# Scuola Normale Superiore di Pisa 

## Classe di Scienze

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# Joel E. Cohen <br> The arithmetic-geometric mean and its generalizations for noncommuting linear operators 

Annali della Scuola Normale Superiore di Pisa, Classe di Scienze 4 e série, tome 15, $\mathrm{n}^{\circ} 2$ (1988), p. 239-308<br>[http://www.numdam.org/item?id=ASNSP_1988_4_15_2_239_0](http://www.numdam.org/item?id=ASNSP_1988_4_15_2_239_0)

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# Echanqes Annales 

# The Arithmetic-Geometric Mean and Its Generalizations for Noncommuting Linear Operators 

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## Introduction

The arithmetic-geometric mean (to be defined in a moment) is the limit of an iterative process that operates recursively on pairs of positive real numbers. For over two centuries, an enormous amount of effort by some great mathematicians has been devoted to understanding and to generalizing the arithmetic-geometric mean. There have been two simple reasons why all this attention has been devoted to what is in essence a very humble idea. First, the limit has an important meaning or use that a priori could hardly be suspected from the definition of the iterative process. Specifically, the limit can be used to compute elliptic integrals, which are of substantial mathematical and scientific interest. Second, the iterative process converges to its limit with exceptional rapidity (quadratically-also to be defined later), so that very few iterative steps are required to approximate the limit very closely.

A large classical literature concerns generalizations of the arithmeticgeometric mean, or what could be called means and their iterations (see [6]). This paper concerns extensions of the arithmetic-geometric mean and of the classical generalizations from the case that the variables are real numbers to the case that the variables are linear operators. As in the case of positive numbers we are interested in three questions (for each generalization): the existence of a limit, the speed of convergence to the limit, and possible explicit formulas for the limit (for example, in terms of elliptic integrals). Though the machinery we have developed and the results we have obtained are substantial, as witnessed by the length of this paper, our success in achieving all three aims is not complete. First, we prove the existence of a limit for all the iterations we consider formally here. But for other interesting iterations, which appear to be plausible operator-theoretic generalizations of the arithmetic-geometric mean for

[^0]positive numbers, we observe numerically an apparent convergence to a limit but are unable to explain the observation mathematically. Hence we do not believe we have the last word on the existence of limits for generalizations of the arithmetic-geometric mean to linear operators. Second, for simplicity we establish a quadratic rate of convergence only for the "monster algorithm" considered in Section 3 below, although we believe a similar analysis can be used to determine, in all the other examples we treat, whether convergence is linear or quadratic. Third, we interpret the limiting linear operator in terms of elliptic integrals only for a small subset of the iterations whose limits we prove to exist. Even classically, explicit integral formulas for limits of iterated means are known only for a few examples which are very close to the arithmeticgeometric mean. However, in our case (see Section 4), we give a family, indexed by real numbers $\lambda \geq 1$, of reasonable definitions of the arithmetic-geometric mean of two linear operators $A$ and $B$; but we only obtain an explicit integral formula when $\lambda=1$.

On balance, the results of this paper are largely foundational: we prove the existence of limits for a wide variety of operator-theoretic generalizations, many apparently new, of the arithmetic-geometric mean. Though our success in finding explicit integral formulas for the limits is limited, it is possible that these results, and future extensions, will prove practically important for numerical algorithms to compute functions of matrices that can be derived from matrix elliptic integrals.

We now sketch the arithmetic-geometric mean and our results more precisely.

If $A$ and $B$ are positive real numbers, define a map $f$ by

$$
\begin{equation*}
f(A, B)=\left((A+B) / 2,(A B)^{1 / 2}\right) \tag{0.1}
\end{equation*}
$$

If $f^{k}$ denotes the $k$ th iterate of $f$, it is not hard to prove that there is a positive number $M=M(A, B)$ such that

$$
\lim _{k \uparrow \infty} f^{k}(A, B)=(M, M)
$$

The number $M$ is called the "arithmetic-geometric mean of $A$ and $B$ " or the " $A G M$ of $A$ and $B$." First Landen, then Lagrange and finally Gauss observed independently that

$$
\int_{0}^{\pi / 2}\left(A^{2} \cos ^{2} \Theta+B^{2} \sin ^{2} \Theta\right)^{-1 / 2} d \Theta=(\pi / 2)(M(A, B))^{-1}
$$

Lagrange and Legendre used this observation to compute elliptic integrals. Historical references to this work and to some of Gauss' deeper work on the $A G M$ can be found in [6] and [15].

An enormous literature concerning "means and their iterations" touches on a wide range of mathematics [6]. For examples, if $A$ and $B$ are positive reals, $0<\alpha, \beta<1$ and [11]

$$
\begin{equation*}
f(A, B)=\left((1-\alpha) A+\alpha B, A^{1-\beta} B^{\beta}\right) \tag{0.2}
\end{equation*}
$$

or $p>0, q>0$ and

$$
\begin{equation*}
f(A, B)=\left(\left(\left[A^{p}+B^{p}\right] / 2\right)^{1 / p}, \quad\left(\left[A^{q}+B^{q}\right] / 2\right)^{1 / q}\right) \tag{0.3}
\end{equation*}
$$

then

$$
\begin{equation*}
\lim _{k \uparrow \infty} f^{k}(A, B)=(M, M) \tag{0.4}
\end{equation*}
$$

where $M$ is a positive number depending on $A$ and $B$, and $\alpha, \beta$ or $p, q$. One can also study functions $f$ which are functions of $m$ variables, $m>2$, and try to prove analogues of (0.4). Borchardt [9] considered the map

$$
\begin{gather*}
f(A, B, C, D)=  \tag{0.5}\\
\left((A+B+C+D) / 4,\left((A B)^{1 / 2}+(C D)^{1 / 2}\right) / 2\right. \\
\left.\left((A C)^{1 / 2}+(B D)^{1 / 2}\right) / 2,\left((A D)^{1 / 2}+(B C)^{1 / 2}\right) / 2\right)
\end{gather*}
$$

for positive reals $A, B, C$ and $D$ and proved (this is the easy part of his work) that

$$
\begin{equation*}
\lim _{k \uparrow \infty} f^{k}(A, B, C, D)=(M, M, M, M) \tag{0.6}
\end{equation*}
$$

where $M$ is a positive number depending on $A, B, C$ and $D$. Many other examples are mentioned in Section 2 below.

