle a, a \rangle_\varphi^{1/2} \langle b^*ba, b^*ba \rangle_\varphi^{1/2} = 0 .
\]

A similar but simpler argument shows that \( \mathcal{N} \) is a subspace.

Now consider the vector space \( \mathcal{A}/\mathcal{N} \). With \( \sim \) denoting equivalence mod \( \mathcal{N} \), we have

\[ a \sim a' \quad \text{and} \quad b \sim b' \implies \langle a, b \rangle_\varphi = \langle a', b' \rangle_\varphi , \]

and hence we may define a non-degenerate inner product on \( \mathcal{A}/\mathcal{N} \) by \( \langle \{a\}, \{b\} \rangle = \langle a, b \rangle_\varphi \). Let \( \mathcal{H} \) be the completion of \( \mathcal{A}/\mathcal{N} \) in the corresponding Hilbertian norm, and let \( \langle \cdot, \cdot \rangle_{\mathcal{H}} \) denote the resulting inner product on \( \mathcal{H} \).

For \( a \in \mathcal{A} \), let \( \pi(a) \) denote the linear operator on \( \mathcal{A}/\mathcal{N} \) defined by \( \pi(a)\{b\} = \{ab\} \) which is well-defined since \( \mathcal{N} \) is a left ideal. Next note that since \( b^*a^*ab = \|a\|^2b^*b + b^*(\|a^*a\|1 - a^*a)b \) and \( b^*(\|a^*a\|1 - a^*a)b \) is positive,

\[ \|\pi(a)\{b\}\|_{\mathcal{H}}^2 = \varphi(b^*a^*ab) \leq \|a\|^2\varphi(b^*b) = \|a\|^2\|\{b\}\|_{\mathcal{H}}^2 . \]

Since \( \mathcal{A}/\mathcal{N} \) is dense in \( \mathcal{H} \), \( \pi(a) \) extends to a bounded operator on \( \mathcal{H} \) with \( \|\pi(a)\| \leq \|a\| \). It is evident that \( \pi \) is a homomorphism of \( \mathcal{A} \) into \( \mathcal{B}(\mathcal{H}) \), and note that for all \( x, y \in \mathcal{A} \),

\[ \langle \{x\}, \pi(a)\{y\} \rangle_{\mathcal{H}} = \varphi(x^*ay) = \varphi((a^*x)^*y) = \langle \pi(a)\{x\}, \{y\} \rangle_{\mathcal{H}} , \]

showing that \( \pi(a^*) = \pi(a)^{*} \), and thus \( \pi \) is a \(*\)-homomorphism.

The representation \( \pi \) is cyclic since for all \( a \in \mathcal{A} \), \( \{a\} = \{a1\} = \pi(a)\{1\} \), showing that \( \eta := \{1\} \) is a cyclic vector for \( \pi \). Finally, note that \( \langle \eta, \pi(a)\eta \rangle_{\mathcal{H}} = \varphi(1^*a1) = \varphi(a) \), and this proves (5.9). The final statement now follows from Theorem 5.24.

5.27 COROLLARY. Let \( \mathcal{A} \) be a \( C^* \) algebra with identity 1. For every non-zero \( a \in \mathcal{A} \), there is a representation \( \pi \) of \( \mathcal{A} \) such that \( \|\pi(a)\| = \|a\| \).
Proof. By Lemma 5.20, there exists \( \varphi \in \mathcal{A}_+^* \) such that \( |\varphi(a^*a)| = \|a\|^2 \). Let \( \pi \) be the GNS representation of \( \mathcal{A} \) associated to \( \varphi \), and \( \eta \) the associated distinguished cyclic unit vector. Then

\[
\|\pi(a)\eta\|_{\mathcal{H}}^2 = \langle \eta \pi(a^*a)\eta \rangle_{\mathcal{H}} = \varphi(a^*a) = \|a\|^2 ,
\]

showing that \( \|\pi(a)\| \geq \|a\| \), and since it is automatic that \( \|\pi(a)\| \leq \|a\| \), \( \|\pi(a)\| = \|a\| \). \( \square \)

We now arrive at the Non-Commutative Gelfand-Naimark Theorem:

5.28 THEOREM (Non-Commutative Gelfand-Naimark Theorem). Every C* algebra \( \mathcal{A} \) with an identity is isometrically *-isomorphic to a C* algebra of operators.

Proof. For each \( a \in \mathcal{A} \) choose an irreducible representation \( \pi \) of \( \mathcal{A} \) such that \( \|\pi(a)\| = \|a\| \). Now form the direct sum of all of these representations.

5.6 The GNS construction for \( M_n(\mathbb{C}) \) and the normalized trace

Fix \( n \in \mathbb{N} \), \( n \geq 2 \), and let \( \mathcal{A} = M_n(\mathbb{C}) \). Define \( \varphi_{\text{tr}} \in \mathcal{A}^* \) by

\[
\varphi_{\text{tr}}(a) = \frac{1}{n} \text{Tr}(a) .
\]

Since \( \varphi_{\text{tr}}(a^*a) = \sum_{i,j=1}^n |a_{i,j}|^2 \), it is evident that \( \varphi_{\text{tr}}(a) \) is a state, called the normalized trace. It is also evident from the same computation that the normalized trace is faithful. It has one more important property: For all \( a, b \in \mathcal{A} \),

\[
\varphi_{\text{tr}}(ab) = \varphi_{\text{tr}}(ba) , \tag{5.10}
\]

as one readily verifies.

Since \( \varphi_{\text{tr}} \) is faithful, the left ideal \( \mathcal{N} \) that arose in the GNS construction is simply \( \{0\} \), and so the Hilbert space \( \mathcal{H} \) is simply \( \mathcal{A} \) itself equipped with the inner product

\[
\langle a, b \rangle_{\mathcal{H}} = \frac{1}{n} \text{Tr}[a^*b] , \tag{5.11}
\]

which is a normalized form of the Hilbert-Schmidt inner product.

In this finite dimensional setting, no completion is needed; \( \mathcal{H} \) is simply \( M_n(\mathbb{C}) \) itself, with the inner product (5.11). For all \( x \in \mathcal{A} \), let \( \zeta_x \) denote \( x \) regarded as an element of \( \mathcal{H} \).

Let \( \pi \) denote the GNS representation of \( \mathcal{A} \) determined by \( \varphi_{\text{tr}} \). Then \( \pi(a)\eta_x = ax \), so that if we define the operator \( L_a \) on \( \mathcal{H} \) by \( L_a\zeta_x = \zeta_{ax} \), then \( \pi(a) = L_a \), the operation of left multiplication by \( a \). Since \( \varphi_{\text{tr}} \) is faithful, \( \ker(\pi) = \{0\} \), and then by Theorem 2.26, \( \pi(\mathcal{A}) \) is a C* algebra, and \( \pi : \mathcal{A} \to \pi(\mathcal{A}) \) is an isometric *-isomorphism.

In this finite dimensional setting \( \pi(\mathcal{A}) \) is not only a C* algebra, but also a von Neumann algebra. Let us use \( \mathcal{M} \) to denote \( \pi(\mathcal{A}) \). Now observe that since the center of \( \mathcal{A} \) is trivial, and since \( \pi \) is a isomorphism of \( \mathcal{A} \) onto \( \mathcal{M} \), the center of \( \mathcal{M} \) is trivial, so that \( \mathcal{M} \) is a factor. By Theorem 5.16, there is a unitary \( u \) from \( \mathcal{B}(\mathcal{H}) \) onto \( \mathbb{C}^n \otimes \mathbb{C}^n \) such that \( u^*u = M_n(\mathbb{C}) \otimes 1 \). (We are also using the fact that \( M_m(\mathbb{C}) \) and \( M_n(\mathbb{C}) \) are not isomorphic for \( m \neq n \).) The commutant \( \mathcal{M}' \) of \( \mathcal{M} \) then consists of all elements of \( \mathcal{B}(\mathcal{H}) \) of the form \( u(1 \otimes b)u^* \), with \( b \in \mathcal{A} \).
We can make this more explicit as follows. For \( \zeta \otimes \xi \in \mathbb{C}^n \otimes \mathbb{C}^n \), define \( v(\zeta \otimes \xi) \) to be the \( n \times n \) matrix \( \sqrt{n}[\zeta_i \xi^*_j] \) which we regard as an element of \( \mathcal{H} \). It is evident that

\[
\|v(\zeta \otimes \xi)\|_{\mathcal{H}} = \|\zeta\|_{\mathbb{C}^n} \|\xi\|_{\mathbb{C}^n} = \|\zeta \otimes \xi\|_{\mathbb{C}^n \otimes \mathbb{C}^n}.
\]

