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# Burnside factors, amenability defects and transitive families of projections in factors of type II(1) 

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# Burnside Factors, Amenability Defects and Transitive Families of Projections in Factors of Type $I I_{1}$ 

BY<br>Jon P. Bannon<br>B.S., University of New Hampshire (1999)

DISSERTATION

Submitted to the University of New Hampshire
in partial fulfillment of the requirements for the degree of

Doctor of Philosophy
in
Mathematics

May 2005

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## Dedication

To Tenaya

## Acknowledgments

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ABSTRACT<br>Burnside Factors, Amenability Defects and Transitive Families of Projections in Factors of Type $I I_{1}$<br>by<br>Jon P. Bannon<br>University of New Hampshire, May, 2005

We introduce a notion of transitive family of projections in a type $I I_{1}$ factor and prove that there exists (i) a 5 element transitive family in the hyperfinite $I I_{1}$ and (ii) a 12 element free transitive family. We then prove that the group von Neumann algebras of the known infinite free Burnside groups are all type $I_{1}$ factors. Our investigation of weak-amenability properties of Burnside groups leads us to consider the Connes theory of correspondences. From this investigation we are able to define a new Følner invariant for type $I_{1}$ factors. We prove a monotonicity result and find a positive lower bound for the free group factor $L\left(\mathbb{F}_{2}\right)$.

## Chapter 1

## Introduction and Background

### 1.1 Introduction

In this dissertation, we investigate various questions about the structure of type $I I_{1}$ factors. The way we present the material in the paper respects the order in which it was investigated, to give the reader a sense of the genesis of the ideas. We begin in Chapter 2 by introducing the notion of transitive family of projections in a type $I I_{1}$ factor. We prove, by generalizing a method of Halmos, that in certain classes of type $I_{1}$ factors there always exists a transitive family of five nontrivial projections. We then prove that one can find a transitive family of five nontrivial projections which are free with respect to the trace of the factor they generate. In Chapter 3 we present a short proof of a theorem proved originally by W. Burnside in 1902 which states that the only homomorphism of an infinite group of finite exponent into the general linear group of a finite dimensional vector space is the zero homomorphism. The main original contribution of our proof is to highlight which parts of the proof are intrinsically operator-algebraic and which are not, helping to more clearly expose the obstruction to finding out whether or not the infinite Burnside groups provide a counterexample to the Connes embedding conjecture. After this, we go on to prove that the group von Neumann algebras of infinite free Burnside groups of large enough odd exponent are type $I I_{1}$ factors. Our proof relies heavily on results from group theory. In Chapter 4 we follow closely the unpublished notes of Sorin Popa on the

Connes theory of correspondences, filling in details in order to provide a readable, introductory account of this theory. The development of Chapter 4 leads us to use the Connes-Følner condition in chapter 5 to define a new Følner invariant $F ø l(M)$ for type $I_{1}$ factors, about which we obtain various results, culminating in a proof that $F \varnothing l\left(L\left(\mathbb{F}_{2}\right)\right) \geq \frac{1}{7}$.

Our main result in the first chapter came about by trying to apply the technique of Murray and von Neumann's proof that the free group factors do not possess Property $\Gamma$ to the generator question for type $I I_{1}$ factors. The generator question of von Neumann asks if every von Neumann algebra acting on a separable Hilbert space can be generated, as a von Neumann algebra, by a single element. This is a central open question in the subject. The work in the second chapter was inspired by two questions of Liming Ge. The first question, paraphrased, asks whether or not the group von Neumann algebras of infinite free Burnside groups provide a counterexample to the Connes embedding conjecture. The proof we give of the theorem Burnside in the first part of Chapter 3 may be a good starting point for attacking Ge's question. The second part of Chapter 3 deals with Ge's question of whether or not the group von Neumann algebra of an infinite free Burnside group is a factor, and whether or not it must contain a noncommutative free group subfactor. This is motivated by the famous question of von Neumann in group theory asking if every non-amenable group must contain a nonabelian free subgroup, which was answered in the negative with the infinite free Burnside groups as a class of counterexamples. We solve the first part of this question in the affirmative. It is a well-known result of Adian that all of the free Burnside groups we consider are non-amenable groups. Recently the work of the group theorists Osin, Arzhantseva, Burillo, Lustig, Reeves, Short and Ventura has shown that there exist various notions of weak amenability for groups, and that the free Burnside groups are not, with respect to these various notions, weakly amenable (c.f.
[19], [4]). These authors construct several invariants that measure the amenability defect of a finitely generated group. This inspired us to consider whether or not there may be analogous notions of weak amenability for finitely generated type $I I_{1}$ factors, and whether we could define new notions of amenability defect for these factors. The work of Osin uses Huliniski's representation-theoretic characterization of amenability, which leads us to consider the non-commutative representation theory provided by Connes's theory of correspondences. In Chapter 4 we closely follow Popa's unpublished notes on correspondences and fill in many details. This work provided us with a powerful point of view for thinking about amenability questions for type $I I_{1}$ factors and, in conjunction with the technique of Murray and von Neumann used in the first chapter, motivated the original work in chapter 5.

This research was partially supported by a University of New Hampshire dissertation fellowship.

### 1.2 Background

The basics of the theory of operator algebras can be found in [16]. In this dissertation, we will provide a brief overview below, of related topics, for the sake of completeness. Other ideas will be introduced later in the text as needed. For Burnside groups, we refer to Adian [1].

Let $H$ be a Hilbert space, and $B(H)$ the algebra of all bounded linear operators from $H$ into itself. By the Riesz representation theorem, there is a natural involution * on $B(H)$, where if $T \in B(H)$ then $T^{*}$ is defined to be the unique operator in $B(H)$ satisfying $\langle T x, y\rangle=\left\langle x, T^{*} y\right\rangle$ for all $x, y \in H$. A subalgebra $\mathcal{A}$ of $B(H)$ is said to be self-adjoint if $T \in \mathcal{A}$ implies that $T^{*} \in \mathcal{A}$. Given $T \in B(H)$ and $x, y \in$ $H$, define $\omega_{x, y}(T)=\langle T x, y\rangle$. The weak-operator topology on $B(H)$ is the coarsest topology with respect to which each of the linear functionals $\omega_{x, y}$ is continuous. A
von Neumann algebra is a self-adjoint subalgebra of $B(H)$ that is closed in the weakoperator topology. John von Neumann discovered that just as one can decompose any semisimple algebra as a direct sum of simple algebras "indexed" by central elements, one can decompose any von Neumann algebra as a direct integral over its center, with the analogues of the simple summands played by von Neumann algebras, each having trivial center $\mathbb{C} I$. For this reason, he defined a factor to be a von Neumann algebra that has a trivial center. From 1935 to 1942 in the series of papers entitled "On Rings of Operators", von Neumann and F.J. Murray developed the theory of von Neumann algebras beginning from the point of view of the Wedderburn structure theory of semisimple algebras.

All von Neumann algebras are generated by the self-adjoint projections they contain, a fact that motivated Murray and von Neumann to use properties of the projection lattice to classify the factors. Murray and von Neumann compared the ranges of two given projections in the factor by a partial isometry also in the factor. From this idea they obtained an equivalence relation on the set of projections, and a total ordering $\preceq$ on the set of equivalence classes. More precisely, two projections $P, Q$ in a factor are equivalent if there exists an operator $V$ in the factor so that $V^{*} V=P$ and $V V^{*}=Q$; we have $P \preceq Q$ if there is a subprojection $Q_{0}$ of $Q$ in the factor so that $P$ is equivalent to $Q_{0}$. Imitating set theory, Murray and von Neumann defined a finite projection in a factor to be one that is not equivalent to any proper subprojection in the factor. They also defined the notion of minimal projection in the natural way. They classified the factors into three broad types, type $I$ factors are those with a minimal projection, type $I I$ factors are those containing no minimal projection but containing a finite projection, and type III factors are those in which every nonzero projection is infinite.

It should be noted that a general von Neumann algebra may also be called type $I$
(respectively $I I$ or $I I I$ ) if it has a direct integral decomposition into factors that are all of type $I$ (respectively $I I$ or $I I I$ ).

All von Neumann algebras are algebras with identity, so a finer classification of factors is available. If the identity of a factor is a finite projection, then we say that the factor is finite. It is not hard to show that every type $I$ factor contains a finite collection of mutually orthogonal equivalent minimal projections that sum to the identity element, and from this we may construct a $*$-isomorphism of any finite type $I$ factor with some $M_{n}(\mathbb{C})$. The only other finite factors are the infinite-dimensional ones, which are called type $I I_{1}$ factors. An important equivalent criterion for a factor to be finite is that there exist a unique faithful tracial state on the factor. The range of this tracial state is $\{0,1,2, \ldots, n\}$ for a type $I$ factor $*$-isomorphic to $M_{n}(\mathbb{C})$ and is $[0,1]$ for any type $I I_{1}$ factor. Two projections in the projection lattice of a finite factor are equivalent if and only if they have the same trace. Any factor that is not finite is called properly infinite, and must be of one of the distinct types $I_{\infty}(\cong \mathcal{B}(H)), I I_{\infty}$ or III. Although there is no trace on an infinite factor, we may define a $[0, \infty]$-valued dimension function on the projection lattice of a factor that behaves like the trace does on a finite factor. In particular two projections are equivalent if and only if they have the same dimension. The range of the dimension function is $\{0,1,2, \ldots, \infty\}$ for a type $I_{\infty}$ factor, $[0, \infty]$ for a type $I_{\infty}$ factor and $\{0, \infty\}$ for a type $I I I$ factor. In the first paper on rings of operators, Murray and von Neumann were able to construct examples of factors of every type using various actions of discrete groups on measure spaces.

Perhaps the most important motivation for studying type $I I_{1}$ factors lies in the further classification of all factors up to $*$-isomorphism, thanks to the remarkable work of A. Connes and M. Takesaki. In his 1973 Ph.D. thesis, Connes classified the type $I I$ factors into type $I I I_{\lambda}$, with $\lambda \in[0,1]$. Takesaki proved his duality theorem
by showing that every type $I I I$ factor is an abelian extension (by $\mathbb{R}$ ) of a type $I I_{\infty}$ factor. Since it can be easily shown that every type $I I_{\infty}$ factor is the tensor product of a type $I I_{1}$ factor by $\mathcal{B}(H)$, the problem of completely classifying von Neumann algebras essentially reduces to classifying the type $I_{1}$ factors.

## Chapter 2

## Transitive Families of Projections in Factors of Type $\mathrm{II}_{1}$

Let $\mathcal{H}$ be a complex, separable Hilbert space, and $\mathcal{M} \subseteq \mathcal{B}(\mathcal{H})$ be a factor of type $I I_{1}$. If $S$ is a non-empty set, we say that a family of norm-closed subspaces $\left\{\mathcal{H}_{i}\right\}_{i \in S}$ of $\mathcal{H}$ is transitive relative to $\mathcal{M}$ if for each $i \in S$, the projection $P_{i}$ of $\mathcal{H}$ onto $\mathcal{H}_{i}$ lies in $\mathcal{M}$ and only the scalar operators leave all of the $\mathcal{H}_{i}$ invariant. In this case, we also say that the family $\left\{P_{i}\right\}_{i \in S}$ is a transitive family of projections relative to $\mathcal{M}$. When $\operatorname{dim}(\mathcal{H}) \geq 3$, a transitive family cannot contain only two nontrivial projections $P$ and $Q$, since in this case for any $0<\lambda<1$,

$$
\lambda P(I-Q)+Q(I-P)
$$

leaves the ranges of both $P$ and $Q$ invariant, but cannot commute with $P$ unless $P Q P=Q P$. In this paper we first prove that if $\mathcal{M}$ is a type $I_{1}$ factor and is generated by two self-adjoint elements, then there is a transitive family of five projections relative to $\mathcal{M} \otimes M_{2}(\mathbb{C})$. This leads us to the question of whether or not there is a transitive family of three or four projections relative to some factor of type $I I_{1}$ ? To shed light on this question we consider free families of projections. A family $\left\{P_{i}\right\}_{i=1}^{n}$ of projections in a factor of type $I I_{1}$ is free if each $P_{i}$ has trace $\frac{1}{2}$ and the $P_{i}$ are free with respect to the trace (in the sense of Voiculescu, see [21] and [5]). We shall exhibit a free transitive family of twelve projections.

In $\mathcal{B}(\mathcal{H})$, a family of norm-closed subspaces is transitive if the only bounded operators on $\mathcal{H}$ that leave every subspace in the family invariant are scalars. Transitive families of subspaces were first considered by Paul Halmos in his 1970 paper "Ten problems in Hilbert space" [13]. In this paper Halmos studied medial subspace lattices, which are families of subspaces that contain $\{0\}, \mathcal{H}$, and at least two nontrivial subspaces of $\mathcal{H}$, with the additional property that any pair of nontrivial subspaces $K_{1}, K_{2}$ in the lattice are topologically complementary (that is, $K_{1} \cap K_{2}=\{0\}$ and $\left.\overline{s p a n}\left\{K_{1}, K_{2}\right\}=\mathcal{H}\right)$. Halmos constructed a finite-dimensional example of a transitive medial subspace lattice having five nontrivial elements, and raised the question of how small a transitive medial subspace lattice could be. In 1971, Harrison, Radjavi and Rosenthal found that, in a separable, infinite-dimensional Hilbert space, there is a transitive medial subspace lattice having four nontrivial elements[14]. It has become apparent since, that the construction of a medial subspace lattice having three elements is a difficult problem. In fact, even finding a transitive family of three nontrivial norm-closed subspaces is hard. Lambrou and Longstaff have shown that in finite ( $\geq 3$ ) dimensional $\mathcal{H}$, the smallest possible cardinality of a transitive family of subspaces is four [17]. Hadwin, Longstaff and Rosenthal have (when $\operatorname{dim} \mathcal{H}$ is infinite) found a transitive family of two norm-closed subspaces and a linear manifold, and have shown that the existence of a three element transitive family of norm-closed subspaces would follow from the existence of two dense operator ranges in $\mathcal{H}$ such that the only bounded operators leaving both of the ranges invariant are scalars[11].

We note that the questions considered in this paper are closely related to the generator question of von Neumann algebras, which asks if every von Neumann algebra acting on a separable Hilbert space is generated by two self-adjoint elements. The last example in this note shows that free families of projections that generate factors can be transitive.

### 2.1 Main results on transitive families

For basic information about von Neumann algebras, we refer the reader to [16].

Definition 1 Let $\mathcal{H}$ be a complex, separable Hilbert space, let I denote the identity in $\mathcal{B}(\mathcal{H})$, and let $I \in \mathcal{M} \subseteq \mathcal{B}(\mathcal{H})$ be a factor of type $I I_{1}$. Let $S$ be a nonempty set. $A$ family $\left\{P_{i}\right\}_{i \in S}$ of projections in $\mathcal{B}(\mathcal{H})$ is transitive relative to $\mathcal{M}$ if each $P_{i}$ is in $\mathcal{M}$ and the only operators $T \in \mathcal{M}$ that satisfy $\left(I-P_{i}\right) T P_{i}=0$ for all $i \in S$ are scalars.

Remark 2 When there is no danger of confusing which factor we are considering, we say that a family of projections $\left\{P_{i}\right\}_{i \in S} \subseteq \mathcal{M}$ is transitive, when $\left\{P_{i}\right\}_{i \in S}$ is transitive relative to $\mathcal{M}$.

Proposition 3 Let $\mathcal{H}$ be a complex, separable Hilbert space, and let $\mathcal{M} \subseteq \mathcal{B}(\mathcal{H})$ be a factor of type $I I_{1}$ such that $\mathcal{M}$ is generated, as a von Neumann algebra, by three projections $P_{1}, P_{2}$ and $P_{3}$. Then the family $\left\{P_{1}, I-P_{1}, P_{2}, I-P_{2}, P_{3}, I-P_{3}\right\}$ is transitive relative to $\mathcal{M}$.

Proof. If $T \in \mathcal{M}$ leaves the ranges of each of these projections invariant, then $T P_{i}-P_{i} T=P_{i} T P_{i}+\left(I-P_{i}\right) T P_{i}-P_{i} T P_{i}-P_{i} T\left(I-P_{i}\right)=0$ for $i=1,2,3$. It follows that $T \in \mathcal{M} \cap \mathcal{M}^{\prime}=\mathbb{C} I$.

We now extend an idea of Halmos[13].

Proposition 4 Let $\mathcal{H}$ be a complex, separable Hilbert space, and let $\mathcal{M} \subseteq \mathcal{B}(\mathcal{H})$ be a factor of type $I I_{1}$ such that $\mathcal{M}$ is generated, as a von Neumann algebra, by two self-adjoint elements $A, B$. There is a transitive family of 5 projections relative to $\mathcal{M} \otimes M_{2}(\mathbb{C})$.

Proof. We realize $\mathcal{M} \otimes M_{2}(\mathbb{C})$ as $M_{2}(\mathcal{M})$ acting on $\mathcal{H} \oplus \mathcal{H}$. Let $I$ denote the identity in $\mathcal{M}$, and $I_{2}$ the identity in $M_{2}(\mathcal{M})$. Now each of the projections

$$
P_{1}=\left[\begin{array}{ll}
I & 0 \\
0 & 0
\end{array}\right], P_{2}=\left[\begin{array}{cc}
0 & 0 \\
0 & I
\end{array}\right] \text { and } P_{3}=\left[\begin{array}{cc}
\frac{1}{2} I & \frac{1}{2} I \\
\frac{1}{2} I & \frac{1}{2} I
\end{array}\right]
$$

lies in $M_{2}(\mathcal{M})$. If an operator $T \in M_{2}(\mathcal{M})$ leaves the ranges of each of these three projections invariant, then $T$ must have the form

$$
\left[\begin{array}{cc}
T_{1} & 0 \\
0 & T_{1}
\end{array}\right]
$$

with $T_{1} \in \mathcal{M}$. We consider the matrix

$$
A_{1}=\left[\begin{array}{cc}
\frac{1}{2} I & \frac{1}{2} I \\
\frac{1}{2} A & \frac{1}{2} A
\end{array}\right] \in M_{2}(\mathcal{M})
$$

Letting $\lambda=\left\|A_{1} A_{1}^{*}\right\|^{-1}$, we note that there is the following equality of range projections $R\left(\lambda A_{1} A_{1}^{*}\right)=R\left(A_{1} A_{1}^{*}\right)=R\left(A_{1}\right)$. Now notice that $\lambda A_{1} A_{1}^{*}$ is a positive element of norm 1 , and therefore $0 \leq \lambda A_{1} A_{1}^{*} \leq I_{2}$, and by Lemma 5.15 in the first volume of [16], the sequence $\left\{\left(\lambda A_{1} A_{1}^{*}\right)^{\frac{1}{n}}\right\}$ converges in the strong-operator topology to $R\left(\lambda A_{1} A_{1}^{*}\right)$ and therefore $R\left(A_{1}\right) \in M_{2}(\mathcal{M})$, since $M_{2}(\mathcal{M})$ is a von Neumann algebra. Note that the range of the operator $A_{1}$ is a closed subspace of $\mathcal{H} \oplus \mathcal{H}$, since it is the graph of the bounded operator $A$. Now if $T$ leaves the ranges of $P_{1}, P_{2}, P_{3}$, and $A_{1}$ invariant, then for any $x \in \mathcal{H}$ it must be that

$$
\left[\begin{array}{cc}
T_{1} & 0 \\
0 & T_{1}
\end{array}\right]\left[\begin{array}{c}
x \\
A x
\end{array}\right]=\left[\begin{array}{c}
T_{1} x \\
T_{1} A x
\end{array}\right]=\left[\begin{array}{c}
T_{1} x \\
A T_{1} x
\end{array}\right]
$$

It follows that $T_{1} A=A T_{1}$.
Similarly, we consider

$$
B_{1}=\left[\begin{array}{cc}
\frac{1}{2} I & \frac{1}{2} I \\
\frac{1}{2} B & \frac{1}{2} B
\end{array}\right] \in M_{2}(\mathcal{M})
$$

and see that $R\left(B_{1}\right) \in M_{2}(\mathcal{M})$, and if $T$ leaves the ranges of $P_{1}, P_{2}, P_{3}$ and $B_{1}$ invariant, then $T_{1} B=B T_{1}$. Hence $T_{1}$ commutes with both generators of the factor $\mathcal{M}$, and hence $T_{1} \in \mathcal{M} \cap \mathcal{M}^{\prime}=\mathbb{C} I$. Thus the family of projections

$$
\left\{P_{1}, P_{2}, P_{3}, R\left(A_{1}\right), R\left(B_{1}\right)\right\} \subseteq M_{2}(\mathcal{M})
$$

is transitive.

Corollary 5 There is a transitive family of 5 projections relative to the hyperfinite $I_{1}$ factor $\mathcal{R}$.

Proof. It is well known that $\mathcal{R} \simeq \mathcal{R} \otimes M_{2}(\mathbb{C})$, and that $\mathcal{R}$ is generated by two selfadjoint elements. From the above proposition, $\mathcal{R} \otimes M_{2}(\mathbb{C})$ contains a transitive family of five projections. It follows that $\mathcal{R}$ contains a transitive family of five projections.

We now exhibit a free transitive family of projections.

Let $\left\{G_{i}\right\}_{i=1}^{n}$ be groups, and $e_{i}$ is the identity of $G_{i}$ for $i=1,2, \ldots n$. Let $\underset{i=1}{*} G_{i}$ denote the group free product of the $G_{i}$, and let $e$ denote its identity element. Recall that elements in $\underset{i=1}{*} G_{i}$ are given, in reduced form, by elements in the set $\{e\} \cup$ $\bigcup_{k \in \mathbb{N}}\left\{g_{i_{1}} g_{i_{2}} g_{i_{3}} \ldots g_{i_{k}}: i_{j} \in\{1,2, . ., n\} ; i_{j} \neq i_{j+1}\right.$ for $\left.j \in\{1,2, \ldots,(n-1)\} ; g_{i_{j}} \in G_{i_{j}} \backslash\left\{e_{i_{j}}\right\}\right\}$.

Let $G$ denote the group free product $\underbrace{\mathbb{Z}_{2} * \mathbb{Z}_{2} \ldots * \mathbb{Z}_{2}}_{12 \text { times }}$, and let $\mathbb{A}=\left\{a_{1}, a_{2}, a_{3}, . ., a_{12}\right\}$ denote the canonical set of generators of $G$. This group is I.C.C., therefore the group von Neumann algebra $\mathcal{L}_{G}$ acting on $l_{2}(G)$ is a factor of type $I I_{1}$. Recall that each element in $\mathcal{L}_{G}$ has the form $L_{x}$, where $x \in l_{2}(G)$ and the action on a function $y \in l_{2}(G)$ is defined by $\left(L_{x} y\right)(g)=\sum_{h \in G} x\left(g h^{-1}\right) y(h)$. With $g \in G$, let $x_{g}$ be the function in $l_{2}(G)$ that takes the value 1 on $g$ and 0 on every other group element. To avoid excessive use of subscripts, we everywhere write $L_{g}$ in place of $L_{x_{g}}$. Since $L_{a_{i}}^{2}=I$ for each $i \in\{1,2,3, \ldots, 12\}$, it is evident that $P_{i}=\frac{I+L_{a_{i}}}{2}$ is a projection in $\mathcal{L}_{G}$, and the family $\left\{P_{1}, P_{2}, \ldots, P_{12}\right\}$ is free with respect to the trace on $\mathcal{L}_{G}$.

Theorem 6 In $\mathcal{L}_{G}$, the family $\left\{P_{1}, P_{2}, P_{3}, \ldots, P_{12}\right\}$ is transitive.

Proof. Suppose that $L_{f} \in \mathcal{L}_{G}$ is a solution to the system $\left(I-P_{i}\right) T P_{i}=0,(i=$ $1,2,3 \ldots 12$ ), and therefore

$$
\left(I-L_{a_{i}}\right) L_{f}\left(I+L_{a_{i}}\right)=0 \quad(i=1,2,3 \ldots 12)
$$

Let both sides of this equation act on $x_{e}$, to obtain

$$
L_{f} x_{e}=L_{a_{i}} L_{f} x_{e}-L_{f} L_{a_{i}} x_{e}+L_{a_{i}} L_{f} L_{a_{i}} x_{e}(i=1,2,3, \ldots, 12)
$$

We see that

$$
\begin{aligned}
\left(L_{f} x_{e}\right)(g) & =\sum_{h \in G} f\left(g h^{-1}\right) x_{e}(h)=f(g), \\
\left(L_{a_{i}}\left(L_{f} x_{e}\right)\right)(g) & =\left(L_{a_{i}} f\right)(g)=\sum_{h \in G} x_{a_{i}}\left(g h^{-1}\right) f(h)=f\left(a_{i} g\right), \\
\left(L_{f}\left(L_{a_{i}} x_{e}\right)\right)(g) & =\left(L_{f} x_{a_{i}}\right)(g)=\sum_{h \in G} f\left(g h^{-1}\right) x_{a_{i}}(h)=f\left(g a_{i}\right), \\
\left(L_{a_{i}}\left(L_{f} L_{a_{i}} x_{e}\right)\right)(g) & =\left(L_{a_{i}} L_{f} x_{a_{i}}\right)(g) \\
& =\sum_{h \in G} x_{a_{i}}\left(g h^{-1}\right)\left(\sum_{k \in G} f\left(h k^{-1}\right) x_{a_{i}}(k)\right)=f\left(a_{i} g a_{i}\right) .
\end{aligned}
$$

From these it follows that for all $g \in G$

$$
f(g)=f\left(a_{i} g\right)-f\left(g a_{i}\right)+f\left(a_{i} g a_{i}\right) \quad(i=1,2,3, \ldots, 12)
$$

By the triangle inequality, we see that

$$
|f(g)| \leq\left|f\left(a_{i} g\right)\right|+\left|f\left(g a_{i}\right)\right|+\left|f\left(a_{i} g a_{i}\right)\right| \quad(i=1,2,3, \ldots, 12)
$$

and by the well known inequality $\left(x_{1}+\ldots+x_{k}\right)^{2} \leq k\left(x_{1}^{2}+\ldots+x_{k}^{2}\right)$ for non-negative real $k$, we see that

$$
|f(g)|^{2} \leq 3\left(\left|f\left(a_{i} g\right)\right|^{2}+\left|f\left(g a_{i}\right)\right|^{2}+\left|f\left(a_{i} g a_{i}\right)\right|^{2}\right) \quad(i=1,2,3, \ldots, 12)
$$

With $i, j \in\{1,2, \ldots, 12\}$ given, let $S_{i j}=\left\{g \in G: g\right.$ begins with $a_{i}$ and ends with $a_{j}$ in its reduced form in the free product $\}$.