A first goal of this paper is to describe reasonable analogues of the $A G M$ and its generalizations when all the variables are positive definite, bounded, selfadjoint linear operators on a Hilbert space. Abbreviating the phrase "positive definite, bounded, self-adjoint linear operator" to "positive definite operator", the first question is: what should be the analogue of $A^{1 / 2} B^{1 / 2}$ (for $A$ and $B$ positive reals) when $A$ and $B$ are positive definite opcrators? More generally, if $\sigma \in \mathbb{Z}^{m}$ satisfies $\sigma_{i}>0,1 \leq i \leq m$, and $\sum_{i=1}^{m} \sigma_{i} \quad 1$ and $A_{i}, 1 \leq i \leq m$, are positive reals, what is a reasonable analogue of $A_{1}^{\sigma_{1}} A_{2}^{\sigma_{2}} \cdots A_{m}^{\sigma_{m}}$ when the variables $A_{i}, \quad 1 \leq i \leq m$, are positive definite operators? We suggest that a reasonable analogue of $\prod_{i=1}^{m} A_{i}^{\sigma_{i}}$ is

$$
\begin{equation*}
\exp \left(\sum_{j=1}^{m} \sigma_{j} \log A_{j}\right) \tag{0.7}
\end{equation*}
$$

If, for positive definite operators $A_{j}, 1 \leq j \leq m$, and a real number $r \neq 0$ we define

$$
M_{r \sigma}\left(A_{1}, A_{2}, \cdots, A_{m}\right)=\left(\sum_{j=1}^{m} \sigma_{j} A_{j}^{r}\right)^{1 / r}
$$

one can prove that

$$
\lim _{r \rightarrow 0} M_{r \sigma}\left(A_{1}, A_{2}, \cdots, A_{m}\right)=\exp \left(\sum_{j=1}^{m} \sigma_{j} \log A_{j}\right)
$$

in the operator norm topology, so our suggestion dovetails nicely with certain reasonable means.

Using (0.7), we give operator-valued analogues of maps $f$ like those mentioned before, and we prove convergence of $f^{k}(A, B)$ in the strong operator topology. For example, a very special case of results in Section 2 is that if $0<\alpha, \beta<1$, and $A$ and $B$ are positive definite operators, and

$$
\begin{equation*}
f(A, B)=((1-\alpha) A+\alpha B, \exp ((1-\beta) \log A+\beta \log B)) \tag{0.8}
\end{equation*}
$$

then there exists a positive definite operator $E$ such that

$$
\lim _{k \uparrow \infty} f^{k}(A, B)=(E, E)
$$

The first two sections of this paper deal with the convergence of very general operator-valued versions of extensions of the $A G M$. In Section 1, we give results (Theorems 1.1 and 1.2) which enable one to prove convergence in the strong operator topology of certain sequences of $n$-tuples of positive definite linear operators. An example would be $\left(A_{k}, B_{k}\right)=f^{k}(A, B)$ with $f$ as in (0.8). The key idea in Sections 1 and 2 is to exploit the concavity of certain maps $A \rightarrow g(A)$, for positive definite $A$, and to use the beautiful classical theory of Loewner. In the applications in Section 2, we use only the concavity of the maps $A \rightarrow \log A$ and $A \rightarrow A^{p}(0<p<1)$ and the convexity of $A \rightarrow A^{-1}$; the full Loewner machinery is not needed.

The arguments simplify in the case of finite-dimensional matrices. Theorem 1.2, in particular, is not needed in the finite-dimensional case.

In Section 2, we use the convergence results of Section 1 to prove operatorvalued versions of convergence theorems for iterates of many classical means. Our convergence theorems suggest that our conventions were reasonable and provide an answer to the question raised in [6, p. 196] of how to extend the usual means to noncommuting variables. The maps $f$ we consider are not usually order-preserving, so the general convergence results in Section 4 of [26] (see [27] for a summary) are not applicable.

In Section 3, we extend the domain of $M(A, B)$, the operator-valued $A G M$ of $A$ and $B$, to pairs of bounded linear operators which are not necessarily positive definite and self-adjoint, and prove that $(A, B) \rightarrow M(A, B)$ is analytic.

The analogues of these questions are considered for a more general "monster algorithm" introduced in [6]. We also consider the commutative case ( $A B=B A$ ) and prove an integral formula for $M(A, B)$ analogous to that when $A$ and $B$ are real. The commutative case was also treated in [32], but the discussion there seems incomplete.

There are already numerous papers concerning operator-valued versions of the $A G M$ and other means. Section 4 of our paper displays the connection between our operator-valued definition of the $A G M$ and one introduced by Fujii [19] and Ando and Kubo [5]. We prove that the two definitions are in general different. However, there is a continuum of "reasonable" definitions of an $A G M$, parametrized by $\lambda \geq 1$, such that $\lambda=1$ corresponds to that of Fujii-Ando-Kubo and $\lambda=\infty$ corresponds to ours. For each $\lambda \geq 1$ and each pair of positive definite operators $A$ and $B$ there exists in the limit a positive definite operator $E_{\lambda}$ which is the $A G M$ of $A$ and $B$ for the algorithm corresponding to $\lambda$, and generally $E_{\lambda} \neq E_{\mu}$ if $\lambda \neq \mu$.

## 1. - Convergence criteria for sequences of linear operators

We recall some standard notation and results. If $X$ and $Y$ are complex Banach spaces, we denote by $\mathcal{L}(X, Y)$ the set of bounded complex linear operators from $X$ to $Y ; X^{*}=\mathcal{L}(X, \mathbb{C})$ will denote the continuous complex linear maps from $X$ to $\mathbb{C}$, the complex numbers. If $X=Y$, we shall write $\mathcal{L}(X)$ instead of $\mathcal{L}(X, X) . \mathcal{L}(X, Y)$ is a Banach space in the standard norm, $\|A\|=\sup \{\|A x\|: x \in X$ and $\|x\| \leq 1\}$. If $A \in \mathcal{L}(X, X), \sigma(A)$ will denote the spectrum of $A$, so

$$
\sigma(A)=\{\lambda \in \mathbb{C}: \lambda-A \text { is not one-one and onto }\}
$$

If $D$ is an open neighborhood of $\sigma(A)$ and $f: D \rightarrow \mathbb{C}$ is analytic, we define $f(A)$ in terms of Cauchy's integral formula:

$$
\begin{equation*}
f(A)=(2 \pi i)^{-1} \int_{\Gamma} f(z)(z-A)^{-1} d z \tag{1.1}
\end{equation*}
$$

where $\Gamma$ is a finite union of simple, closed rectifiable curves in $D$ which contains $\sigma(A)$ in its interior. An exposition of the basic results about this functional calculus can be found in [17], [33] or [36].

Recall that if $A$ denotes the algebra of functions which are analytic on an open neighborhood of $\sigma(A)$, then the map $f \rightarrow f(A)$ defined by (1.1) is an algebra homomorphism and $\sigma(f(A))=f(\sigma(A))$. If $g$ is analytic on an open neighborhood of $\sigma(B)$, where $B=f(A)$, then $g(B)=(\mathrm{g} \circ \mathrm{f})(A)$.