Extending \( v \) by linearity, we obtain an isometry from \( \mathbb{C}^n \otimes \mathbb{C}^n \) into \( \mathcal{H} \), which is necessarily unitary since the dimensions of the domain and range are equal. We again denote the extension by \( v \). Now observe that for all \( a \in \mathcal{A} \), \( L_av(\zeta \otimes \xi) = v(a\zeta \otimes \xi) \), or what is the same thing,

\[
v^*L_av = a \otimes 1_{\mathbb{C}^n}.
\]

In particular, \( v^*Mv = M_n(\mathbb{C}) \otimes 1_{\mathbb{C}^n} \), and thus \( u = v^* \) is one choice of the unitary provided by Theorem 5.16. In the same way we see that

\[
v^*R_av = 1_{\mathbb{C}^n} \otimes a^*.
\]

Since \((M_n(\mathbb{C}) \otimes 1_{\mathbb{C}^n})' = 1_{\mathbb{C}^n} \otimes M_n(\mathbb{C})\), it follows that \( v^*Mv = \mathcal{M}' \).

Define a conjugate linear transformation \( J \) from \( \mathcal{H} \) to itself by

\[
J\zeta_a = \zeta_a^*.
\]

for all \( a \in \mathcal{A} \). Note that \( \|J\zeta_a\|_{\mathcal{H}}^2 = \frac{1}{n} \text{Tr}[aa^*] = \frac{1}{n} \text{Tr}[a^*a] = \|\zeta_a\|_{\mathcal{H}}. \) That is, because of (5.10), \( J \) is an isometry. Moreover, \( J^2 = 1 \) and so \( J = J^* = J^{-1} \).

For each \( b \in \mathcal{A} \) define the operator \( R_b \) on \( \mathcal{H} \) by

\[
R_b\zeta_x = \zeta_{xb}
\]

for all \( c \in \mathcal{A} \). That is, \( R_b \) is the operator of right multiplication by \( b \). Now observe that for all \( a, x \in \mathcal{A} \), \( J(L_ao_x) = J\zeta_{ax} = \zeta_{x^*a^*} = R_{a^*}J\zeta_x. \) In short,

\[
JL_ao = R_{a^*},
\]

and hence \( J\mathcal{M}J = \mathcal{M}' \).

To bring out the symmetry between \( \mathcal{M} \) and \( \mathcal{M}' \), let us introduce \( \mathcal{H}^* \) to be the Hilbert space that is the same set as \( \mathcal{H} \) with the same law of vector addition, but with the scalar multiplication \( (\lambda, \zeta) \mapsto \lambda^*\zeta \) and the inner product \( (\zeta, \xi) \mapsto (\zeta, \xi)^*_\mathcal{H} =: (\zeta, \xi)_{\mathcal{H}^*} \). Note that \( \mathcal{B}(\mathcal{H}^*) = \mathcal{B}(\mathcal{H}) \), and so we may regard each \( R_{a^*} \) as an element of \( \mathcal{B}(\mathcal{H}^*) \). Then it is easy to check that the map \( a \mapsto R_{a^*} =: \pi'(a) \) is a representation of \( \mathcal{A} \) on \( \mathcal{H}^* \). The map \( J \) is then unitary from \( \mathcal{H} \) to \( \mathcal{H}^* \) (though no longer self adjoint), and now (5.12) can be written as

\[
J\pi J^* = \pi',
\]

and we have that \((\pi(\mathcal{A}))'=\pi'(\mathcal{A}) \). The fact that the GNS construction in this simple case yields not one, but two commuting isometric representations of \( \mathcal{A} \) will turn out to be very useful later on.
6 Quantum operations and completely positive maps

6.1 Some important isomorphisms

Let $\mathcal{H}$ and $\mathcal{K}$ be two Hilbert spaces with $\mathcal{K}$ separable. Let $\{\eta_j\}$ be an orthonormal basis for $\mathcal{K}$. Then the general element $\xi$ of the Hilbert space $\mathcal{H} \otimes \mathcal{K}$ has the form

$$\xi = \sum_{j=1}^{\dim \mathcal{K}} \zeta_j \otimes \eta_j \quad \text{and} \quad \|\xi\|_{\mathcal{H} \otimes \mathcal{K}}^2 = \sum_{j=1}^{\dim \mathcal{K}} \|\zeta_j\|_{\mathcal{H}}^2. \quad (6.1)$$

For $n \in \mathbb{N}$, let $\mathcal{H}_n$ denote the direct sum of $n$ copies of $\mathcal{H}$, $\mathcal{H}_n := \mathcal{H} \oplus \cdots \oplus \mathcal{H}$, $n$ times.

Let $\mathcal{H}_\aleph$ denote the direct sum of countably infinitely many copies of $\mathcal{H}$. We write $\{\zeta_j\}_{1 \leq j \leq n}$ to denote the general element of $\mathcal{H}_n$, and we write $\{\zeta_j\}_{j \in \mathbb{N}}$ to denote the general element of $\mathcal{H}_\aleph$.

When $\dim(\mathcal{K}) = n < \infty$, we define an isomorphism from $\mathcal{H}_n$ onto $\mathcal{H} \otimes \mathcal{K}$ by choosing an orthonormal basis $\{\eta_1, \ldots, \eta_n\}$ of $\mathcal{K}$ and then define the map

$$\{\zeta_j\}_{j \in \mathbb{N}} \mapsto \sum_{j=1}^{n} \zeta_j \otimes \eta_j. \quad (6.2)$$

This gives a unitary map from $\mathcal{H}_n$ onto $\mathcal{H} \otimes \mathcal{K}$. When $\mathcal{K}$ is infinite dimensional and $\{\eta_j\}$ is an orthonormal basis of $\mathcal{K}$, the map given in (6.2) is unitary from $\mathcal{H}_\aleph$ onto $\mathcal{H} \otimes \mathcal{K}$.

6.1 Definition. Let $\mathcal{H}$ be any Hilbert space and let $\mathcal{K}$ be a separable Hilbert space. Let $\{\eta_j\}$ be an orthonormal basis of $\mathcal{K}$. Define $V_j : \mathcal{K} \to \mathcal{H} \otimes \mathcal{K}$ by

$$V_j \zeta = \zeta \otimes \eta_j. \quad (6.3)$$

Note that $V_j$ is an isometry from $\mathcal{K}$ into $\mathcal{H} \otimes \mathcal{K}$, and that

$$\sum_{j=1}^{\dim(\mathcal{K})} V_j^* V_j = 1_{\mathcal{H}}. \quad (6.4)$$

Next, for every $\zeta \in \mathcal{H}$ and $\eta \in \mathcal{K}$ we define the rank-one operator $|\zeta\rangle \langle \eta|$ from $\mathcal{K}$ to $\mathcal{H}$ by

$$|\zeta\rangle \langle \eta| = (\langle \eta, \zeta \rangle) \zeta.$$

Let $\{\eta_j\}$ be an orthonormal basis for $\mathcal{K}$. Define the sesquilinear map

$$\sum_{j=1}^{\dim \mathcal{K}} \zeta_j \otimes \eta_j \mapsto \sum_{j=1}^{\dim \mathcal{K}} |\zeta_j\rangle \langle \eta_j|.$$

On the right we have the general element of $\mathcal{C}_2(\mathcal{K}, \mathcal{H})$, the Hilbert space of Hilbert-Schmidt linear maps from $\mathcal{K}$ to $\mathcal{H}$; that is, the space of linear maps $x : \mathcal{K} \to \mathcal{H}$ such that $\text{Tr}[x^* x] < \infty$. Moreover, this map is easily seen to be an isometry; i.e.,

$$\left\| \sum_{j=1}^{\dim \mathcal{K}} \zeta_j \otimes \eta_j \right\|_{\mathcal{H} \otimes \mathcal{K}}^2 = \sum_{j=1}^{\infty} \|\zeta_j\|_{\mathcal{H}}^2 = \left\| \sum_{j=1}^{\dim \mathcal{K}} |\zeta_j\rangle \langle \eta_j| \right\|_{\mathcal{C}_2(\mathcal{K}, \mathcal{H})}^2.$$
Next, still under the assumption that \( \mathcal{K} \) is separable, consider the algebraic tensor product \( \mathcal{B}(\mathcal{H}) \otimes \mathcal{B}(\mathcal{H}) \). The general element of \( \mathcal{B}(\mathcal{H}) \otimes \mathcal{B}(\mathcal{H}) \) is a linear combination of elements of the form \( x \otimes y \). We may regard these as operators on \( \mathcal{H} \otimes \mathcal{H} \) through
\[
(x \otimes y) \zeta \otimes \eta = (x \zeta) \otimes (y \eta) .
\]
This gives us a natural embedding of \( \mathcal{B}(\mathcal{H}) \otimes \mathcal{B}(\mathcal{H}) \) into \( \mathcal{B}(\mathcal{H} \otimes \mathcal{H}) \).