With $g \in S_{i j}$, let $\left\{b_{i j}^{1}, \ldots, b_{i j}^{10}\right\}=\mathbb{A} \backslash\left\{a_{i}, a_{j}\right\}$. We then have

$$
\begin{aligned}
|f(g)|^{2} & \leq 3\left(\left|f\left(b_{i j}^{1} g\right)\right|^{2}+\left|f\left(g b_{i j}^{1}\right)\right|^{2}+\left|f\left(b_{i j}^{1} g b_{i j}^{1}\right)\right|^{2}\right), \\
|f(g)|^{2} & \leq 3\left(\left|f\left(b_{i j}^{2} g\right)\right|^{2}+\left|f\left(g b_{i j}^{2}\right)\right|^{2}+\left|f\left(b_{i j}^{2} g b_{i j}^{2}\right)\right|^{2}\right), \\
& \vdots \\
|f(g)|^{2} & \leq 3\left(\left|f\left(b_{i j}^{10} g\right)\right|^{2}+\left|f\left(g b_{i j}^{10}\right)\right|^{2}+\left|f\left(b_{i j}^{10} g b_{i j}^{10}\right)\right|^{2}\right) .
\end{aligned}
$$

Adding these inequalities, we obtain that

$$
10|f(g)|^{2} \leq 3 \sum_{k=1}^{10}\left(\left|f\left(b_{i j}^{k} g\right)\right|^{2}+\left|f\left(g b_{i j}^{k}\right)\right|^{2}+\left|f\left(b_{i j}^{k} g b_{i j}^{k}\right)\right|^{2}\right)
$$

Summing over $g \in S_{i j}$, we have

$$
\sum_{g \in S_{i j}}|f(g)|^{2} \leq \frac{3}{10} \sum_{g \in S_{i j}} \sum_{k=1}^{10}\left(\left|f\left(b_{i j}^{k} g\right)\right|^{2}+\left|f\left(g b_{i j}^{k}\right)\right|^{2}+\left|f\left(b_{i j}^{k} g b_{i j}^{k}\right)\right|^{2}\right)
$$

Suppose that $g_{1} \neq g_{2}$ are elements in $S_{i j}$. Note that by construction, all elements of the form $b_{i j}^{k} g_{1}, b_{i j}^{k} g_{1} b_{i j}^{l}$ or $g_{2} b_{i j}^{k}$ are in reduced form in the free product, for $k$, $l \in\{1,2, \ldots, 10\}$. Consider any $k_{1}, k_{2}, k_{3}, k_{4} \in\{1,2, \ldots, 10\}$. Since $g_{1} \neq g_{2}$, it follows that $b_{i j}^{k_{1}} g_{1} \neq b_{i j}^{k_{2}} g_{2}, b_{i j}^{k_{1}} g_{1} b_{i j}^{k_{2}} \neq b_{i j}^{k_{3}} g_{2} b_{i j}^{k_{4}}, b_{i j}^{k_{1}} g_{1} \neq g_{2} b_{i j}^{k_{2}}, b_{i j}^{k_{1}} g_{1} \neq b_{i j}^{k_{2}} g_{2} b_{i j}^{k_{3}}$, and $g_{1} b_{i j}^{k_{1}} \neq$ $b_{i j}^{k_{2}} g_{2} b_{i j}^{k_{3}}$. Therefore the right hand sum above can have no repeated terms, meaning for any $g_{0} \in G$, the term $\left|f\left(g_{0}\right)\right|^{2}$ shows up at most once on the right hand side of the inequality. We sum over all $i, j$ to obtain

$$
\sum_{i, j=1 g \in S_{i j}}^{12}|f(g)|^{2} \leq \frac{3}{10} \sum_{i, j=1}^{12} \sum_{g \in S_{i j}} \sum_{k=1}^{10}\left(\left|f\left(b_{i j}^{k} g\right)\right|^{2}+\left|f\left(g b_{i j}^{k}\right)\right|^{2}+\left|f\left(b_{i j}^{k} g b_{i j}^{k}\right)\right|^{2}\right)
$$

Let $S=\sum_{i, j=1}^{12} \sum_{g \in S_{i j}} \sum_{k=1}^{10}\left(\left|f\left(b_{i j}^{k} g\right)\right|^{2}+\left|f\left(g b_{i j}^{k}\right)\right|^{2}+\left|f\left(b_{i j}^{k} g b_{i j}^{k}\right)\right|^{2}\right)$. We now note that there can be repeated terms in $S$. We shall list the ways that a given term $\left|f\left(g_{0}\right)\right|^{2}$ may be repeated in the sum $S$. Suppose that $a, b \in \mathbb{A}$, and that $g_{0}$ begins with $a$ and ends with $b$ in its reduced form. Each occurrence of the term $\left|f\left(g_{0}\right)\right|^{2}$ in $S$ corresponds to an appearance of $\left|f\left(g_{0}\right)\right|^{2}$ on the right side of an inequality of the form $\left|f\left(g^{\prime}\right)\right|^{2} \leq$ $3\left(\left|f\left(a_{i} g^{\prime}\right)\right|^{2}+\left|f\left(g^{\prime} a_{i}\right)\right|^{2}+\left|f\left(a_{i} g^{\prime} a_{i}\right)\right|^{2}\right)$, where $a_{i} \in \mathbb{A}$ and $g^{\prime}$ is one of the group elements
$a g_{0}, g_{0} b$ or $a g_{0} b$. If $a \neq b$, then there can only be two occurrences of $\left|f\left(g_{0}\right)\right|^{2}$ in the sum $S$, one coming from the inequality

$$
\left|f\left(a g_{0}\right)\right|^{2} \leq 3\left(\left|f\left(g_{0}\right)\right|^{2}+\left|f\left(a g_{0} a\right)\right|^{2}+\left|f\left(g_{0} a\right)\right|^{2}\right)
$$

and one from the inequality

$$
\left|f\left(g_{0} b\right)\right|^{2} \leq 3\left(\left|f\left(b g_{0} b\right)\right|^{2}+\left|f\left(g_{0}\right)\right|^{2}+\left|f\left(b g_{0}\right)\right|^{2}\right)
$$

If $a=b$, then $\left|f\left(g_{0}\right)\right|^{2}$ may occur three times. Once in

$$
\left|f\left(a g_{0}\right)\right|^{2} \leq 3\left(\left|f\left(g_{0}\right)\right|^{2}+\left|f\left(a g_{0} a\right)\right|^{2}+\left|f\left(g_{0} a\right)\right|^{2}\right)
$$

again in the inequality

$$
\left|f\left(g_{0} a\right)\right|^{2} \leq 3\left(\left|f\left(a g_{0} a\right)\right|^{2}+\left|f\left(g_{0}\right)\right|^{2}+\left|f\left(a g_{0}\right)\right|^{2}\right),
$$

and finally, in the inequality

$$
\left|f\left(a g_{0} a\right)\right|^{2} \leq 3\left(\left|f\left(g_{0} a\right)\right|^{2}+\left|f\left(a g_{0}\right)\right|^{2}+\left|f\left(g_{0}\right)\right|^{2}\right)
$$

We therefore note that any term $\left|f\left(g_{0}\right)\right|^{2}$ in $S$ may occur at most three times. We call the number of times the term $\left|f\left(g_{0}\right)\right|^{2}$ appears in the sum $S$ the multiplicity of $\left|f\left(g_{0}\right)\right|^{2}$ in $S$.

Let script $\mathcal{T}$ denote $\{t: t$ is a term in $S\}$. Then, since all terms in $S$ are nonnegative, $S=\sum_{t \in \mathcal{T}} n_{t} t$ where $n_{t}$ is the multiplicity of the term $t$ in $S$.

We now have that

$$
\begin{aligned}
\sum_{g \in G \backslash\{e\}}|f(g)|^{2} & =\sum_{i, j=1}^{12} \sum_{g \in S_{i j}}|f(g)|^{2} \\
& \leq \frac{3}{10} S=\frac{3}{10} \sum_{t \in \mathcal{T}} n_{t} t \leq \frac{9}{10} \sum_{t \in \mathcal{T}} t \\
& \leq \frac{9}{10} \sum_{g \in G \backslash\{e\}}|f(g)|^{2} .
\end{aligned}
$$

Therefore $\sum_{g \in G \backslash\{e\}}|f(g)|^{2}$ is necessarily zero and $f(g)=0$ when $g \neq e$. We have now that only $f(e)$ may be nonzero, and hence $L_{f}$ must be a scalar. It follows that the family $\left\{P_{1}, \ldots, P_{12}\right\}$ is transitive.

Remark 7 The number of projections in the above theorem may be reduced. We believe that 4 such free projections should form a transitive family, but new techniques may be needed to prove this.

## Chapter 3

## Burnside Group Factors

In 1902, William Burnside raised the question of whether a finitely generated group must be finite if each of its elements has order dividing a given natural number $n$, called the exponent of the group [3]. Along with this question Burnside provided cases in which it had an affirmative answer, namely for any group of exponent 2 or 3 and for all groups of exponent 4 that can be generated by 2 elements.

No further progress was made on this problem until 1940, when I.N. Sanov showed that all finitely generated groups of exponent 4 must be finite [20]. Seventeen years later, Marshall Hall [12]demonstrated that this was also true for groups of exponent 6.

In 1964 Golod discovered the first example of a finitely generated infinite group with the property that every element in the group has finite order. This finding suggested the existence of infinite groups of large exponent.

In 1968, Novikov and Adian published a ground breaking series of papers [2] in which they proved that there are infinite periodic groups with odd exponent $n \geq 4381$. Their proof followed from a complicated inductive method to present the free Burnside groups $B(m, n)=\mathbb{F}_{m} / \mathbb{F}_{m}^{n}$ by relations of the form $A^{n}=1$ with specially chosen elements $A$ in $\mathbb{F}_{m}$. In 1975, Adian [1] improved the method and showed that there are infinite periodic groups of odd exponent $n \geq 665$. Beyond proving that the groups $B(m, n)$ are infinite, Adian and Novikov were able to prove much more. For example, they determined that the word and conjugacy problems are solvable in $B(m, n)$, that
any finite or abelian subgroup of $B(m, n)$ is cyclic, and that the centralizer of any non-identity element in $B(m, n)$ is a cyclic group of order $n$. This last result will be the main ingredient in the proof that the group $B(m, n)$ is an I.C.C. group.

### 3.1 An operator algebra view of a theorem of Burnside

In 1905 , Burnside proved that if $G$ is a subgroup of $G l(k, \mathbb{C})$ having finite exponent $n$, then the order of $G$ must be finite. The proof of this result is over 90 pages long but contains some deep ideas that we should consider. We include here a short proof of this theorem with an operator-algebraic flavor. The main reason to consider an old question like this is that it may provide insight into a question of Ge which asks whether or not infinite Burnside group von Neumann algebras can be embedded into an ultrapower of the hyperfinite type $I I_{1}$ factor.

Below, regard $M_{k}(\mathbb{C})$ as acting on $V=\mathbb{C}^{k}=\operatorname{span}\left\{e_{1}, \ldots, e_{k}\right\}$, where

$$
e_{i}=\left(\begin{array}{c}
0 \\
\vdots \\
1 \\
\vdots \\
0
\end{array}\right) \leftarrow\left(i^{\prime} \text { th position }\right)
$$

All groups $G \leq G l(k, \mathbb{C}) \subseteq M_{k}(\mathbb{C})$, i.e. we consider the groups represented already for simplicity.

Definition $8 A$ group $G$ is reducible if $V$ has a nontrivial $G$-invariant subspace, otherwise $G$ is said to be irreducible.

Definition 9 Let $\operatorname{End}(V)\left(\cong M_{k}(\mathbb{C})\right)$ denote the ring of all linear maps of $V$ to itself. We denote by $E n d_{G}(V)$ the subring of $E n d(V)$ consisting of linear maps that are also $G$-linear.

Remark 10 If $\psi: \operatorname{End}(V) \rightarrow M_{k}(\mathbb{C})$ is an isomorphism, then $\psi\left(\operatorname{End}_{G}(V)\right)=$ $(\operatorname{span}(G))^{\prime}$.

Lemma 11 (Schur) If $G$ is irreducible, then $E n d_{G}(V)$ is a division ring.

Proof. Suppose that $G$ is irreducible and $0 \neq \varphi \in \operatorname{End}_{G}(V)$. Since $\operatorname{ker} \varphi$ and $\varphi(V)$ are $G$-invariant subspaces of $V, \operatorname{ker} \varphi=\{0\}$ and $\varphi(V)=V$ and $\varphi$ is a $G$-linear automorphism of $V$.

Theorem 12 (Wedderburn Reciprocity) If $G$ is irreducible, then $(\operatorname{span}(G))^{\prime \prime}=\operatorname{span}(G)$.

Proof. It is clear that $\operatorname{span}(G) \subseteq(\operatorname{span}(G))^{\prime \prime}$. The bijection $\psi: M_{k}(\mathbb{C}) \rightarrow \mathbb{C}^{k^{2}}$ sending

$$
\left[\overrightarrow{v_{1}}\left|\overrightarrow{v_{2}}\right| \ldots \mid \overrightarrow{v_{k}}\right] \mapsto\left(\begin{array}{c}
\overrightarrow{v_{1}} \\
\overrightarrow{v_{2}} \\
\vdots \\
\overrightarrow{v_{k}}
\end{array}\right)
$$

gives us that $\psi T \psi^{-1}$ agrees with

$$
\left[\begin{array}{ccc}
T & 0 & 0 \\
0 & \ddots & 0 \\
0 & 0 & T
\end{array}\right]=T \otimes I \in M_{k^{2}}(\mathbb{C})
$$

in its action on $\mathbb{C}^{k^{2}}$. Let

$$
G^{\perp}=\left\{S \in M_{k}(\mathbb{C}) \mid \operatorname{Tr}\left(S^{*} g\right)=0 \text { for all } g \in G\right\}
$$

Decompose $M_{k}(\mathbb{C})=\operatorname{span}(G) \oplus G^{\perp}$ as an orthogonal direct sum with respect to the inner product $\langle A, B\rangle=\operatorname{Tr}\left(B^{*} A\right)$. Let $P$ be the orthogonal projection in $M_{k^{2}}(\mathbb{C})$ of $\mathbb{C}^{k^{2}}$ onto $\psi(\operatorname{span}(G))$. Suppose $v \in M_{k}(\mathbb{C}), v=v_{0}+v_{1}$, where $v_{0} \in \operatorname{span}(G)$ and $v_{1} \in G^{\perp}$. Note that $\operatorname{span}(G)$ and $G^{\perp}$ are both $G$-invariant subspaces. If $T \in \operatorname{span}(G)$,
then

$$
\begin{aligned}
P \psi T \psi^{-1} \psi(v) & =P \psi T\left(v_{0}+v_{1}\right)=P \psi\left(T v_{0}+T v_{1}\right) \\
& =P \psi T v_{0}=\psi T v_{0}
\end{aligned}
$$

and

$$
\begin{aligned}
\psi T \psi^{-1} P \psi(v) & =\psi T \psi^{-1}\left(P \psi\left(v_{0}\right)+P \psi\left(v_{1}\right)\right) \\
& =\psi T \psi^{-1}\left(P \psi\left(v_{0}\right)\right) \\
& =\psi T \psi^{-1} \psi\left(v_{0}\right)=\psi T v_{0}
\end{aligned}
$$

It follows that $P$ commutes with $\psi T \psi^{-1}=T \otimes I$ for all $T \in \operatorname{span}(G)$, hence $P \in$ $(\operatorname{span}(G))^{\prime} \otimes M_{k}(\mathbb{C}) \subseteq M_{k^{2}}(\mathbb{C})$.

Given $S \in(\operatorname{span}(G))^{\prime \prime}, S \otimes I=\psi S \psi^{-1}$ commutes with $P$, hence leaves the range $\psi(\operatorname{span}(G))$ of $P$ invariant. However, $I \in \operatorname{span}(G)$, so

$$
\psi S \psi^{-1} \psi(I)=\psi(S) \in \psi(\operatorname{span}(G))
$$

and therefore

$$
S \in \operatorname{span}(G)
$$

and $(\operatorname{span}(G))^{\prime \prime} \subseteq \operatorname{span}(G)$.

Lemma 13 (Burnside) If $G$ is irreducible then $\operatorname{span}(G)=M_{k}(\mathbb{C})$.

Proof. It is clear that $\operatorname{span}(G) \subseteq M_{k}(\mathbb{C})$. By Schur's lemma, $E n d_{G}(V)$ is a division ring, so $(\operatorname{span}(G))^{\prime}$ is as well. If $0 \neq T \in(\operatorname{span}(G))^{\prime}$ and $\lambda \in \sigma(T)(\neq \varnothing)$, then $T-\lambda I=0$ and $T=\lambda I$. Hence $(\operatorname{span}(G))^{\prime}=\mathbb{C} I$, and $(\operatorname{span}(G))^{\prime \prime}=M_{k}(\mathbb{C})$. By the Wedderburn Reciprocity Theorem, $\operatorname{span}(G)=(\operatorname{span}(G))^{\prime \prime}$, and the proof is complete.

Remark 14 If we considered only subgroups of unitary elements in $M_{k}(\mathbb{C})$, the above proof would follow from the von Neumann Double Commutant Theorem, which is a strengthening of Weddeburn Reciprocity to the infinite-dimensional seiting.

Proposition 15 (Burnside) If $G$ is irreducible, and there is a positive integer $n$ such that for all $g \in G, g^{n}=I$, then $|G| \leq n^{k^{3}}$.

Proof. Given $h \in G$, if $\lambda \in \sigma(h)$ then $x=h^{n} x=\lambda^{n} x$ for any eigenvector $x$ and therefore $\left(\lambda^{n}-1\right)=0$, so $\lambda$ is an $n^{\prime}$ th root of unity. It follows that $\operatorname{Tr}(h)$ can take on at most $n^{k}$ different values. (let $\operatorname{Tr}(h)$ be the usual non-normalized trace on $M_{k}(\mathbb{C})$ ). Choose a basis $\left(g_{1}, \ldots, g_{k^{2}}\right) \subseteq G$ for $M_{k}(\mathbb{C})$. We prove that

$$
\begin{gathered}
x=y \text { in } G \Leftrightarrow \\
\left(\operatorname{Tr}\left(x^{*} g_{1}\right), \ldots, \operatorname{Tr}\left(x^{*} g_{k^{2}}\right)\right)=\left(\operatorname{Tr}\left(y^{*} g_{1}\right), \ldots, \operatorname{Tr}\left(y^{*} g_{k^{2}}\right)\right) \text { in } \mathbb{C}^{k^{2}} .
\end{gathered}
$$

The " $\Rightarrow$ " direction is trivial. To prove the " $\Leftarrow$ " direction, note that if

$$
\operatorname{Tr}\left(x^{*} g_{i}\right)=\operatorname{Tr}\left(y^{*} g_{i}\right) \text { for all } i
$$

then $\operatorname{Tr}\left((x-y)^{*} g_{i}\right)=0$ for all $i$ and hence $\operatorname{Tr}\left((x-y)^{*} z\right)=0$ for all $z \in \operatorname{span}(G)=$ $M_{k}(\mathbb{C})$, and therefore $x-y=0$ and $x=y$. It follows that

$$
|G|=\#\left\{\left(\operatorname{Tr}\left(x^{*} g_{1}\right), \ldots, \operatorname{Tr}\left(x^{*} g_{k^{2}}\right)\right) \mid x \in G\right\} \leq\left(n^{k}\right)^{k^{2}}=n^{k^{3}}
$$

Theorem 16 (Burnside) If $G \leq G l(k, \mathbb{C})$ such that for some positive integer $n$ for every $g \in G, g^{n}=I$, then $|G| \leq n^{k^{3}}$.

The case where $G$ is irreducible was the subject of the previous proposition. Suppose that $V$ has a nontrivial $G$-invariant subspace. The proof of the result will be by induction on $k$. The case where $k=1$ is trivial. Suppose that all cases less than $k$ have been settled. If $W \leq V$ is a nontrivial $G$-invariant subspace, that $\operatorname{dim}(W)=r$,
and there is an $S \in M_{k}(\mathbb{C})$ such that $S^{-1} G S=G_{0} \leq G l(k, \mathbb{C})$ such that each element of $G_{0}$ has the form

$$
\left[\begin{array}{cc}
x_{1} & x_{3} \\
0 & x_{2}
\end{array}\right]
$$

where $x_{1}$ is an $r \times r$ matrix, $x_{2}$ is a $(k-r) \times(k-r)$ matrix. The matrices $x_{1} \in G_{1}$ (a representation of $G$ on $W$ ) and the $x_{2} \in G_{2}$ (a representation of $G$ on a complement of $W$ ). Since $0<r<k, G_{1}$ and $G_{2}$ satisfy the induction hypothesis, so $\left|G_{1}\right| \leq n^{r^{3}}$ and $\left|G_{2}\right| \leq n^{(k-r)^{3}}$. Given $A, B \in G_{0}$ with

$$
A=\left[\begin{array}{ll}
x_{1} & x_{3} \\
0 & x_{2}
\end{array}\right]
$$

and

$$
B=\left[\begin{array}{cc}
x_{1} & x_{3}^{\prime} \\
0 & x_{2}
\end{array}\right]
$$

we have that

$$
B^{-1} A=\left[\begin{array}{cc}
I_{r} & T \\
0 & I_{k-r}
\end{array}\right]
$$

for some matrix $T$. We have that

$$
I=\left(B^{-1} A\right)^{n}=\left[\begin{array}{ll}
I_{r} & n T \\
0 & I_{k-r}
\end{array}\right]
$$

implies that $T=0$ and hence that $B^{-1} A=I$. We have proven that given $x_{1} \in G_{1}$ and $x_{2} \in G_{2}$ that there exists a unique $x_{3}$ so that

$$
\left[\begin{array}{cc}
x_{1} & x_{3} \\
0 & x_{2}
\end{array}\right] \in G_{0} .
$$

It follows that

$$
|G|=\left|G_{0}\right| \leq\left|G_{1}\right|\left|G_{2}\right| \leq n^{r^{3}+(k-r)^{3}} \leq n^{r^{3}}
$$

since

$$
k^{3}=(r+k-r)^{3}=r^{3}+(k-r)^{3}+3 r^{2}(k-r)+3(k-r)^{2} r \geq r^{3}+(k-r)^{3}
$$

### 3.2 The Burnside groups are I.C.C.

In this section we prove a result that we're sure is known to group theorists, but we cannot find in the literature. This result is the first that ties together two deep areas of mathematics, the study of infinite Burnside groups and that of type $I_{1}$ factors.

A group $G$ is said to be an I.C.C. group if the conjugacy class of each non-identity element of $G$ is infinite. In what follows $G \backslash S$ will denote the set theoretic complement of $S$ in $G$. If $S \subseteq G$, and $g \in G$, we define $g S=\{g s: s \in S\}$. The centralizer of $g$ in $G$ is $C(g)=\{h \in G \mid h g=g h\}$.

If $m, n$ are positive integers, the group $B(m, n)$ will denote the free Burnside group of exponent $n$ on $m$ generators, and let 1 be its identity element. This group is given by a set of generators $a_{1}, a_{2} \ldots, a_{m}$ with defining relations of the form $g^{n}=1$ for every word $g$ in the group alphabet $\left\{a_{1}^{ \pm 1}, a_{2}^{ \pm 1}, \ldots, a_{m}^{ \pm 1}\right\}$.

We now quote a result of Adian [1].

Theorem 17 (Adian) The centralizer of an arbitrary non-identity element in $B(m, n)$ is a cyclic group of order $n$, for odd $n \geq 665$.

Lemma 18 Let $G$ be an infinite group with identity element $e$. If $C(g)$ is finite for every $g \in G \backslash\{e\}$, then $G$ is I.C.C.

Proof. Let $g \neq e$, and suppose that $C(g)$ is finite. Define $g_{0}$ to be $e$. There is an element $g_{1} \in G \backslash C(g)$, since $G$ is infinite. Suppose $k>1$ and that we have found $g_{0}, \ldots, g_{k}$ so that whenever $i, j \in\{0, \ldots, k\}$ and $i \neq j$ we have that $g_{i} g g_{i}^{-1} \neq g_{j} g g_{j}^{-1}$, or equivalently, $g_{i} \notin g_{j} C(g)$. For each $i \in\{0, \ldots, k\}$ the set $g_{i} S_{1}$ is finite, and hence the set $S=\bigcup_{i=0}^{k} g_{i} S_{1}$ is a finite union of finite sets, and is finite. It follows that there exists $g_{n+1} \in G \backslash S$ so that $g_{0}, \ldots, g_{n}, g_{n+1}$ have the property that $g_{i} g g_{i}^{-1} \neq g_{j} g g_{j}^{-1}$ whenever $i, j \in\{0, \ldots, n+1\}$ and $i \neq j$. In this way, we construct a sequence $\left\{g_{n}\right\}_{n=0}^{\infty}$
of elements in $G$ such that $g_{i} g g_{i}^{-1} \neq g_{j} g g_{j}^{-1}$ whenever $i \neq j$. Consequently, there are infinitely many distinct conjugates of $g$ in $G$.

Theorem 19 If $n \geq 665$ is odd, the group $B(m, n)$ is an I.C.C. group.

Proof. By Adian's result the centralizer of any non-identity element in $B(m, n)$ is a cyclic group of order $n$. By the lemma, it follows that $B(m, n)$ is an I.C.C. group.

## Chapter 4

## Correspondences

### 4.0.1 $W^{*}$-algebras and von Neumann algebras

A von Neumann algebra is a unital $C^{*}$-subalgebra of $\mathcal{B}(\mathcal{H})$ that is closed in the weak operator topology. A $W^{*}$-algebra is an (abstract) $C^{*}$-algebra that is a dual Banach space when viewed as a Banach space with respect to its $C^{*}$-norm. In other words, a $C^{*}$-algebra $\mathcal{A}$ is a $W^{*}$-algebra if there exists a Banach space $\mathcal{B}$, so that $\mathcal{B}^{\#}=\mathcal{A}$. Note that, if $N \subseteq \mathcal{B}(\mathcal{H})$ is a von Neumann algebra, the set $N_{\#}$ of all linear functionals $\rho$ on $N$ that are continuous on $(N)_{1}$ with respect to the weak operator topology is a normclosed subspace of the Banach space $N^{\#}$, and is therefore a Banach space. We call $N_{\#}$ the predual of $N$. Since a * isomorphism of one von Neumann algebra onto another induces a linear isometric isomorphism between their preduals, the Banach space $N_{\#}$ is, up to isometric isomorphism, independent of the faithful representation of the von Neumann algebra $N$. The importance of the predual is that any von Neumann algebra $N$ is isomorphic, as a Banach space, to $\left(N_{\#}\right)^{\#}$ (every norm-continuous linear functional on $N_{\#}$ has the form $\widehat{T}$, for some $T \in N$, where $\widehat{T}(\rho)=\rho(T)$ for all $\rho \in N_{\#}$ ). It follows that every von Neumann algebra is a $W^{*}$-algebra. The well-known stronger result of Sakai states that a $C^{*}$-algebra $\mathcal{A}$ is $*$-isomorphic to a von Neumann algebra if and only if $\mathcal{A}$ is a $W^{*}$-algebra. This theorem enables us to speak interchangeably of $W^{*}$ algebras and von Neumann algebras in what follows.

The ultraweak topology on a $W^{*}$-algebra $N$ is the weak* topology on $N$ obtained from $N_{\#}$. In terms of the convergence of nets, $T_{\lambda} \rightarrow T$ ultraweakly if and only
if $T_{\lambda}(\rho) \rightarrow T(\rho)$ for each $\rho \in N_{\#}$. If we regard $N$ as a von Neumann algebra, then $T_{\lambda} \rightarrow T$ ultraweakly if and only if $\rho\left(T_{\lambda}\right) \rightarrow \rho(T)$ for each linear functional $\rho$ on $N$ that is weak operator continuous on the unit ball $(N)_{1}$ of $N$. A $*$-homomorphism between two $W^{*}$-algebras $N, M$ shall be called normal if it is continuous with respect to the ultraweak topology on $N$ and the ultraweak topology on $M$. For a more extensive treatment on the ultraweak topology and normal functionals, we refer the reader to [16].

### 4.0.2 First definition of correspondence

We now introduce the first notion of correspondence.

Definition 20 Let $N, M$ be $W^{*}$-algebras. A correspondence between $N$ and $M$ (an $N$-M Hilbert $W^{*}$-bimodule) is a Hilbert space $H$ equipped with bilinear product maps

$$
\begin{array}{ll}
N \times H \rightarrow H:(T, \xi) \mapsto T \xi & (\text { ultraweak } \times\| \|)-\| \| \text { continuous } \\
H \times M \rightarrow H:(\xi, S) \mapsto \xi S & (\|\| \times \text { ultraweak })-\| \| \text { continuous }
\end{array}
$$

such that
(i) $I_{N} \xi=\xi I_{M}=\xi$
(ii) $T_{1}\left(T_{2} \xi\right)=T_{1} T_{2} \xi$
(iii) $\left(\xi S_{1}\right) S_{2}=\xi\left(S_{1} S_{2}\right)$
$(i v)(T \xi) S=T(\xi S)$
for all $\xi \in H$, all $T, T_{1}, T_{2} \in N$, and all $S, S_{1}, S_{2} \in M$.
We now make a few comments about this definition. If $\xi_{\lambda} \xrightarrow{\| \|} \xi$ in $H$, then for any $T \in N$, we have that $\left(T, \xi_{\lambda}\right) \xrightarrow{\text { u.w.x\| } \|}(T, \xi)$, so by continuity, $T \xi_{\lambda} \xrightarrow{\| \|} T \xi$ in $H$. Therefore the operator $L_{T}$ on $H$ defined by $L_{T} \xi=T \xi$ is bounded, and clearly preserves the unit.

If $T \in N$ is unitary, then for all $\xi \in H$,

$$
\begin{aligned}
\langle T \xi, \xi\rangle & =\left\langle T \xi, T T^{*} \xi\right\rangle=\left\langle L_{T} \xi, L_{T} T^{*} \xi\right\rangle \\
& =\left\langle L_{T}^{*}\left(L_{T} \xi\right), T^{*} \xi\right\rangle=\left\langle T^{*} T \xi, T^{*} \xi\right\rangle=\left\langle\xi, T^{*} \xi\right\rangle .
\end{aligned}
$$

therefore $L_{T}^{*}=L_{T^{*}}$ in this case. since we may write any element in $N$ as a finite linear combination of unitary elements, this also holds in general as well, by the bilinearity of the inner product. Now if $T_{\lambda} \xrightarrow{\text { u.w. }} T$, then for a given $\xi \in H$ we have that $T_{\lambda} \xi \xrightarrow{\| \|} T \xi$ and therefore $L_{T_{\lambda}} \xrightarrow{\text { S.O.T. }} L_{T}$, so if $T_{\lambda}$ is a monotone increasing net of positive operators with least upper bound $T$, then $L_{T_{\lambda}}$ is a monotone increasing net of positive operators with least upper bound $L_{T}$. We note that the ultraweakly continuous states $\omega$ on range $(L)^{\prime \prime}$ are exactly the normal states, i.e. those for which $\omega\left(H_{\lambda}\right) \rightarrow \omega(H)$ whenever $\left\{H_{\lambda}\right\}$ is a monotone increasing net of self-adjoint operators with least upper bound $H$. We may use this fact to prove that the map $L: T \rightarrow L_{T}$ is u.w.-u.w. continuous from $N$ into range $(L)^{\prime \prime}$ since for every normal state $\varphi$ on range $(L)^{\prime \prime}$, we have that $L \circ \varphi$ is a normal state on $N$. Every ultraweakly continuous linear functional on range $(L)^{\prime \prime}$ can be written as a linear combination of at most four normal states, a fact which follows from the polarization identity and the fact that each normal state on $B(H)$ can be written $\omega=\sum_{i=1}^{\infty} \omega_{x_{i}}$ with $\sum_{i=1}^{\infty}\left\|x_{i}\right\|^{2}=1$. We obtain that for every u.w. continuous linear functional $\varphi$ on $\operatorname{range}(L)^{\prime \prime}$, that $L o \varphi$ is an u.w. continuous linear functional on $N$, and this gives us that the map $L: N \rightarrow \operatorname{range}(L)^{\prime \prime}$ is u.w.-u.w. continuous.