If $X$ and $Y$ are real Banach spaces, $\mathcal{L}(X, Y)$ denotes the bounded, real linear maps from $X$ to $Y$. If $A \in \mathcal{L}(X)$ and $\tilde{X}$ denotes the complexification of $X$, then $A$ can be extended uniquely to a complex linear map $\tilde{A}: \tilde{X} \rightarrow \tilde{X}$, and we define $\sigma(A)$, the spectrum of $A$, to be $\sigma(\tilde{A})$. If $f$ is analytic on a neighborhood
of $\sigma(\tilde{A})$ and $f(\bar{z})=\overline{f(z)}$ (where $\bar{w}$ denotes the complex conjugate of $w$ ), then $f(\tilde{A})(X) \subseteq X$, so we define $f(A)=f(\tilde{A}) \mid X$ in this case.

Aside from the norm topology on $\mathcal{L}(X, X)$, there are also locally convex topologies called the "strong operator topology" and the "weak operator topology" (see [17], Chapter 6). If $\left(A_{k}\right)$ is a sequence of bounded linear operators in $\mathcal{L}(X, X)$ and $A \in \mathcal{L}(X, X)$, then $\left(A_{k}\right)$ approaches $A$ in the strong operator topology as $k \rightarrow \infty$ if, for all $x \in X$,

$$
\lim _{k \uparrow \infty}\left\|A_{k}(x)-A(x)\right\|=0
$$

and $A_{k}$ approaches $A$ in the weak operator topology if, for all $x \in X$ and $\Psi \in \boldsymbol{X}^{*}$,

$$
\lim _{k \uparrow \infty} \Psi\left(A_{k}(x)-A(x)\right)=0
$$

If $\left(A_{k}\right)$ is a sequence of bounded linear operators which approaches a bounded linear operator $A$ in the weak operator topology, we shall write

$$
w-\lim _{k \uparrow \infty} A_{k}=A \text { or } A_{k} \rightharpoonup A
$$

Similarly, if $\left(\boldsymbol{A}_{\boldsymbol{k}}\right)$ converges to $\boldsymbol{A}$ in the strong operator topology we shall write

$$
s-\lim _{k \uparrow \infty} A_{k}=A \text { or } A_{k} \rightarrow A
$$

and if $\lim _{k \uparrow \infty}\left\|A_{k}-A\right\|=0$ we shall write

$$
n-\lim _{k \uparrow \infty} A_{k}=A \text { or } A_{k} \Rightarrow A
$$

In this paper we shall also deal with sequences $\left(A^{(k)}\right)$ of ordered $m$-tuples of bounded linear operators, so

$$
A^{(k)}=\left(A_{1}^{(k)}, A_{2}^{(k)}, \cdots, A_{m}^{(k)}\right)
$$

where $A_{j}^{(k)} \in \mathcal{L}(X, Y)$ for $1 \leq j \leq m$. We shall say the $A^{(k)}$ converges in the strong operator topology to the $m$-tuple $A=\left(A_{1}, A_{2}, \cdots, A_{m}\right)$ and write $A^{(k)} \rightarrow A$ if

$$
s-\lim _{k \uparrow \infty} A_{j}^{(k)}=A_{j} \quad \text { for } 1 \leq j \leq m
$$

Similarly, we shall write $A^{(k)} \rightharpoonup A$ if $A_{j}^{(k)} \rightharpoonup A_{j}$ for $1 \leq j \leq m$ and $A^{(k)} \Rightarrow A$ if $A_{j}^{(k)} \Rightarrow A$ for $1 \leq j \leq m$.

If $H$ is a complex Hilbert space with inner product $\langle x, y\rangle$ and $A \in \mathcal{L}(H)$, then $A$ is self-adjoint if $\langle A x, y\rangle=<x, A y>$ for all $x, y \in H$. If $A$ is self-adjoint and $f: \sigma(A) \subseteq \mathbb{Z} \rightarrow \mathbb{~ i s ~ a ~ c o n t i n u o u s ~ m a p , ~ o n e ~ c a n ~ d e f i n e ~}$ $f(A)$. This definition agrees with that in (1.1) when $f$ is analytic on an open
neighborhood of $A$. If $A \in \mathcal{L}(H)$ is self-adjoint, we shall say that $A$ is "positive semidefinite" (sometimes called nonnegative definite) if

$$
\begin{equation*}
<A x, x>\geq 0 \quad \text { for all } x \in H \tag{1.2}
\end{equation*}
$$

and $A$ is "positive definite" if there exists $\varepsilon>0$ such that

$$
\begin{equation*}
<A x, x>\geq \varepsilon \quad \text { for all } x \in H \text { with }\|x\|=1 \tag{1.3}
\end{equation*}
$$

We abbreviate "positive semidefinite" as "p.s.d.", "nonnegative definite" as "n.n.d." and "positive definite" as "p.d.". Positive semidefiniteness induces a partial ordering on the set of self-adjoint operators $A \in \mathcal{L}(H)$ : if $A$ and $B$ are bounded, self-adjoint operators, we write $A \leq B$ if $B-A$ is positive semidefinite.

Henceforth, whenever we say that $A \in \mathcal{L}(H)$ is positive definite or positive semidefinite, it will be assumed that $A$ is self-adjoint.

If $K$ denotes the set of bounded, self-adjoint p.s.d. linear maps in. $\mathcal{L}(H)$, then $K$ is an example of a cone (with vertex at 0 ) in a Banach space $Y=\mathcal{L}(H)$; and $K^{0}$, the interior of $K$, is the set of self-adjoint, p.d. operators in $\mathcal{L}(H)$. In general, if $C$ is a subset of a Banach space $Z$, we say that $C$ is a cone (with vertex at 0 ) if $C$ is a closed, convex subset of $Z$ and (a) if $x \in C$, then $t x \in C$ for all real numbers $t \geq 0$ and (b) if $x \in C-\{0\}$, then $-x \notin C$. A cone $C$ induces a partial ordering on $Z$ by $x \leq y$ if and only if $y-x \in C$. If $D$ is a subset of a Banach space $Z_{1}, C_{1}$ is a cone in $Z_{1}$ and $C_{2}$ is a cone in a Banach space $Z_{2}$ and $f: D \rightarrow Z_{2}$ is a map, we say that $f$ is order-preserving (with respect to the partial orderings induced in $Z_{j}$ by $C_{j}$ ) if for all $x$ and $y$ in $D$ such that $x \leq y$ (in the partial ordering induced by $C_{1}$ ) one has $f(x) \leq f(y)$ (in the partial ordering induced by $C_{2}$ ). Usually we shall have $Z_{1}=Z_{2}$ and $C_{1}=C_{2}$. If $D$ is convex, we say that $f: D \rightarrow Z_{2}$ is "convex" (with respect to the partial ordering induced by $C_{2}$ ) if for all $x$ and $y$ in $D$ and all real numbers $t$ with $0 \leq t \leq 1$, one has

$$
\begin{equation*}
f((1-t) x+t y) \leq(1-t) f(x)+t f(y) . \tag{1.4}
\end{equation*}
$$

We shall say that $f$ is strictly convex if $f$ is convex and for all $x \neq y$ in $D$

$$
f((1-t) x+t y) \neq(1-t) f(x)+t f(y) \quad \text { for } 0<t<1
$$

We shall say that $f$ is concave (strictly concave) if $-f$ is convex (strictly convex).