6.2 Lemma. For Hilbert spaces \( \mathcal{H} \) and \( \mathcal{K} \), and any \( \zeta_1, \zeta_2 \in \mathcal{H} \), and any \( \eta_1, \eta_2 \in \mathcal{K} \),
\[
|\zeta_1 \otimes \eta_1 \rangle \langle \zeta_2 \otimes \eta_2| = |\zeta_1 \rangle \langle \eta_1| \otimes |\zeta_2 \rangle \langle \eta_2| ,
\]
where the right hand side is regarded as an element of \( \mathcal{B}(\mathcal{H} \otimes \mathcal{K}) \) through the natural embedding of \( \mathcal{B}(\mathcal{H}) \otimes \mathcal{B}(\mathcal{K}) \) into \( \mathcal{B}(\mathcal{H} \otimes \mathcal{K}) \).

Proof. It suffices to check that for all \( \zeta_3 \in \mathcal{H} \) and \( \eta_3 \in \mathcal{K} \), both sides have the same action on \( \zeta_3 \otimes \eta_3 \). By the definitions,
\[
|\zeta_3 \otimes \eta_3 \rangle \langle \zeta_3 \otimes \eta_3| = |\zeta_3 \rangle \langle \zeta_3| \otimes |\eta_3 \rangle \langle \eta_3| = |\zeta_3 \rangle \langle \zeta_3| \otimes |\eta_3 \rangle \langle \eta_3| .
\]

Now we specialize to a case that will be important in what follows. Let \( \mathcal{K} \) be finite dimensional, and identify it with \( \mathbb{C}^n \) for \( n = \text{dim}(\mathcal{K}) \). We may then identify \( \mathcal{B}(\mathcal{K}) \) with \( M_n(\mathbb{C}) \). Let \( \{\eta_i, \ldots, \eta_n\} \) be any orthonormal basis of \( \mathbb{C}^n \). Then \( \{|\eta_i\rangle \langle \eta_j| : 1 \leq i, j \leq n\} \) is a basis for for \( M_n(\mathbb{C}) \). It is easy to check that
\[
|\eta_i\rangle \langle \eta_j| |\eta_k\rangle \langle \eta_l| = \delta_{j,k} |\eta_i\rangle \langle \eta_l| .
\]

(6.7)

For any \( a_1 \otimes m_1, a_2 \otimes m_2 \in \mathcal{B}(\mathcal{H}) \times \mathcal{B}(\mathcal{K}) \), we define the product \( (a_1 \otimes m_1)(a_2 \otimes m_2) \) by
\[
(a_1 \otimes m_1)(a_2 \otimes m_2) = a_1 a_2 \otimes m_1 m_2 ,
\]
and extend this by linearity. Likewise, for any \( a \in \mathcal{B}(\mathcal{H}) \) and any \( m \in M_n(\mathbb{C}) \), the involution \( (a \otimes m)^* = a^* \otimes m^* \), and also extend this by linearity.

Since \( \{ |\eta_i\rangle \langle \eta_j| : 1 \leq i, j \leq n\} \) is a basis of \( M_n(\mathbb{C}) \), the general element \( \tilde{a} \) of \( \mathcal{B}(\mathcal{H}) \otimes M_n(\mathbb{C}) \) can be written as \( \tilde{a} = \sum_{i,j=1}^n a_{i,j} \otimes |\eta_i\rangle \langle \eta_j| \). By (6.7),
\[
\left( \sum_{i,j=1}^n a_{i,j} \otimes |\eta_i\rangle \langle \eta_j| \right) \left( \sum_{k,\ell=1}^n b_{k,\ell} \otimes |\eta_k\rangle \langle \eta_\ell| \right) = \sum_{i,j=1}^n \left( \sum_{k,\ell=1}^n a_{i,j} b_{k,\ell} \right) \otimes |\eta_i\rangle \langle \eta_\ell| .
\]

(6.8)

We define \( M_n(\mathcal{B}(\mathcal{H})) \) to be the set of \( n \times n \) matrices with entries in \( \mathcal{B}(\mathcal{H}) \). Let \( [a_{i,j}] \) denote the element of \( M_n(\mathbb{C}) \) with \( i, j \) entry is \( a_{i,j} \). Define a linear transformation from \( \mathcal{B}(\mathcal{H}) \otimes M_n(\mathbb{C}) \) onto \( M_n(\mathcal{B}(\mathcal{H})) \) by
\[
\tilde{a} = \sum_{i,j=1}^n a_{i,j} \otimes |\eta_i\rangle \langle \eta_j| \mapsto [a_{i,j}] .
\]

(6.9)
The transformation in (6.9) is evidently injective, and hence is a vector space isomorphism. In fact, the inverse map is simply given by

\[ [a_{i,j}] \mapsto \sum_{i,j=1}^{n} a_{i,j} |\eta_i\rangle\langle\eta_j| . \]  

(6.10)

By (6.8), vector space isomorphism is also a algebra isomorphism where \( M_n(\mathcal{B}(\mathcal{H})) \) is given the natural product, and one easily checks that \([a_{i,j}]^* = [a_{j,i}^*] \), so that it is a \(*\)-isomorphism.

Another identification will be useful in what follows: There is a natural \(*\)-isomorphism of \( \mathcal{B}(\mathcal{H} \otimes \mathbb{C}^n) \) with \( M_n(\mathcal{B}(\mathcal{H})) \). Let \( \{\eta_1, \ldots, \eta_n\} \) be an orthonormal basis of \( \mathbb{C}^n \). For \( j = 1, \ldots, n \), let \( V_j \) be the isometry from \( \mathcal{H} \) into \( \mathcal{H} \otimes \mathbb{C}^n \) given by (6.3). Define a linear transformation from \( \mathcal{B}(\mathcal{H} \otimes \mathbb{C}^n) \) to \( M_n(\mathcal{B}(H)) \) by

\[ \hat{a} \mapsto [V_i^*\hat{a}V_j] . \]  

(6.11)

for all \( \hat{a} \in \mathcal{B}(\mathcal{H} \otimes \mathbb{C}^n) \).

This too is also a vector space isomorphism. To see this, let \([a_{i,j}] \in M_n(\mathcal{B}(H))\), and \( \{\eta_1, \ldots, \eta_n\} \) an orthonormal basis of \( \mathbb{C}^n \), consider \( \sum_{i,j=1}^{n} [a_{i,j}] \otimes |\eta_i\rangle\langle\eta_j| \) as an element of \( \mathcal{B}(\mathcal{H} \otimes \mathbb{C}^n) \) through the natural embedding of \( \mathcal{B}(\mathcal{H}) \otimes M_n(\mathbb{C}) \) into \( \mathcal{B}(\mathcal{H} \otimes \mathbb{C}^n) \) that is given by (6.5). Then for all \( \hat{a} \in \mathcal{B}(\mathcal{H} \otimes \mathbb{C}^n) \), and all \( \zeta \in \mathcal{H} \) and all \( \eta \in \mathbb{C}^n \),

\[
\left( \sum_{i,j=1}^{n} [V_i^*\hat{a}V_j] \otimes |\eta_i\rangle\langle\eta_j| \right) \zeta \otimes \eta = \sum_{i,j=1}^{n} \langle\eta_j, \eta|V_i^*\hat{a}V_j\zeta \rangle \otimes \eta_i \\
= \sum_{i,j=1}^{n} \langle\eta_j, \eta|[V_i^*\hat{a}\langle\zeta \otimes \eta_j]\rangle \otimes \eta_i \\
= \sum_{i,j=1}^{n} [V_i^*\hat{a}\langle\zeta \otimes \eta\rangle] \otimes \eta_i = \hat{a}\langle\zeta \otimes \eta\rangle 
\]