Note that $(N)_{1}=\left(\left(N_{\#}\right)^{\#}\right)_{1}$ is weak*-compact, and therefore compact in the ultraweak topology. If $\varphi: N \rightarrow \mathcal{B}(H)$ is a normal $*$-representation, then $\varphi$ is continuous from the ultraweak topology on $N$ to the weak operator topology on $\mathcal{B}(H)$. It follows that $\varphi\left((N)_{1}\right)$ is W.O.T. compact, and hence W.O.T. closed. Since $\varphi$ takes $N$ onto $\varphi(N)$, which is a $C^{*}$-algebra and therefore a Banach space, it follows that $\varphi$ is an open mapping, so there exists $r>0$ such that $(\varphi(N))_{r} \subseteq \varphi\left((N)_{1}\right)$. By the Kaplansky
density theorem,

$$
\left(\varphi(N)^{-}\right)_{r} \subseteq(\varphi(N))_{r}^{-} \subseteq(\varphi(N))_{1}^{-}=(\varphi(N))_{1} \subseteq \varphi(N)
$$

where ( $)^{-}$denotes closure in the weak-operator topology. Hence $\varphi(N)^{-} \subseteq \varphi(N)$, and $\varphi(N)$ is W.O.T. closed, and therefore a von Neumann algebra. In particular, $\operatorname{range}(L)^{\prime \prime}=\operatorname{range}(L)$ in the above discussion.

In summary, we have that the map $L: T \mapsto L_{T}$ defines a normal, unital *representation of $N$ on $\mathcal{B}(H)$.

We define $M^{o p}$ to be equal to $M$ as a Banach space, but with the product

$$
S_{1} \circ S_{2}=S_{2} S_{1}
$$

where $S_{2} S_{1}$ is the product in the $W^{*}$-algebra $M$. Natural Banach and $C^{*}$-algebra structures on $M^{o p}$ are inherited from the Banach and $C^{*}$-algebra structures on $M$. Since $M=M^{o p}$ as a Banach space, $M_{\#}=\left(M^{o p}\right)_{\#}$, and therefore $M^{o p}$ is a $W^{*}$-algebra. We call $M^{o p}$ the opposite $W^{*}$-algebra of $M$.

We see, via an argument nearly identical to the one above, that the map $S \mapsto R_{S}$, where $R_{S} \xi=\xi S$ defines a normal *-representation of $M^{o p}$ on $\mathcal{B}(H)$.

Finally, property (iv) gives that $(T \xi) S=T(\xi S)$, and hence $R_{S} L_{T} \xi=L_{T} R_{S} \xi$ for all $\xi \in H, T \in N$, and $S \in M$. Therefore the representations $S \mapsto R_{S}$ and $T \mapsto L_{T}$ commute.

### 4.0.3 Second definition of correspondence

This brings us to our second definition of correspondence.

Definition 21 Let $N, M$ be $W^{*}$-algebras. A correspondence between $N$ and $M$ is a pair $\left(\pi_{N}, \pi_{M^{o p}}\right)$, where $\pi_{N}: N \rightarrow \mathcal{B}(H)$ and $\pi_{M^{o p}}: M^{o p} \rightarrow \mathcal{B}(H)$ are normal, unital *-representations of $N$ and $M^{o p}$ on the same Hilbert space $H$ such that $\pi_{N}(N) \subseteq$ $\pi_{M^{o p}}\left(M^{o p}\right)^{\prime}$.

Note that the Hilbert space $H$ in the above definition is naturally an $N$ - $M$ Hilbert $W^{*}$-bimodule with multiplications defined by

$$
\begin{aligned}
& (T, \xi) \mapsto \pi_{N}(T) \xi, \\
& (\xi, S) \mapsto \pi_{M^{\circ p}}(S) \xi .
\end{aligned}
$$

Given a correspondence ( $\pi_{N}, \pi_{M^{\circ p}}$ ), we may construct a unital *-representation $\pi_{N} \otimes$ $\pi_{M^{o p}}$ of the algebraic tensor product $*$-algebra $N \otimes M^{o p}$ on $\mathcal{B}(H) \otimes \mathcal{B}(H)$ :

$$
\left[\left(\pi_{N} \otimes \pi_{M^{\circ p}}\right)(T \otimes S)\right](\xi \otimes \eta)=\left(\pi_{N}(T) \otimes \pi_{M^{o p}}(S)\right)(\xi \otimes \eta)=\pi_{N}(T) \xi \otimes \pi_{M^{o p}}(S) \eta
$$

which is normal when restricted to $N \otimes I_{M^{\circ}}$ and also normal when restricted to $I_{N} \otimes M^{o p}$. Normality here means that if $T_{\lambda} \xrightarrow{u . w .} T$ in $N$, then $\pi_{N}\left(T_{\lambda} \otimes I_{M^{o p}}\right) \xrightarrow{\text { u.w. }}$ $\pi_{N}\left(T \otimes I_{M^{o p}}\right)$, and similarly in the other coordinate.

Conversely, given a unital *-representation $\pi: N \otimes M^{o p} \rightarrow \mathcal{B}(H)$ such that $\left.\pi\right|_{N \otimes I_{M^{o p}}}$ and $\left.\pi\right|_{I_{N} \otimes M^{o p}}$ are each normal, the pair $\left(\left.\pi\right|_{N \otimes I_{M^{o p}}},\left.\pi\right|_{I_{N \otimes M^{o p}}}\right)$ is a correspondence. In this manner we obtain a third definition of correspondence.

### 4.0.4 Third definition of correspondence

Definition 22 Let $N, M$ be $W^{*}$-algebras. A correspondence between $N$ and $M$ is a unital $*$-representation of the algebraic tensor product $N \otimes M^{o p}$ that is normal when restricted to $N \otimes I_{M^{o p}}$ and is normal when restricted to $I_{N} \otimes M^{o p}$.

We shall now motivate a fourth definition of correspondence, in the case where $N$ is a $W^{*}$-algebra, and $M$ is a countably-decomposable factor von Neumann algebra. This fourth point of view will help us gain some intuition for Connes' view of correspondences as morphisms in the category of von Neumann algebras in his book [7].

### 4.0.5 Normal left $W^{*}$-modules

In order to motivate the fourth definition of correspondence, we revisit the theory of normal left, $M$ Hilbert $W^{*}$-modules, which we shall simply call left $M$-modules. Please note that, although it is in most cases natural to first introduce modules and then bimodules, we chose to introduce bimodules first because they shall be the primary objects of our study.

Definition 23 Let $M$ be a $W^{*}$-algebra. A (left) $M$-module is a Hilbert space $H$ equipped with an (ultraweak $\times\| \|$ ) - \| \| continuous, bilinear product map

$$
\begin{aligned}
M \times H & \rightarrow H \\
(T, \xi) & \mapsto T \xi
\end{aligned}
$$

such that
i) $I_{M} \xi=\xi$
ii) $T_{1}\left(T_{2} \xi\right)=\left(T_{1} T_{2}\right) \xi$
for all $\xi \in H$, and all $T_{1}, T_{2} \in M$.

Proceeding as we did above in the case of bimodules, we have the more convenient equivalent definition of left module. This next definition is the one we will use.

Definition 24 A (left) $M$-module is a pair $(H, \pi)$, where $H$ is a Hibert space, and $\pi: M \rightarrow \mathcal{B}(H)$ is a unital, normal $*$-representation of $M$ on $H$.

We shall now define the concept of $M$-submodule, and the concept of isomorphism of $M$-modules.

Definition 25 An $M$-module ( $K, \pi_{K}$ ) is an $M$-submodule of the $M$-module $(H, \pi)$, if $K=P H$ for some projection $P \in(\pi(M))^{\prime} \subseteq \mathcal{B}(H)$, and $\pi_{K}=\pi_{P}$, where $\pi_{P}(T)=$ $\pi(T) P$ is the natural representation of $M$ on $P H$.

Definition $26 \operatorname{Let}\left(H_{1}, \pi_{H_{1}}\right)$ and $\left(H_{2}, \pi_{H_{2}}\right)$ be $M$-modules. A linear map $\varphi: H_{1} \rightarrow$ $\mathrm{H}_{2}$ is an $M$-module homomorphism if it is $M$-linear, that is,

$$
\varphi\left(\pi_{H_{1}}(T) \xi\right)=\pi_{H_{2}}(T) \varphi(\xi)
$$

for all $\xi \in H_{1}$ and all $T \in M$. An $M$-module isomorphism is a unitary $M$-module homomorphism, meaning that as a linear map it is norm-preserving and invertible.

We now demonstrate a basic link between $M$-modules and Murray-von Neumann equivalence of projections.

Proposition 27 Let $(H, \pi)$ be an $M$-module. Two $M$-submodules ( $P H, \pi_{P}$ ) and $\left(Q H, \pi_{Q}\right)$ of $(H, \pi)$ are isomorphic if and only if $P \sim Q$ in $\pi(M)^{\prime}$.

Proof. Suppose that $P \sim Q$ in $\pi(M)^{\prime}$. Let $V \in \pi(M)^{\prime}$ satisfy $V^{*} V=P$ and $V V^{*}=Q$, then $V$ is a partial isometry of $P H$ onto $Q H$, and the map $\varphi$ defined by $\varphi(P \xi)=V P \xi$ is an $M$-module isomorphism. Conversely, suppose that $\varphi: P H \rightarrow Q H$ is an $M$-module isomorphism, then by definition, $\varphi$ is onto $Q$ and $\|\varphi(P \xi)\|=\|P \xi\|$ for all $\xi \in H$, so $V \xi=\varphi(P \xi)$ defines a partial isometry $V$ in $\mathcal{B}(H)$ with initial space $P H$ and final space $Q H$, hence $P \sim Q$ in $\pi(M)^{\prime}$.

We now give a basic example of an $M$-module, and then show that any separable $M$ module can be described using this example. A von Neumann algebra $M$ is countably decomposable if any family of mutually orthogonal projections in $M$ has cardinality at most $\aleph_{0}$. Recall that a von Neumann algebra $M$ is countably decomposable if and only if there is a faithful, normal state on $M$. The proof of this result can be found in [15], but we include it for completeness.

Example 28 Let $M$ be a countably decomposable von Neumann algebra. If $\rho$ is a faithful normal state on $M$, then $\left(L^{2}(M, \rho), \pi_{\rho}\right)$ is a left $M$-module, where $\pi_{\rho}$ is the GNS representation of $M$ on $L^{2}(M, \rho)$ associated to the state $\rho$.

Let $l^{2}(\mathbb{N})$ be the Hilbert space of all square summable sequences of complex numbers. Note that the Hilbert space $L^{2}(M, \varphi) \otimes l^{2}(\mathbb{N})$ is naturally an $M$-module when the action of $M$ is given, for $T \in M$ and $\xi \otimes \eta \in L^{2}(M, \varphi) \otimes l^{2}(\mathbb{N})$, by $T(\xi \otimes \eta)=(T \otimes I)(\xi \otimes \eta)=T \xi \otimes \eta$. If $\left\{\left(H_{i}, \pi_{i}\right)\right\}_{i=1}^{\infty}$ are $M$-modules, then $\left(\oplus_{i=1}^{\infty} H_{i}, \oplus_{i=1}^{\infty} \pi_{i}\right)$ is an $M$-module, called the direct sum of the $\left\{\left(H_{i}, \pi_{i}\right)\right\}_{i=1}^{\infty}$. Here, the action of $M$ on $\bigoplus_{i=1}^{\infty} H_{i}$ is given by $\left(\bigoplus_{i=1}^{\infty} \pi_{i}(T)\right)\left(\xi_{i}\right)_{i=1}^{\infty}=\left(\pi_{i}(T) \xi_{i}\right)_{i=1}^{\infty}$. An $M$-module is called separable if it is separable as a Hilbert space. For notational purposes, in the following proposition we shall not write ( $\left.L^{2}(M, \rho), \pi_{\rho}\right)$ for the GNS $M$-module associated to the state $\rho$, but instead we shall refer to "the $M$-module $L^{2}(M, \rho)$ ". Also, we shall write the GNS action of $M$ on $L^{2}(M, \rho)$, for each $T \in M$ and $\xi \in L^{2}(M, \rho)$, by $T \xi$, rather than $\pi_{\rho}(T) \xi$. The same conventions will be held for modules closely related to $L^{2}(M, \rho)$, for example direct sums of copies of $L^{2}(M, \rho)$.

Proposition 29 Let $M$ be a von Neumann algebra, and $\rho$ be a faithful, normal state on $M$. If $H$ is a separable $M$-module, then there exists an $M$-submodule of $L^{2}(M, \rho) \otimes$ $l^{2}(\mathbb{N})$ that is isomorphic to $H$ as an $M$-module.

Proof. Decompose $H$ as $\bigoplus_{i=1}^{\infty}\left[\pi(M) \xi_{i}\right]$ for a sequence of vectors in $H$, and we see that $\pi=\left.\bigoplus_{i=1}^{\infty} \pi\right|_{\left[\pi(M) \xi_{i}\right]}$. If each $\left(\left[\pi(M) \xi_{i}\right],\left.\pi\right|_{\left[\pi(M) \xi_{i}\right]}\right)$ were isomorphic to an $M$-submodule of $L^{2}(M, \rho) \otimes l^{2}(\mathbb{N})$, then $H$ would be isomorphic to an $M$-submodule of $\bigoplus_{i=1}^{\infty}\left(L^{2}(M, \rho) \otimes\right.$ $l^{2}(\mathbb{N})$ ), where the action of $M$ on this Hilbert space would be the direct sum of the standard (GNS) action. But with this action of $M$, any Hilbert space isomorphism of $\bigoplus_{i=1}^{\infty}\left(L^{2}(M, \rho) \otimes l^{2}(\mathbb{N})\right)$ onto $L^{2}(M, \rho) \otimes l^{2}(\mathbb{N})$ naturally gives an $M$-module isomorphism from $\bigoplus_{i=1}^{\infty}\left(L^{2}(M, \rho) \otimes l^{2}(\mathbb{N})\right)$ onto $L^{2}(M, \rho) \otimes l^{2}(\mathbb{N})$. So we need to prove the theorem only in the case where $H$ is cyclic.

We introduce the following notation: if $\phi$ is a state on $M$, then $\langle,\rangle_{\phi}$ (resp. $\left\|\left\|\|_{\phi}\right.\right.$ ) shall denote the GNS inner product (resp. norm), associated to the state on $L^{2}(M, \omega)$.

Suppose now that $H$ is cyclic. In this case, there is a unit vector $\eta \in H$ such that $[\pi(M) \eta]=H$, and $(H, \pi) \cong L^{2}(M, \psi)$, where $\psi(T)=\langle\pi(T) \eta, \eta\rangle_{\psi}$ is the normal state on $M$ associated to $\eta$. We also may regard $M$ as represented on $\mathcal{B}\left(L^{2}(M, \rho)\right)$. From this viewpoint there exists a sequence $\left\{\xi_{i}\right\}_{i=1}^{\infty}$ of unit vectors in $L^{2}(M, \rho)$, such that $\sum_{i=1}^{\infty}\left\|\xi_{i}\right\|_{\rho}^{2}=1$, and $\psi(T)=\sum_{i=1}^{\infty}\left\langle T \xi_{i}, \xi_{i}\right\rangle_{\rho}$ for all $T \in M$. We shall also by $\psi$ denote the obvious extension of $\psi$ to all of $\mathcal{B}\left(L^{2}(M, \rho)\right)$.

Note that for all $i \in \mathbb{N}$,

$$
\left\|T \xi_{i}\right\|_{\rho}^{2}=\left\langle T^{*} T \xi_{i}, \xi_{i}\right\rangle_{\rho} \leq \psi\left(T^{*} T\right)=\left\langle\pi\left(T^{*} T\right) \eta, \eta\right\rangle_{\psi}=\|\pi(T) \eta\|_{\psi}^{2}
$$

It follows that $\pi(T) \eta \mapsto T \xi_{i}$ defines a bounded $M$-linear map from $\pi(M) \eta$ into $L^{2}(M, \rho)$ that extends to a map $R_{i}: L^{2}(M, \psi) \rightarrow L^{2}(M, \rho)$.

Let $\left\{e_{i}\right\}_{i=1}^{\infty}$ be the standard orthonormal basis for $l^{2}(\mathbb{N})$, and define

$$
\vartheta: L^{2}(M, \psi) \rightarrow L^{2}(M, \rho) \otimes l^{2}(\mathbb{N})
$$

by $\vartheta(\xi)=\sum_{i=1}^{\infty}\left(R_{i} \xi \otimes e_{i}\right)$. This map is $M$-linear since each $R_{i}$ is $M$-linear. Let the norm and inner product on $L^{2}(M, \rho) \otimes l^{2}(\mathbb{N})$ be written as $\|\|$ and $\langle$,$\rangle respectively.$ Now for all $T \in M$,

$$
\begin{aligned}
\|\vartheta(\pi(T) \eta)\| & =\left\|\sum_{i=1}^{\infty}\left(R_{i}(T \eta) \otimes e_{i}\right)\right\|=\left\|\sum_{i=1}^{\infty}\left(T \xi_{i} \otimes e_{i}\right)\right\| \\
& =\left\langle\sum_{i=1}^{\infty}\left(T \xi_{i} \otimes e_{i}\right), \sum_{j=1}^{\infty}\left(T \xi_{j} \otimes e_{j}\right)\right\rangle^{1 / 2} \\
& =\left(\sum_{i=1}^{\infty}\left\langle T \xi_{i}, T \xi_{i}\right\rangle_{\rho}\right)^{1 / 2}=\left(\sum_{i=1}^{\infty}\left\langle T^{*} T \xi_{i}, \xi_{i}\right\rangle_{\rho}\right)^{1 / 2} \\
& =\left(\psi\left(T^{*} T\right)\right)^{1 / 2}=\left(\left\langle\pi\left(T^{*} T\right) \eta, \eta\right\rangle_{\psi}\right)^{1 / 2} \\
& =\left(\left\langle\pi(T)^{*} \pi(T) \eta, \eta\right\rangle_{\psi}\right)^{1 / 2}=\left(\langle\pi(T) \eta, \pi(T) \eta\rangle_{\psi}\right)^{1 / 2} \\
& =\|\pi(T) \eta\|_{\psi} .
\end{aligned}
$$

It follows that $\vartheta$ is isometric, and $H \cong L^{2}(M, \psi) \cong \vartheta\left(L^{2}(M, \psi)\right)$ which is an $M$ submodule of $L^{2}(M, \rho) \otimes l^{2}(\mathbb{N})$.

By the previous two propositions, classifying the separable $M$-modules up to isomorphism amounts to classifying all $M$-submodules of the single $M$-module $L^{2}(M, \rho) \otimes$ $l^{2}(\mathbb{N})$ up to isomorphism, which amounts to classifying the projections in $(M \bar{\otimes} \mathbb{C} I)^{\prime}=$ $M^{\prime} \bar{\otimes} \mathcal{B}\left(l^{2}(\mathbb{N})\right) \subseteq \mathcal{B}\left(L^{2}(M, \rho) \otimes l^{2}(\mathbb{N})\right)$ up to Murray-von Neumann equivalence (Here we denote by $M^{\prime}$ the commutant of $M$ in $\mathcal{B}\left(L^{2}(M, \rho)\right)$ ). If $M$ is a factor von Neumann algebra, then either $M$ is of type $I I I$, in which case there is only one separable $M$-module up to isomorphism, or there is a faithful normal semi-finite tracial weight $\tau_{\infty}$ on $\left(M^{\prime} \otimes \mathcal{B}\left(l^{2}(\mathbb{N})\right)\right)_{+}$such that $\tau_{\infty}\left(I_{M} \otimes E_{11}\right)=1$ where $E_{11}=\left\langle, e_{1}\right\rangle e_{1} \in \mathcal{B}\left(l^{2}(\mathbb{N})\right)$. In the latter case, the isomorphism classes of separable $M$ modules are classified by the values of $\tau_{\infty}$ taken on the projection lattice in $M^{\prime} \bar{\otimes} \mathcal{B}\left(l^{2}(\mathbb{N})\right)$. Thus, the classification of separable modules over factors comes directly from the type classification of factors.

In particular, if $M$ is a finite factor, we define the $M$-dimension, or coupling constant, of the $M$-module ( $H, \pi$ ) by

$$
\operatorname{dim}_{M}((H, \pi))=\tau_{\infty}(P)
$$

where $P$ is any projection in $M^{\prime} \bar{\otimes} \mathcal{B}\left(l^{2}(\mathbb{N})\right)$ such that $(H, \pi) \cong P\left(L^{2}(M, \rho) \otimes l^{2}(\mathbb{N})\right)$ as an $M$-module. In the case where $N \subseteq M$ is an inclusion of a subfactor $N$ of a finite factor $M$ with trace $\tau$, one may naturally view $L^{2}(M, \tau)$ as an $N$ module, and in this situation we define the Jones index $[M, N]$ of the inclusion to be $\operatorname{dim}_{N}\left(L^{2}(M, \tau)\right)$.

In general, it is not true that if two von Neumann algebras are $*$-isomorphic, their commutants are as well. The following proposition will allow us to see that in the above situation, $\pi(M)$ is $*$-isomorphic to $(M \bar{\otimes} \mathbb{C} I) P$ (since $M$ is a factor, it is algebraically simple and therefore $\pi(M)$ is $*$-isomorphic to $M$ ) and also $\pi(M)^{\prime} \subseteq$ $\mathcal{B}(H)$ is $*$-isomorphic to $((M \bar{\otimes} \mathbb{C} I) P)^{\prime}=P((M \bar{\otimes} \mathbb{C} I)) P=P\left(M^{\prime} \bar{\otimes} \mathcal{B}\left(l^{2}(\mathbb{N})\right)\right) P \subseteq$ $\mathcal{B}\left(P\left(L^{2}(M, \rho) \otimes l^{2}(\mathbb{N})\right)\right)$.

Proposition 30 Let $M$ be a von Neumann algebra. If two $M$-modules $\left(H_{1}, \pi_{1}\right)$ and
$\left(H_{2}, \pi_{2}\right)$ are isomorphic as $M$-modules, then $\pi_{1}(M)^{\prime}$ is $*$-isomorphic to $\pi_{2}(M)^{\prime}$.

Proof. Suppose that $\varphi: H_{1} \rightarrow H_{2}$ is an $M$-module isomorphism. It follows that $\varphi\left(\pi_{1}(T) \xi\right)=\pi_{2}(T) \varphi(\xi)$ for all $T \in M$, hence that $\pi_{1}(T)=\varphi^{-1} \pi_{2}(T) \varphi=(A d(\varphi) \circ$ $\left.\pi_{2}\right)(T)$ for every $T \in M$. Since $\varphi$ is a unitary linear map from $H_{1}$ onto $H_{2}$, the mapping $A d(\varphi): \mathcal{B}\left(H_{2}\right) \rightarrow \mathcal{B}\left(H_{1}\right)$ sending $A \in \mathcal{B}\left(H_{2}\right)$ to $\varphi^{-1} A \varphi \in \mathcal{B}\left(H_{1}\right)$ is a surjective $*$-isomorphism. We shall show that $\left.A d(\varphi)\right|_{\pi_{2}(M)^{\prime}}$ gives a $*$-isomorphism of $\pi_{2}(M)^{\prime}$ onto $\pi_{1}(M)^{\prime}$. If $T^{\prime} \in \pi_{2}(M)^{\prime}$, then $T^{\prime} \pi_{2}(T)=\pi_{2}(T) T^{\prime}$ for all $T \in M$. We now show that $\left.\operatorname{Ad}(\varphi)\right|_{\pi_{2}(M)^{\prime}}\left(T^{\prime}\right) \in \pi_{1}(M)^{\prime}$. Let $\pi_{1}(T) \in \pi_{1}(M)$, then

$$
\begin{aligned}
\left.A d(\varphi)\right|_{\pi_{2}(M)^{\prime}}\left(T^{\prime}\right) \pi_{1}(T) & =\varphi^{-1} T^{\prime} \varphi \pi_{1}(T) \\
& =\varphi^{-1} T^{\prime} \varphi \varphi^{-1} \pi_{2}(T) \varphi \\
& =\varphi^{-1} T^{\prime} \pi_{2}(T) \varphi=\varphi^{-1} \pi_{2}(T) T^{\prime} \varphi \\
& =\varphi^{-1} \pi_{2}(T) \varphi \varphi^{-1} T^{\prime} \varphi=\pi_{1}(T) \varphi^{-1} T^{\prime} \varphi \\
& =\left.\pi_{1}(T) A d(\varphi)\right|_{\pi_{2}(M)^{\prime}}\left(T^{\prime}\right)
\end{aligned}
$$

It is easy to show that the map $\left.A d(\varphi)\right|_{\pi_{2}(M)^{\prime}}$ preserves adjoints, and is bijective.

## Amplifications of Factors

Suppose that $M$ is a countably decomposable factor, and let $H$ be a separable Hilbert space. A factor $*$-isomorphic to $P(M \bar{\otimes} \mathcal{B}(H)) P$ with $P$ some projection in $M \bar{\otimes} \mathcal{B}(H)$ is called an amplification of $M$. If $P$ is an infinite projection in $M \bar{\otimes} \mathcal{B}(H)$, then $P(M \bar{\otimes} \mathcal{B}(H)) P$ and $M \bar{\otimes} \mathcal{B}(H)$ are $*$-isomorphic. To see this, note that if $V V^{*}=I$ and $V^{*} V=P$ in $M \bar{\otimes} \mathcal{B}(H)$, then the linear map

$$
\theta: T \mapsto P V^{*} T V P
$$

satisfies $\theta\left(T_{1}\right) \theta\left(T_{2}\right)=P V^{*} T_{1} V P V^{*} T_{2} V P=P V^{*} T_{1} V V^{*} V V^{*} T_{2} V P=P V^{*} T_{1} T_{2} V P=$ $\theta\left(T_{1} T_{2}\right)$ and preserves adjoints. Since $M \bar{\otimes} \mathcal{B}(H)$ is a factor, $\theta$ is an injective map.

Given $P T P \in P(M \bar{\otimes} \mathcal{B}(H)) P$, we have that $V T V^{*} \in M \bar{\otimes} \mathcal{B}(H)$ and $\theta\left(V T V^{*}\right)=$ $P T P$ and hence $\theta$ is onto. From the reasoning above, we see that in the case where $M$ is properly infinite, every amplification of $M$ is isomorphic to $M \bar{\otimes} \mathcal{B}(H)$.