Our first lemma is well-known for real-valued functions. The proof in our generality is essentially the same and we omit it.

Lemma 1.1. Let $D$ be a convex subset of a Banach space $Z_{1}, C_{2}$ a cone in a Banach space $Z_{2}$ and $f: D \rightarrow Z_{2}$ a map which is continuous on line segments in $D$ (so the map $t \rightarrow f((1-t) x+t y), 0 \leq t \leq 1$, is continuous for
all $x$ and $y$ in $D$ ). If, with respect to the partial ordering induced by $C_{2}$, one has

$$
\begin{equation*}
f((1 / 2) x+(1 / 2) y) \leq(1 / 2) f(x)+(1 / 2) f(y) \tag{1.5}
\end{equation*}
$$

for all $x$ and $y$ in $D$, then $f$ is convex. If $f: D \rightarrow Z_{2}$ is convex and for all $x$ and $y$ in $D$ with $x \neq y$ one has

$$
f((1 / 2) x+(1 / 2) y) \neq(1 / 2) f(x)+(1 / 2) f(y)
$$

then $f$ is strictly convex. If $s_{j}, 1 \leq j \leq n$, are nonnegative real numbers such that $\sum_{j=1}^{n} s_{j}=1$, and $x_{j}, 1 \leq j \leq n$, are any points in $D$ and $f$ is convex, then

$$
\begin{equation*}
f\left(\sum_{j=1}^{n} s_{j} x_{j}\right) \leq \sum_{j=1}^{n} s_{j} f\left(x_{j}\right) \tag{1.6}
\end{equation*}
$$

If $f$ is strictly convex and $s_{j}>0$ for $1 \leq j \leq n$, then equality holds in (1.6) if and only if all the points $x_{j}$ are equal for $1 \leq j \leq n$.

We shall eventually need some continuity results for the strong operator topology.

Lemma 1.2. Suppose that $\left(L_{k}\right)$ is a sequence of bounded linear maps of a complex Banach space $\boldsymbol{X}$ to itself and that $\left(L_{k}\right)$ converges to a bounded linear operator $L$ in the strong operator topology. Assume that $\sigma\left(L_{k}\right) \subseteq B$ and $\sigma(L) \subseteq B$, where $B$ is a compact subset of the complex numbers, that $D$ is a bounded open neighborhood of $B$ such that $\Gamma=\partial D$ consists of a finite number of simple, closed rectifiable curves and that $f$ is a complex-valued function which is defined and analytic on an open neighborhood of $\bar{D}$. If there exists $a$ constant $M$ such that

$$
\begin{equation*}
\left\|\left(\lambda-L_{k}\right)^{-1}\right\| \leq M \text { for all } \lambda \in \Gamma \text { and all } k \geq 1 \tag{1.7}
\end{equation*}
$$

then $f\left(L_{k}\right) \rightarrow f(L)$. If $f(z)$ is an entire function, then $f\left(L_{k}\right) \rightarrow f(L)$. If $X$ is a Hilbert space and all the operators $L_{k}$ are normal (or self-adjoint), $f\left(L_{k}\right) \rightarrow f(L)$.

Proof. For $\lambda \in \Gamma$ and a fixed $x \in X$ one has

$$
\left(\lambda-L_{k}\right)^{-1} x-(\lambda-L)^{-1} x=\left(\lambda-L_{k}\right)^{-1}\left(L_{k}-L\right)(\lambda-L)^{-1} x .
$$

Applying the estimate in (1.7) yields

$$
\begin{equation*}
\left\|\left(\lambda-L_{k}\right)^{-1} x-(\lambda-L)^{-1} x\right\| \leq M\left\|\left(L_{k}-L\right)(\lambda-L)^{-1} x\right\| \tag{1.8}
\end{equation*}
$$

The uniform boundedness principle implies that there is a constant $M_{1}$ such that $\left\|L_{k}-L\right\| \leq M_{1}$ for all $k \geq 1$. Since $\lambda \rightarrow(\lambda-L)^{-1}$ is continuous in the
operator norm, (1.8) implies that there is a constant $M_{2}$, independent of $\lambda \in \Gamma$, such that

$$
\begin{equation*}
\left\|\left(\lambda-L_{k}\right)^{-1} x-(\lambda-L)^{-1} x\right\| \leq M_{2} \quad \text { for all } \lambda \in \Gamma \tag{1.9}
\end{equation*}
$$

Inequality (1.8) and the fact that $\left(L_{k}\right) \rightarrow L$ imply that

$$
\lim _{k \uparrow \infty}\left\|\left(\lambda-L_{k}\right)^{-1} x-(\lambda-L)^{-1} x\right\|=0
$$

A version of the Lebesgue dominated convergence theorem now implies that

$$
\begin{aligned}
& \lim _{k \uparrow \infty}(2 \pi i)^{-1} \int_{\Gamma} f(\lambda)\left(\lambda-L_{k}\right)^{-1} x d \lambda=\lim _{k \uparrow \infty} f\left(L_{k}\right) x \\
& =(2 \pi i)^{-1} \int_{\Gamma} f(\lambda)(\lambda-L)^{-1} x d \lambda=f(L) x
\end{aligned}
$$

so $f\left(L_{k}\right) \rightarrow f(L)$.
If $X$ is a Hilbert space, $\lambda \in \Gamma$ and $L_{k}$ is normal, it follows that $\lambda-L_{k}$ and $\left(\lambda-L_{k}\right)^{-1}$ are normal, so (see [36])

$$
\begin{equation*}
\left\|\left(\lambda-L_{k}\right)^{-1}\right\|=\sup \left\{\left|(\lambda-z)^{-1}\right|: z \in \sigma\left(L_{k}\right)\right\} \tag{1.10}
\end{equation*}
$$

Because $\sigma\left(L_{k}\right) \subseteq B$ and $\Gamma$ are disjoint compact sets, (1.10) implies that (1.7) is satisfied, so $f\left(L_{k}\right) \rightarrow f(L)$ by the first part of the lemma.