That is, the transformation (6.10), followed by the natural embedding of \( \mathcal{B}(\mathcal{H}) \otimes M_n(\mathbb{C}) \) into \( \mathcal{B}(\mathcal{H} \otimes \mathbb{C}^n) \), is inverse to the map defined in (6.11). Finally, using (6.4), one readily checks that the map in (6.11) is a \(*\)-algebra isomorphism as well as a vector space isomorphism. We have proved:

6.3 THEOREM. For any Hilbert space \( \mathcal{H} \) and any \( n \in \mathbb{N} \), \( \mathcal{B}(H) \otimes M_n(\mathbb{C}), M_n(\mathcal{B}(\mathcal{H})) \) and \( \mathcal{B}(\mathcal{H} \otimes \mathbb{C}^n) \), equipped with their natural \(*\)-algebra structures, are all \(*\)-algebra isomorphic. Moreover, for any orthonormal basis \( \{\eta_1, \ldots, \eta_n\} \) of \( \mathbb{C}^n \), and:

(1) The map in (6.9) is a \(*\)-isomorphism of \( \mathcal{B}(\mathcal{H}) \otimes M_n(\mathbb{C}) \) onto \( M_n(\mathcal{B}(\mathcal{H})) \), and the map in (6.10) is its inverse.

(2) The map in (6.11) is a \(*\)-isomorphism of \( \mathcal{B}(\mathcal{H} \otimes \mathbb{C}^n) \) onto \( M_n(\mathcal{B}(H)) \), and the map in (6.10), followed by the natural embedding of \( \mathcal{B}(\mathcal{H}) \otimes M_n(\mathbb{C}) \) into \( \mathcal{B}(\mathcal{H} \otimes \mathbb{C}^n) \), is its inverse.

Since \( \mathcal{H} \otimes \mathbb{C}^n \) is a Hilbert space, \( \mathcal{B}(\mathcal{H} \otimes \mathbb{C}^n) \) is a \( C^* \)-algebra. We may use the \(*\)-algebra isometries provided by Theorem 6.3 to transfer the operator norm in \( \mathcal{B}(\mathcal{H} \otimes \mathbb{C}^n) \) to \( \mathcal{B}(\mathcal{H}) \otimes M_n(\mathbb{C}) \) and to \( M_n(\mathcal{B}(\mathcal{H}) \), thus making them \( C^* \)-algebras, isomorphic to \( \mathcal{B}(\mathcal{H} \otimes \mathbb{C}^n) \).

The isomorphism is basis dependent, but the norm is not since the norm on a \( C^* \) algebra is unique.
6.2 The $C^*$ algebras $\mathcal{A} \otimes M_n(\mathbb{C})$.

There is a natural family of $C^*$ algebras associated to every $C^*$ algebra $\mathcal{A}$, namely the $C^*$ algebras $\mathcal{A} \otimes M_n(\mathbb{C})$ for each $n \in \mathbb{N}$, which we now define. As a vector space, $\mathcal{A} \otimes M_n(\mathbb{C})$ is the algebraic tensor product of the vector spaces $\mathcal{A}$ and $M_n(\mathbb{C})$. Let $\{\eta_1, \ldots, \eta_n\}$ be any orthonormal basis for $\mathbb{C}^n$. Since the $\{|\eta_i\rangle\langle\eta_j| : 1 \leq i, j \leq n\}$ is a basis for $M_n(\mathbb{C})$, the general element $\tilde{a}$ of $\mathcal{A} \otimes M_n(\mathbb{C})$ has the form

$$\tilde{a} = \sum_{i,j=1}^{n} a_{i,j} \otimes |\eta_i\rangle\langle\eta_j|,$$

where for each $i, j$, $a_{i,j} \in \mathcal{A}$.

We give it the natural algebraic structure by defining $(a_1 \otimes m_1)(a_2 \otimes m_2) = (a_1a_2 \otimes m_1m_2)$, and then extending this by linearity. If $\tilde{a} = \sum_{i,j=1}^{n} a_{i,j} \otimes |\eta_i\rangle\langle\eta_j|$ and $\tilde{b} = \sum_{i,j=1}^{n} b_{i,j} \otimes |\eta_i\rangle\langle\eta_j|$ as in (6.12), then one checks as above that

$$\tilde{a}\tilde{b} = \left( \sum_{j=1}^{n} a_{i,j}b_{j,\ell} \right) \otimes |\eta_i\rangle\langle\eta_\ell|.$$

Defining $(a \otimes m)^* = (a^* \otimes b^*)$, makes $\mathcal{A} \otimes M_n(\mathbb{C})$ a $*$-algebra.

Let $M_n(\mathcal{A})$ denote the set of $n \times n$ matrices with entries in $\mathcal{A}$. We write $[a_{i,j}]$ to denote the element of $M_n(\mathcal{A})$ whose $i, j$ entry is $a_{i,j}$. $M_n(\mathcal{A})$ is a $*$-algebra with the obvious operations. Then by (6.13), the map

$$\sum_{i,j=1}^{n} a_{i,j} \otimes e^{(i,j)} = \tilde{a} \mapsto [a_{i,j}]$$

is a $*$ isomorphism of $\mathcal{A} \otimes M_n(\mathbb{C})$ onto $M_n(\mathcal{A})$, and we use this isomorphism to identify $\mathcal{A} \otimes M_n(\mathbb{C})$ with $M_n(\mathcal{A})$.

We now claim that there is a norm on $\mathcal{A} \otimes M_n(\mathbb{C})$ that makes it a $C^*$ algebra. To see this easily, suppose first that $\mathcal{A}$ is a $C^*$ sub algebra of $\mathcal{B}$ for some Hilbert space $\mathcal{H}$. Then $M_n(\mathcal{A})$ is evidently a closed subspace of $M_n(\mathcal{B}(\mathcal{H}))$ on which we have a natural $C^*$-algebra norm through the identification of $M_n(\mathcal{B}(\mathcal{H}))$ with $\mathcal{B}(\mathcal{H}_n)$ as explained in the previous section. Thus, $M_n(\mathcal{A})$ is a $C^*$-subalgebra of $M_n(\mathcal{B}(\mathcal{H}))$.

In general, we can always use the Gelfand-Naimark Theorem to identify $\mathcal{A}$ with a $C^*$-algebra of operators, and thus the above discussion applies to the general case.

6.4 DEFINITION. Let $\mathcal{A}$ and $\mathcal{B}$ be $C^*$-algebras, and let $\Phi$ be a bounded linear map from $\mathcal{A}$ to $\mathcal{B}$. Then for all $n \in \mathbb{N}$, define $\Phi_n := \Phi \otimes 1_{\mathbb{C}^n}$, so that

$$\Phi_n(a \otimes m) = \Phi(a) \otimes m$$

for all $a \in \mathcal{A}$ and all $m \in M_n(\mathbb{C})$. In particular, for any orthonormal basis $\{\eta_1, \ldots, \eta_n\}$ of $\mathbb{C}^n$,

$$\Phi_n \left( \sum_{i,j=1}^{n} a_{i,j} \otimes |\eta_i\rangle\langle\eta_j| \right) = \sum_{i,j=1}^{n} \Phi(a_{i,j}) \otimes |\eta_i\rangle\langle\eta_j| .$$

so that for $\tilde{a}$ given by (6.12), $\Phi_n(\tilde{a})$, considered as an element of $M_n(\mathcal{A})$, is given by

$$[\Phi_n(\tilde{a})]_{i,j} = [\Phi(a_{i,j})].$$

That is, the action of $\Phi_n$ on $[a_{i,j}] \in M_n(\mathcal{A})$ is given by the action of $\Phi$ on each entry of $[a_{i,j}]$. 
6.3 Positive and completely positive maps

6.5 DEFINITION. Let \( \mathcal{A} \) and \( \mathcal{B} \) be \( C^* \) algebras. A linear map \( \Phi : \mathcal{A} \to \mathcal{B} \) is positive in case \( \Phi(a^*a) \geq 0 \) for all \( a \in \mathcal{A} \). If \( \mathcal{A} \) and \( \mathcal{B} \) have identities \( 1_\mathcal{A} \) and \( 1_\mathcal{B} \) respectively, then \( \Phi \) is unital in case \( \Phi(1_\mathcal{A}) = 1_\mathcal{B} \).