In the case where $M$ is a finite factor, we have that $M \bar{\otimes} \mathcal{B}(H)$ possesses a faithful normal tracial weight, $\tau_{\infty}=\tau \otimes T r$. If $P$ is a finite projection in $M \otimes \mathcal{B}(H)$, suppose that $\tau_{\infty}(P)=\alpha<\infty$. Choose $k \geq \alpha$ so that we may find a projection $Q_{1} \in M$ with $\tau\left(Q_{1}\right)=\frac{\alpha}{k}$. Find a finite projection $Q_{2} \in \mathcal{B}(H)$ so that $\operatorname{Tr}\left(Q_{2}\right)=\operatorname{rank}\left(Q_{2}\right)=k$. We now show that $Q_{2} \mathcal{B}(H) Q_{2}$ must be a finite type $I$ factor. Finiteness follows since $Q_{2}$ is the identity of $Q_{2} \mathcal{B}(H) Q_{2}$, and is a finite projection. We know that a projection $E$ in a factor $M$ is minimal if and only if $E M E=\mathbb{C} E$. Suppose that $E$ is a minimal projection in $\mathcal{B}(H)$. Since $\mathcal{B}(H)$ is a factor, the Murray-von Neumann order $\preceq$ on its projection lattice is a total ordering. Since $E$ is a minimal projection, we have that $E \preceq Q_{2}$, that is, there is a subprojection $E_{0}$ of $Q_{2}$ such that $E \sim E_{0}$ in $\mathcal{B}(H)$. It follows that $E_{0}$ is also a minimal projection in $\mathcal{B}(H)$, since for any non-zero projection $Q \in \mathcal{B}(H)$, we have that $E_{0} \sim E \preceq Q$. We claim the $Q_{2} E_{0} Q_{2}\left(=E_{0}\right)$ is a minimal projection in $Q_{2} \mathcal{B}(H) Q_{2}$. It is clear that $E_{0}$ is a projection, we must show that it is minimal in $Q_{2} \mathcal{B}(H) Q_{2}$. To show that $Q_{2} E_{0} Q_{2}$ is minimal, consider $Q_{2} T Q_{2} \in Q_{2} \mathcal{B}(H) Q_{2}$, , then

$$
\left(Q_{2} E_{0} Q_{2}\right)\left(Q_{2} T Q_{2}\right)\left(Q_{2} E_{0} Q_{2}\right)=Q_{2} \underbrace{E_{0} Q_{2} T Q_{2} E_{0} Q_{2}}_{\lambda E_{0}})=\lambda Q_{2} E_{0} Q_{2}
$$

It follows that $\left(Q_{2} E_{0} Q_{2}\right)\left(Q_{2} \mathcal{B}(H) Q_{2}\right)\left(Q_{2} E_{0} Q_{2}\right)=\mathbb{C}\left(Q_{2} E_{0} Q_{2}\right)$, and $Q_{2} E_{0} Q_{2}$ is a minimal projection. We see that $Q_{2} \mathcal{B}(H) Q_{2}$ is a finite type $I$ factor with identity $Q_{2}$. We have chosen $\operatorname{Tr}$ so that if $E_{0}$ is a minimal projection in $\mathcal{B}(H)$, then $\operatorname{Tr}\left(E_{0}\right)=1$. We can therefore write $Q_{2}$ as the sum of $k$ equivalent, pairwise orthogonal projections $E_{1}, \ldots, E_{k}$ in $\mathcal{B}(H)$, and therefore $\sum_{i=1}^{k} Q_{2} E_{i} Q_{2}=Q_{2}\left(\sum_{i=1}^{k} E_{i}\right) Q_{2}=Q_{2}$ in $Q_{2} \mathcal{B}(H) Q_{2}$, and hence $Q_{2}$ is the sum of $k$ such projections in $Q_{2} \mathcal{B}(H) Q_{2}$, and hence $Q_{2} \mathcal{B}(H) Q_{2}$ is a factor of type $I_{k}$, and $*$-isomorphic to $M_{k}(\mathbb{C})$. It follows
that if $P$ is a finite projection such that $\tau_{\infty}(P)=\alpha$, then $P(M \bar{\otimes} \mathcal{B}(H)) P \cong\left(Q_{1} \otimes\right.$ $\left.Q_{2}\right)(M \otimes \mathcal{B}(H))\left(Q_{1} \otimes Q_{2}\right) \cong\left(Q_{1} \otimes I\right)\left(M \otimes M_{k}(\mathbb{C})\right)\left(Q_{1} \otimes I\right)$. For this reason, we call a factor *-isomorphic to

$$
Q\left(M \bar{\otimes} M_{k}(\mathbb{C})\right) Q
$$

an amplification of $M$ when $0<\alpha \leq k$, and $\tau_{k}(Q)=\frac{\alpha}{k}$, where $\tau_{k}=\tau \otimes \frac{1}{k} \operatorname{Tr}$ is the normalized trace on the factor $M \bar{\otimes} M_{k}(\mathbb{C})$.

## Tomita-Takesaki Theory

We now recall some basic facts about the Tomita-Takesaki modular theory that will be needed in the sequel. Let $M \subseteq \mathcal{B}(\mathcal{H})$ be a von Neumann algebra, with commutant $M^{\prime}$. Suppose, in addition, that there is a joint cyclic and separating unit vector $\xi \in \mathcal{H}$ for $M$ and $M^{\prime}$. In this case, the mapping

$$
S: A \xi \mapsto A^{*} \xi \quad(A \in M)
$$

extends to a closable operator on $\mathcal{H}$, having polar decomposition $S=J \Delta^{1 / 2}$, where $\Delta$ is a positive invertible operator on $\mathcal{H}$, and $J$ is an order two antilinear isometry of $\mathcal{H}$ onto itself. The mapping

$$
\phi: A \mapsto J A^{*} J
$$

defines a *-anti-isomorphism from $M$ onto $M^{\prime}$. It follows that $M^{o p}$ and $M^{\prime}$ are *isomorphic under these circumstances, an isomorphism given by $\phi \circ i d^{-1}$.

### 4.0.6 Fourth definition of correspondence

We assume in this section that $N$ is a $W^{*}$-algebra and that $M$ is a countably decomposable factor. Let $\rho$ be a faithful normal state on $M^{o p}$. Suppose that ( $\pi_{N}, \pi_{M^{\circ p}}$ ) is a correspondence between $N$ and $M$, such that the underlying Hilbert space $H$ of the correspondence to be separable. We may consider only the action of
$M^{o p}$ for a moment, and view the underlying Hilbert space $H$ as a left $M^{o p}$-module ( $H, \pi_{M^{\circ P}}$ ). From our work on separable left modules, there is a projection $P \in$ $\left(M^{o p}\right)^{\prime} \otimes \mathcal{B}\left(l^{2}(\mathbb{N})\right) \subseteq \mathcal{B}\left(L^{2}\left(M^{o p}, \rho\right) \otimes l^{2}(\mathbb{N})\right)$ so that $\left(H, \pi_{M^{\circ p}}\right)$ is isomorphic as an $M^{o p}$ module to $P\left(L^{2}\left(M^{o p}, \rho\right) \otimes l^{2}(\mathbb{N})\right.$ ), and $\left(\pi_{o p}\left(M^{o p}\right)\right)^{\prime} \subseteq \mathcal{B}(H)$ is *-isomorphic to $P\left(\left(M^{o p}\right)^{\prime} \bar{\otimes} \mathcal{B}\left(l^{2}(\mathbb{N})\right)\right) P \subseteq \mathcal{B}\left(L^{2}\left(M^{o p}, \rho\right) \otimes l^{2}(\mathbb{N})\right)$, say $\phi$ gives a $*$-isomorphism of $\left(\pi_{o p}\left(M^{o p}\right)\right)^{\prime}$ onto $P\left(\left(M^{o p}\right)^{\prime} \bar{\otimes} \mathcal{B}\left(l^{2}(\mathbb{N})\right)\right) P$. Note that this is a $*$-isomorphism between von Neumann algebras, and is hence automatically normal. Since $\pi_{N}(N) \subseteq$ $\left(\pi_{o p}\left(M^{o p}\right)\right)^{\prime}$, the map $\phi$ restricted to $\pi_{N}(N)$ gives a normal $*$-homomorphism from $\pi_{N}(N)$ into $P\left(\left(M^{o p}\right)^{\prime} \bar{\otimes} \mathcal{B}\left(l^{2}(\mathbb{N})\right)\right) P$.

Now by the Tomita-Takesaki theory, whenever a von Neumann algebra has a joint cyclic and separating vector, then it is *-anti-isomorphic to $\left(M^{o p}\right)^{\prime}$. We know that $M^{o p}$ is also *-anti-isomorphic to $M$. If $\varphi_{1}: M^{o p} \rightarrow\left(M^{o p}\right)^{\prime}$ and $\varphi_{2}: M^{o p} \rightarrow M$ are *-antiisomorphisms, then $\varphi_{2}^{-1} \circ \varphi_{1}$ is a $*$-isomorphism of $\left(M^{o p}\right)^{\prime}$ onto $M$. From these we obtain a $*$-isomorphism $\Phi$ from $P\left(\left(M^{o p}\right)^{r} \bar{\otimes} \mathcal{B}\left(l^{2}(\mathbb{N})\right) P\right.$ onto $P_{1}\left(M \bar{\otimes} \mathcal{B}\left(l^{2}(\mathbb{N})\right) P_{1}\right.$ for some $P_{1}$ (Note that $L^{2}\left(M^{o p}, \rho\right)$ may be naturally identified with $\left.L^{2}(M, \rho)\right)$. As a result, we have that $\left.\Phi \circ \phi\right|_{\pi_{N}(N)}$ is a $*$-homomorphism from $\pi_{N}(N)$ into $P_{1}\left(M \bar{\otimes} \mathcal{B}\left(l^{2}(\mathbb{N})\right) P_{1}\right.$. This motivates the following definition.

Definition 31 Let $N$ be a $W^{*}$-algebra, and $M$ be a countably decomposable factor. A correspondence between $N$ and $M$ is a normal, unital $*$-homomorphism from $N$ into an amplification of $M$.

In motivating this definition, we have already shown how to get such a *-homomorphism from a (separable) correspondence. We should show how to get a (separable) correspondence from such a $*$-homomorphism. Suppose that $M^{\alpha}=P(M \overline{8} \mathcal{B}(H)) P$ is an amplification of the factor $M$ (with $H$ separable) and that $\varpi: N \rightarrow M^{\alpha}$ is a normal, unital $*$-homomorphism. Reversing the process above, we get that $P(M \bar{\otimes} \mathcal{B}(H)) P \cong$ $P\left(\left(M^{\circ p}\right)^{\prime} \bar{\otimes} \mathcal{B}(H)\right) P$ via a $*$ - isomorphism $\phi_{1}$. We have that $\phi_{1} \circ \varpi=\pi_{N}$ gives a nor-
mal representation of $N$ on $P\left(L^{2}\left(M^{o p}, \rho\right) \otimes H\right)$, and $\pi_{M^{o p}}=\phi_{2}$ sending $T \in M^{o p}$ to $(T \otimes I) P \in\left(M^{o p} \otimes \mathbb{C} I\right) P$ gives the normal representation of $M^{o p}$ on $P\left(L^{2}\left(M^{o p}, \rho\right) \otimes H\right)$. These two representations commute by construction, since $P\left(\left(M^{o p}\right)^{\prime} \bar{\otimes} \mathcal{B}(H)\right) P$ is the commutant of $\left(M^{o p} \otimes \mathbb{C} I\right) P$.

### 4.0.7 Fifth definition of correspondence

We now set out to show that if $N$ is a $W^{*}$-algebra and $M$ is a countably decomposable finite factor then one can, via Stinespring dilation, define a correspondence between $N$ and $M$ as a unital, normal, completely positive map between $N$ and $M$. For this we include a short discussion of completely positive maps which can be found, for the most part, in [8].

## Completely positive maps

Let $N, M$ be $C^{*}$-algebras. A linear map

$$
T: N \rightarrow M
$$

naturally induces a linear map $T^{(n)}\left(=T \otimes i d_{n}\right)$ from $N \otimes M_{n}(\mathbb{C})$ into $M \otimes M_{n}(\mathbb{C})$ defined by $T^{(n)}(a \otimes A)=T a \otimes A$ for $a \in N$ and $A \in M_{n}(\mathbb{C})$. Alternatively, viewing $N \otimes M_{n}(\mathbb{C})$ and $M \otimes M_{n}(\mathbb{C})$ as $M_{n}(N)$ and $M_{n}(M)$ respectively, we see that

$$
T^{(n)}\left(a_{i j}\right)_{i, j=1}^{n}=\left(T a_{i j}\right)_{i, j=1}^{n}
$$

for $\left(a_{i j}\right)_{i, j=1}^{n} \in M_{n}(N)$.
A linear map $T: N \rightarrow M$ is said to be completely positive if $T^{(n)}$ is positive for all $n \in \mathbb{N}$.

The transpose map on $M_{2}(\mathbb{C})$ is the standard example of a positive linear mapping that is not completely positive, since if $A \in M_{2}(\mathbb{C})$ has the property that $\langle A x, x\rangle \geq 0$
for all $x \in \mathbb{C}^{2}$, then $\left\langle A^{T} x, x\right\rangle \geq 0$ for all $x \in \mathbb{C}^{2}$, but

$$
B=\left[\begin{array}{llll}
0 & 0 & 0 & 0 \\
0 & 1 & 1 & 0 \\
0 & 1 & 1 & 0 \\
0 & 0 & 0 & 0
\end{array}\right]
$$

has non-negative eigenvalues $0,0,0,2$, but the map $\left(-^{T}\right)^{(2)}$ applied to $B$ yields the matrix

$$
\left[\begin{array}{ll}
{\left[\begin{array}{ll}
0 & 0 \\
0 & 1
\end{array}\right]^{T}} & {\left[\begin{array}{ll}
0 & 0 \\
1 & 0
\end{array}\right]^{T}} \\
{\left[\begin{array}{ll}
0 & 1 \\
0 & 0
\end{array}\right]^{T}} & {\left[\begin{array}{ll}
1 & 0 \\
0 & 0
\end{array}\right]^{T}}
\end{array}\right]=\left[\begin{array}{llll}
0 & 0 & 0 & 1 \\
0 & 1 & 0 & 0 \\
0 & 0 & 1 & 0 \\
1 & 0 & 0 & 0
\end{array}\right]
$$

that has eigenvalues $-1,1,1,1$, one of which is negative.
Since the positive elements in $N \otimes M_{n}(\mathbb{C})\left(\right.$ and $M \otimes M_{n}(\mathbb{C})$ ) are those that can be written as $A^{*} A$, and since $*$-homomorphisms are multiplicative and $*$-preserving, it is routine to verify that any $*$-homomorphism from $N$ into $M$ is a completely positive map.

Another example of a completely positive map, this time from a $C^{*}$-algebra $N$ to itself, is given by *-conjugation $A d(T)(-)=T^{*} T$ by any element $T$ in $N$. The complete positivity of this map can be seen as follows: regard $N$ as faithfully represented on a Hilbert space $H$, and regard $N \otimes M_{n}(\mathbb{C})$ as acting on the Hilbert space $H \otimes \mathbb{C}^{n}$. If $\sum\left(A_{i} \otimes B_{i}\right) \in N \otimes M_{n}(\mathbb{C})$ is such that

$$
\left\langle\sum\left(A_{i} \otimes B_{i}\right)(\xi \otimes \eta),(\xi \otimes \eta)\right\rangle=\sum\left\langle A_{i} \xi, \xi\right\rangle\left\langle B_{i} \eta, \eta\right\rangle \geq 0
$$

for all $\xi \otimes \eta$ in $H \otimes \mathbb{C}^{n}$, then

$$
\begin{aligned}
& \left\langle\left(A d(T) \otimes i d_{n}\right)\left(\sum\left(A_{i} \otimes B_{i}\right)\right)(\xi \otimes \eta),(\xi \otimes \eta)\right\rangle \\
& =\left\langle\sum\left(T^{*} A_{i} T \otimes B_{i}\right)(\xi \otimes \eta),(\xi \otimes \eta)\right\rangle \\
& =\sum\left\langle A_{i} T \xi, T \xi\right\rangle\left\langle B_{i} \eta, \eta\right\rangle \geq 0
\end{aligned}
$$

It follows that $A d(T)$ is completely positive. Note that if $N$ is a $C^{*}$-algebra with unit $I_{N}$, that $\operatorname{Ad}(T)\left(I_{N}\right)=T^{*} T$ does not necessarily equal $I_{N}$.

Given a representation $\pi$ of a $C^{*}$-algebra on a Hilbert space $H$, and $P$ a projection in $\mathcal{B}(H)$, the mapping $\pi_{P}=\left.\operatorname{Ad}(p) \circ \pi\right|_{P H}$ is called the compression of the representation $\pi$ to the closed subspace $P H$. Since the map $\left.T \mapsto T\right|_{P H}$, by an argument similar to the one above, is completely positive, the map $\pi_{P}$ is the composition of three completely positive maps, hence is completely positive.

The compression of a representation is not generally a representation. In fact, any state $\rho$ on a unital $C^{*}$-algebra $N$ may be regarded as the compression of the universal $G N S$ representation obtained from $\rho$ to the one-dimensional subspace spanned by the vector $v$ for which $\rho(A)=\langle\pi(A) v, v\rangle$ for all $A \in N$. The following computation shows why. Suppose that $w \in \operatorname{span}\{v\}$, and that $P$ is the projection onto $\operatorname{span}\{v\}$. We have that

$$
\begin{aligned}
P \pi(A) P w & =\langle w, v\rangle P \pi(A) v \\
& =\langle\pi(A) v, v\rangle\langle w, v\rangle v \\
& =\langle\pi(A) v, v\rangle P w \\
& =\rho(A) P w \\
& =\rho(A) w .
\end{aligned}
$$

It follows from the Stinespring theorem (a special case of which we prove in the next section) that every unital completely positive map from a unital $C^{*}$-algebra $N$
into some $\mathcal{B}(H)$ must be the compression of some representation of $\pi$ on a Hilbert, space containing $H$ as a subspace.

## Correspondences and completely positive maps

In this section we show how to associate a correspondence between a $W^{*}$-algebra $N$ and a countably decomposable finite factor $M$ with a trace $\tau$ to a normal, unital, completely positive map from $N$ into $M$ and how to obtain a correspondence between $N$ and $M$ from such a completely positive map. The construction below should work for any factor, but having a faithful trace $\tau$ instead of just a state $\rho$ considerably shortens the construction. We shall mainly concern ourselves with the case where $N$ and $M$ are type $I I_{1}$ factors later on however, so the cost of this assumption is not, too great. First, we shall show how to make $N \otimes M$ into a correspondence between $N$ and $M$ using an inner product obtained from a normal, unital completely positive $\operatorname{map} \phi: N \rightarrow M$.

Lemma 32 Let $N$ be a $W^{*}$-algebra and $M$ a countably decomposable factor with faithful normal state $\rho$. Regard $M \subseteq \mathcal{B}\left(L^{2}(M, \rho)\right)$ and let $\phi: N \rightarrow M$ be a unital completely positive map. The form $\langle,\rangle_{\phi}$ defined on simple tensors by

$$
\left\langle S_{1} \otimes T_{1}, S_{2} \otimes T_{2}\right\rangle_{\phi}=\left\langle\phi\left(S_{2}^{*} S_{1}\right) T_{1}, T_{2}\right\rangle
$$

extends to a positive conjugate-bilinear form $\langle,\rangle_{\phi}: N \otimes M \rightarrow \mathbb{C}$ on the algebraic tensor product. Here $\langle$,$\rangle denotes the inner product on the GNS Hilbert space L^{2}(M, \rho)$.

Proof. The fact that the form $\langle,\rangle_{\phi}$ extends to a conjugate-bilinear form on the algebraic tensor product $N \otimes M$ is straightforward. To prove that this map is positive we must show that, for any element $\sum_{i=1}^{n}\left(S_{i} \otimes T_{i}\right) \in N \otimes M$, that

$$
\left\langle\sum_{i=1}^{n}\left(S_{i} \otimes T_{i}\right), \sum_{i=1}^{n}\left(S_{i} \otimes T_{i}\right)\right\rangle_{\phi}=\sum_{i, j}\left\langle\phi\left(S_{j}^{*} S_{i}\right) T_{i}, T_{j}\right\rangle \geq 0
$$

To do this, consider the algebraic tensor product $M \otimes M_{n}(\mathbb{C})$ and the algebraic tensor product $N \otimes M_{n}(\mathbb{C})$. The algebraic tensor products admit $C^{*}$-norms from those of $N$ and $M_{n}(\mathbb{C})$, and are complete with respect to these norms, and may hence be viewed as $C^{*}$-algebras. Therefore, to show that an element $A$ in $M \otimes M_{n}(\mathbb{C})$ is positive, it suffices to show that we may write $A=B^{*} B$ for some $B \in M \otimes M_{n}(\mathbb{C})$. Let $\left\{E_{i j}\right\}_{i, j=1}^{n}$ denote the standard system of $n \times n$ matrix units in $M_{n}(\mathbb{C})$. The element $\sum_{l, k=1}^{n}\left(S_{l}^{*} S_{k}\right) \otimes E_{l k} \in N \otimes M_{n}(\mathbb{C})$ is positive, since

$$
\sum_{l, k=1}^{n}\left(S_{l}^{*} S_{k}\right) \otimes E_{l k}=\left(\sum_{l=1}^{n} S_{l} \otimes E_{1 l}\right)^{*}\left(\sum_{k=1}^{n} S_{k} \otimes E_{1 k}\right)
$$

Since the map $\phi: N \rightarrow M$ is completely positive, it follows that $\phi \otimes i d_{n}: N \otimes$ $M_{n}(\mathbb{C}) \rightarrow M \otimes M_{n}(\mathbb{C})$ is positive, and hence the element $\sum_{l, k=1}^{n} \phi\left(S_{l}^{*} S_{k}\right) \otimes E_{l k}$ is positive in $M \otimes M_{n}(\mathbb{C})$. Regard $M_{n}(\mathbb{C})$ as acting on $\mathbb{C}^{n}$ with standard orthonormal basis $\left\{e_{1}, \ldots, e_{n}\right\}$. It follows that for any $\sum_{i=1}^{n}\left(T_{i} \otimes e_{i}\right)$ in $M \otimes \mathbb{C}^{n} \subseteq L^{2}(M, \rho) \otimes \mathbb{C}^{n}=H$ we have

$$
\begin{aligned}
\left\langle\left(\sum_{l, k=1}^{n} \phi\left(S_{l}^{*} S_{k}\right) \otimes E_{l k}\right)\left(\sum_{i=1}^{n}\left(T_{i} \otimes e_{i}\right)\right),\left(\sum_{i=1}^{n}\left(T_{i} \otimes e_{i}\right)\right)\right\rangle_{H} & =\sum_{i, j} \sum_{k, l}\left\langle\phi\left(S_{j}^{*} S_{i}\right) T_{i}, T_{j}\right\rangle\left\langle E_{l k} e_{i}, e_{j}\right\rangle_{\mathbb{C}^{n}} \\
& =\sum_{k, l}\left\langle\phi\left(S_{l}^{*} S_{k}\right) T_{k}, T_{l}\right\rangle \\
& =\left\langle\sum_{k=1}^{n}\left(S_{k} \otimes T_{k}\right), \sum_{l=1}^{n}\left(S_{l} \otimes T_{l}\right)\right\rangle_{\phi} .
\end{aligned}
$$

This shows that the conjugate bilinear form $\langle,\rangle_{\phi}$ is positive.
Assume now that $\phi$ is normal, and let $H_{\phi}$ denote the completion of the normed space $(N \otimes M) / W$, where $W$ is the subspace $\left\{v \mid v \in N \otimes M\right.$ and $\left.\langle v, v\rangle_{\phi}=0\right\}$. This is a subspace because the form $\langle,\rangle_{\phi}$ is positive, and hence admits a Cauchy-Schwartz inequality. The subspace is closed due to the continuity of the norm $\|\cdot\|_{\phi}=\langle\cdot, \cdot\rangle_{\phi}^{1 / 2}$.

We consider the natural action of $N \otimes M^{o p}$ on the space $N \otimes M$ defined by $(S \otimes T)\left(\sum_{i=1}^{n}\left(S_{i} \otimes T_{i}\right)\right)=\sum_{i=1}^{n}\left(S S_{i} \otimes T_{i} T\right)$. Here the products $S S_{i}$ are taken in $N$ and the products $T_{i} T$ in $M$.

We now show that the subspace $W$ of $N \otimes M$ is invariant under this action of $N \otimes M^{o p}$ and that this action induces an action of $N \otimes M^{o p}$ on $H_{\phi}$. Note, if $A \in W$ and $B \in N \otimes M^{o p}$, regard $B$ as in $N \otimes M$, and then $\left|\langle A, B\rangle_{\phi}\right| \leq\langle A, A\rangle_{\phi}^{1 / 2}\langle B, B\rangle_{\phi}^{1 / 2}=0$. Now for $S \otimes T \in N \otimes M^{o p}$ and $v=\sum_{k=1}^{n}\left(S_{k} \otimes T_{k}\right) \in N \otimes M$,

$$
\begin{aligned}
\langle(S \otimes T) v,(S \otimes T) v\rangle_{\phi} & =\left\langle\sum_{k}\left(S S_{k} \otimes T_{k} T\right), \sum_{l}\left(S S_{l} \otimes T_{l} T\right)\right\rangle_{\phi} \\
& =\sum_{l, k}\left\langle\phi\left(S_{l}^{*} S^{*} S S_{k}\right) T_{k} T, T_{l} T\right\rangle=\sum_{l, k} \tau\left(T^{*} T_{l}^{*} \phi\left(S_{l}^{*} S^{*} S S_{k}\right) T_{k} T\right) \\
& =\sum_{l, k} \tau\left(T_{l}^{*} \phi\left(S_{l}^{*} S^{*} S S_{k}\right) T_{k} T T^{*}\right)=\sum_{l, k}\left\langle\phi\left(S_{l}^{*} S^{*} S S_{k}\right) T_{k} T T^{*}, T_{l}\right\rangle \\
& =\left\langle\sum_{k}\left(S^{*} S S_{k} \otimes T_{k} T T^{*}\right), \sum_{l}\left(S_{l} \otimes T_{l}\right)\right\rangle_{\phi}=\left\langle\left(S^{*} S \otimes T T^{*}\right) v, v\right\rangle_{\phi} \\
& =\left\langle(S \otimes T)^{*}(S \otimes T) v, v\right\rangle_{\phi}
\end{aligned}
$$

(Note: This computation was shortened by the traciality assumption.)
Now we see that if $v \in W$, then

$$
\begin{aligned}
0 & \leq\langle(S \otimes T) v,(S \otimes T) v\rangle_{\phi}=\left\langle(S \otimes T)^{*}(S \otimes T) v, v\right\rangle_{\phi} \\
& \leq\left\langle(S \otimes T)^{*}(S \otimes T) v,(S \otimes T)^{*}(S \otimes T) v\right\rangle_{\phi}^{1 / 2}\langle v, v\rangle_{\phi}^{1 / 2}=0,
\end{aligned}
$$

therefore $\left|\langle(S \otimes T) v,(S \otimes T) v\rangle_{\phi}\right|=0$ and $(S \otimes T) v \in W$. We have proven that $W$ is invariant under the action of $N \otimes M^{o p}$. It follows that the action of $N \otimes M^{o p}$ on $(N \otimes M) / W$ given by

$$
(S \otimes T)(v+W)=(S \otimes T) v+W
$$

is well-defined.
We are able to obtain more from the above computations, however. Given $v=$ $\sum_{k=1}^{n}\left(S_{k} \otimes T_{k}\right)$, the positivity of the form $\langle,\rangle_{\phi}$ tells us that the linear functional on $N$ defined by

$$
S \otimes I_{M^{o p}} \mapsto \sum_{l, k}\left\langle\phi\left(S_{l}^{*} S S_{k}\right) T_{k}, T_{l}\right\rangle=\left\langle\left(S \otimes I_{M^{o p}}\right) v, v\right\rangle_{\phi}
$$

is positive, and hence bounded with norm $\langle v, v\rangle_{\phi}$. Since $\tau$ and $\phi$ are both normal, this linear functional is again normal. Similarly the linear functional on $M^{o p}$ given by

$$
I_{N} \otimes T \mapsto \sum_{l, k}\left\langle\phi\left(S_{l}^{*} S_{k}\right) T_{k} T, T_{l}\right\rangle=\left\langle v\left(I_{N} \otimes T\right), v\right\rangle_{\phi}
$$

is positive, normal and has norm $\langle v, v\rangle_{\phi}$.
From these, we have that for any $v=\sum_{k=1}^{n}\left(S_{k} \otimes T_{k}\right) \in N \otimes M$, we have that

$$
\begin{aligned}
\left\|\left(S \otimes I_{M^{o p}}\right) v\right\|_{\phi}^{2} & =\left\langle\left(S \otimes I_{M^{o p}}\right) v,\left(S \otimes I_{M^{o p}}\right) v\right\rangle_{\phi} \\
& =\left\langle\left(S \otimes I_{M^{o p}}\right)^{*}\left(S \otimes I_{M^{o p}}\right) v, v\right\rangle_{\phi}=\left\langle\left(S^{*} S \otimes I_{M^{o p}}\right) v, v\right\rangle_{\phi} \\
& \leq\langle v, v\rangle_{\phi}\left\|S^{*} S\right\|_{N}
\end{aligned}
$$

so that each $S \otimes I_{M^{\circ p}}$ is a bounded operator with norm at most $\left\|S^{*} S\right\|_{N}^{1 / 2}=\|S\|_{N}$, with $\|\cdot\|_{N}$ the operator norm on $N$. Similarly, we have that each $I_{N} \otimes T$ is bounded with norm at most $\|T\|_{M^{\circ p}}$. It follows that each $S \otimes T=\left(S \otimes I_{M^{\circ p}}\right)\left(I_{N} \otimes T\right)$ acts as a bounded operator on $N \otimes M$, and therefore all finite linear combinations do as well, so the action extends by continuity to all of $H_{\phi}$.