Finally, suppose $\boldsymbol{X}$ is a complex Banach space and $f$ is entire. If $R=\sup \left\{\left\|L_{k}\right\|: k \geq 1\right\}$, we can take $D=\{z:|z|<2 R\}$ and for $\lambda \in \partial D$, we have

$$
\left(\lambda-L_{k}\right)^{-1}=\lambda^{-1} \sum_{j=0}^{\infty}\left(\lambda^{-1} L_{k}\right)^{j}
$$

so

$$
\left\|\left(\lambda-L_{k}\right)^{-1}\right\| \leq|\lambda|^{-1} \sum_{j=0}^{\infty}\left|\lambda^{-j}\right|\left\|L_{k}\right\|^{j} \leq 2|\lambda|^{-1}=2 R^{-1}
$$

Thus the first part of the lemma implies that $f\left(L_{k}\right) \rightarrow f(L)$ in this case also.
REMARK 1.1. The obvious analogue of Lemma 1.2 for the weak operator topology is false. Let $H$ be $1^{2}$ and let $\left\{e_{j}: j \geq 1\right\}$ be the standard orthonormal basis for $1^{2}$. For $n \geq 1$, define a self-adjoint operator $A_{n}: H \rightarrow H$ by

$$
\begin{aligned}
A_{n}\left(e_{j}\right)=e_{n-j+1} & \text { for } 1 \leq j \leq n \text { and } \\
A_{n}\left(e_{j}\right)=0 & \text { for } j>n .
\end{aligned}
$$

One can easily prove that $\left(A_{n}\right) \rightharpoonup 0$, but $\left(A_{n}^{2}\right) \rightharpoonup I$, the identity.

Before stating our first theorem we recall some basic facts about matrices with nonnegative entries. If $M$ is an $n \times n$ matrix all of whose entries are nonnegative, $M$ is called "irreducible" if, for each ordered pair $(i, j)$ with $1 \leq i, j \leq n$, there exists an integer $p \geq 1$ (possibly dependent on $(i, j)$ ) such that the entry in row $i$ and column $j$ of $M^{p}$ is strictly positive. The matrix $M$ is called "primitive" if there exists an integer $p \geq 1$ such that all entries of $M^{p}$ are strictly positive. If $M$ is an irreducible matrix with nonnegative entries and $r$ denotes the spectral radius of $M$, then $r>0$ and there exists a unique (within scalar multiples) column vector $u$ such that all entries of $u$ are positive and

$$
\begin{equation*}
M u=r u \tag{1.11}
\end{equation*}
$$

If $M$ is primitive and if one defines $M_{1}=r^{-1} M$ (where $r$ is the spectral radius of $M$ ), then for any nonzero vector $x$, all of whose components are nonnegative, one has

$$
\begin{equation*}
\lim _{k \uparrow \infty} M_{1}^{k} x=\alpha u \tag{1.12}
\end{equation*}
$$

where $u$ is the eigenvector in (1.11) and $\alpha$ is a positive number depending on $x$.

If $M=\left(m_{i j}\right)$ is a matrix with nonnegative entries, $M$ is called "columnstochastic" if

$$
\sum_{i=1}^{n} m_{i j}=1 \quad \text { for } 1 \leq j \leq n
$$

and $M$ is "row-stochastic" if

$$
\sum_{j=1}^{n} m_{i j}=1 \quad \text { for } 1 \leq i \leq n
$$

It is an elementary exercise in the theory of nonnegative matrices that the spectral radius of any column-stochastic (or row-stochastic) matrix equals one. Furthermore, a trivial argument shows that the product of column-stochastic (or row-stochastic) matrices is column-stochastic (or row-stochastic).

Now suppose that $M$ is a column-stochastic, primitive matrix with nonnegative entries and let $u$ be a column vector, all of whose entries are positive, such that

$$
M u=u
$$

If we normalize $u$ by demanding

$$
\sum_{j=1}^{n} u_{j}=1
$$

we know that $u$ is unique. Define $M_{\infty}$ to be the $n \times n$ matrix all of whose columns equal $u$. If $\epsilon_{j}, 1 \leq j \leq n$, denotes the standard orthonormal basis
of $\mathbb{\mathbb { N }}^{n}$, we know that $M^{k} e_{j}$ is the $j$ th column of $M^{k}$, and because $M^{k}$ is column-stochastic, (1.12) implies that

$$
\lim _{k \nmid \infty} M^{k} e_{j}=u, \quad 1 \leq j \leq n
$$

We conclude from the previous equation that, if $M$ is column-stochastic and primitive, then

$$
\begin{equation*}
\lim _{k \uparrow \infty} M^{k}=M_{\infty} \tag{1.13}
\end{equation*}
$$

If $K$ is a cone in a Banach space $X$, let $C$ denote the cone which is the $n$-fold Cartesian product of $K$. Let $\boldsymbol{Y}$ denote the $n$-fold Cartesian product of $X$ with any of the standard norms. If $M$ is an $n \times n$ matrix with nonnegative entries, $M$ induces a bounded linear map $W$ of $Y$ to $\boldsymbol{Y}$ by

$$
\begin{gathered}
W(x)=x M=y, \text { where } \\
x=\left(x_{1}, x_{2}, \cdots, x_{n}\right), y=\left(y_{1}, y_{2}, \cdots, y_{n}\right), y_{j}=\sum_{i=1}^{n} m_{i j} x_{i}
\end{gathered}
$$

It is easy to check that $W(C) \subseteq C$, that $W(C-\{0\}) \subseteq C-\{0\}$ if no row of $M$ is identically zero, and that (if $K^{0}$ is nonempty) $W\left(C^{0}\right) \subseteq C^{0}$ if no column of $M$ is identically zero.

We are in a position to state our first theorem. For simplicity, we restrict ourselves to the cone of p.s.d. bounded linear operators on a Hilbert space, but versions of the following theorem can be given for more general cones.

THEOREM 1.1. Let $K$ denote the cone of positive semidefinite, self-adjoint bounded linear operators on a Hilbert space $H$. Let $C$ denote the $n$-fold Cartesian product of $K, C=K \times K \times \cdots \times K$. Let $\boldsymbol{Y}$ denote the $n$-fold Cartesian product of $\boldsymbol{X}=\mathcal{L}(H, H)$ with itself, $\boldsymbol{Y}=\boldsymbol{X} \times \boldsymbol{X} \times \cdots \times \boldsymbol{X}$. Suppose that $f: C^{0} \rightarrow C^{0}$ is a continuous map and $\phi: K^{0} \rightarrow \boldsymbol{X}$ is a continuous map and define $\Phi: C^{0} \rightarrow Y$ by

$$
\Phi\left(A_{1}, A_{2}, \cdots, A_{n}\right)=\left(\phi\left(A_{1}\right), \phi\left(A_{2}\right), \cdots, \phi\left(A_{n}\right)\right)
$$