Let \( 6.8 \) LEMMA. Such a form, at least when the Hilbert spaces are finite dimensional.

In any \( C^* \) algebra \( \mathcal{A} \), we can write the general element \( a \) as \( a = (a_1 - a_2) + i(b_1 - b_2) \) where \( a_1, a_2, b_1, b_2 \in \mathcal{A}_+ \). Then

\[
\Phi(a^*) = \Phi((a_1 - a_2) - i(b_1 - b_2)) = (\Phi(a_1) - \Phi(a_2)) - i(\Phi(b_1) - \Phi(b_2)) = \Phi(a)^*.
\]

That is, \( \Phi \) automatically respects the involutions on \( \mathcal{A} \) and \( \mathcal{B} \).

If \( \Phi \) is any *-homomorphism of \( \mathcal{A} \) into \( \mathcal{B} \), then for all \( a \in \mathcal{A} \), \( \Phi(a^*a) = \Phi(a)^*\Phi(a) \geq 0 \), and since every element of \( \mathcal{A}_+ \) if is the form \( a^*a \), it follows that every *-homomorphism is positive.

Here is another important example: Let \( \mathcal{A} = \mathcal{B} = M_n(\mathbb{C}) \), and for \( a \in \mathcal{A} \), let \( \Phi(a) = a^T \), the transpose of \( a \). Then evidently \( \Phi : a \to a^T \) is positive. Since for all \( a_1, a_2 \in \mathcal{A} \), \( (a_1a_2)^T = a_2^Ta_1^T \), \( \Phi \) is not a *-homomorphism.

6.6 DEFINITION \((n\text{-positive and completely positive})\). Let \( \mathcal{A} \) and \( \mathcal{B} \) be \( C^* \) algebras. A linear map \( \Phi : \mathcal{A} \to \mathcal{B} \) is \( n \)-positive in case \( \Phi_n : \mathcal{A} \otimes \mathbb{C}^n \to \mathcal{B} \otimes \mathbb{C}^n \) is positive. A linear map \( \Phi : \mathcal{A} \to \mathcal{B} \) is completely positive in case \( \Phi_n \) is positive for all \( n \in \mathbb{N} \).

6.7 THEOREM. Let \( \mathcal{A} \) be a \( C^* \)-subalgebra of \( \mathcal{B}(\mathcal{H}) \) for some Hilbert space \( \mathcal{H} \). Let \( \mathcal{K} \) be a second Hilbert space, and let \( \Phi : \mathcal{A} \to \mathcal{B}(\mathcal{K}) \) be given by

\[
\Phi(a) = \sum_{j=1}^m W_j^*aW_j \tag{6.14}
\]

where for each \( j = 1, \ldots, m \), \( W_j \) is a bounded linear transformation from \( \mathcal{K} \) to \( \mathcal{H} \). Then \( \Phi \) is completely positive.

Proof. Since a sum of completely positive maps is evidently completely positive, it suffices to consider the case \( \Phi(a) = W^*aW \) for a bounded linear transformation \( W \) from \( \mathcal{K} \) to \( \mathcal{H} \). But for any \( n \in \mathbb{N} \), if \( \bar{a} = [a_{i,j}] \) is any element of \( M_n(\mathcal{A}) \),

\[
\Phi_n(\bar{a}) = [W^*a_{i,j}W] = \left( \sum_{i=1}^n W^* \otimes |\eta_i\rangle \langle \eta_i| \right) \left( \sum_{k,l} a_{i,j} \otimes |\eta_k\rangle \langle \eta_l| \right) \left( \sum_{j=1}^n W \otimes |\eta_j\rangle \langle \eta_j| \right)^*.
\]

which is clearly positive.

We will see later that this is essentially the only example: All completely positive maps have such a form, at least when the Hilbert spaces are finite dimensional.

6.8 LEMMA. Let \( \mathcal{A} \) be a \( C^* \) algebra with identity 1. Then for all \( a, b \in \mathcal{A} \), \( a^*a \leq b \) if and only if

\[
\begin{bmatrix}
1 & a^* \\
1 & b
\end{bmatrix}
\]

is positive in \( M_2(\mathcal{A}) \).
Proof. By the Gelfand-Naimark Theorem, we may suppose that \( \mathcal{A} \) is a \( C^* \) subalgebra of \( \mathcal{B}(\mathcal{H}) \) for some Hilbert space \( \mathcal{H} \). If \( a = 0 \), or if \( b \) is not positive, the claim is trivially true, so suppose this is not the case. Choose any \( \eta \in \mathcal{H} \) with \( \|a\eta\|_\mathcal{H} = 1 \), and let \( \zeta = -a\eta \). Then
\[
\left\langle \left( \begin{array}{c} \xi \\ \eta \end{array} \right), \left[ \begin{array}{cc} 1 & a \\ a^* & b \end{array} \right] \left( \begin{array}{c} \xi \\ \eta \end{array} \right) \right\rangle_{\mathcal{H} \oplus \mathcal{H}} = 1 - 2\|a\eta\|_\mathcal{H} + \langle \eta, b\eta \rangle_\mathcal{H} = \langle \eta, \eta \rangle_\mathcal{H} - \|a\eta\|_\mathcal{H}^2.
\]
Thus positivity of \( \left[ \begin{array}{cc} 1 & a \\ a^* & b \end{array} \right] \) is positive in \( M_2(\mathcal{A}) \) implies that \( \langle \eta, [b - a^*a]\eta \rangle_\mathcal{H} \geq 0 \) for all \( \eta \) such that \( a\eta \neq 0 \), and hence for all \( \eta \in \mathcal{H} \), so that \( b \geq a^*a \).

Conversely suppose \( b \geq a^*a \), and define \( c = b - a^*a \). Then
\[
\left[ \begin{array}{cc} 1 & a \\ a^* & b \end{array} \right] = \left[ \begin{array}{cc} 1 & a \\ a^* & a^*a \end{array} \right] + \left[ \begin{array}{cc} 0 & 0 \\ 0 & c \end{array} \right] = xx^* + \left[ \begin{array}{cc} 0 & 0 \\ 0 & c \end{array} \right]
\]
where \( x \) is the linear map from \( \mathcal{H} \) to \( \mathcal{H} \oplus \mathcal{H} \) given by \( x\eta = \left( \begin{array}{c} \eta \\ a\eta \end{array} \right) \) displays the left hand side as a sum of positive operators. \( \square \)

6.9 THEOREM. Let \( \mathcal{A} \) and \( \mathcal{B} \) be \( C^* \) algebras with identities \( 1_\mathcal{A} \) and \( 1_\mathcal{B} \) respectively. Let \( \Phi \) be a unital 2-positive map from \( \mathcal{A} \) to \( \mathcal{B}(\mathcal{H}) \) for a Hilbert space \( \mathcal{H} \). Then for all \( a \in \mathcal{A} \),
\[
\Phi(a)^*\Phi(a) \leq \Phi(a^*a).
\]
(6.15)

Proof. Since \( \left[ \begin{array}{cc} 1 & a \\ a^* & b \end{array} \right] \geq 0 \) in \( M_2(\mathcal{A}) \),
\[
\Phi_2\left( \left[ \begin{array}{cc} 1 & a \\ a^* & a^*a \end{array} \right] \right) = \left[ \begin{array}{cc} 1 & \Phi(a) \\ \Phi(a)^* & \Phi(a^*a) \end{array} \right] \geq 0,
\]
and by Lemma 6.8, this implies (6.15). \( \square \)

Associated to every completely positive map \( \Phi \) from a \( C^* \) algebra to \( \mathcal{B}(\mathcal{H}) \) for a Hilbert space \( \mathcal{H} \) is a non-negative sesquilinear form on \( \mathcal{A} \otimes \mathcal{H} \) that we now describe.