By the normality of the functionals above, if $S_{\lambda} \otimes I_{M^{o p}} \xrightarrow{\text { ultraweak }} S \otimes I_{M^{o p}}$ in $N$, then for every vector $v \in H_{\phi}$, we have that

$$
\left\langle\left(S_{\lambda} \otimes I_{M^{\circ p}}\right) v, v\right\rangle_{\phi} \rightarrow\left\langle\left(S \otimes I_{M^{o p}}\right) v, v\right\rangle_{\phi}
$$

and hence $S_{\lambda} \otimes I_{M^{\circ p}} \xrightarrow{\text { W.O.T. }} S \otimes I_{M^{o p}}$ as acting on $H_{\phi}$, so the action is normal when restricted to $N \otimes I_{M^{o p}}$, and similarly normal when restricted to $I_{N} \otimes M^{o p}$.

It follows that $H_{\phi}$ is a correspondence between $N$ and $M$.
We now show how to recover $\phi$ from $H_{\phi}$. Given $T \in M \subseteq L^{2}\left(M, \tau_{M}\right)$, define $\Xi: M \rightarrow H_{\phi}$ such that $\Xi(T)=I_{N} \otimes T$. By the above remarks this extends to a contractive linear operator on all of $L^{2}\left(M, \tau_{M}\right)$, which we shall also call $\Xi$. We have,
for each $T_{1}, T_{2} \in M$, and $S \in N$, that

$$
\begin{aligned}
\left\langle\Xi^{*}\left(S \otimes I_{M^{o p}}\right) \Xi T_{1}, T_{2}\right\rangle_{\tau_{M}} & =\left\langle\left(S \otimes I_{M^{o p}}\right)\left(I_{N} \otimes T_{1}\right),\left(I_{N} \otimes T_{2}\right)\right\rangle_{\phi} \\
& =\left\langle\left(S \otimes T_{1}\right),\left(I_{N} \otimes T_{2}\right)\right\rangle_{\phi}=\tau_{M}\left(T_{2}^{*} \phi(S) T_{1}\right)=\left\langle\phi(S) T_{1}, T_{2}\right\rangle_{\tau_{M}}
\end{aligned}
$$

Hence the "coefficients" of $\Xi^{*}\left(S \otimes I_{M^{o p}}\right) \Xi$ match those of $\phi(S)$, and $\phi(S)=\Xi^{*}(S \otimes$ $\left.I_{M^{\circ p}}\right) \Xi$.

The previous paragraph suggests how we may canonically associate a completely positive map to a given correspondence $H$. Now choose a bounded vector $\xi$ in $H$, that is, let $\xi$ be a unit vector with the property that there is a $c>0$ so that for all $T \in M,\left\langle\xi T T^{*}, \xi\right\rangle \leq c \tau_{M}\left(T T^{*}\right)$. To see that such bounded vectors exist, consider first the case where $H=L^{2}\left(M, \tau_{M}\right)$, and choose $\xi \in M \subseteq H$ unitary, then appeal to the classification of right modules parallel to the earlier treated classification of left modules to obtain the general case. (Rough-hewn hint: We know that $H$ is isomorphic to $\left(L^{2}(M) \otimes L^{2}(\mathbb{N})\right) P$, and that $P$ cannot kill all simple tensors of the form $U \otimes S$ with $U$ unitary in $M^{o p}$ and $S \in L^{2}(\mathbb{N})$. Choose such a simple tensor, and note that given $T \in M$,

$$
\begin{aligned}
\|(U \otimes S) P(T \otimes I) P\|_{H}^{2} & \leq\|(U \otimes S)(T \otimes I)\|_{H}^{2} \\
& =\|T U\|_{L^{2}(M)}^{2}\|S\|_{L^{2}(\mathbb{N})}^{2}=c\|T\|_{L^{2}(M)}^{2}
\end{aligned}
$$

where $\left.c=\|S\|_{L^{2}(\mathbb{N})}^{2}\right)$

Define $\Xi: L^{2}\left(M, \tau_{M}\right) \rightarrow H$ by extending the map on $M$ defined by $\Xi(T)=\xi T$. Since $H$ is a correspondence, $\Xi$ is a bounded operator such that

$$
\begin{aligned}
\left\langle\Xi^{*} \pi_{N}(S) \Xi\left(J T^{*} J\right) T_{1}, T_{2}\right\rangle_{\tau_{M}} & =\left\langle\pi_{N}(S)\left(\xi T_{1} T\right),\left(\xi T_{2}\right)\right\rangle_{\tau_{M}} \\
& =\left\langle\pi_{M^{\circ p}}(T) \pi_{N}(S)\left(\xi T_{1}\right),\left(\xi T_{2}\right)\right\rangle_{\tau_{M}} \\
& =\left\langle\pi_{N}(S)\left(\xi T_{1}\right), \pi_{M^{o p}}(T)^{*}\left(\xi T_{2}\right)\right\rangle_{\tau_{M}} \\
& =\left\langle\pi_{N}(S)\left(\xi T_{1}\right), \pi_{M^{o p}}\left(T^{*}\right)\left(\xi T_{2}\right)\right\rangle_{\tau_{M}} \\
& =\left\langle\pi_{N}(S)\left(\xi T_{1}\right),\left(\xi T_{2} T^{*}\right)\right\rangle_{\tau_{M}}=\left\langle\Xi^{*} \pi_{N}(S) \Xi T_{1}, T_{2} T^{*}\right\rangle_{\tau_{M}} \\
& =\left\langle\left(J T^{*} J\right) \Xi^{*} \pi_{N}(S) \Xi T_{1}, T_{2}\right\rangle_{\tau_{M}}
\end{aligned}
$$

This shows that $\Xi^{*} \pi_{N}(S) \Xi$ commutes with $J M J=M^{\prime}$, and hence $\Xi^{*} \pi_{N}(S) \Xi \in M^{\prime \prime}=$ $M$. It follows that the range of the map $A d(\Xi) \circ \pi_{N}: N \rightarrow \mathcal{B}\left(L^{2}\left(M, \tau_{M}\right)\right)$ is contained in $M$. Note also that the map defined by $\phi(S)=\Xi^{*} \pi_{N}(S) \Xi$ is a composition $\operatorname{Ad}(\Xi) \circ$ $\pi_{N}$ of two normal completely positive maps, $*$-conjugation by $\Xi$ and the normal representation $\pi_{N}$, hence $\phi$ is also completely positive and normal. We have thus obtained from the correspondence $H$ between $N$ and $M$ a normal, unital, completely positive map $\phi$ from $N$ into $M$. This motivates the following fifth interpretation of correspondence.

Definition 33 Let $N$ be a $W^{*}$-algebra, and $M$ be a countably decomposable factor. A correspondence between $N$ and $M$ is a normal, unital completely positive map from $N$ into $M$.

### 4.0.8 Equivalence and containment of correspondences

In this section we define what it means for two correspondences between $W^{*}$-algebras $N$ and $M$ to be equivalent. Suppose that $H, H^{\prime}$ are correspondences between $N$ and $M$, and that $\pi_{H}$ and $\pi_{H^{\prime}}$ are the representations of $N \otimes M^{o p}$ on $H$ and $H^{\prime}$
respectively. Recall that $\pi_{H}$ is unitarily equivalent to $\pi_{H^{\prime}}$ if there exists a unitary linear map $U: H \rightarrow H^{\prime}$ so that $\pi_{H}=A d(U) \circ \pi_{H^{\prime}}$.

Definition 34 Let $N, M$ be $W^{*}$-algebras and let $H, H^{\prime}$ be correspondences between $N$ and $M$. We say that $H$ is equivalent to $H^{\prime}$, written $H \sim H^{\prime}$, if the associated representations of $N \otimes M^{o p}$ are unitarily equivalent.

If $H$ is a correspondence, we denote by $\widehat{H}$ its equivalence class under $\sim$. We denote the set of all equivalence classes of correspondences by $\operatorname{Corr}(N, M)$. When no confusion is likely, we will sometimes omit the hat ${ }^{\text {a }}$ and use the same notation for a correspondence and its class.

We now introduce the notion of subcorrespondence as an $N \otimes M^{o p}$-invariant subspace of $H$, i.e. a subrepresentation of $N \otimes M^{o p}$ on $H$.

Definition 35 A subcorrespondence of a correspondence $H$ is a subrepresentation of $N \otimes M^{o p}$ on $H$. A correspondence $K$ is said to be subequivalent to $H$ if $K$ is equivalent to a subcorrespondence of $H$.

### 4.0.9 Examples of correspondences

We now give a few examples of correspondences. The first two shall play a key role in what follows. The last example demonstrates that the "category" of correspondences is, in some sense, much richer than that of left-modules. Although these examples may be considered in full generality (using a faithful normal weight in place of a faithful normal state), we shall only include what we need below.

Example 36 (the identity correspondence, or trivial correspondence) Let $M$ be a $W^{*}$-algebra with faithful normal state $\rho$, and let $L^{2}(M, \rho)$ denote the standard $G N S$ Hilbert space. By the Tomita-Takesaki theory, $M$ acts on this Hilbert space by left
and right multiplication,

$$
\begin{aligned}
& (T, \xi) \mapsto T \xi \\
& (\xi, T) \mapsto J T^{*} J \xi=\xi T
\end{aligned}
$$

The left multiplication gives a normal representation of $M$, and the right multiplication gives a normal representation of $M^{o p}$. These representations commute, and hence $H_{i d}=L^{2}(M, \rho)$ becomes a correspondence between $M$ and itself, called the trivial or identity correspondence. By basic GNS considerations, using any other state at the outset shall yield an equivalent correspondence, hence the identity correspondence is unique up to equivalence.

It should be noted that if $M$ is finite and countably decomposable, then a correspondence $H$ between $M$ and itself contains the identity correspondence $H_{i d}$ as a direct summand if and only if there exists a separating central vector for $M$ in $H$, that is, a vector $\xi \in H$ such that for any $T \in M$, if $T \xi=0$ then $T=0$. To see this, note that if $H=H_{i d} \oplus K$, then the vector $\left(\xi_{0}, 0\right)$ is the desired separating vector with $\xi_{0}$ the separating vector for the action of $M$ on $H_{i d}$. Conversely, if $\xi$ is a separating vector for the action of $M$ on $H$, then we may write $H=[M \xi] \oplus[M \xi]^{\perp}$, and $[M \xi]$ is isomorphic to $H_{i d}$.

Example 37 (The coarse correspondence) Let $M$ and $N$ be $W^{*}$-algebras with faithful, normal states $\rho_{N}, \rho_{M}$ respectively. The coarse correspondence between $N$ and $M$ is the Hilbert space $L^{2}\left(N, \rho_{N}\right) \otimes L^{2}\left(M, \rho_{M}\right)$, such that $N$ acts via left multiplication in the first coordinate, and $M$ acts via right multiplication in the second coordinate, that is, for $S \in N$ and $T \in M$,

$$
(S \otimes T) \xi=S\left(J T^{*} J\right) \xi=S \xi T
$$

We may also identify the coarse correspondence with the Hilbert space of all HilbertSchmidt operators $L^{2}\left(L^{2}\left(N, \rho_{N}\right), L^{2}\left(M, \rho_{M}\right)\right)$ by identifying simple tensors with "rank-
one operators". Let $\xi \in L^{2}\left(N, \rho_{N}\right), S \in N, T \in M$, then

$$
(S \otimes T)(\xi) \equiv\langle\xi, S\rangle_{\rho_{N}} T \in L^{2}\left(M, \rho_{M}\right)
$$

The left action of $N$ is given by $S_{0}(S \otimes T)(\xi) \equiv\left\langle\xi, S_{0} S\right\rangle_{\rho_{N}} T$ and the right action of $M$ by $(S \otimes T) T_{0}(\xi) \equiv\langle\xi, S\rangle_{\rho_{N}} T T_{0}$, which are the natural choices.

If $M$ is a factor, let $A u t(M)$ denote the group of all $*$-automorphisms of $M$. Two such automorphisms $\theta_{1}$ and $\theta_{2}$ are said to be outer conjugate if there exists a unitary element $W \in M$ and $\phi \in \operatorname{Aut}(M)$ so that $\phi^{-1} \circ \theta_{1} \circ \phi=A d(W) \circ \theta_{2}$. We say that $\theta_{1}$ and $\theta_{2}$ are trivially outer conjugate if $\phi=i d$.

Example 38 Let $M$ be a factor of type $I I_{1}$ with faithful normal trace $\tau$ and let $\theta \in$ Aut $(M)$. Consider the Hilbert space $H_{\theta}=L^{2}(M, \tau)$. We define an $M-M$ bimodule structure on $H_{\theta}$ as follows. Let $S, T \in N$ and $\xi \in H_{\theta}$. Define the left and right actions by

$$
S \cdot \xi \cdot T=S\left(J \theta\left(T^{*}\right) J\right) \xi
$$

We claim that if $\theta_{1}$ and $\theta_{2}$ are distinct automorphisms of $M$, then $H_{\theta_{1}}$ is isomorphic to $H_{\theta_{2}}$ if and only if $\theta_{1}$ is trivially outer conjugate to $\theta_{2}$. If $\varphi: H_{\theta_{1}} \rightarrow H_{\theta_{2}}$ is an isomorphism of correspondences, then for $T \in M$ and $\xi \in H_{\theta_{1}}$ we have that

$$
\varphi\left(J \theta_{1}\left(T^{*}\right) J \xi\right)=J \theta_{2}\left(T^{*}\right) J \varphi(\xi)
$$

and hence that $J \theta_{1}\left(T^{*}\right) J=\varphi^{-1} J \theta_{2}\left(T^{*}\right) J \varphi$. Multiplying on the left and right by $J$ and using the facts that $J^{2}=I$ and $M$ is a self-adjoint algebra, we have that for all $T \in M$,

$$
\theta_{1}(T)=(J \varphi J)^{-1} \theta_{2}(T)(J \varphi J)
$$

Note that $J \varphi J \in M^{\prime \prime}$, since $\varphi^{*}$ (as well as $\varphi$ ) commutes with the left action of $M$, and hence lies in $M^{\prime} \subseteq \mathcal{B}\left(L^{2}(M, \tau)\right)$. Conversely, suppose that there is a unitary element $U \in M$ so that $\theta_{1}(T)=\theta_{2}\left(U^{*} T U\right)$ for every $T \in M$. We then have that the unitary
operator $J \theta_{2}\left(U^{*}\right) J: H_{\theta_{2}} \rightarrow H_{\theta_{1}}$ is an $M$-module isomorphism, since $\theta_{2}(U) \theta_{1}(T)=$ $\theta_{2}(T U)$

$$
\begin{aligned}
\xi \theta_{2}(T) \theta_{2}(U) & =\xi \theta_{2}(T U) \\
& =\xi \theta_{2}(U) \theta_{1}(T)
\end{aligned}
$$

and this action commutes with left multiplications.
From this example, we see that there are at least as many isomorphism classes of correspondences between a factor of type $I I_{1}$ and itself as there are trivially outer conjugacy classes of automorphisms of that factor.

### 4.1 Amenability and correspondences

### 4.1.1 Topology on correspondences

Unless otherwise explicitly noted, from now on assume that all $W^{*}$-algebras considered have separable predual. We now define a topology on the set of equivalence classes $\operatorname{Corr}(N, M)$ of separable correspondences between $W^{*}$-algebras $N$ and $M$. This topology is defined in a way analogous to the standard topology on the space of unitary representations of a group.

Definition 39 Let $\widehat{H}_{0} \in \operatorname{Corr}(N, M)$ and $\varepsilon>0$. Let $F \subseteq N, E \subseteq M$ and $X=$ $\left\{\xi_{1}, \ldots, \xi_{p}\right\} \subseteq H_{0}$ be finite sets. We define

$$
U\left(\widehat{H}_{0} ; \varepsilon, F, E, X\right) \subseteq \operatorname{Corr}(N, M)
$$

to be the set of classes of correspondences $\widehat{H} \in \operatorname{Corr}(N, M)$ for which there exists $\left\{\eta_{1}, \ldots, \eta_{n}\right\} \subseteq H$ such that

$$
\left|\left\langle S \xi_{i} T, \xi_{j}\right\rangle-\left\langle S \eta_{i} T, \eta_{j}\right\rangle\right|<\varepsilon
$$

for all $S \in F, T \in E$ and $1 \leq i, j \leq p$. The $U$-topology on $\operatorname{Corr}(N, M)$ is the one for which the sets $U$ form a basis of open neighborhoods.

We may define another topology on this class that turns out to be equivalent to the $U$-topology on $\operatorname{Corr}(N, M)$.

Definition 40 Let $\widehat{H}_{0} \in \operatorname{Corr}(N, M)$ and $\varepsilon>0$. Let $F \subseteq N, E \subseteq M$ and $X=$ $\left\{\xi_{1}, \ldots, \xi_{p}\right\} \subseteq H_{0}$ be finite sets. We define

$$
V\left(\widehat{H}_{0} ; \varepsilon, F, E, X\right) \subseteq \operatorname{Corr}(N, M)
$$

to be the set of classes of correspondences $\widehat{H} \in \operatorname{Corr}(N, M)$ for which there exists a correspondence $H_{1}$ in the class $\widehat{H}$ such that $H_{1}=H_{0}$ as a Hilbert space and such that if $S \cdot \xi \cdot T$ denotes the bimodule structure on $H_{1}$ (with $\xi \in H_{1}=H_{0}$ ) and $S \xi T$ the bimodule structure on $H_{0}$, then

$$
\|S \cdot \xi \cdot T-S \xi T\|<\varepsilon
$$

for all $S \in F, T \in E$ and $\xi \in X$. The $V$-topology on $\operatorname{Corr}(N, M)$ is the one for which the sets $V$ form a basis of open neighborhoods.

Proposition 41 The $U$ and $V$ topologies on $\operatorname{Corr}(N, M)$ are equivalent.

As a result of the above proposition, we shall not need to specify which topology on $\operatorname{Corr}(N, M)$ we are speaking of from now on. We now introduce an important notion for our consideration of amenability in terms of correspondences.

Definition 42 Let $H_{0}, H$ be equivalence classes of separable correspondences between $W^{*}$-algebras $N$ and $M$. We say that $H_{0}$ is weakly contained in $H$ if $H$ is in the closure of $H_{0}$, i.e. there exist bimodule structures $\cdot \lambda$ on $H_{0}$ such that in the $V$-topology picture

$$
\left\|T \cdot{ }_{\lambda} \xi \cdot{ }_{\lambda} S-T \xi S\right\| \rightarrow 0
$$

We write this as $H \Subset H_{0}$.

## Neighborhoods of $H_{i d}$

In this section we describe the neighborhoods of the identity correspondence in more detail. Let $M$ be a finite factor throughout this section. Given a positive real number $\varepsilon$ and a finite subset $E$ of $M$, we define $W(\varepsilon, E)$ to be the set of equivalence classes of correspondences $\widehat{H}$ between $M$ and itself such that there exists a unit vector $\xi \in H$ with the property that $\|T \xi-\xi T\|<\varepsilon$ for all $T \in E$. This is to say that $W(\varepsilon, E)$ is the set of classes of correspondences that have a vector which almost commutes with every element in the finite set $E$.

Proposition 43 The sets $W$ form a basis of neighborhoods of $\widehat{H}_{i d}$ in $\operatorname{Corr}(M, M)$.

Proof. Let $F$ be a finite subset of $M$ so that $I_{M} \in F$, and $\xi_{0}$ the trace vector in $L^{2}(M, \tau)=H_{i d}$. If $\widehat{H} \in V\left(\widehat{H}_{i d} ; \frac{\epsilon}{2}, F, F,\left\{\xi_{0}\right\}\right)$, then there exists a vector $\eta \in H$ such that $\left\|T_{1} \xi_{0} T_{2}-T_{1} \cdot \eta \cdot T_{2}\right\|<\frac{\varepsilon}{2}$ for all $T_{1}, T_{2} \in F$. Recall that $H=H_{i d}$ as a Hilbert space, but with different left and right actions. We use a "." to denote the actions in $H$. Since $I_{M} \in F$, we have that $\left\|T \xi_{0}-T \cdot \eta\right\|<\frac{\varepsilon}{2}$ and $\left\|\xi_{0} T-\eta \cdot T\right\|<\frac{\varepsilon}{2}$ for all $T \in F$, therefore by the fact that $\xi_{0}$ is a trace vector,

$$
\begin{aligned}
\|T \cdot \eta-\eta \cdot T\| & =\left\|T \cdot \eta-T \xi_{0}+\xi_{0} T-\eta \cdot T\right\| \\
& \leq\left\|T \cdot \eta-T \xi_{0}\right\|+\left\|\xi_{0} T-\eta \cdot T\right\| \\
& <\frac{\varepsilon}{2}+\frac{\varepsilon}{2}=\varepsilon,
\end{aligned}
$$

therefore $\widehat{H} \in W(\varepsilon, F)$. We have shown that $V\left(\widehat{H}_{i d} ; \frac{\varepsilon}{2}, F, F,\left\{\xi_{0}\right\}\right) \subseteq W(\varepsilon, F)$.
We shall now show that for any $\varepsilon>0$ and finite subset $F$ of $M$, there exist, $\varepsilon^{\prime}>0$ and a finite subset $F^{\prime}$ of $M$ such that $W\left(\varepsilon^{\prime}, F^{\prime}\right) \subseteq U\left(\widehat{H}_{i d} ; \varepsilon, F, F,\left\{\xi_{0}\right\}\right)$.

To this end, suppose that $\varepsilon$ and $F$ are given. The Dixmier approximation theorem tells us that for any $\varepsilon^{\prime}>0$, there exist unitary operators $U_{1}, \ldots, U_{m}$ in $M$ such that

$$
\left\|\frac{1}{m} \sum_{i=1}^{m} U_{i}^{*} S_{1} S_{2} U_{i}-\tau\left(S_{1} S_{2}\right) I_{M}\right\|<\varepsilon^{\prime}
$$

for every $S_{1}$ and $S_{2}$ in $F$. Define the set $F_{0}=F \cup\left\{U_{i}^{*} S_{1} S_{2}: S_{1}, S_{2} \in F\right\}_{i=1}^{m} \cup\left\{U_{i}\right\}_{i=1}^{m}$, and then $F^{\prime}=F_{0} \cup F_{0}^{*}$. Now suppose that $\widehat{H} \in W\left(\varepsilon^{\prime}, F^{\prime}\right)$. In this case we have that there exists a unit vector $\xi \in H$ so that $\|T \cdot \xi-\xi \cdot T\|<\varepsilon^{\prime}$ for all $T \in F^{\prime}$. If $T_{1}, T_{2} \in F^{\prime}$, then
(i)
(ii)

$$
\begin{aligned}
& \left|\left\langle T_{1} T_{2} \cdot \xi, \xi\right\rangle-\left\langle T_{2} T_{1} \cdot \xi, \xi\right\rangle\right| \\
& =\left|\left\langle T_{1} T_{2} \cdot \xi, \xi\right\rangle-\left\langle T_{2} \cdot \xi \cdot T_{1}, \xi\right\rangle+\left\langle T_{2} \cdot \xi \cdot T_{1}, \xi\right\rangle-\left\langle T_{2} T_{1} \cdot \xi, \xi\right\rangle\right| \\
& \leq\left|\left\langle T_{1} T_{2} \cdot \xi, \xi\right\rangle-\left\langle T_{2} \cdot \xi \cdot T_{1}, \xi\right\rangle\right|+\left|\left\langle T_{2} \cdot \xi \cdot T_{1}, \xi\right\rangle-\left\langle T_{2} \cdot\left(T_{1} \cdot \xi\right), \xi\right\rangle\right| \\
& =\left|\left\langle T_{1} T_{2} \cdot \xi, \xi\right\rangle-\left\langle T_{2} \cdot \xi \cdot T_{1}, \xi\right\rangle\right|+\left|\left\langle T_{2} \cdot\left(\xi \cdot T_{1}-T_{1} \cdot \xi\right), \xi\right\rangle\right| \\
& \leq\left|\left\langle T_{1} T_{2} \cdot \xi, \xi\right\rangle-\left\langle T_{2} \cdot \xi \cdot T_{1}, \xi\right\rangle\right|+\left\|T_{2}\right\| \varepsilon^{\prime} \\
& =\left|\left\langle T_{1} T_{2} \cdot \xi, \xi\right\rangle-\left\langle T_{2} \cdot \xi, \xi \cdot T_{1}^{* *}\right\rangle\right|+| | T_{2} \| \varepsilon^{\prime} \\
& =\left|\left\langle T_{1} T_{2} \cdot \xi, \xi\right\rangle-\left\langle T_{2} \cdot \xi \cdot T_{1}^{*}, \xi\right\rangle+\left\langle T_{2} \cdot \xi \cdot T_{1}^{*}, \xi\right\rangle-\left\langle T_{2} \cdot \xi, \xi \cdot T_{1}^{*}\right\rangle\right|+\left\|T_{2}\right\| \varepsilon^{\prime} \\
& \leq\left|\left\langle T_{1} T_{2} \cdot \xi, \xi\right\rangle-\left\langle T_{2} \cdot \xi \cdot T_{1}^{*}, \xi\right\rangle\right|+\left|\left\langle T_{2} \cdot \xi \cdot T_{1}^{*}, \xi\right\rangle-\left\langle T_{2} \cdot \xi, \xi \cdot T_{1}^{*}\right\rangle\right|+\left\|T_{2}\right\| \varepsilon^{\prime} \\
& =\left|\left\langle T_{1} T_{2} \cdot \xi, \xi\right\rangle-\left\langle T_{2} \cdot \xi \cdot T_{1}^{*}, \xi\right\rangle\right|+\left|\left\langle T_{2} \cdot \xi \cdot\left(T_{1}^{*}, \xi-\xi \cdot T_{1}^{*}\right)\right\rangle\right|+\left\|T_{2}\right\| \varepsilon^{\prime} \\
& \leq\left|\left\langle T_{1} T_{2} \cdot \xi, \xi\right\rangle-\left\langle T_{2} \cdot \xi \cdot T_{1}^{*}, \xi\right\rangle\right|+2\left\|T_{2}\right\| \varepsilon^{\prime} \\
& =2\left\|T_{2}\right\| \varepsilon^{\prime}
\end{aligned}
$$

Now we let $T_{1}=U_{i}^{*} S_{1} S_{2}$ and $T_{2}=U_{i}$ in (ii) to yield, for each $i=1,2, \ldots, m$

$$
\begin{aligned}
& \left|\left\langle U_{i}^{*} S_{1} S_{2} U_{i} \cdot \xi, \xi\right\rangle-\left\langle U_{i} U_{i}^{*} S_{1} S_{2} \cdot \xi, \xi\right\rangle\right| \\
& =\left|\left\langle U_{i}^{*} S_{1} S_{2} U_{i} \cdot \xi, \xi\right\rangle-\left\langle S_{1} S_{2} \cdot \xi, \xi\right\rangle\right| \\
& \leq 2 \varepsilon^{\prime} \mid\left\|U_{i}\right\|=2 \varepsilon^{\prime} \leq 2 \varepsilon^{\prime} \max _{T \in F^{\prime}}\|T\|=2 \varepsilon^{\prime} c .
\end{aligned}
$$

Note that

$$
\left|\sum_{i=1}^{m}\left\langle U_{i}^{*} S_{1} S_{2} U_{i} \cdot \xi, \xi\right\rangle-m\left\langle S_{1} S_{2} \cdot \xi, \xi\right\rangle\right| \leq \sum_{i=1}^{m}\left|\left\langle U_{i}^{*} S_{1} S_{2} U_{i} \cdot \xi, \xi\right\rangle-\left\langle S_{1} S_{2} \cdot \xi, \xi\right\rangle\right| \leq 2 m \varepsilon^{\prime} c
$$

and therefore

$$
\left|\frac{1}{m} \sum_{i=1}^{m}\left\langle U_{i}^{*} S_{1} S_{2} U_{i} \cdot \xi, \xi\right\rangle-\left\langle S_{1} S_{2} \cdot \xi, \xi\right\rangle\right| \leq 2 \varepsilon^{\prime} c
$$

From this follows

$$
\begin{aligned}
& \quad(i i i) \\
& \left|\tau\left(S_{1} S_{2}\right)-\left\langle S_{1} S_{2} \cdot \xi, \xi\right\rangle\right| \\
& \leq\left|\tau\left(S_{1} S_{2}\right)\left\langle I_{M} \cdot \xi, \xi\right\rangle-\frac{1}{m} \sum_{i=1}^{m}\left\langle U_{i}^{*} S_{1} S_{2} U_{i} \cdot \xi, \xi\right\rangle\right|+\left|\frac{1}{m} \sum_{i=1}^{m}\left\langle U_{i}^{*} S_{1} S_{2} U_{i} \cdot \xi, \xi\right\rangle-\left\langle S_{1} S_{2} \cdot \xi, \xi\right\rangle\right| \\
& \leq\left\|\frac{1}{m} \sum_{i=1}^{m} U_{i}^{*} S_{1} S_{2} U_{i}-\tau\left(S_{1} S_{2}\right) I_{M}\right\|+2 \varepsilon^{\prime} c \\
& \leq \varepsilon^{\prime}+2 \varepsilon^{\prime} c \\
& \leq \varepsilon^{\prime} c+2 \varepsilon^{\prime} c=3 \varepsilon^{\prime} c
\end{aligned}
$$

where the last inequality follows because there are unitary operators in $F^{\prime}$.
Combining (i) and (iii) with $\varepsilon^{\prime}=\frac{\varepsilon}{4 c}$, we obtain

$$
\begin{aligned}
\left|\left\langle T_{1} \cdot \xi \cdot T_{2}, \xi\right\rangle-\left\langle T_{1} \xi_{0} T_{2}, \xi_{0}\right\rangle\right| & \leq\|T\| \varepsilon^{\prime}+\left|\left\langle T_{1} T_{2} \cdot \xi, \xi\right\rangle-\tau\left(T_{1} T_{2}\right)\right| \\
& \leq\|T\| \varepsilon^{\prime}+3 \varepsilon^{\prime} c \\
& \leq 4 \varepsilon^{\prime} c=\varepsilon
\end{aligned}
$$

for every $T_{1}$ and $T_{2}$ in $F^{\prime}$. It follows that $\widehat{H} \in U\left(\widehat{H}_{i d} ; \varepsilon, F, F,\left\{\xi_{0}\right\}\right)$ and the proposition is proven.