Assume that for every $A \in C^{0}$ there exist $B \in C^{0}$ and positive reals $\alpha$ and $\beta$ such that

$$
\begin{equation*}
\alpha B \leq f^{j}(A) \leq \beta B \tag{1.14}
\end{equation*}
$$

and

$$
\begin{equation*}
\Phi\left(f^{j}(A)\right) \leq \Phi(\beta B) \tag{1.15}
\end{equation*}
$$

for all $j \geq 0$, where the partial ordering in (1.14) and (1.15) is induced by $C$ and $f^{j}$ is the $j$ th iterate of $f$. Assume that there exists an $n \times n$, primitive, column-stochastic matrix $M$ such that, for every $A \in C^{0}$,

$$
\begin{equation*}
\Phi(f(A)) \geq \Phi(A) M \tag{1.16}
\end{equation*}
$$

Let $u$ denote the unique column vector in $\mathbb{R}^{n}$ such that all components of $u_{i}$ of $u$ are positive and

$$
M u=u
$$

and

$$
\sum_{i=1}^{n} u_{i}=1
$$

and let $\pi_{i}$ denote the projection of $Y$ onto its ith coordinate. If, for $A \in C^{0}$, we define

$$
f^{k}(A)=\left(A_{1}^{(k)}, A_{2}^{(k)}, \cdots, A_{n}^{(k)}\right)
$$

so $A_{i}^{(k)}=\pi_{i}\left(f^{k}(A)\right)$, there exists $E \in K^{0}$ such that

$$
\begin{equation*}
s-\lim _{k \uparrow \infty} \sum_{i=1}^{n} u_{i} \phi\left(A_{i}^{(k)}\right)=E . \tag{1.17}
\end{equation*}
$$

Furthermore, for $1 \leq i \leq n$, one has

$$
\begin{equation*}
w-\lim _{k \uparrow \infty} \phi\left(A_{i}^{(k)}\right)=E \tag{1.18}
\end{equation*}
$$

Proof. If $u$ is the eigenvector in the statement of the theorem, define, for $A=\left(A_{1}, A_{2}, \cdots, A_{n}\right) \in C^{0}$, a function $\Psi: C^{0} \rightarrow K^{0}$ by the formula

$$
\begin{equation*}
\Psi(A)=\sum_{j=1}^{n} u_{j} \phi\left(A_{j}\right) \tag{1.19}
\end{equation*}
$$

Inequality (1.16) implies, if $A=\left(A_{1}, A_{2}, \cdots, A_{n}\right)$, that

$$
\begin{align*}
\Psi(f(A)) & =\sum_{j=1}^{n} u_{j} \phi\left(\pi_{j} f(A)\right) \\
\geq \sum_{j=1}^{n} u_{j}\left(\sum_{i=1}^{n} m_{i j} \phi\left(A_{i}\right)\right) & =\sum_{i=1}^{n} \phi\left(A_{i}\right)\left(\sum_{j=1}^{n} m_{i j} u_{j}\right)  \tag{1.20}\\
& =\sum_{i=1}^{n} u_{i} \phi\left(A_{i}\right)=\Psi(A)
\end{align*}
$$

If we define $E_{k}=\Psi\left(f^{k}(A)\right), E_{k}-E_{1}$ is an increasing sequence of bounded, self-adjoint operators, and (1.14) and (1.15) imply that $E_{k}-E_{1}$ is bounded
above (in the partial ordering on $K$ ). Thus as $k \rightarrow \infty, E_{k} \rightarrow E$ with $E-E_{1}$ self-adjoint. By iterating inequality (1.16) we see that, for any $k \geq 1$,

$$
\begin{equation*}
\Phi\left(f^{k}(A)\right) \geq \Phi(A) M^{k} \tag{1.21}
\end{equation*}
$$

The remarks preceding the theorem imply that given any $\varepsilon>0$, there exists $N_{1} \geq 1$ such that for all $k \geq N_{1}$,

$$
\begin{equation*}
(1+\varepsilon) M_{\infty} \geq M^{k} \geq(1-\varepsilon) M_{\infty} \tag{1.22}
\end{equation*}
$$

where $M_{\infty}$ is the matrix with all columns equal to $u$, and (1.22) means that for $1 \leq i, j \leq n$, the $i, j$ entry of $M^{k}$ is greater than or equal to the $i, j$ entry of $(1-\varepsilon) M_{\infty}$. If $k \geq N_{1}$, and $1 \leq i \leq n$, it follows that

$$
\begin{align*}
\pi_{i}\left(\Phi\left(f^{k}(A)\right)\right) & \geq \pi_{i}\left(\Phi\left(f^{k-N_{1}}(A)\right) M^{N_{1}}\right)=\pi_{i}\left(\Phi\left(f^{k-N_{1}}(A)\right) M_{\infty}\right) \\
& +\pi_{i}\left(\Phi\left(f^{k-N_{1}}(A)\right)\left(M^{N_{1}}-M_{\infty}\right)\right)  \tag{1.23}\\
& =E_{k-N_{1}}+\pi_{i}\left(\Phi\left(f^{k-N_{1}}(A)\right)\left(M^{N_{1}}-M_{\infty}\right)\right)
\end{align*}
$$

For a given $x \in H$ and $\varepsilon>0$ there exists $N_{2}$ such that

$$
\begin{equation*}
<E_{j} x, x>\geq<E x, x>-\varepsilon \quad \text { for } j \geq N_{2} \tag{1.24}
\end{equation*}
$$

because $E_{j} \rightarrow E$. Combining inequalities (1.23) and (1.24) yields for $k \geq$ $N_{1}+N_{2}$,

$$
\begin{align*}
<\phi\left(A_{i}^{(k)} x\right), x> & \geq<E x, x>-\varepsilon  \tag{1.25}\\
& +<\pi_{i}\left(\Phi\left(f^{k-N_{1}}(A)\right)\left(M^{N_{1}}-M_{\infty}\right)\right) x, x>
\end{align*}
$$

which implies that (using inequality (1.22) and recalling that $\Phi\left(f^{k-N_{1}}(A)\right)$ is bounded in norm)

$$
\begin{equation*}
\lim _{k \rightarrow \infty} \inf <\phi\left(A_{i}^{(k)}\right) x, x>\geq<E x, x> \tag{1.26}
\end{equation*}
$$

If, for some $x \in H$ and some $i$, one has

$$
\begin{equation*}
\lim _{k \rightarrow \infty} \sup <\phi\left(A_{i}^{(k)} x\right), x \gg<E x, x> \tag{1.27}
\end{equation*}
$$

then inequalities (1.26) and (1.27) imply

$$
\lim _{k \rightarrow \infty} \sup <E_{k} x, x>=\lim _{k \rightarrow \infty} \sup \sum_{j=1}^{n} u_{j}<\phi\left(A_{i}^{(k)}\right) x, x \gg<E x, x>
$$

The above inequality contradicts the fact that

$$
\lim _{k \uparrow \infty}<E_{k} x, x>=<E x, x>
$$

so inequality (1.27) must be false. We conclude from (1.26) that

$$
\begin{equation*}
\lim _{k \uparrow \infty}<\phi\left(A_{i}^{(k)}\right) x, x>=<E x, x> \tag{1.28}
\end{equation*}
$$

Standard arguments using the polarization now imply that

$$
w-\lim _{k \uparrow \infty} \phi\left(A_{i}^{(k)}\right)=E
$$

There are several obstacles to using Theorem 1.1. The first problem, of course, is to prove the existence of $\phi$ and $M$ as in Theorem 1.1 for examples of interest to us. We shall use the classical results of C. Loewner concerning the concavity of order-preserving maps from the cone $K$ of p.s.d. bounded linear operators of a Hilbert space $H$ to $\mathcal{L}(H)$.