6.10 DEFINITION (The Stinespring inner product). Let \( \mathcal{A} \) be a \( C^* \)-algebra and let \( \mathcal{H} \) be a Hilbert space. Let \( \Phi : \mathcal{A} \to \mathcal{B}(\mathcal{H}) \) be completely positive. Consider two arbitrary elements of \( \mathcal{A} \otimes \mathcal{H} \) which we may take to be of the form
\[
\sum_{j=1}^n a_j \otimes \eta_j \quad \text{and} \quad \sum_{j=1}^n b_j \otimes \xi_j
\]
for some common value of \( n \in \mathbb{N} \) by allowing some terms to be zero.
Then the sesquilinear form \( \langle \cdot, \cdot \rangle_\Phi \) on \( \mathcal{A} \otimes \mathcal{H} \) given by

\[
\left\langle \sum_{j=1}^n a_j \otimes \eta_j, \sum_{j=1}^n b_j \otimes \xi_j \right\rangle = \sum_{i,j=1}^n \langle \eta_i, \Phi(a_i^* b_j) \xi_j \rangle_\mathcal{H} \quad (6.16)
\]

is the Stinespring inner product on \( \mathcal{A} \otimes \mathcal{H} \).

The Stinespring inner product is non-negative (as the name suggests): Let \( \sum_{j=1}^n a_j \otimes \eta_j \in \mathcal{A} \otimes \mathcal{H} \), and define \( \tilde{a} = [a_i^* a_j] \in M_n(\mathcal{A}) \). Also define \( \tilde{\eta} = \begin{pmatrix} \eta_1 \\ \vdots \\ \eta_n \end{pmatrix} \in \mathcal{H}_n \), the direct sum of \( n \) copies of \( \mathcal{H} \). Then from (6.16),

\[
\left\langle \sum_{j=1}^n a_j \otimes \eta_j, \sum_{j=1}^n a_j \otimes \eta_j \right\rangle = \langle \tilde{\eta}, \Phi_n(\tilde{a}) \tilde{\eta} \rangle_{\mathcal{H}_n}.
\]

The right hand side is non-negative since \( \Phi_n(\tilde{a}) \geq 0 \).

### 6.4 The partial trace

Recall that every linear functional \( \phi \) on \( M_n(\mathbb{C}) \) is of the form \( \phi(x) = \text{Tr}[xz] \) for some uniquely determined \( z \in M_n(\mathbb{C}) \). This is because if we equip \( M_n(\mathbb{C}) \) with the Hilbert-Schmidt inner product \( \langle x, y \rangle = \text{Tr}[x^* y] \), it is a (finite dimensional) Hilbert space, and then we may apply the Riesz Lemma.

Now suppose that \( a \in \mathcal{B}(\mathbb{C}^n \otimes \mathbb{C}^p) \). Define a linear functional \( \phi_a \) on \( M_n(\mathbb{C}) \) by

\[
x \mapsto \text{Tr}[(x \otimes 1)a] =: \phi_a(x)
\]

Since this is a linear functional on \( M_n(\mathbb{C}) \), it has the form \( \phi_a(x) = \text{Tr}[bx] \) for some uniquely determined \( b \in M_n(\mathbb{C}) \). This brings us to the following definition:

### 6.11 Definition. Let \( a \in \mathcal{B}(\mathbb{C}^m \otimes \mathbb{C}^n) \). The partial traces of \( a \) on \( \mathbb{C}^m \) and \( \mathbb{C}^n \) respectively are the operators \( \text{Tr}_2[a] \in M_m(\mathbb{C}) \) and \( \text{Tr}_1[a] \in M_n(\mathbb{C}) \) such that for all \( x \in M_m(\mathbb{C}) \) and all \( y \in M_n(\mathbb{C}) \),

\[
\text{Tr}[x \text{Tr}_2[a]] = \text{Tr}[(x \otimes 1)a] \quad \text{and} \quad \text{Tr}[y \text{Tr}_1[a]] = \text{Tr}[(1 \otimes y)a] \quad . \quad (6.17)
\]

By using an orthonormal basis of \( \mathbb{C}^m \otimes \mathbb{C}^n \) of the form \( \{ \eta_i \otimes \zeta_j : 1 \leq i \leq m, 1 \leq j \leq n \} \), it is evident that for all \( y \in M_n(\mathbb{C}) \) and all \( z \in M_m(\mathbb{C}) \),

\[
\text{Tr}[y \otimes z] = \text{Tr}[y] \text{Tr}[z] \quad . \quad (6.18)
\]

Therefore, when \( a = y \otimes z \) with \( y \in M_m(\mathbb{C}) \) and \( z \in M_n(\mathbb{C}) \),

\[
\text{Tr}[(x \otimes 1)a] = \text{Tr}[x y \otimes z] = \text{Tr}[x y] \text{Tr}[z]
\]

where on the right the traces are taken in \( M_m(\mathbb{C}) \) and \( M_n(\mathbb{C}) \) respectively. Thus, \( \text{Tr}_2[y \otimes z] = \text{Tr}[z]y \). Likewise, \( \text{Tr}_1[y \otimes z] = \text{Tr}[y]z \).
By Theorem 6.3, the general element $a$ of $\mathcal{B}(C^m \otimes C^n)$ has the form

$$a = \sum_{i,j=1}^{n} V_i^* a V_j \otimes |\eta_i\rangle \langle \eta_j|$$

where $\{\eta_1, \ldots, \eta_n\}$ is an orthonormal basis of $C^n$, the the isometries $V_j$, $j = 1, \ldots, n$, are defined in (6.3). Then by linearity of the trace and the obvious identity $\text{Tr}[|\eta_i\rangle \langle \eta_j|] = \delta_{i,j}$,

$$\text{Tr}_2[a] = \sum_{j=1}^{n} V_j^* a V_j.$$

(6.19)

By symmetry, the same reasoning applies to $a \mapsto \text{Tr}_1[a]$. Let $\{\zeta_1, \ldots, \zeta_m\}$ be an orthonormal basis of $C^m$. For $j = 1, \ldots, m$, define $W_j : C^n \to C^m \otimes C^n$ by $W_j \eta = \zeta_j \otimes \eta$, then we have $\text{Tr}_1[a] = \sum_{j=1}^{n} W_j^* a W_j$. By Theorem 6.7, the maps $a \mapsto \text{Tr}_1[a]$ and $a \mapsto \text{Tr}_2[a]$ are both completely positive.

6.12 EXAMPLE (The partial transpose). As our terminology suggests, not every positive map is completely positive. Here is an important example. Let $\mathcal{A} = M_2(C)$, and let $\Psi$ be the transpose map $\Psi(a) = a^T$. Then $\Psi$ is positive, but $\Psi_2$ is not. Indeed, identifying $\mathcal{A} \otimes M_2(C) = M_2(C \otimes M_2(C)$ with $M_2(M_2())$ ad above, we have that for any $a, b, c, d \in M_2(C)$,

$$\Psi_2\left(\begin{bmatrix} a & b \\ c & d \end{bmatrix}\right) = \begin{bmatrix} \Psi(a) & \Psi(b) \\ \Psi(c) & \Psi(d) \end{bmatrix} = \begin{bmatrix} a^T & b^T \\ c^T & d^T \end{bmatrix}$$

need not be positive. To see this, consider the choice $a = e^{(1,1)}$, $b = e^{(1,2)}$, $c = e^{(2,1)}$ and $d = e^{(2,2)}$. Then

$$\begin{bmatrix} e^{(1,1)} & e^{(1,2)} \\ e^{(2,1)} & e^{(2,2)} \end{bmatrix} = \begin{bmatrix} 1 & 0 & 0 & 1 \\ 0 & 0 & 0 & 0 \\ 0 & 0 & 0 & 0 \\ 1 & 0 & 0 & 1 \end{bmatrix}$$

is positive, but

$$\Psi_2\left(\begin{bmatrix} e^{(1,1)} & e^{(1,2)} \\ e^{(2,1)} & e^{(2,2)} \end{bmatrix}\right) = \begin{bmatrix} e^{(1,1)} & e^{(2,1)} \\ e^{(1,2)} & e^{(2,2)} \end{bmatrix} = \begin{bmatrix} 1 & 0 & 0 & 0 \\ 0 & 0 & 1 & 0 \\ 0 & 1 & 0 & 0 \\ 0 & 0 & 0 & 1 \end{bmatrix}$$

is not positive.