## Amenable groups

A locally compact group $G$ is said to be amenable if there exists a finitely additive measure $\mu$ on the Borel subsets of $G$ which is invariant under the left translation action of $G$ on itself and satisfies $\mu(G)=1$. It is a well-known fact that the amenability of a group can be characterized by its representation theory, in the sense that a group is amenable if and only if the left regular representation of the group weakly contains the trivial representation. We explain what is meant by this below. In what follows, $L^{2}(G)$ shall denote the Hilbert space of square-integrable functions with respect to normalized Haar measure on $G$. Recall that the left regular representation of $G$ on $L^{2}(G)$ is defined by $h f(g)=f\left(h^{-1} g\right)$ for all $g, h \in G$ and $f \in L^{2}(G)$.

Definition 44 The left regular representation of a locally compact group $G$ on the Hilbert space $L^{2}(G)$ is said to weakly contain the trivial representation if for any $\varepsilon>0$ and any compact subset $S \subseteq G$, there exists $v \in L^{2}(G)$ such that $\|v\|=1$ and

$$
|\langle s v, v\rangle-1|<\varepsilon
$$

for any $s \in S$.

Theorem 45 (Huliniski) A locally compact group $G$ is amenable if and only if the left regular representation of $G$ weakly contains the trivial representation.

Proof. (see A. Hulaniski, Means and Følner conditions on locally compact groups, Studia Math., 27 (1966), 87-104.)

## Correspondences and group representations

Let $N$ be a von Neumann algebra. If $M=N \rtimes G$ is the von Neumann algebra crossed product of $N$ by an action of a locally compact group $G$, then there is a canonical way to associate a correspondence between $M$ and itself to a given strong-operator continuous unitary representation $\pi$ of $G$ on a Hilbert space $H_{\pi}$.

Let $H=H_{\pi} \otimes L^{2}(M)$. We take the actions on $L^{2}(M)$ from the left and right regular representations of $M$ along with the action of $G$ on $H_{\pi}$ and define a bimodule action on $H$. The right action is given, for $T \in M, \xi \in L^{2}(M)$, and $\eta \in H_{\pi}$ by

$$
(\eta \otimes \xi) T=\eta \otimes(\xi T)
$$

The left action is given, for $S \in N$, by

$$
S(\eta \otimes \xi)=\eta \otimes(S \xi)
$$

and for $g \in G$ by

$$
g(\eta \otimes \xi)=(\pi(g) \eta) \otimes \xi
$$

Now that we have a canonical way to assign correspondences to group representations, let us look at some examples.

Example 46 Recall that the trivial representation of $G$ is the representation $\pi_{i d}$ : $G \rightarrow \mathbb{C}$ such that $\pi_{i d}(g)=1$ for all $g \in G$. We see that the correspondence canonically assigned to the trivial representation is the identity correspondence $\mathbb{C} \otimes L^{2}(M)=$ $L^{2}(M)=H_{i d}$.

Example 47 Now consider the left regular representation $\lambda: G \rightarrow L^{2}(G)$ such that $\lambda(g) f(h)=f\left(g^{-1} h\right)$ for all $g, h \in G$. In this case we get the correspondence $L^{2}(G) \otimes$ $L^{2}(N \rtimes G)$. If $N=\mathbb{C}$ and $G$ is a discrete group, then since the elements of $G$ form an orthonormal basis for the dense subspace $M$ of $L^{2}(M)$, we have that $L^{2}(G)=$ $L^{2}(\mathbb{C} G)=L^{2}(M)$. Under the assumption $N=\mathbb{C}$ we also obtain that the left action of $N$ on $L^{2}(G) \otimes L^{2}(N \rtimes G)$ becomes trivial. From the previous two sentences it follows that the correspondence canonically associated to the left regular representation is the coarse correspondence $L^{2}(M) \otimes L^{2}(M)=H_{c}$.

### 4.2 Correspondences and the amenability of von Neumann algebras

In this section we restrict our attention to type $I_{1}$ factors since our goal is to develop an invariant for type $I I_{1}$ factors which will measure the "amenability defect" of such a factor. The examples of the previous section suggest that the natural definition for amenability for a type $I I_{1}$ factor should shadow the representation-theoretic definition for groups we have given. This is to say that a von Neumann algebra $M$ should be called amenable if the identity correspondence is weakly contained in the coarse correspondence. Although this is only one of myriad equivalent characterizations for the amenable von Neumann algebras, it seems to provide the right point of view for bootstrapping notions of amenability defect for groups to get notions in the type $I I_{1}$ factor setting.

In this section we prove the equivalence of this notion of amenability with injectivity in the case of type $I I_{1}$ factors. This essentially recasts the celebrated theorem of Connes in the language of correspondences.

Definition $48 A$ von Neumann algebra $M$ is said to be amenable if the identity correspondence $H_{i d}$ of $M$ is weakly contained in the coarse correspondence $H_{c o}$ of $M$.

Recall that a von Neumann algebra $M \subseteq \mathcal{B}(\mathcal{H})$ is called injective if there is a norm-one Banach space projection (conditional expectation) of $\mathcal{B}(\mathcal{H})$ onto $M$. A special case of the famous theorem of Alain Connes states that a type $I I_{1}$ factor is injective if and only if the factor is hyperfinite. We now give a proof, due to Sorin Popa, that a type $I I_{1}$ factor is amenable if and only if it is hyperfinite. To prove the theorem, we cite part of Theorem 5.1 of [6], but do not prove it here.

Proposition 49 Let $M$ be a factor of type $I I_{1}$ with trace $\tau$ acting standardly on $\mathcal{H}$ $\left(=L^{2}(M, \tau)\right)$. The following are equivalent:
(i) $M$ is injective.
(ii) There is a hypertrace on $M$, that is a state $\Phi$ on $B(\mathcal{H})$ that restricts to a normal trace-state on $M$.
(iii) Given $\left\{x_{1}, x_{2}, \ldots, x_{n}\right\} \subset M$ and $\varepsilon>0$, there exists a nonzero finite-rank projection $e \in B(\mathcal{H})$ such that $\forall j \in\{1,2, \ldots, n\}$

$$
\left\|\left[x_{j}, e\right]\right\|_{H . S} \leq \varepsilon\|e\|_{H . S .} \text { and }\left|\tau\left(x_{j}\right)-\frac{\left\langle x_{j} e, e\right\rangle_{H . S}}{\langle e, e\rangle_{H . S .}}\right| \leq \varepsilon .
$$

Proposition 50 A type $I_{1}$ factor $M$ is amenable if and only if $M$ is injective.

Proof. Suppose that $M$ is injective. Given any finite subset $F \subseteq M$, by the result of Connes, there exist finite-rank projections $\eta_{n}$ on $L^{2}(M)$ so that

$$
\left\|T \eta_{n}-\eta_{n} T\right\|_{\text {H.S. }} \rightarrow 0
$$

for any $T \in F$. These finite rank projections are Hilbert-Schmidt operators on $L^{2}(M)$, and hence may be regarded as elements of the coarse correspondence $H_{c}$. As a result, we see that $H_{i d}$ is weakly contained in $H_{c}$.

Conversely, suppose that $M$ is amenable. Let $\Lambda$ denote the directed set of finite subsets of $M$. Using the hypothesis, find a net $\left\{\eta_{i}\right\}_{i \in \Lambda} \subseteq H_{c o}=L^{2}(M) \otimes L^{2}(M)$ of unit vectors so that $\left\|T \eta_{i}-\eta_{i} T\right\| \rightarrow 0$ for all $T \in M$. Let $\Phi(T)=\lim _{\Lambda}\left\langle T \eta_{i}, \eta_{i}\right\rangle$ for $T \in \mathcal{B}\left(L^{2}(M)\right)$ denote a Banach limit over $\Lambda$. It follows that $\Phi$ is a state on $\mathcal{B}\left(L^{2}(M)\right)$, and that given any unitary element $U \in M$, we have that

$$
\begin{aligned}
\Phi\left(U T U^{*}\right) & =\lim _{\Lambda}\left\langle U T U^{*} \eta_{i}, \eta_{i}\right\rangle \\
& =\lim _{\Lambda}\left\langle T U^{*} \eta_{i}, U^{*} \eta_{i}\right\rangle=\lim _{\Lambda}\left\langle T \eta_{i} U^{*}, \eta_{i} U^{*}\right\rangle \\
& =\lim _{\Lambda}\left\langle T \eta_{i}, \eta_{i}\right\rangle=\Phi(T)
\end{aligned}
$$

Hence $\Phi$ is a hypertrace on $M \subseteq \mathcal{B}\left(L^{2}(M)\right)$, and therefore $M$ is injective.

## Chapter 5

## Følner Invariants

In the previous chapter we saw how the condition in theorem 5.1 of [6] could be used to show that injectivity of a type $I I_{1}$ factor is equivalent to the amenability of that factor. In this chapter we will more closely examine this condition and then use it to define a nontrivial invariant for type $I I_{1}$ factors. The main task after having defined this number is to show that it is computable and can distinguish two non-hyperfinite factors from one another. The latter task will provide us with a lot of work in the future.

In [9], Følner used combinatorial methods to obtain a condition on a discrete group that is equivalent to amenability of the group. Later I. Namioka was able to obtain Følner's condition using functional analytic methods due to Day [18]. In his classification of injective factors, Alain Connes exploits an analogy between an invariant mean on a group and a hypertrace on a type $I I_{1}$ factor $M$ acting standardly on a Hilbert space $\mathcal{H}$. More specifically, Connes follows Namioka and applies Day's idea to the hypertrace of the $I_{1}$ factor to get a Folner-type condition for type $I I_{1}$ factors. This condition is satisfied for the factor $M$ if and only if $M$ is injective, that is, there is a norm-one Banach space projection from $B(\mathcal{H})$ onto $M$, vindicating the analogy between injectivity of factors of type $I I_{1}$ and amenability of discrete groups. Recently, group theorists have devised ways to measure the degree of non-amenability of a group. In particular Arzhantseva, Burillo, Lustig, Reeves, Short and Ventura [4] have discovered the notion of universal Folner invariant for a finitely generated group
$G$, denoted by $F \varnothing l(G)$, which for an $n$-generated group satisfies $0 \leq F \varnothing l(G) \leq \frac{2 n-2}{2 n-1}$. In particular, $F \varnothing l(G)=0$ whenever $G$ is amenable and $F \varnothing l(G)=\frac{2 n-2}{2 n-1}$ if and only if $G=\mathbb{F}_{n}$. This group invariant is based on the Følner condition, so one hopes that via the connection with functional analysis the universal Folner invariant has an analogue for type $I I_{1}$ factors. In this chapter, we develop such an analogue.

Heuristically, a space satisfies a Følner-type condition if it can be exhausted by a family $\left\{A_{n}\right\}$ of sets of finite volume, with boundaries $\left\{\partial A_{n}\right\}$ of finite volume, such that

$$
\begin{equation*}
\lim _{n \rightarrow \infty} \frac{\operatorname{vol}\left(\partial A_{n}\right)}{\operatorname{vol}\left(A_{n}\right)}=0 \tag{5.1}
\end{equation*}
$$

In the case of discrete groups, the volume of a subset is just its cardinality. Various notions of boundary have been considered for subsets of finitely generated discrete groups, among which are the Cheeger boundary, the interior boundary and the exterior boundary. Remarkably, the usual notion of amenability of a finitely generated discrete group $G$, i.e. the existence of an invariant mean on $G$, is equivalent to $G$ satisfying condition 5.1 for any of the competing definitions of boundary.

The notion of boundary of a subset of a finitely generated group $G$ generally depends on a given finite generating subset $X$. Arzhantseva, Burillo, Lustig, Reeves, Short and Ventura define

$$
\begin{equation*}
F ø l(G, X)=\inf _{\substack{A \subseteq G \\ \text { finite }}} \frac{\# \partial_{X} A}{\# A} \tag{5.2}
\end{equation*}
$$

where $\partial_{X} A=\left\{a \in A \mid a x \notin A\right.$ for some $\left.x \in X^{ \pm 1}\right\}$ is the interior boundary of $A$ with respect to $X$ in $G$. They go on to define the universal Følner invariant

$$
\begin{equation*}
F \varnothing l(G)=\inf _{X} F \varnothing l(G, X) \tag{5.3}
\end{equation*}
$$

where the infimum is taken over all finite generating subsets $X$ of $G$. If $F \varnothing l(G)=0$, the group $G$ is said to be weakly-amenable and if not, $G$ is said to be uniformly non-amenable.

Our motivation for considering these numbers for type $I_{1}$ factors comes from the fact that in the above group theory paper, the authors are able to prove that for $n \geq 665$ odd and $m>1, B(m, n)$ is uniformly non-amenable. Our hope is to use the corresponding invariant for type $I I_{1}$ factors to glean some information about Burnside factors.

### 5.1 Connes' Følner-type condition

From Theorem 2.5 in [6], the first condition in the following proposition is equivalent, without the stipulation that $M$ is finitely generated, to injectivity of a $I I_{1}$ factor $M$ with trace $\tau$ acting standardly on $\mathcal{H}\left(=L^{2}(M, \tau)\right)$. Let $\mathcal{U}(M)$ denote the group of unitary elements in $M$. Throughout this paper we only consider von Neumann algebras that can be generated by finitely many elements; we do this in order to consider the lower Følner number.

Proposition 51 Let $M$ be a factor of type $I I_{1}$ with trace $\tau$ acting standardly on $\mathcal{H}$ $\left(=L^{2}(M, \tau)\right)$. The following are equivalent:
i Given $\left\{x_{1}, x_{2}, \ldots, x_{n}\right\} \subset M$ and $\varepsilon>0$, there exists a nonzero finite-rank projection $e \in B(\mathcal{H})$ such that $\forall j \in\{1,2, \ldots, n\}$

$$
\left\|\left[x_{j}, e\right]\right\|_{H . S .} \leq \varepsilon\|e\|_{H . S .} \text { and }\left|\tau\left(x_{j}\right)-\frac{\left\langle x_{j} e, e\right\rangle_{H . S}}{\langle e, e\rangle_{H . S .}}\right| \leq \varepsilon .
$$

ii Given $\left\{U_{1}, U_{2}, \ldots, U_{n}\right\} \subset \mathcal{U}(M) M$ and $\varepsilon>0$, there exists a nonzero finite-rank projection $e \in B(\mathcal{H})$ such that $\forall j \in\{1,2, \ldots, n\}$

$$
\left\|\left[U_{j}, e\right]\right\|_{H . S .} \leq \varepsilon\|e\|_{H . S .} \text { and }\left|\tau\left(U_{j}\right)-\frac{\left\langle U_{j} e, e\right\rangle_{H . S}}{\langle e, e\rangle_{H . S}}\right| \leq \varepsilon .
$$

Proof. ( $i \Rightarrow i i$ ) Suppose $i$ holds and $i i$ follows directly.
( $i i \Rightarrow i$ ) Suppose $i i$ holds. Let $\left\{x_{1}, x_{2}, \ldots, x_{n}\right\} \subset M$ and $\varepsilon>0$. Each $x_{i}$ can be written as a linear combination of at most four unitary elements, as $x_{i}=\sum_{j=1}^{4} \lambda_{j}^{(i)} U_{j}^{(i)}$, with $\lambda_{j}^{(i)} \in \mathbb{C}$. Let $S=\bigcup_{i, j}\left\{U_{j}^{(i)}\right\} \subset \mathcal{U}(M)$. Let $C=\max \left\{\lambda_{j}^{(i)} \mid i \in\{1,2, \ldots, n\}\right.$; $j \in\{1,2,3,4\}\}$. By $i i$, given $\varepsilon^{\prime}=\frac{\varepsilon}{4 C}$ we can find a finite rank projection $e$ such that $\forall U \in S,\|[U, e]\|_{H . S .} \leq \varepsilon^{\prime}\|e\|_{H . S .}$ and $\left|\tau(U)-\frac{\langle U e, e\rangle_{H . S} \mid}{\langle e, e\rangle_{H . S}}\right| \leq \varepsilon^{\prime}$. Note that if we apply the triangle inequality

$$
\begin{aligned}
\left\|\left[x_{i}, e\right]\right\|_{H . S} & =\left\|\left(\sum_{j=1}^{4} \lambda_{j}^{(i)} U_{j}^{(i)}\right) e-e\left(\sum_{j=1}^{4} \lambda_{j}^{(i)} U_{j}^{(i)}\right)\right\|_{H . S .} \\
& =\left\|\sum_{j=1}^{4} \lambda_{j}^{(i)}\left(U_{j}^{(i)} e-e U_{j}^{(i)}\right)\right\|_{\text {H.S. }} \leq 4 C\left\|\left[U_{j}^{(i)}, e\right]\right\|_{\text {H.S. }} \leq 4 C \varepsilon^{\prime}\|e\|_{\text {H.S. }} \\
& =\varepsilon\|e\|_{\text {H.S. }}
\end{aligned}
$$

and

$$
\begin{aligned}
\left|\tau\left(x_{j}\right)-\frac{\left\langle x_{j} e, e\right\rangle_{H . S}}{\langle e, e\rangle_{H . S}}\right| & =\left|\tau\left(\sum_{j=1}^{4} \lambda_{j}^{(i)} U_{j}^{(i)}\right)-\frac{\left\langle\sum_{j=1}^{4} \lambda_{j}^{(i)} U_{j}^{(i)} e, e\right\rangle_{H . S .}}{\langle e, e\rangle_{H . S .}}\right| \\
& \leq 4 C\left|\tau\left(U_{j}^{(i)}\right)-\frac{\left\langle U_{j}^{(i)} e, e\right\rangle_{H . S}}{\langle e, e\rangle_{H . S .}}\right| \leq 4 C \varepsilon^{\prime} \\
& =\varepsilon .
\end{aligned}
$$

therefore $i$ holds.

Remark 52 For the remainder of this paper, we shall refer to property (ii) in the above theorem as the Connes-Følner condition.

Definition 53 Let $M$ be a von Neumann algebra, and $X \subset \mathcal{U}(M)$ be a finite subset of $M$. We define the property $Q(X, \varepsilon)$ to be "there exists a nonzero finiterank projection $e \in B(\mathcal{H})$ such that $\forall j \in\{1,2, \ldots, n\},\left\|\left[U_{j}, e\right]\right\|_{\text {H.S. }} \leq \varepsilon\|e\|_{\text {H.S. }}$ and $\left|\tau\left(U_{j}\right)-\frac{\left\langle U_{j e}, e\right\rangle_{\text {F.S. }}}{\langle e, e\rangle_{H . S}}\right| \leq \varepsilon . "$

Definition 54 Let $M$ be a von Neumann algebra, and $X \subset \mathcal{U}(M)$ be a finite subset of $M$. Define

$$
F \phi l(M, X)=\inf \{\varepsilon>0: Q(X, \varepsilon)\} .
$$

Remark 55 Note that for the element $I \in M$, we have that $[I, e]=0$ and $\tau(I)-$ $\frac{\langle I e, e\rangle_{H . S}}{\langle e, e\rangle_{H . S}}=0$, so we may disregard the element $I$ in the sets $X$ for which we consider $Q(X, \varepsilon)$. More specifically, $Q(X, \varepsilon)$ holds if and only if $Q(X \backslash\{I\}, \varepsilon)$ holds.

Definition 56 Let $M$ be a von Neumann algebra. We define the universal Følner invariant $F ø l(M)=\sup _{X} F \emptyset l(M, X)$, where the supremum is taken over all finite sets $X \subset \mathcal{U}(M)$.

Proposition $57 M$ is injective if and only if $F ø l(M)=0$.

Proof. By Theorem 2.5 of [6] and Proposition $51, F ø l(M, X)=0$ for all finite sets $X$ of unitary elements of $M$ if and only if $M$ is injective.

We now prove a monotonicity result.

Proposition 58 Let $M$ be a von Neumann algebra. If $X_{1}$ and $X_{2}$ are finite subsets of $\mathcal{U}(M)$ that generate $M$, and $X_{1} \subseteq X_{2}$, then $F \varnothing l\left(M, X_{1}\right) \leq F ø l\left(M, X_{2}\right)$.

Proof. We have that for any $\varepsilon>0$ that $Q\left(X_{2}, \varepsilon\right) \Rightarrow Q\left(X_{1}, \varepsilon\right)$, hence $\inf \{\varepsilon>0$ : $\left.Q\left(X_{1}, \varepsilon\right)\right\} \leq \inf \left\{\varepsilon>0: Q\left(X_{2}, \varepsilon\right)\right\}$.

Remark 59 Note that if the type $I I_{1}$ factor $M$ is not hyperfinite, then there is a finite subset $X$ for which $\{\varepsilon>0: \sqcap Q(X, \varepsilon)\}$ is not empty, and hence

$$
\inf \{\varepsilon>0: Q(X, \varepsilon)\}=\sup \{\varepsilon>0: \neg Q(X, \varepsilon)\}
$$

under these circumstances.

Suppose $L^{2}(M)$ is separable. For a positive operator $T \in B\left(L^{2}(M)\right)$, let $\operatorname{Tr}(T)=$ $\sum_{i=1}^{\infty}\left\langle T e_{i}, e_{i}\right\rangle$, where $\left\{e_{i}\right\}_{i=1}^{\infty}$ is any orthonormal basis for $L^{2}(M)$. The Hilbert-Schmidt norm of an operator $T \in B\left(L^{2}(M)\right)$ is given by $\|T\|_{H . S}=\operatorname{Tr}\left(T^{*} T\right)^{1 / 2}$. We say that $T \in B\left(L^{2}(M)\right)$ is in the Hilbert-Schmidt class when $\|T\|_{\text {H.S. }}<\infty$. The class of all such operators in $B\left(L^{2}(M)\right)$ may be regarded as a Hilbert space when equipped with the inner product $\langle A, B\rangle=\operatorname{Tr}\left(B^{*} A\right)$.

Proposition 60 Let $X$ be a finite subset of $U(M)$ and e a rank $l(<\infty)$ projection in $B\left(L^{2}(M)\right)$. For all $U \in X$,

$$
\frac{\|\left[U, e \|_{H . S}\right.}{\|e\|_{\text {H.S. }}}=\sqrt{2} \sqrt{1-\|e U e\|_{\tau_{l}}^{2}} .
$$

where $\|\cdot\|_{\tau_{l}}$ is the Hilbert-Schmidt norm on $M_{l}(\mathbb{C})$ with respect to the normalized trace $\tau_{l} \equiv \frac{1}{l} T r_{l}$ on $M_{l}(\mathbb{C})$.

Proof. First note that $\|e\|_{\text {H.S. }}^{2}=\langle e, e\rangle_{H . S .}=\left(\operatorname{Tr}\left(e^{*} e\right)^{1 / 2}\right)^{2}=\operatorname{Tr}(e)=l$. The following computation proves the proposition:

$$
\begin{aligned}
\|[U, e]\|_{H . S} & =\|U e-e U\|_{H . S .} \\
& =\left(\operatorname{Tr}\left((U e-e U)^{*}(U e-e U)\right)\right)^{1 / 2} \\
& =\left(\operatorname{Tr}\left(\left(e U^{*}-U^{*} e\right)(U e-e U)\right)\right)^{1 / 2} \\
& =\left(\operatorname{Tr}(e)-\operatorname{Tr}\left(U^{*} e U e\right)-\operatorname{Tr}\left(e U^{*} e U\right)+\operatorname{Tr}\left(U^{*} e U\right)\right)^{1 / 2} \\
& =\sqrt{2}\left(\operatorname{Tr}(e)-\operatorname{Tr}\left(e U^{*} e U e\right)\right)^{1 / 2} \\
& =\sqrt{2} \sqrt{l-\|e U e\|_{H . S .}^{2}}=\left(\sqrt{2} \sqrt{1-\|e U e\|_{\tau_{l}}^{2}}\right)\|e\|_{H . S .}
\end{aligned}
$$

where in the fourth equality we have used the fact that $\operatorname{Tr}$ is a trace (the reason this is justified may be found in Kadison and Ringrose vol. II).

Proposition 61 For any type $I I_{1}$ factor $M, F \phi l(M) \leq 2$.

Proof. First suppose that $X$ is a finite set of unitary elements in $M$, such that $\varepsilon>2$ and $\neg Q_{0}(X, \varepsilon)$ holds. In particular, $\varepsilon>\sqrt{2}$ under these circumstances. If $\sqrt{2} \sqrt{1-\|e U e\|_{\tau_{l}}^{2}}>\varepsilon$ then $\|e U e\|_{\tau_{l}}^{2}<0$, which cannot happen. It therefore must be that for every finite-rank projection $e$ in $B\left(L^{2}(M)\right)$, there exists $U \in X$ such that

$$
\left|\tau(U)-\tau_{l}(e U e)\right|>\varepsilon
$$

However, using the triangle and Cauchy-Schwartz inequalities,

$$
2<\varepsilon<\left|\tau(U)-\tau_{l}(e U e)\right| \leq|\tau(U)|+\left|\tau_{l}(e U e)\right| \leq 2
$$

a contradiction.

### 5.2 Group von Neumann algebras

Let $\left\{g_{i}\right\}_{i=1}^{\infty}=G$ be a countable I.C.C. discrete group, and $1 \in G$ be its identity element. In this section we shall explore $F \varnothing l\left(\mathcal{L}_{G}, X\right)$ for sets $X$ of unitary generators of the factor von Neumamn algebra $\mathcal{L}_{G}$ associated to $G$.

Given $g \in G$, let $L_{g} \in B\left(L^{2}(G)\right)$ be the operator defined by $L_{g} h=g h$ for $h \in G$. Assume $X=\left\{U_{1}, U_{2}, \ldots, U_{k}\right\}$ and that $U_{j}=\sum_{t \in G} \mu_{t}^{(j)} L_{t}$ for each $j \in\{1,2, \ldots, k\}$, where the sum is a strong-operator limit of finitely indexed sub-sums. Let $e$ be a finite-rank projection in $B\left(L^{2}(G)\right)$, and suppose that $\left\{\xi_{1}, \ldots, \xi_{l}\right\}$ is an orthonormal basis for the closure of the range of $e$ in $L^{2}(G)$. Define the operator $\xi_{i} \otimes \xi_{j} \in B\left(L^{2}(G)\right)$ so that for $v \in L^{2}(G),\left(\xi_{i} \otimes \xi_{j}\right) v=\left\langle v, \xi_{j}\right\rangle \xi_{i}$. It is straightforward to check that $e=\sum_{i=1}^{l} \xi_{i} \otimes \xi_{i}$. Furthermore, suppose that $\xi_{i}=\sum_{g \in G} \lambda_{g}^{(i)} g$, so that $\sum_{g \in G} \overline{\lambda_{g}^{(j)}} \lambda_{g}^{(i)}=\delta_{i j}$ for $i, j \in\{1,2, \ldots, l\}$. Lemma 62 For all $j \in\{1,2, \ldots, k\}$,

$$
\left\|e U_{j} e\right\|_{H . S .}^{2}=\left\|\left[\left\langle\xi_{q}, U_{j} \xi_{p}\right\rangle\right\rangle_{q, p=1}^{l}\right\|_{T r_{l}}^{2},
$$

where $\|\cdot\|_{r_{l}}$ is the Hilbert-Schmidt norm on $M_{l}(\mathbb{C})$ with respect to the (non-normalized) trace $\operatorname{Tr}_{l}$ on $M_{l}(\mathbb{C})$.