However, even assuming that we can establish the hypotheses of Theorem 1.1 in examples of interest, Theorem 1.1 provides inadequate information when $H$ is infinite dimensional. If $H$ is finite dimensional, the weak, strong and operator norm topologies on $\mathcal{L}(H)$ are identical, so Theorem 1.1 implies

$$
\lim _{k \nmid \infty}\left\|\phi\left(A_{i}^{(k)}\right)-E\right\|=0 \quad \text { for } 1 \leq i \leq n
$$

If $\phi$ is one-one with norm-continuous inverse, one concludes that

$$
\lim _{k \nmid \infty}\left\|A_{i}^{(k)}-\phi^{-1}(E)\right\|=0
$$

If $H$ is infinite dimensional, one would hope that there exists a p.d. bounded linear operator $G$ such that

$$
\begin{equation*}
s-\lim _{k \uparrow \infty} A_{i}^{(k)}=G \quad \text { for } 1 \leq i \leq n \tag{1.29}
\end{equation*}
$$

However, as Remark 1.1 shows, the weak operator convergence in (1.18) may be very far from the strong operator convergence hoped for in (1.29).

We now begin to address there deficiencies.
If $K$ is the cone of p.s.d. bounded linear operators in $\mathcal{L}(H)$ and $H$ is a Hilbert space, we need to know when certain maps defined on $D=K^{0}$ are order-preserving or concave or strictly concave. The first lemma is Loewner's theorem [23, 24] concerning order-preserving maps on $K^{0}$; an exposition of Loewner's theory is given in [16].

LEMMA 1.3. (C. Loewner) Let $K$ denote the cone of positive semidefinite, bounded linear maps of a Hilbert space $H$ to itself. Suppose that $f:(0, \infty) \rightarrow \mathbb{Z}$ is a continuous, real-valued map and that $f$ has an analytic extension to $U=\{z \in \mathbb{\Xi}: \operatorname{Im}(z) \neq 0$ or $(\operatorname{Im}(z)=0$ and $\operatorname{Re}(z)>0)\}$ such that $\operatorname{Im}(f(z))>0$ for all $z$ such that $\operatorname{Im}(z)>0$. Then the map $A \rightarrow f(A)$ for $A \in D=K^{0}$ is order-preserving with respect to the partial order induced by $K$.

Functions that satisfy the hypotheses of Lemma 1.2 are $f(z)=\log (z)$ and $f(z)=z^{p}, 0<p \leq 1$, and $\log (z)$ will be the example of most interest here. One can give a simple and self-contained proof that if $A, B \in K^{0}$ and $A \leq B$, then $B^{-1} \leq A^{-1}$ (although we shall not do so). Using this fact, we now prove directly that the maps $A \rightarrow \log (A)$ and $A \rightarrow A^{p}, 0<p \leq 1$, are order-preserving on $K^{0}$. If $E \in K^{0}$ and $E_{t} \equiv(1-t) I+t E$, where $I$ denotes the identity map, then one can easily prove that

$$
\begin{equation*}
\log (E)=\int_{0}^{1}(d / d t) \log \left(E_{t}\right) d t=\int_{0}^{1} E_{t}^{-1}(E-I) d t \tag{1.30}
\end{equation*}
$$

An algebraic manipulation gives

$$
\begin{equation*}
\log (E)=\int_{0}^{1}\left[t^{-1} I-t^{-1} E_{t}^{-1}\right] d t \tag{1.31}
\end{equation*}
$$

If $A$ and $B$ are in $K^{0}$ and $A \leq B$ and $A_{t}=(1-t) I+t A$ and $B_{t}=(1-t) I+t B$ for $0 \leq t \leq 1$, then $A_{t} \leq B_{t}$ for $0 \leq t \leq 1$, so $t^{-1} A_{t}^{-1} \geq t^{-1} B_{t}^{-1}$ for $0<t \leq 1$ and

$$
\begin{equation*}
t^{-1} I-t^{-1} A_{t}^{-1} \leq t^{-1} I-t^{-1} B_{t}^{-1}, \text { for } 0<t \leq 1 \tag{1.32}
\end{equation*}
$$

Using (1.31) and (1.32) one finds that $\log (A) \leq \log (B)$. That $A^{p} \leq B^{p}$ if $A \leq B$ and $0<p<1$ follows by a similar argument from the formula (see [21], p. 286)

$$
\begin{equation*}
E^{p}=[\sin (\pi p) / \pi] \int_{0}^{\infty} \lambda^{p-1} E(\lambda+E)^{-1} d \lambda, \quad E \in K^{0}, 0<p<1 \tag{1.33}
\end{equation*}
$$

We also need concavity results for maps of $K^{0}$ to $\mathcal{L}(H)$.
Lemma 1.4. (Ando [3]) Suppose that $K, H$ and $f$ are as in Lemma 1.3. Then, for $A \in K^{0}$, the map $A \rightarrow f(A)$ is concave and the map $A \rightarrow A f(A)$ is convex (with respect to the partial ordering from $K$ ).

Ando states Lemma 1.4 for finite-dimensional Hilbert spaces (Theorem 4 in [3]), but the same argument, based on Loewner's theory, works for general Hilbert spaces.

Lemma 1.4 is not quite adeguate for our purposes. We need strict concavity and convexity results, and in fact we shall need a property (see Theorem 1.2 below) analogous to the property of uniform convexity for norms. Such results will follow from the strict convexity of the map $A \rightarrow A^{-1}$ for $A \in K^{0}$, and this strict convexity was proved independently by P. Whittle [34, Lemma 1] and I. Olkin and J. Pratt [29]. Whittle's lemma is stated for finite-dimensional Hilbert spaces, but the proof applies in general and yields the following lemma.

Lemmma 1.5. ([34] and [29]) Let $H$ be a Banach space and suppose that $A, B \in \mathcal{L}(H)$. If $\lambda$ is a real number such that $0<\lambda<1$ and $A, B$ and $(1-\lambda) A+\lambda B$ are one-one and onto, then

$$
\begin{gather*}
(1-\lambda) A^{-1}+\lambda B^{-1}-((1-\lambda) A+\lambda B)^{-1}= \\
\lambda(1-\lambda)\left(A^{-1}-B^{-1}\right)\left(\lambda A^{-1}+(1-\lambda) B^{-1}\right)^{-1}\left(A^{-1}-B^{-1}\right) . \tag{1.34}
\end{gather*}
$$

If $H$ is a Hilbert space and $K$ denotes the cone of positive semidefinite operators in $\mathcal{L}(H)$, then the map $A \rightarrow A^{-1}$ of $K^{0}$ to $K^{0}$ is strictly convex.