The fact that

$$\begin{bmatrix} e^{(1,1)} & e^{(1,2)} \\ e^{(2,1)} & e^{(2,2)} \end{bmatrix} = \sum_{i,j=1}^{2} e^{(i.j)} \otimes e^{(1.j)}$$

revealed the failure of $\Psi$ to be 2-positive is no accident, as we explain in the the next section.

6.5 Choi’s Theorem

Choi’s Theorem gives a complete description of completely positive maps in finite dimensions. Let $\mathcal{L}(M_n(C), M_p(C))$ denote the space of linear transformations from $M_n(C)$ to $M_p(C)$. Let $\Phi \in \mathcal{L}(M_n(C), M_p(C))$ for some $n, p \in \mathbb{N}$, and then for any $m \in \mathbb{N}$, let $\Phi_m : M_n(C) \otimes M_m(C) \to M_p(C) \otimes M_m(C)$ be defined as above.
We may identify $M_n(\mathbb{C}) \otimes M_m(\mathbb{C})$ with $\mathcal{B}(\mathbb{C}^n \otimes \mathbb{C}^m)$ as in Theorem 6.3, and thus may regard $\Phi_m$ as a linear map from $\mathcal{B}(\mathbb{C}^n \otimes \mathbb{C}^m)$ to $M_p(\mathbb{C}) \otimes M_m(\mathbb{C})$. To show that $\Phi_m$ is positive, one has to show that for all positive $a \in \mathcal{B}(\mathbb{C}^n \otimes \mathbb{C}^m)$, $\Phi_m(a)$ is positive in $M_p(\mathbb{C}) \otimes M_m(\mathbb{C})$. By the Spectral Theorem, every positive element $a$ of $\mathcal{B}(\mathbb{C}^n \otimes \mathbb{C}^m)$ may be decomposed as a sum of rank-one projections, and thus $\Phi_m$ is positive if and only if it positive on every rank-one projection in $\mathcal{B}(\mathbb{C}^n \otimes \mathbb{C}^m)$. Let $|\xi\rangle\langle\xi|$ be such a rank-one projection.

The general element of $\xi$ of $\mathbb{C}^n \otimes \mathbb{C}^m$ has the form

$$\xi = \sum_{j=1}^{n} \eta_j \otimes \zeta_j$$

for some set $\{\zeta_1, \ldots, \zeta_n\}$ of $n$ vectors in $\mathbb{C}^m$.

Now define a linear transformation $a$ from $\mathbb{C}^m$ to $\mathbb{C}^m$ by $a\eta_j = \sqrt{n}\zeta_j$ for $j = 1, \ldots, n$, and define the vector $\omega \in \mathbb{C}^n \otimes \mathbb{C}^n$ by

$$\omega = \frac{1}{\sqrt{n}} \sum_{j=1}^{n} \eta_j \otimes \eta_j .$$

Note that $\omega$ is a unit vector in $\mathbb{C}^n \otimes \mathbb{C}^n$ and that

$$\xi = (1_n \otimes a)\omega .$$

We now claim that

$$\Phi_m(|\xi\rangle\langle\xi|) = (1_n \otimes a)\Phi_n(|\omega\rangle\langle\omega|)(1_n \otimes a)^* .$$

This is true since $\Phi_m$ acts on the first factor of $M_n(\mathbb{C}) \otimes M_m(\mathbb{C})$, while multiplication by $(1_n \otimes a)$ on the left and by $(1_n \otimes a)^*$ on the right act on the second factor of $M_n(\mathbb{C}) \otimes M_m(\mathbb{C})$. To write it out explicitly, note that by (6.21), $|\xi\rangle\langle\xi| = |(1_n \otimes a)\omega\rangle\langle(1_n \otimes a)\omega| = (1_n \otimes a)|\omega\rangle\langle\omega|(1_n \otimes a)^*$. By Lemma 6.2, (6.20) and (6.21), it follows that

$$|\xi\rangle\langle\xi| = \sum_{i,j=1}^{n} |\eta_i\rangle\langle\eta_j| \otimes |\zeta_i\rangle\langle\zeta_j|$$

and

$$|\omega\rangle\langle\omega| = \frac{1}{n} \sum_{i,j=1}^{n} |\eta_i\rangle\langle\eta_j| \otimes |\eta_i\rangle\langle\eta_j| .$$

Then

$$\Phi_m(|\xi\rangle\langle\xi|) = \sum_{i,j=1}^{n} \Phi(|\eta_i\rangle\langle\eta_j|) \otimes |\zeta_i\rangle\langle\zeta_j|$$

$$= \frac{1}{n} \sum_{i,j=1}^{n} \Phi(|\eta_i\rangle\langle\eta_j|) \otimes |a\eta_i\rangle\langle a\eta_j|$$

$$= (1_n \otimes a) \left( \frac{1}{n} \sum_{i,j=1}^{n} \Phi(|\eta_i\rangle\langle\eta_j|) \otimes |\eta_i\rangle\langle\eta_j| \right) (1_n \otimes a)^*$$

$$= (1_n \otimes a)\Phi_n(|\omega\rangle\langle\omega|)(1_n \otimes a)^* ,$$

which proves (6.23). By (6.23), whenever $\Phi_n(|\omega\rangle\langle\omega|) \geq 0$, then for all $\xi \in \mathbb{C}^n \otimes \mathbb{C}^m$, $\Phi_m(|\xi\rangle\langle\xi|) \geq 0$.

Note that $\omega$ is a unit vector in $\mathbb{C}^n \otimes \mathbb{C}^n$, and so $\frac{1}{n} \sum_{i,j=1}^{n} |\eta_i\rangle\langle\eta_j| \otimes |\eta_i\rangle\langle\eta_j| = |\omega\rangle\langle\omega|$ is a rank-one projector, and in particular, is positive.
We can use any orthonormal basis for form $\omega$ and the projector onto its span, but at this level of generality, we might as well use the standard basis:

6.13 DEFINITION (Choi projector and Choi matrix). Let $\{\eta_1, \ldots, \eta_n\}$ be the standard basis of $\mathbb{C}^n$ so that $|\eta_i\rangle \langle \eta_j|$ is the $i,j$th matrix unit; i.e., the element of $M_n(\mathbb{C})$ with 1 in the $i,j$ place and 0 elsewhere. The Choi projector in $M_n(\mathbb{C}^n) \otimes M_n(\mathbb{C}^n)$ is the element

$$p_C = \frac{1}{n} \sum_{i,j=1}^n |\eta_i\rangle \langle \eta_j| \otimes |\eta_i\rangle \langle \eta_j|.$$  \hspace{1cm} (6.24)

As we have seen above, it is the orthogonal projection $|\omega\rangle \langle \omega|$ onto the span of the unit vector $\omega$ given in (6.21).

Let $\Phi$ be any linear transformation of $M_n(\mathbb{C})$ to $M_p(\mathbb{C})$. The Choi matrix of $\Phi$ is the element of $M_n(M_p(\mathbb{C}))$ given by

$$\Phi_n(p_C) = \frac{1}{n} \sum_{i,j=1}^n \Phi(|\eta_i\rangle \langle \eta_j|) \otimes |\eta_i\rangle \langle \eta_j|.$$  \hspace{1cm} (6.25)

We can now restate the conclusion that whenever $\Phi_n(|\omega\rangle \langle \omega|) \geq 0$, then for all $\xi \in \mathbb{C}^n \otimes \mathbb{C}^m$, $\Phi_m(|\xi\rangle \langle \xi|) \geq 0$:

6.14 THEOREM (Choi’s Theorem). Let $\Phi$ be a linear transformation from $M_n(\mathbb{C})$ to $M_p(\mathbb{C})$. Then $\Phi$ is completely positive if and only if $\Phi_n(p_C)$ is positive where $p_C$ is the Choi projector given by (6.24).

Choi’s Theorem says that $\Phi \in \mathcal{L}(M_n(\mathbb{C}), M_p(\mathbb{C}))$ is completely positive if and only if $\Phi_n$ is positive, but it says much more than that: To check this, we need only look at $\Phi_n$ applied to the single positive element $p_C$.