Proof. To begin with,

$$
\begin{aligned}
\operatorname{Tr}\left(U_{j}^{*} e U_{j} e\right) & =\sum_{i=1}^{\infty}\left\langle U_{j}^{*} e U_{j} e g_{i}, g_{i}\right\rangle=\sum_{i=1}^{\infty}\left\langle U_{j} e g_{i}, e U_{j} g_{i}\right\rangle \\
& =\sum_{i=1}^{\infty} \sum_{t \in G} \sum_{s \in G} \mu_{t}^{(j)} \overline{\mu_{s}^{(j)}}\left\langle L_{t} e g_{i}, e s g_{i}\right\rangle .
\end{aligned}
$$

However for any $h \in G$, we have

$$
\begin{aligned}
e h & =\sum_{j=1}^{l}\left(\xi_{j} \otimes \xi_{j}\right)(h)=\sum_{j=1}^{l}\left\langle h, \xi_{j}\right\rangle \xi_{j}=\sum_{j=1}^{l}\left\langle h, \sum_{g \in G} \lambda_{g}^{(j)} g\right\rangle \xi_{j} \\
& =\sum_{j=1}^{l} \sum_{g \in G} \overline{\lambda_{g}^{(j)}}\langle h, g\rangle \xi_{j}=\sum_{j=1}^{l} \overline{\lambda_{h}^{(j)}} \xi_{j} .
\end{aligned}
$$

Hence $e g_{i}=\sum_{p=1}^{l} \overline{\lambda_{g_{i}}^{(p)}} \xi_{p}$ and $e s g_{i}=\sum_{q=1}^{l} \overline{\lambda_{s g_{i}}^{(q)}} \xi_{q}$. It follows that

$$
\sum_{i=1}^{\infty}\left\langle U_{j} e g_{i}, e U_{j} g_{i}\right\rangle=\sum_{i=1}^{\infty} \sum_{t \in G} \sum_{s \in G} \sum_{p=1}^{l} \sum_{q=1}^{l} \mu_{t}^{(j)} \overline{\mu_{s}^{(j)} \lambda_{g_{i}}^{(p)}} \lambda_{s g_{i}}^{(q)}\left\langle L_{t} \xi_{p}, \xi_{q}\right\rangle
$$

But

$$
\left\langle L_{t} \xi_{p}, \xi_{q}\right\rangle=\left\langle L_{t} \sum_{h \in G} \lambda_{h}^{(p)} h, \sum_{v \in G} \lambda_{v}^{(q)} v\right\rangle=\sum_{h \in G v \in G} \sum_{h} \lambda^{(p)} \overline{\lambda_{v}^{(q)}}\langle t h, v\rangle=\sum_{h \in G} \lambda_{h}^{(p)} \overline{\lambda_{t h}^{(q)}},
$$

and therefore

$$
\begin{aligned}
& \sum_{i=1}^{\infty} \sum_{t \in G} \sum_{s \in G} \sum_{p=1}^{l} \sum_{q=1}^{l} \mu_{t}^{(j)} \overline{\mu_{s}^{(j)} \lambda_{g_{i}}^{(p)}} \lambda_{s g_{i}}^{(q)}\left\langle L_{t} \xi_{p}, \xi_{q}\right\rangle \\
& =\sum_{i=1}^{\infty} \sum_{t \in G} \sum_{s \in G} \sum_{p=1}^{l} \sum_{q=1}^{l} \sum_{h \in G} \mu_{t}^{(j)} \overline{\mu_{s}^{(j)} \lambda_{g_{i}}^{(p)}} \lambda_{s g_{i}}^{(q)} \lambda_{h}^{(p)} \overline{\lambda_{t h}^{(q)}} \\
& =\sum_{g \in G} \sum_{t \in G} \sum_{s \in G} \sum_{p=1}^{l} \sum_{q=1}^{l} \sum_{h \in G} \mu_{t}^{(j)} \overline{\mu_{s}^{(j)} \lambda_{g}^{(p)}} \lambda_{s g}^{(q)} \lambda_{h}^{(p)} \overline{\lambda_{t h}^{(q)}}
\end{aligned}
$$

Changing the order of summation,

$$
\begin{aligned}
& \sum_{g \in G} \sum_{t \in G} \sum_{s \in G} \sum_{p=1}^{l} \sum_{q=1}^{l} \sum_{h \in G} \mu_{t}^{(j)} \overline{\mu_{s}^{(j)}} \lambda_{g}^{(p)} \lambda_{s g}^{(q)} \lambda_{h}^{(p)} \overline{\lambda_{t h}^{(q)}} \\
& =\sum_{p=1}^{l} \sum_{q=1}^{l}\left(\overline{\sum_{h \in G} \sum_{t \in G} \overline{\mu_{t}^{(j)}} \lambda_{h}^{(p)}} \lambda_{t h}^{(q)}\right)\left(\sum_{g \in G} \sum_{s \in G} \overline{\mu_{s}^{(j)} \lambda_{g}^{(p)}} \lambda_{s g}^{(q)}\right) \\
& =\sum_{p=1}^{l} \sum_{q=1}^{l}\left|\sum_{h \in G} \sum_{t \in G}^{\mu_{t}^{(j)} \lambda_{h}^{(p)}} \lambda_{t h}^{(q)}\right|^{2} \\
& =\sum_{p=1}^{l} \sum_{q=1}^{l}\left|\left\langle\xi_{q}, U_{j} \xi_{p}\right\rangle\right|^{2}
\end{aligned}
$$

since

$$
\begin{aligned}
\sum_{h \in G} \sum_{t \in G} \overline{\mu_{t}^{(j)} \lambda_{h}^{(p)}} \lambda_{t h}^{(q)} & =\left\langle\left(\sum_{t \in G} \overline{\mu_{t}^{(j)}} L_{t^{-1}}\right) \sum_{h \in G} \lambda_{h}^{(q)} h, \sum_{g \in G} \lambda_{g}^{(p)} g\right\rangle \\
& =\left\langle U_{j}^{*} \xi_{q}, \xi_{p}\right\rangle=\left\langle\xi_{q}, U_{j} \xi_{p}\right\rangle
\end{aligned}
$$

Proposition 63 For all $j \in\{1,2, \ldots, k\}$,

$$
\left\|\left[U_{j}, e\right]\right\|_{H . S .}=\sqrt{2} \sqrt{l-\left\|\left[\left\langle\xi_{q}, U_{j} \xi_{p}\right\rangle\right]_{q, p=1}^{l}\right\|_{T r_{l}}^{2}}
$$

Proof. We combine the previous lemma with an earlier result, and this is immediate.

Remark 64 Suppose for a moment that $G$ is generated by $S=\left\{g_{1}, g_{2}, \ldots, g_{m}\right\}$ In the above proof, if we consider the finite rank projection $e=g_{1} \otimes g_{1}+\ldots+g_{k} \otimes g_{k}$ and the unitary element $L_{g_{j}}$, we compute $\operatorname{Tr}\left(L_{g_{j}^{-1}} e L_{g_{j}} e\right)=\# S_{1}^{(j)}$, where

$$
S_{1}^{(j)}=\left\{g_{i} \in\left\{g_{1}, g_{2}, \ldots, g_{m}\right\}: g_{j} g_{i} \in\left\{g_{1}, g_{2}, \ldots, g_{m}\right\}\right\}
$$

We obtain that $\sqrt{2}\left(\operatorname{Tr}(e)-\operatorname{Tr}\left(U_{j}^{*} e U_{j} e\right)\right)^{1 / 2}=\sqrt{2}\left(m-\# S_{1}^{(j)}\right)^{1 / 2}$ in this case. Defining $S_{2}^{(j)} \subseteq\left\{g_{1}, \ldots, g_{m}\right\}$ as

$$
S_{2}^{(j)}=\left\{g_{i} \in\left\{g_{1}, g_{2}, \ldots, g_{m}\right\}: g_{j} g_{i} \in g_{j} S \Delta S\right\}
$$

since $S=S_{1}^{(j)} \dot{\cup} S_{2}^{(j)}$ we get that $\sqrt{2}\left(m-\# S_{1}^{(j)}\right)^{1 / 2}=\sqrt{2} \sqrt{\# S_{2}^{(j)}}=\sqrt{2} \sqrt{\frac{\# S_{2}^{(j)}}{\# S}}\|e\|_{H . S .}$. Note that

$$
\bigcup_{j=1}^{k} S_{2}^{(j)}=\left\{g_{i} \in\left\{g_{1}, g_{2}, \ldots, g_{m}\right\}: \exists g_{j} \in\left\{g_{1}, g_{2}, \ldots, g_{m}\right\} \text { s.t. } g_{j} \in g_{j} S \Delta S\right\}=\partial_{\left\{g_{1}, \ldots, g_{k}\right\}} S
$$

the (left) interior boundary of $S$. This is a relation of our concept with the classical Følner sets.

Proposition 65 For all $j \in\{1,2, \ldots, k\},\left|\tau\left(U_{j}\right)-\frac{\left\langle U_{j e, e}\right\rangle_{H . S}}{\langle e, e\rangle_{H . S}}\right|=\left|\tau\left(U_{j}\right)-\tau_{l}\left(\left[\left\langle\xi_{q}, U_{j}^{*} \xi_{p}\right\rangle\right]_{q, p=1}^{l}\right)\right|$, where $\tau_{l}=\frac{1}{l} \operatorname{Tr}_{l}$, the normalized trace on $M_{l}(\mathbb{C})$.

Proof. We show that $\frac{\left\langle U_{j} e, e\right\rangle_{H . S}}{\langle e, e\rangle_{H . S}}=\tau_{l}\left(\left[\left\langle\xi_{q}, U_{j}^{*} \xi_{p}\right\rangle\right]_{q, p=1}^{l}\right)$. Note that

$$
\langle e, e\rangle_{H . S .}=\left(\operatorname{Tr}\left(e^{*} e\right)^{1 / 2}\right)^{2}=\operatorname{Tr}(e)=l
$$

and that

$$
\begin{aligned}
\left\langle U_{j} e, e\right\rangle_{H . S} & =\operatorname{Tr}\left(e^{*} U_{j} e\right) \\
& =\operatorname{Tr}\left(U_{j} e\right)=\sum_{i=1}^{\infty}\left\langle U_{j} e g_{i}, g_{i}\right\rangle \\
& =\sum_{i=1}^{\infty}\left\langle U_{j} \sum_{k=1}^{l} \overline{\lambda_{g_{i}}^{(k)}} \xi_{k}, g_{i}\right\rangle \\
& =\sum_{i=1}^{\infty} \sum_{k=1}^{l} \overline{\lambda_{g_{i}}^{(k)}}\left\langle U_{j} \xi_{k}, g_{i}\right\rangle \\
& =\sum_{i=1}^{\infty} \sum_{k=1}^{l} \overline{\lambda_{g_{i}}^{(k)}}\left\langle U_{j} \xi_{k}, g_{i}\right\rangle \\
= & \sum_{i=1}^{\infty} \sum_{k=1}^{l} \overline{\lambda_{g_{i}}^{(k)}}\left\langle U_{j} \sum_{g \in G} \lambda_{g}^{(k)} g, g_{i}\right\rangle \\
= & \sum_{i=1}^{\infty} \sum_{k=1}^{l} \sum_{g \in G} \lambda_{g}^{(k)} \overline{\lambda_{g_{i}}^{(k)}}\left\langle U_{j} g, g_{i}\right\rangle
\end{aligned}
$$

$$
\begin{aligned}
& =\sum_{i=1}^{\infty} \sum_{k=1}^{l} \sum_{g \in G} \sum_{t \in G} \mu_{t}^{(j)} \lambda_{g}^{(k)} \overline{\lambda_{g_{i}}^{(k)}}\left\langle t g, g_{i}\right\rangle \\
& =\sum_{g \in G} \sum_{k=1}^{l} \sum_{t \in G} \mu_{t}^{(j)} \lambda_{i^{-1} g}^{(k)} \overline{\lambda_{g}^{(k)}} \\
& =\sum_{k=1}^{l} \sum_{g \in G} \sum_{t \in G} \mu_{t}^{(j)} \lambda_{t^{-1} g}^{(k)} \overline{\lambda_{g}^{(k)}}=\sum_{k=1}^{1}\left\langle\xi_{k}, U_{j}^{*} \xi_{k}\right\rangle,
\end{aligned}
$$

since

$$
\begin{aligned}
\sum_{g \in G} \sum_{t \in G} \mu_{t}^{(j)} \lambda_{t^{-1} g}^{(k)} \overline{\lambda_{g}^{(k)}} & =\left\langle\left(\sum_{t \in G} \mu_{t}^{(j)} L_{t}\right) \sum_{h \in G} \lambda_{h}^{(k)} h, \sum_{t \in G} \lambda_{t}^{(k)} t\right\rangle \\
& =\left\langle U_{j} \xi_{k}, \xi_{k}\right\rangle=\left\langle\xi_{k}, U_{j}^{*} \xi_{k}\right\rangle
\end{aligned}
$$

Corollary 66 For all $j \in\{1,2, \ldots, k\}$,

$$
\left|\tau\left(U_{j}\right)-\frac{\left\langle U_{j} e, e\right\rangle_{H . S}}{\langle e, e\rangle_{H . S .}}\right|=\left|\tau\left(U_{j}\right)-\overline{\tau_{l}\left(\left[\left\langle\xi_{q}, U_{j} \xi_{p}\right\rangle\right]_{q, p=1}^{l}\right)}\right|
$$

where $\tau_{l}=\frac{1}{l} \operatorname{Tr}_{l}$, the normalized trace on $M_{l}(\mathbb{C})$.
Proof. Since $a_{i j}=\left\langle\xi_{j}, U^{*} \xi_{i}\right\rangle=\overline{\left\langle\xi_{i}, U \xi_{j}\right\rangle}=\overline{b_{j i}}$, we have that $\operatorname{Tr}_{l}\left(\left[a_{i j}\right]_{i, j}\right)=\sum_{i=1}^{l} a_{i i}=$ $\sum_{i=1}^{l} \overline{b_{i i}}=\overline{\sum_{i=1}^{l} b_{i i}}=\overline{\operatorname{Tr}_{l}\left(\left[b_{i j}\right]_{i, j}\right)}$.
Remark 67 Let $\varepsilon>0$. Since $\|e\|_{\text {H.S. }}=\sqrt{l} \neq 0$, we have that

$$
\begin{gathered}
\frac{\sqrt{2} \sqrt{l-\left\|\left[\left\langle\xi_{q}, U_{j} \xi_{p}\right\rangle\right]_{q, p=1}^{l}\right\|_{T r_{l}}^{2}}}{\sqrt{l}} \\
=\sqrt{2} \sqrt{1-\frac{1}{l}\left\|\left[\left\langle\xi_{q}, U_{j} \xi_{p}\right\rangle\right]_{q, p=1}^{l}\right\|_{T r_{l}}^{2}} \\
=\sqrt{2} \sqrt{\left.\left.1-\frac{1}{l} \operatorname{Tr}\left(\left(\left[\left\langle\xi_{q}, U_{j} \xi_{p}\right\rangle\right]_{q, p=1}^{l}\right)^{*([ }\left[\xi_{q}, U_{j} \xi_{p}\right\rangle\right]_{q, p=1}^{l}\right)\right)} \\
=\sqrt{2} \sqrt{1-\frac{1}{l}\left\|\left[\left\langle\xi_{q}, U_{j} \xi_{p}\right\rangle\right]_{q, p=1}^{l}\right\|_{\tau_{l}}^{2}},
\end{gathered}
$$

therefore $\left\|\left[U_{j}, e\right]\right\|_{\text {H.S. }} \leq \varepsilon\|e\|_{\text {H.S. }}$ if and ondy if

$$
\sqrt{2} \sqrt{1-\left\|\left[\left\langle\xi_{q}, U_{j} \xi_{p}\right\rangle\right]_{q, p=1}^{l}\right\|_{\tau_{l}}^{2}} \leq \varepsilon .
$$

We see also that $\left|\tau\left(U_{j}\right)-\frac{\left\langle U_{j e, e}\right\rangle_{H . S} .}{\langle e,\rangle_{H . S}}\right| \leq \varepsilon$ if and only if

$$
\left|\tau\left(U_{j}\right)-\overline{\tau_{l}\left(\left[\left\langle\xi_{q}, U_{j} \xi_{p}\right\rangle\right]_{q, p=1}^{l}\right)}\right| .
$$

This gives us the following definitions and proposition.

Definition 68 Given a finite set $X$ of unitary elements in a $I I_{1}$ factor $M$ acting standardly on $\mathcal{H}$, and $\varepsilon>0$, we define the property $Q_{0}(X, \varepsilon)$ to be "there exists a finite orthonormal set $\left\{\xi_{1}, \ldots, \xi_{l}\right\} \in \mathcal{H}$ such that $\forall U \in X$,

$$
\sqrt{2} \sqrt{1-\left\|\left[\left\langle\xi_{q}, U \xi_{p}\right\rangle\right]_{q, p=1}^{l}\right\|_{\tau_{l}}^{2}} \leq \varepsilon \text { and }\left|\tau(U)-\overline{\tau_{l}\left(\left[\left\langle\xi_{q}, U \xi_{p}\right\rangle\right]_{q, p=1}^{l}\right)}\right| \leq \varepsilon . "
$$

Remark 69 Since we may in general assume that $I \notin X$, we have in the case of a group von Neumann algebras where the generating set $X$ are group elements that $\tau\left(L_{g_{i}}\right)=0$ for all $g_{i} \in X$, and that property $Q_{0}(X, \varepsilon)$ reduces to "there exists a finite orthonormal set $\left\{\xi_{1}, \ldots, \xi_{l}\right\} \in \mathcal{H}$ such that $\forall U \in X$,

$$
\sqrt{2} \sqrt{1-\left\|\left[\left\langle\xi_{q}, U \xi_{p}\right\rangle\right]_{q, p=1}^{l}\right\|_{\tau_{l}}^{2}} \leq \varepsilon \text { and }\left|\tau_{l}\left(\left[\left\langle\xi_{q}, U \xi_{p}\right\rangle\right]_{q, p=1}^{l}\right)\right| \leq \varepsilon^{\prime \prime}
$$

Remark 70 Since $0 \leq \frac{\|U, e\|_{H . S}}{\|\epsilon\|_{H . S}}=\sqrt{2} \sqrt{1-\left\|\left[\left\langle\xi_{q}, U \xi_{p}\right\rangle\right]_{q, p=1}^{l}\right\|_{\tau_{i}}^{2}}$, we have that $0 \leq 1-$ $\left\|\left[\left\langle\xi_{q}, U \xi_{p}\right\rangle\right]_{q, p=1}^{l}\right\|_{\tau_{l}}^{2}$, hence if $\varepsilon>0$, then $\sqrt{2} \sqrt{1-\left\|\left[\left\langle\xi_{q}, U \xi_{p}\right\rangle\right]_{q, p=1}^{l}\right\|_{\tau_{l}}^{2}} \leq \varepsilon$ is equivalent to $\left\|\left[\left\langle\xi_{q}, U \xi_{p}\right\rangle\right]_{q, p=1}^{l}\right\|_{\tau_{l}} \geq \sqrt{1-\frac{\varepsilon^{2}}{2}}$. We therefore may rewrite property $Q_{0}(X, \varepsilon)$ as "there exists a finite orthonormal set $\left\{\xi_{1}, \ldots, \xi_{l}\right\} \in \mathcal{H}$ such that $\forall U \in X$,

$$
\left\|\left[\left\langle\xi_{q}, U \xi_{p}\right\rangle\right]_{q, p=1}^{l}\right\|_{\tau_{l}} \geq \sqrt{1-\frac{\varepsilon^{2}}{2}} \text { and }\left|\tau(U)-\overline{\tau_{l}\left(\left[\left\langle\xi_{q}, U \xi_{p}\right\rangle\right]_{q, p=1}^{l}\right)}\right| \leq \varepsilon "
$$

Furthermore, imitating Voiculescu, we may define $\Omega_{R}^{0}\left(\left\{U_{1}, \ldots, U_{k}\right\}: l, n, \varepsilon\right)=\left\{\left(W_{1}\right.\right.$, $\left.W_{2}, \ldots, W_{k}\right) \in \mathcal{U}\left(M_{n}(\mathbb{C})\right)^{k}\left|R \leq\left\|W_{i}\right\|_{\tau_{n}} ;\left|\tau\left(U_{i_{1}} U_{i_{2}}, \ldots, U_{i_{j}}\right)-\overline{\tau_{n}\left(W_{i_{1}} W_{i_{2}}, \ldots, W_{i_{j}}\right)}\right| \leq \varepsilon\right.$
for $i_{j} \in\{1,2, \ldots, k\}$ and $\left.1 \leq j \leq l\right\}$. We may then rewrite property $Q_{0}(X, \varepsilon)$ as "there exists a finite orthonormal set $\left\{\xi_{1}, \ldots, \xi_{n}\right\} \in \mathcal{H}$ such that $\forall U \in X$,

$$
\left[\left\langle\xi_{q}, U \xi_{p}\right\rangle\right]_{q, p=1}^{n} \in \Omega_{\sqrt{1-\frac{\epsilon^{2}}{2}}}^{0}\left(\left\{U_{1}, \ldots, U_{k}\right\}: 1, n, \varepsilon\right)^{\prime \prime}
$$

Notice, that the quantity we have introduced here is not directly related to the $\Gamma_{R}$ sets of Voiculescu, we simply imitate his notation. We easily have the following proposition, and omit the proof.

Proposition 71 If $M$ is a type $I I_{1}$ factor associated to an I.C.C. discrete group, then for all finite sets $X$ of unitary elements that generate $M$ and all $\varepsilon>0$, property $Q(X, \varepsilon)$ is equivalent to property $Q_{0}(X, \varepsilon)$.

We now attempt to compute explicit lower bounds for $F \varnothing l\left(\mathcal{L}_{\mathbb{F}_{2}}, X\right)$, for various finite subsets $X$.

Remark 72 Let $X=\left\{L_{a}, L_{b}\right\}$ is the set of standard unitary generators of $\mathcal{L}_{\mathfrak{F}_{2}}$, If $\xi_{1}, \ldots, \xi_{k}$ is an orthonormal set in $L^{2}(M)$, then we know that if the quantity

$$
\frac{1}{k} \sum_{i, j=1}^{k}\left|\left\langle L_{a} \xi_{j}, \xi_{i}\right\rangle\right|^{2}
$$

is sufficiently close to 1 , then the corresponding quantity

$$
\frac{1}{k} \sum_{i, j=1}^{k}\left|\left\langle L_{b} \xi_{j}, \xi_{i}\right\rangle\right|^{2}
$$

is close to 0 and vice-versa. To see why, note that if we could write each $\xi_{i}=$ $\xi_{i}^{a}+\xi_{i}^{b}$, where $\xi^{a}$ denotes the part of $\xi$ with support $a \ldots$ and $\xi^{b}$ the part of $\xi$ with support b.... Notice that $\left\langle L_{a} \xi_{j}, \xi_{i}\right\rangle=\left\langle L_{a} \xi_{j}, \xi_{i}^{a}\right\rangle$, so that $\left|\left\langle L_{a} \xi_{j}, \xi_{i}\right\rangle\right|^{2} \leq\left\|\xi_{i}^{a}\right\|^{2}$. Since $\left\|\xi_{i}^{a}\right\|^{2}+\left\|\xi_{i}^{b}\right\|^{2}=1$, we have that if the $\xi_{i}$ are concentrated on the $\xi_{i}^{a}$ parts then the $\left\|\xi_{i}^{b}\right\|^{2}$ parts are small. It follows that if $\frac{1}{k} \sum_{i, j=1}^{k}\left|\left\langle L_{a} \xi_{j}, \xi_{i}^{a}\right\rangle\right|^{2}$ is large (near 1 ), then $\frac{1}{k} \sum_{i, j=1}^{k}\left|\left\langle L_{b} \xi_{j}, \xi_{i}^{b}\right\rangle\right|^{2} \leq C \max \left(\left\|\xi_{i}^{b}\right\|^{2}\right)$ will be small (near 0 ). This suggests a "balancing type" result should hold for the standard generators of $\mathcal{L}_{\mathbb{F}_{2}}$. The least
amount of information is obtained by the above reasoning when the "weights" $\left\|\xi_{i}^{a}\right\|^{2}$ and $\left\|\xi_{i}^{b}\right\|^{2}$ are all $\frac{1}{2}$.

Lemma 73 Let $\xi_{1}$ and $\xi_{2}$ be vectors in $L^{2}\left(\mathbb{F}_{2}\right)$, then $\left\langle L_{a} \xi_{1}, \xi_{2}\right\rangle=\left\langle L_{a} \xi_{1}, \xi_{2}^{a}\right\rangle+\left\langle\left(L_{a} \xi_{1}^{a^{-1}}\right)^{b}, \xi_{2}^{b}\right\rangle+$ $\left\langle\left(L_{a} \xi_{1}^{a^{-1}}\right)^{b^{-1}}, \xi_{2}^{b^{-1}}\right\rangle+\left\langle\left(L_{a} \xi_{1}^{a^{-1}}\right)^{e}, \xi_{2}^{e}\right\rangle$.

Proof. Let $S_{a} \subseteq \mathbb{F}_{2}$ be the set of reduced words in $\mathbb{F}_{2}$ that begin with the letter $a$. Similarly define $S_{a^{-1}}, S_{b}, S_{b^{-1}}$. Writing $\xi_{1}$ as $\xi_{1}^{a}+\xi_{1}^{b}+\xi_{1}^{a^{-1}}+\xi_{1}^{b^{-1}}+\xi_{1}^{e}$, we get that $L_{a}$ sends $\xi_{1}^{a}+\xi_{1}^{b}+\xi^{b^{-1}}+\xi_{1}^{e}$ to vectors with supports in $\overline{\operatorname{span}}\left(S_{a}\right)$, and $\xi_{1}^{a^{-1}}$ to vectors with supports in the orthocomplement of this set. This gives the desired decomposition.

Lemma 74 Let $\xi_{1}$ and $\xi_{2}$ be vectors in $L^{2}\left(\mathbb{F}_{2}\right)$, then $\left\langle L_{a^{-1}} \xi_{2}, \xi_{1}\right\rangle=\left\langle L_{a^{-1}} \xi_{2}, \xi_{1}^{a^{-1}}\right\rangle+$ $\left\langle\left(L_{a^{-1}} \xi_{2}^{a}\right)^{b}, \xi_{1}^{b}\right\rangle+\left\langle\left(L_{a^{-1}} \xi_{2}^{a}\right)^{b^{-1}}, \xi_{1}^{b^{-1}}\right\rangle+\left\langle\left(L_{a^{-1}} \xi_{2}^{a}\right)^{e}, \xi_{1}^{e}\right\rangle$.

Lemma 75 Let $\xi_{1}$ and $\xi_{2}$ be vectors in $L^{2}\left(\mathbb{F}_{2}\right)$, then

$$
\begin{aligned}
& \left\langle L_{a} \xi_{1}, \xi_{2}^{a}\right\rangle+\left\langle\left(L_{a} \xi_{1}^{a^{-1}}\right)^{b}, \xi_{2}^{b}\right\rangle+\left\langle\left(L_{a} \xi_{1}^{a^{-1}}\right)^{b^{-1}}, \xi_{2}^{b^{-1}}\right\rangle+\left\langle\left(L_{a} \xi_{1}^{a^{-1}}\right)^{e}, \xi_{2}^{e}\right\rangle \\
& =\left\langle L_{a} \xi_{1}^{a^{-1}}, \xi_{2}\right\rangle+\left\langle\xi_{1}^{b},\left(L_{a^{-1}} \xi_{2}^{a}\right)^{b}\right\rangle+\left\langle\xi_{1}^{b^{-1}},\left(L_{a^{-1}} \xi_{2}^{a}\right)^{b-1}\right\rangle+\left\langle\xi_{1}^{\epsilon},\left(L_{a^{-1}} \xi_{2}^{a}\right)^{e}\right\rangle
\end{aligned}
$$

Proof. Follows from the properties of the inner product, adjoint, and the previous two lemmas.

Lemma 76 Let $\xi_{1}$ and $\xi_{2}$ be vectors in $L^{2}\left(\mathbb{F}_{2}\right)$, then $\left\langle L_{a} \xi_{1}, \xi_{2}^{a}\right\rangle$ and $\left\langle L_{a} \xi_{1}^{a^{-1}}, \xi_{2}\right\rangle$ share only the term

$$
\left\langle L_{a} \xi_{1}^{a^{-1}}, \xi_{2}^{a}\right\rangle
$$

Proof. We have

$$
\left\langle L_{a} \xi_{1}, \xi_{2}^{a}\right\rangle=\left\langle L_{a} \xi_{1}^{a}, \xi_{2}^{a}\right\rangle+\left\langle L_{a} \xi_{1}^{b}, \xi_{2}^{a}\right\rangle+\left\langle L_{a} \xi_{1}^{a^{-1}}, \xi_{2}^{a}\right\rangle+\left\langle L_{a} \xi_{1}^{b^{-1}}, \xi_{2}^{a}\right\rangle+\left\langle L_{a} \xi_{1}^{e}, \xi_{2}^{a}\right\rangle
$$

and

$$
\left\langle L_{a} \xi_{1}^{a^{-1}}, \xi_{2}\right\rangle=\left\langle L_{a} \xi_{1}^{a^{-1}}, \xi_{2}^{a}\right\rangle+\left\langle L_{a} \xi_{1}^{a^{-1}}, \xi_{2}^{b}\right\rangle+\left\langle L_{a} \xi_{1}^{a^{-1}}, \xi_{2}^{a^{-1}}\right\rangle+\left\langle L_{a} \xi_{1}^{a^{-1}}, \xi_{2}^{b^{-1}}\right\rangle+\left\langle L_{a} \xi_{1}^{a^{-1}}, \xi_{2}^{e}\right\rangle
$$

Lemma 77 Let $\xi_{1}$ and $\xi_{2}$ be unit vectors in $L^{2}\left(\mathbb{F}_{2}\right)$, then

$$
\begin{aligned}
\left|\left\langle L_{a} \xi_{1}, \xi_{2}\right\rangle\right| & \leq\left\|\xi_{2}^{a}\right\|+\min \left(\left\|\left(L_{a} \xi_{1}^{a^{-1}}\right)^{b}\right\|,\left\|\xi_{2}^{b}\right\|\right)+ \\
& \min \left(\left\|\left(L_{a} \xi_{1}^{a^{-1}}\right)^{b^{-1}}\right\|,\left\|\xi_{2}^{b^{-1}}\right\|\right)+\min \left(\left\|\left(L_{a} \xi_{1}^{a^{-1}}\right)^{e}\right\|,\left\|\xi_{2}^{\epsilon}\right\|\right)
\end{aligned}
$$

Proof. This is an application of the triangle and Cauchy-Schwartz inequalities to the above decomposition, along with the fact that $L_{a}$ is unitary.

Lemma 78 Let $\xi_{1}$ and $\xi_{2}$ be unit vectors in $L^{2}\left(\mathbb{F}_{2}\right)$, then

$$
\left|\left\langle L_{a} \xi_{1}, \xi_{2}\right\rangle\right| \leq\left\|\xi_{2}^{a}\right\|+\left\|\xi_{2}^{b}\right\|+\left\|\xi_{2}^{b^{-1}}\right\|+\left\|\xi_{2}^{e}\right\|
$$

Proof. This is an easy consequence of the fact that $L_{a}$ is unitary, and the above lemma.

Lemma 79 Let $\xi_{1}$ and $\xi_{2}$ be unit vectors in $L^{2}\left(\mathbb{F}_{2}\right)$, then

$$
\frac{1}{4}\left|\left\langle L_{a} \xi_{1}, \xi_{2}\right\rangle\right|^{2} \leq\left\|\xi_{2}^{a}+\xi_{2}^{b}+\xi_{2}^{b^{-1}}+\xi_{2}^{e}\right\|^{2}
$$

Proof. We have that $\left\|\xi_{2}^{a}\right\|^{2}+\left\|\xi_{2}^{b}\right\|^{2}+\left\|\xi_{2}^{b-1}\right\|^{2}+\left\|\xi_{2}^{e}\right\|^{2}=\left\|\xi_{2}^{a}+\xi_{2}^{b}+\xi_{2}^{b^{-1}}+\xi_{2}^{e}\right\|^{2}$, since these are orthonormal pieces.

Remark 80 If we project all vectors onto the orthocomplement of the identity, these bounds all become smaller. We cannot do this without changing the property we consider, though.

Lemma 81 Let $\xi_{1}$ and $\xi_{2}$ be unit vectors in $L^{2}\left(\mathbb{F}_{2}\right)$, then

$$
\left\langle L_{a} \xi_{1}, \xi_{2}\right\rangle=\sum_{\varepsilon_{1}, \varepsilon_{2} \in\left\{a, a^{-1}, b, b^{-1}, e\right\}}\left\langle L_{a} \xi_{1}^{\varepsilon_{1}}, \xi_{2}^{\varepsilon_{2}}\right\rangle
$$

Proof. This is straightforward.

Lemma 82 Let conditions be as above, then

$$
\left|\left\langle L_{a} \xi_{1}^{b}, \xi_{2}^{a}\right\rangle\right| \leq\left\|\left|\xi_{1}^{b}\right|\right\| \mid \xi_{2}^{a} \| \leq \min \left(\left\|\xi_{1}^{b}\right\|,\left\|\xi_{2}^{a}\right\|\right) .
$$

Similarly for all other terms in the sum of the above lemma.

Proof. Note, since $L_{a}$ is unitary, it is an isometry, so that $\left\|L_{a} \xi_{1}^{b}\right\|=\left\|\xi_{1}^{b}\right\|$. The rest follows from the Cauchy-Schwartz inequality and the fact that both $\left\|\xi_{1}^{b}\right\|$ and $\left\|\xi_{2}^{a}\right\|$ are less than or equal to 1.

Lemma 83 Let $\xi_{1}$ and $\xi_{2}$ be unit vectors in $L^{2}\left(\mathbb{F}_{2}\right)$, then

$$
\begin{gathered}
\left|\left\langle L_{a} \xi_{1}, \xi_{2}\right\rangle\right|^{2}=25 \sum_{\varepsilon_{1}, \varepsilon_{2} \in\left\{a, a^{-1}, b, b^{-1}, e\right\}}\left|\left\langle L_{a} \xi_{1}^{\varepsilon_{1}}, \xi_{2}^{\varepsilon_{2}}\right\rangle\right|^{2} \text {, and } \\
\left|\left\langle L_{a} \xi_{1}, \xi_{2}\right\rangle\right|^{2} \leq 25 \sum_{\varepsilon_{1}, \varepsilon_{2} \in\left\{a, a^{-1}, b, b^{-1}, e\right\}}\left\|\xi_{1}^{\varepsilon_{1}}\right\|\left\|^{2}\right\| \xi_{2}^{\varepsilon_{2}} \|^{2}
\end{gathered}
$$

. Tighter bounds hold if we replace each eligible term like $\left\|\xi_{1}^{a^{-1}}\right\|^{2}\left\|\xi_{2}^{b}\right\|^{2} b y\left\|\left(L_{a} \xi_{1}^{a^{-1}}\right)^{b}\right\|^{2}\left\|\xi_{2}^{b}\right\|^{2}$, and we get looser bounds if we replace all terms on the RHS according to $\left\|\xi_{1}^{a}\right\|^{2}\left\|\xi_{2}^{a}\right\|^{2} \leq$ $\min \left(\left\|\xi_{1}^{a}\right\|,\left\|\xi_{2}^{a}\right\|\right)$.

Proof. Uses an elementary inequality, and the previous lemmas.

Definition 84 Let $M_{n}$ be finite factors with traces $\tau_{n}$, and let $\prod M_{n}$ denote their $C^{*}$-product, i.e. the $C^{*}$-algebra of uniformly norm-bounded sequences equipped with pointwise operations and the supremum norm. Let $\omega \in \beta \mathbb{N} \backslash \mathbb{N}$ be a free ultrafilter. Then

$$
\mathcal{I}_{\omega}=\left\{\left(A_{i}\right)_{i} \in \prod M_{n}: \lim _{i \rightarrow \omega} \tau_{i}\left(A_{i}^{*} A_{i}\right)=0\right\}
$$

is a closed two-sided ideal in $\Pi M_{n}$, and by a result of Sakai, the quotient $(\Pi M) / \mathcal{I}_{\omega}$ is a factor von Neumann algebra $\prod^{\omega} M_{n}$ with a faithful, normal trace $\tau_{\omega}$ defined by $\tau_{\omega}\left(\left(A_{i}\right)_{i}+\mathcal{I}_{\omega}\right)=\lim _{i \rightarrow \omega} \tau_{i}\left(A_{i}\right)$. The factor $\Pi^{\omega} M_{n}$ will be called an ultraproduct of the $M_{n}$ with respect to the free ultrafilter $\omega$, or simply an ultraproduct of the $M_{n}$.

Lemma 85 If $F ø l(M, X)=0$, then whenever $Q(X, \varepsilon)$ via a rank-n projection $e$, the following is true: for each $U \in X$ and $\delta>0$ there exists a unitary $n \times n$ matrix $W \in e B\left(L^{2}(M)\right) e$ such that

$$
\|e U e-W\|_{\tau_{n}} \leq \delta
$$

Proof. Suppose that $F \varnothing l(M, X)=0$. Thus for all $\varepsilon>0$, there exists $n \in \mathbb{N}$ so that $0<\frac{1}{n} \leq \varepsilon$ and $Q\left(X, \frac{1}{n}\right)$ and therefore there is an $e_{n}=\sum_{i=1}^{l(n)} \xi_{i}^{(n)} \otimes \xi_{i}^{(n)}$, where the $\left\{\xi_{i}^{(n)}\right\}_{i=1}^{l(n)}$ are an orthonormal system in $L^{2}(M)$, with

$$
0 \leq \frac{\left\|\left[U, e_{n}\right]\right\|_{H . S}}{\left\|e_{n}\right\|_{H . S .}}=\sqrt{2} \sqrt{1-\left\|e_{n} U e_{n}\right\|_{\tau_{l(n)}}^{2}} \leq \frac{1}{n}
$$

for all $U \in X$. With $e_{n} U e_{n}=A_{n}=\left[\left\langle\xi_{q}^{(n)}, U \xi_{p}^{(n)}\right\rangle\right]_{q, p=1}^{l(n)}$, we have

$$
1-\frac{1}{2 n^{2}} \leq \tau_{l(n)}\left(A_{n}^{*} A_{n}\right)=\tau_{l(n)}\left(A_{n} A_{n}^{*}\right)=\left\|e_{n} U e_{n}\right\|_{\tau_{l(n)}}^{2}
$$

Furthermore, since $e_{n}$ is a projection, $\left\|e_{n}\right\| \leq 1$ and hence

$$
\left\|A_{n}\right\|=\left\|e_{n} U e_{n}\right\| \leq\|U\|\left\|e_{n}\right\|^{2} \leq 1,
$$

and hence $\left\|A_{n} A_{n}^{*}\right\|=\left\|A_{n}\right\|^{2} \leq 1$. Let $\omega$ be a free ultrafilter, and $\prod^{\omega} M_{l(n)}(\mathbb{C})$ denote the ultraproduct factor as defined above. We have a sequence

$$
\left(A_{n}\right) \equiv\left\{A_{n} \mid n_{1}<n_{2} \Rightarrow l\left(n_{1}\right)<l\left(n_{2}\right)\right\}_{n=1}^{\infty}
$$

of matrices satisfying

$$
\begin{aligned}
\tau_{\omega}\left(\left(A_{n}^{*} A_{n}\right)+\mathcal{I}_{\omega}\right) & =\tau_{\omega}\left(\left(A_{n} A_{n}^{*}\right)+\mathcal{I}_{\omega}\right)=1 \\
& =\tau_{\omega}\left(\left(I_{n}\right)+\mathcal{I}_{\omega}\right)
\end{aligned}
$$

so by faithfulness of $\tau_{\omega}$ and the fact that $\left(I_{n}-A_{n} A_{n}^{*}\right)_{n} \geq 0$ for all $n$,

$$
\tau_{\omega}\left(\left(I_{n}-A_{n} A_{n}^{*}\right)_{n}+\mathcal{I}_{\omega}\right)=0
$$

so indeed $\left(A_{n}\right)$ represents a unitary element in the ultraproduct $\prod^{\omega} M_{l(n)}(\mathbb{C})$. Recall that if

$$
\left(A_{n}\right)+\mathcal{I}_{\omega} \text { and }\left(B_{n}\right)+\mathcal{I}_{\omega}
$$

represent distinct elements of $\prod^{\omega} M_{l(n)}(\mathbb{C})$, then the 2-norm distance between them is given by

$$
\begin{aligned}
& \left\|\left(A_{n}-B_{n}\right)+\mathcal{I}_{\omega}\right\|_{2} \\
& =\tau_{\omega}\left(\left(\left(A_{n}^{*}-B_{n}^{*}\right)+\mathcal{I}_{\omega}\right)\left(\left(A_{n}-B_{n}\right)+\mathcal{I}_{\omega}\right)\right)^{1 / 2} \\
& =\left[\lim _{l(n) \rightarrow \omega} \tau_{l(n)}\left(\left(A_{n}^{*}-B_{n}^{*}\right)\left(A_{n}-B_{n}\right)\right)\right]^{1 / 2} \\
& =\left[\lim _{l(n) \rightarrow \omega}\left\|A_{n}-B_{n}\right\|_{\tau_{l(n)}}^{2}\right]^{1 / 2} \\
& =\lim _{l(n) \rightarrow \omega}\left\|A_{n}-B_{n}\right\|_{\tau_{l(n)}} .
\end{aligned}
$$

Suppose that $\delta>0$ and that for every unitary $l(n) \times l(n)$ matrix $W_{n},\left\|A_{n}-W_{n}\right\|_{\tau_{l(n)}}>$ $\delta$, it then follows that $\left\|\left(A_{n}-W_{n}\right)+\mathcal{I}_{\omega}\right\|_{2}>\delta$ in $L^{2}\left(\prod^{\omega} M_{l(n)}(\mathbb{C}), \tau_{\omega}\right)$. Since every sequence $\left(W_{n}\right)$ represents a unitary element in $\prod^{\omega} M_{l(n)}(\mathbb{C})$, and every unitary element. is represented by such a sequence, a contradiction follows, since $\left(A_{n}\right)$ represents a unitary element in $\prod^{\omega} M_{l(n)}(\mathbb{C})$. Therefore, for all $\delta>0$ there exists a unitary $l(n) \times l(n)$ matrix $W_{n}$ so that $\left\|A_{n}-W_{n}\right\|_{\tau_{l(n)}} \leq \delta$, hence we may view $W_{n}$ as a unitary element of $e_{n} B\left(L^{2}(M)\right) e_{n}$ (that is, a unitary operator on $\operatorname{span}\left\{\xi_{i}^{(n)}\right\}_{i=1}^{l(n)} \cong \mathbb{C}^{l(n)}$ ).

We can, by a method of Ravichandran, prove the following theorem.
Theorem 86 If $X=\left\{L_{a}, L_{b}\right\}$ is the set of standard unitary generators of $\mathcal{L}_{\mathbb{F}_{2}}$, then $F \varnothing l\left(\mathcal{L}_{\mathbb{F}_{2}}, X\right)>0$.

Proof. Suppose that $F \emptyset l\left(\mathcal{L}_{\mathbb{F}_{2}}, X\right)=0$. It follows that $Q\left(X, \frac{1}{7}\right)$, so there exists a positive integer $l(=l(n))$ and a rank $l$ projection $e=\sum_{i=1}^{l} \xi_{i} \otimes \xi_{i}$, where the $\left\{\xi_{i}\right\}_{i=1}^{l}$ is an orthonormal system in $L^{2}\left(\mathbb{F}_{2}\right)$, such that for both $U \in X$

$$
0 \leq \frac{\|[U, e]\|_{H . S .}}{\|e\|_{F . S .}}=\sqrt{2} \sqrt{1-\|e U e\|_{\tau_{l}}^{2}} \leq \frac{1}{7}
$$

Henceforth let $U$ be the generator $L_{a}$ and $U^{\prime}$ the generator $L_{b}$. Let, $A=e U e$ and $U e=B$. We have that $B^{*} A=A^{*} A=A^{*} B$ and we may formally write

$$
\begin{aligned}
& \|e U e-U e\|_{\tau_{l}}^{2} \\
& =\|A-B\|_{\tau_{l}}^{2} \\
& =\tau_{l}\left(\left(A^{*}-B^{*}\right)(A-B)\right) \\
& =\tau_{l}\left(A^{*} A-B^{*} A-A^{*} B+B^{*} B\right) \\
& =\tau_{l}\left(A^{*} A\right)-\tau_{l}\left(B^{*} A\right) \\
& -\tau_{l}\left(A^{*} B\right)+\tau\left(B^{*} B\right) \\
& =1-\tau_{l}\left(A^{*} A\right)
\end{aligned}
$$

and since $\frac{97}{98} \leq\|e U e\|_{r_{l}}^{2}=\tau_{l}\left(A^{*} A\right)$, we have that

$$
\frac{1}{98} \geq 1-\tau_{l}\left(A^{*} A\right)=\|e U e-U e\|_{\tau_{i}}^{2}
$$

By the above lemma, there is an $l \times l$ unitary matrix $W \in e B\left(L^{2}\left(\mathbb{F}_{2}\right)\right) e$ such that

$$
\|A-W\|_{\tau_{n}} \leq\left(1-\frac{1}{\sqrt{2}}\right) \frac{1}{7}
$$

Now we have that $\|A-U e\|_{T r} \leq \frac{\sqrt{1}}{7 \sqrt{2}}$ and $\|A-W\|_{T r} \leq\left(1-\frac{1}{\sqrt{2}}\right) \frac{\sqrt{l}}{7}$, so that

$$
\|U e-W\|_{T_{r}} \leq \frac{\sqrt{l}}{7}
$$

so we may write formally,

$$
\|U e-W\|_{\pi_{b}}^{2} \leq \frac{1}{49}
$$

Note that since $e$ is the projection onto $\xi_{1}, \ldots, \xi_{l}$ and $W \in e B\left(L^{2}\left(\mathbb{F}_{2}\right)\right) e$ we have

$$
\|U e-W\|_{\tau_{i}}^{2}=\frac{1}{l} \sum_{i=1}^{l}\left\|(U e-W) \xi_{i}\right\|^{2}
$$

Suppose that $i \in\{1,2, \ldots, l\}$, and that $\xi_{i}=\sum_{g \in \mathbb{F}_{2}} \lambda_{g}^{(i)} g$, where we write $g$ in place of
$\chi_{\{g\}}$ to keep notation simpler, we obtain

$$
\begin{aligned}
\left\|(U e-W) \xi_{i}\right\|^{2} & =\left\|(U-W) \sum_{g \in \mathbb{F}_{2}} \lambda_{g}^{(i)} g\right\|^{2} \\
& =\sum_{g \in \mathbb{F}_{2}}\left|\lambda_{a^{-1} g}^{(i)}-\sum_{i=1}^{l(n)} w_{i k} \lambda_{g}^{(k)}\right|^{2}
\end{aligned}
$$

For $S$ a non-empty subset of $\mathbb{F}_{2}$ and $\eta=\sum_{g \in \mathbb{F}_{2}} \mu_{g} g \in L^{2}\left(\mathbb{F}_{2}\right)$, define $\|\eta\|_{S}^{2}=$ $\sum_{g \in \mathbb{S}}\left|\mu_{g}\right|^{2}$. It follows that

$$
\left\|(U-W) \xi_{i}\right\|_{S}^{2}=\sum_{g \in S}\left|\lambda_{a^{-1} g}^{(i)}-\sum_{k=1}^{l(n)} w_{i k} \lambda_{g}^{(k)}\right|^{2}
$$

We have that

$$
\begin{aligned}
\left|\left(\left\|U \xi_{i}\right\|_{S}-\left\|W \xi_{i}\right\|_{S}\right)\right| & \leq\left\|(U-W) \xi_{i}\right\|_{S} \\
& \leq\left\|(U-W) \xi_{i}\right\|
\end{aligned}
$$

and

$$
\begin{aligned}
& \left|\left(\left\|U \xi_{i}\right\|_{S}^{2}-\left\|W \xi_{i}\right\|_{S}^{2}\right)\right| \\
& \leq\left|\left(\left\|U \xi_{i}\right\|_{S}-\left\|W \xi_{i}\right\|_{S}\right)\right|\left(\left\|U \xi_{i}\right\|_{S}+\left\|W \xi_{i}\right\|_{S}\right) \\
& \leq 2\left|\left(\left\|U \xi_{i}\right\|_{S}-\left\|W \xi_{i}\right\|_{S}\right)\right| \\
& \leq 2\left\|(U-W) \xi_{i}\right\|_{S} \\
& \leq 2\left\|(U-W) \xi_{i}\right\|
\end{aligned}
$$

so that by the triangle inequality,

$$
\begin{aligned}
& \left.\left\lvert\, \frac{1}{l} \sum_{i=1}^{l}\left\|U \xi_{i}\right\|_{S}^{2}-\frac{1}{l} \sum_{i=1}^{l}\left\|W \xi_{i}\right\|_{S}^{2}\right.\right) \mid \\
& \leq \frac{1}{l} \sum_{i=1}^{l}\left|\left(\left\|U \xi_{i}\right\|_{S}^{2}-\left\|W \xi_{i}\right\|_{S}^{2}\right)\right| \\
& \leq \frac{2}{l} \sum_{i=1}^{l}\left\|(U-W) \xi_{i}\right\|
\end{aligned}
$$

$$
\begin{aligned}
& \left.\left\lvert\, \frac{1}{l} \sum_{i=1}^{l}\left\|U \xi_{i}\right\|_{S}^{2}-\frac{1}{l} \sum_{i=1}^{l}\left\|W \xi_{i}\right\|_{S}^{2}\right.\right)\left.\right|^{2} \\
& \leq \frac{4}{l^{2}}\left(\sum_{i=1}^{l}\left\|(U-W) \xi_{i}\right\|\right)^{2} \\
& \leq \frac{4 l}{l^{2}} \sum_{i=1}^{l}\left\|(U-W) \xi_{i}\right\|^{2} \\
& =\frac{4}{l} \sum_{i=1}^{l}\left\|(U-W) \xi_{i}\right\|^{2} \\
& =4\|U-W\|_{\tau_{l}}^{2} \leq \frac{4}{49}
\end{aligned}
$$

If $\eta=\sum_{g \in \mathbb{F}_{2}} \mu_{g} g \in L^{2}\left(\mathbb{F}_{2}\right)$, define $\left.\eta\right|_{S}=\sum_{g \in S} \mu_{g} g \Omega \in L^{2}(S)$. Note that $\left\|\left.\eta\right|_{S}\right\|_{L^{2}(S)}=$ $\|\eta\|_{S}$. We have that

$$
\begin{aligned}
\left.W \xi_{i}\right|_{S} & =\left.\sum_{k=1}^{l} w_{i k} \xi_{k}\right|_{S}=\sum_{g \in S}\left(\sum_{k=1}^{l} w_{i k} \lambda_{g}^{(k)}\right) g \\
& =\left.\left(W \xi_{i}\right)\right|_{S}
\end{aligned}
$$

so that

$$
\left[\begin{array}{ccc}
W & 0 & 0 \\
0 & \ddots & 0 \\
0 & 0 & W
\end{array}\right]\left[\begin{array}{c}
\left.\xi_{1}\right|_{S} \\
\vdots \\
\left.\xi_{l}\right|_{S}
\end{array}\right]=\sum_{g \in S}\left(W\left[\begin{array}{c}
\lambda_{g}^{(1)} \\
\vdots \\
\lambda_{g}^{(l)}
\end{array}\right]\right) g
$$

Since $W$ is a unitary operator on $\mathbb{C}^{l}$, we have that

$$
\begin{aligned}
\left\|\left[\begin{array}{ccc}
W & 0 & 0 \\
0 & \ddots & 0 \\
0 & 0 & W
\end{array}\right]\left[\begin{array}{c}
\left.\xi_{1}\right|_{S} \\
\vdots \\
\left.\xi_{l}\right|_{S}
\end{array}\right]\right\|_{L^{2}(S)^{l}}^{2} & =\sum_{g \in S}\left\|W\left[\begin{array}{c}
\lambda_{g}^{(1)} \\
\vdots \\
\lambda_{g}^{(l)}
\end{array}\right]\right\|_{\mathbb{C}^{l}}^{2} \\
& =\sum_{g \in S}\left\|\left[\begin{array}{c}
\lambda_{g}^{(1)} \\
\vdots \\
\lambda_{g}^{(l)}
\end{array}\right]\right\|_{\mathbb{C}^{l}}^{2} \\
& =\|\left[\begin{array}{c}
\left.\xi_{1}\right|_{S} \\
\vdots \\
\left.\xi_{l}\right|_{S}
\end{array}\right]
\end{aligned}
$$

We may conclude that

$$
\begin{aligned}
\sum_{i=1}^{l}\left\|W \xi_{i}\right\|_{S}^{2} & =\sum_{i=1}^{l}\left\|\left(W \xi_{i}\right)\right\|_{S} \|_{L^{2}(S)}^{2} \\
& =\sum_{i=1}^{l}\left\|\left.W \xi_{i}\right|_{S}\right\|_{L^{2}(S)}^{2} \\
& =\sum_{i=1}^{l}\left\|\left.\xi_{i}\right|_{S}\right\|_{L^{2}(S)}^{2}=\sum_{i=1}^{l}\left\|\xi_{i}\right\|_{S}^{2}
\end{aligned}
$$

We also have that for each $i$,

$$
\begin{aligned}
\left.\left(U \xi_{i}\right)\right|_{S} & =\left.\left(\sum_{g \in \mathbb{F}_{2}} \lambda_{a-1}^{(i)} g\right)\right|_{S} \\
& =\sum_{g \in S} \lambda_{a^{-1} g}^{(i)} g \\
& =\left.U \xi_{i}\right|_{S} .
\end{aligned}
$$

It follows that

$$
\begin{aligned}
\sum_{i=1}^{l}\left\|\left.U \xi_{i}\right|_{S}\right\|_{L^{2}(S)}^{2} & =\sum_{i=1}^{l}\left\|\left.\left(U \xi_{i}\right)\right|_{S}\right\|_{L^{2}(S)}^{2} \\
& =\sum_{i=1}^{l}\left\|U \xi_{i}\right\|_{S}^{2} .
\end{aligned}
$$

Notice that

$$
\begin{aligned}
\left\|\left.\left(U \xi_{i}\right)\right|_{S}\right\|_{\mathcal{L}^{2}(S)}^{2} & =\sum_{g \in S}\left|\lambda_{a^{-1} g}^{(i)}\right|^{2} \\
& =\sum_{g \in a^{-1} S}\left|\lambda_{g}^{(i)}\right|^{2} \\
& =\left\|\left.\xi_{i}\right|_{a^{-1} S}\right\|_{L^{2}\left(a^{-1} S\right)}^{2}=\left\|\xi_{i}\right\|_{a^{-1} S}^{2}
\end{aligned}
$$

We have that

$$
\begin{aligned}
\left|\frac{1}{l} \sum_{i=1}^{l}\left(\left\|\xi_{i}\right\|_{a^{-1} S}^{2}-\left\|\xi_{i}\right\|_{S}^{2}\right)\right| & = \\
\mid & \left.\frac{1}{l} \sum_{i=1}^{l}\left\|U \xi_{i}\right\|_{S}^{2}-\frac{1}{l} \sum_{i=1}^{l}\left\|\xi_{i}\right\|_{S}^{2}\right)\left.\right|^{2} \\
& \leq \frac{4}{49} .
\end{aligned}
$$

Now we shall choose a subset $S$ for which the above inequality will give us a contradiction. For simplicity of notation, let us define

$$
c_{S} \equiv \frac{1}{l} \sum_{i=1}^{l}\left\|\xi_{i}\right\|_{S}^{2}
$$

The above inequality becomes

$$
\left|c_{a-1}-c_{S}\right| \leq \frac{4}{49}
$$

If we carry out the above analysis using $U^{\prime}$ in place of $U$, we obtain

$$
\left|c_{b-1} S-c_{S}\right| \leq \frac{4}{49}
$$

Since $S$ was arbitrary, we could replace $S$ by $a S$ (resp. $b S$ ) to get

$$
\begin{gathered}
\left|c_{S}-c_{a S}\right| \leq \frac{4}{49} \\
\left(\text { resp. }\left|c_{S}-c_{b S}\right| \leq \frac{4}{49}\right)
\end{gathered}
$$

Choose the set $S$ to be all reduced words in $\mathbb{F}_{2}$ that begin with $a^{-1}$. Then $S \cup a S=\mathbb{F}_{2}$ and also $S, b S$ and $b^{-1} S$ are pairwise disjoint. Since $S \cup a S=\mathbb{F}_{2}$, we have that $c_{S}$
or $c_{a S}$ exceeds $\frac{1}{2}$. Since $S, b S$ and $b^{-1} S$ are pairwise disjoint, at least one of $c_{S}, c_{b S}$ or $c_{b-1} S$ must be $\leq \frac{1}{3}$. With no loss of generality, we may assume that $\frac{1}{2} \leq c_{a S}$. It follows that

$$
\frac{1}{2} \leq c_{a S}=\left|c_{S}\right| \leq\left|c_{S}-c_{a S}\right|+\left|c_{S}\right| \leq \frac{4}{49}+c_{S}
$$

so that

$$
\frac{1}{2}-\frac{4}{49} \leq c_{S}
$$

Let us assume, again with no loss of generality, that $c_{b S} \leq \frac{1}{3}$, then

$$
c_{S} \leq\left|c_{S}-c_{b S}\right|+c_{b S} \leq \frac{1}{3}+\frac{4}{49}
$$

It follows that

$$
\frac{5}{12}<\frac{1}{2}-\frac{4}{49} \leq c_{S} \leq \frac{1}{3}+\frac{4}{49}<\frac{5}{12}
$$

which is a contradiction.

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