This proof turns on two essential points: First the extreme points of the unit ball of $\mathcal{B}(\mathbb{C}^n \otimes \mathbb{C}^m)$ are the rank one projectors in $\mathcal{B}(\mathbb{C}^n \otimes \mathbb{C}^m)$, and thus it suffices to consider $\Phi_m(|\xi\rangle \langle \xi|)$ for such a projector. Next, every vector $\xi$ in $\mathbb{C}^n \otimes \mathbb{C}^m$ can be written in the form $\xi = \langle 1_n \otimes a \rangle \omega$ for some $a \in \mathcal{L}(\mathbb{C}^n, \mathbb{C}^m)$. That is, acting only on the right factor, one can “steer” $\omega$ into a general position in $\mathbb{C}^n \otimes \mathbb{C}^m$. This fact has important consequences in quantum mechanics to which we shall return, but it for $m = n$ it is something familiar to us: We know that the representation $\pi$ of $M_n(\mathbb{C})$ on $\mathbb{C}^n \otimes \mathbb{C}^n$ given by $\pi(a) = 1_n \otimes a$ is irreducible, and $\omega$ is a cyclic vector for it. We now give a second proof of Choi’s Theorem that provides additional information.

Second proof of Theorem 6.14. Suppose that a linear map $\Phi$ from $M_n(\mathbb{C})$ to $M_p(\mathbb{C})$ is such that $\Phi(p_C)$ is positive. We then apply the Spectral Theorem to write

$$\Phi_n(p_C) = \sum_{j=1}^{np} \lambda_j |\zeta_j\rangle \langle \zeta_j|$$

where the $\lambda_j$ are the (non-negative) eigenvalues of $\Phi(e_{(c)})$, and the $\zeta_j$ are the eigenvectors.

Each $\zeta_j$ has an expansion $\zeta_j = \sum_{k=1}^n \zeta_{j,k} \otimes \eta_j$ for vectors $\zeta_{j,k} \in \mathbb{C}^p$ and where $\{\eta_k\}$ is the standard basis of $\mathbb{C}^n$. Then by Lemma 6.2,

$$|\zeta_j\rangle \langle \zeta_j| = \sum_{k,\ell=1}^n |\zeta_{j,k}\rangle \langle \zeta_{j,\ell}| \otimes |\eta_k\rangle \langle \eta_\ell|.$$
Therefore, $\Phi_n(p_C) = \sum_{k,\ell=1}^{n} \sum_{j=1}^{np} \lambda_j |\zeta_{j,k}\rangle \langle \zeta_{j,\ell}| \otimes |\eta_k\rangle \langle \eta_\ell|$. It then follows from (6.25) that

$$\Phi(|\eta_k\rangle \langle \eta_\ell|) = n \sum_{j=1}^{np} \lambda_j |\zeta_{j,k}\rangle \langle \zeta_{j,\ell}|$$

(6.26)

Now define $V_j$ to be the $p \times n$ matrix whose $k$ column is $\zeta_{j,k}$, so that $V_j \eta_k = \zeta_{j,k}$. Then $|\zeta_{j,k}\rangle \langle \zeta_{j,\ell}| = V_j |\eta_k\rangle \langle \eta_\ell| V_j^*$. Therefore, if we define $W_j = \sqrt{n \lambda_j} V_j^*$ for each $j$, we can rewrite (6.26) as $\Phi(|\eta_k\rangle \langle \eta_\ell|) = \sum_{j=1}^{np} W_j^* |\eta_k\rangle \langle \eta_\ell| W_j$. But then by linearity, for all $a \in M_n(\mathbb{C})$,

$$\Phi(a) = \sum_{j=1}^{np} W_j^* a W_j .$$

(6.27)

such maps are completely positive by Theorem 6.7.

The second proof has also yielded another result of Choi:

**6.15 THEOREM.** Let $\Phi \in \mathcal{L}(M_n(\mathbb{C}), M_p(\mathbb{C}))$ be completely positive. Then there is a set \{W_1, \ldots, W_{np}\} of $n \times p$ matrices such that for all $a \in M_n(\mathbb{C})$, $\Phi(a)$ is given by (6.27).

**6.16 DEFINITION** (Krauss operators). If $\Phi \in \mathcal{L}(M_n(\mathbb{C}), M_p(\mathbb{C}))$ is completely positive, then a set of of $n \times p$ matrices \{W_1, \ldots, W_m\} such that

$$\Phi(a) = \sum_{j=1}^{m} W_j^* a W_j$$

(6.28)

are called a set of Krauss operator for $\Phi$, and (6.28) is a Krauss representation of $\Phi$.

We shall discuss minimality of Krauss representations in the next section. Note that if the completely positive map $\Phi$ given by (6.28) is unital; i.e., $\Phi(1_m) = 1_p$ if and only if

$$\sum_{j=1}^{m} W_j^* W_j = 1_p$$

(6.29)

**6.17 EXAMPLE** (The partial transpose in $\mathcal{B}(\mathbb{C}^n \otimes \mathbb{C}^n)$). Let $\Psi$ be the transpose map on $M_n(\mathbb{C})$, and let $\Psi_n$ be its extension to $M_n(\mathbb{C}) \otimes M_n(\mathbb{C})$ which, upon identifying $M_n(\mathbb{C}) \otimes M_n(\mathbb{C})$ with $\mathcal{B}(\mathbb{C}^n \otimes \mathbb{C}^n)$, we refer to as the partial transpose on $\mathcal{B}(\mathbb{C}^n \otimes \mathbb{C}^n)$. We compute

$$\Psi_n(p_C) = \frac{1}{n} \sum_{i,j=1}^{n} |\eta_i\rangle \langle \eta_j| \otimes |\eta_j\rangle \langle \eta_i| .$$

(6.30)

We now show that $\Psi_n(p_C)$, while self-adjoint, is not positive. Here is an easy way to see this: Using (6.7) and (6.30), we compute that

$$(\Psi_n(p_C))^2 = \frac{1}{n^2} \sum_{i,j=1}^{n} |\eta_i\rangle \langle \eta_j| \otimes |\eta_j\rangle \langle \eta_j| = \frac{1}{n^2} 1_n .$$
Hence all of the eigenvalues of $\Psi_n(p_C)$ are all $\pm 1/n$. By (6.18) and (6.30),
\[
\text{Tr}[\Psi_n(p_C)] = \frac{1}{n} \sum_{i,j=1}^{n} \text{Tr}[[\eta_i]\langle \eta_j]]^2 = 1.
\]
Hence $1/n$ is an eigenvalue of multiplicity $n(n + 1)/2$ and $-1/n$ is an eigenvalue of multiplicity $n(n - 1)/2$.

6.6 Stinespring’s Theorem

6.18 THEOREM. Let $\mathcal{A}$ be a $C^*$ algebra with identity 1, and let $\Phi : \mathcal{A} \to \mathcal{B}(\mathcal{H})$ be a completely positive map. Then there exists a Hilbert space $\mathcal{K}$ and a unital $\ast$-homomorphism $\pi : \mathcal{A} \to \mathcal{B}(\mathcal{K})$ and a bounded operator $V : \mathcal{H} \to cK$ such that $\|V\|^2 = \|\Phi(1)\|$ and for all $a \in \mathcal{A}$,
\[
\Phi(a) = V^* \pi(a)V.
\]
(6.31)

Proof. Equip the vector space $\mathcal{A} \otimes \mathcal{H}$ with the Stinespring inner product $\langle \cdot, \cdot \rangle_\Phi$. Define
\[
\mathcal{N} = \{\xi \in \mathcal{A} \otimes \mathcal{H} : \langle \xi, \xi \rangle_\Phi = 0\}.
\]
Then $\mathcal{N}$ is a subspace of $\mathcal{A} \otimes \mathcal{H}$, and for all $\xi, \xi' \in \mathcal{A} \otimes \mathcal{H}$ we write $\xi \sim \xi'$ in case $\xi - \xi' \in \mathcal{N}$. We define an inner product, again denoted $\langle \cdot, \cdot \rangle_\Phi$ on the quotient space $\mathcal{A} \otimes \mathcal{H}/\mathcal{N}$ by
\[
\langle \{\xi\}, \{\zeta\} \rangle_\Phi = \langle \xi, \zeta \rangle_\Phi
\]
for all $\xi, \zeta \in \mathcal{A} \otimes \mathcal{H}/\mathcal{N}$ and note that by the Cauchy-Schwarz inequality, the inner product is independent of the choice of representatives. Let $\mathcal{K}$ be the Hilbert space completion of $\mathcal{A} \otimes \mathcal{H}/\mathcal{N}$ in the metric associated to this inner product.

References