By exploiting Lemma 1.5 we obtain the following sharpening of Ando's theorem (Lemma 1.4). See also Bendat and Sherman [8].

Lemma 1.6. Let $f, K$ and $H$ be as in Lemma 1.3. If there do not exist real constants $\alpha$ and $\beta$ such that $f(x)=\alpha+\beta x$ for all $x>0$, then for $A \in K^{0}$, the map $A \rightarrow f(A)$ is strictly concave. If there do not exist real constants $\alpha$ and $\gamma$ such that $f(x)=\alpha+\gamma x^{-1}$ for all $x>0$, then $A \rightarrow A f(A)$ is strictly convex on $K^{0}$.

Proof. Loewner's theory implies that for $\lambda$ not a negative real, $\lambda \neq 0$,

$$
\begin{equation*}
f(\lambda)=\alpha+\beta \lambda+\int_{-\infty}^{0}(1+\lambda t)(t-\lambda)^{-1} d \mu(t) \tag{1.35}
\end{equation*}
$$

where $\alpha$ is real, $\beta \geq 0$ and $\mu$ is a finite nonnegative Borel measure with support in $(-\infty, 0]$. If we assume that $f(x)$ is not an affine map, then $\mu$ is not the zero measure. From (1.35) and the identity

$$
(I+t A)(t I-A)^{-1}=-t I-\left(1+t^{2}\right)(A-t I)^{-1}
$$

we obtain

$$
\begin{equation*}
f(A)=\alpha I+\beta A+\int_{-\infty}^{0}\left[-t I-\left(1+t^{2}\right)(A-t I)^{-1}\right] d \mu(t) \tag{1.36}
\end{equation*}
$$

The map $A \rightarrow \alpha I+\beta A$ is obviously concave and Lemma 1.5 implies that for each $t \leq 0$ and for $A \in K^{0}$ the map

$$
A \rightarrow-t I-\left(1+t^{2}\right)(A-t I)^{-1}
$$

is strictly concave so (1.36) implies $A \rightarrow f(A)$ is concave (Lemma 1.4). To prove that $A \rightarrow f(A)$ is strictly concave, take $A, B \in K^{0}$ with $A \neq B$ and define

$$
E=(A+B) / 2
$$

Equation (1.36) gives

$$
\begin{align*}
(1 / 2) f(A) & +(1 / 2) f(B)-f(E) \\
& =\int_{-\infty}^{0}\left[-(1 / 2)(A-t I)^{-1}-(1 / 2)(B-t I)^{-1}\right.  \tag{1.37}\\
& \left.+(E-t I)^{-1}\right]\left(1+t^{2}\right) d \mu(t)
\end{align*}
$$

The strict concavity of $A \rightarrow-A^{-1}$ implies that there exist $t_{0} \leq 0$, with $t_{0}$ in the support of $\mu, x \in H$ and $\delta>0$ such that

$$
\begin{equation*}
<(-1 / 2)(A-t I)^{-1} x-(1 / 2)(B-t I)^{-1} x+(E-t I)^{-1} x, x>\leq 0 \tag{1.38}
\end{equation*}
$$

for all $t \leq 0$ and strict inequality holds in inequality (1.38) for $\left|t-t_{0}\right| \leq \delta$. Using this information in (1.37) we find

$$
<(1 / 2) f(A) x+(1 / 2) f(B) x-f(E) x, x><0
$$

which proves strict concavity.
Starting from (1.36) we see that

$$
\begin{equation*}
A f(A)=\alpha A+\beta A^{2}+\int_{-\infty}^{0}\left[-t A-\left(1+t^{2}\right) A(A-t I)^{-1}\right] d \mu(t) \tag{1.39}
\end{equation*}
$$

It is easy to prove directly that $A \rightarrow A^{2}$ is strictly convex on $K$. Therefore (because $\beta \geq 0$ ) the map $A \rightarrow \alpha A+\beta A^{2}$ is convex on $K$ and strictly convex if $\beta>0$. Some algebraic manipulation shows that

$$
-t A-\left(1+t^{2}\right) A(A-t I)^{-1}=-t A-\left(1+t^{2}\right) I+|t|\left(1+t^{2}\right)(A-t I)^{-1}
$$

for $t \leq 0$. Lemma 1.5 implies that for each $t \leq 0$, the map

$$
A \rightarrow-t A-\left(1+t^{2}\right) I+|t|\left(1+t^{2}\right)(A-t I)^{-1}
$$

is convex and, in fact, strictly convex if $t<0$. If the support of $\mu$ has nonempty intersection with $(-\infty, 0)$, the same kind of proof as used before shows that $A \rightarrow A f(A)$ is strictly convex $\left(A \in K^{0}\right)$. If the support of $\mu$ is $\{0\}$, (1.35) implies that

$$
f(\lambda)=\alpha+\beta \lambda-\gamma \lambda^{-1}
$$

for some $\gamma \geq 0$. If $\beta>0$, the map $A \rightarrow A f(A)$ is given by $A \rightarrow \alpha A+\beta A^{2}-\gamma I$ in this case and hence is strictly convex. If $\beta=0, f(\lambda)=\alpha-\gamma \lambda^{-1}$, contrary to our assumption.

When $f(z)=\log (z)$ or $z^{p}, 0<p<1$, Lemma 1.6 can be proved directly by using (1.31) and (1.33) and Lemma 1.5 .

An immediate consequence of Lemmas 1.1 and 1.6 is:
Corollary 1.1. Let $H$ be a Hilbert space. Suppose that $A_{j}, 1 \leq j \leq m$, are bounded, self-adjoint, p.d. linear maps of $H$ to $H$. If $s_{j}, 1 \leq j \leq m$, are positive numbers such that $\sum_{j=1}^{m} s_{j}=1$, it follows that

$$
\begin{equation*}
\log \left(\sum_{j=1}^{m} s_{j} A_{j}\right) \geq \sum_{j=1}^{m} s_{j} \log \left(A_{j}\right) \tag{1.40}
\end{equation*}
$$

Equality holds in (1.40) if and only if all the operators $\boldsymbol{A}_{j}$ are equal. If $0<\alpha<1$,

$$
\begin{equation*}
\left(\sum_{j=1}^{m} s_{j} A_{j}\right)^{\alpha} \geq \sum_{j=1}^{m} s_{j} A_{j}^{\alpha} \tag{1.41}
\end{equation*}
$$

Equality holds in (1.41) if and only if all the operators $A_{j}$ are equal.
REMARK 1.2. Suppose that $A_{j}$ and $s_{j}$ are as above and that $p$ and $q$ are real numbers such that $1 \leq p \leq q$. Defining $\alpha=p q^{-1}$ and $B_{j}=A_{j}^{q}$, Corollary 1.1 implies

$$
\begin{equation*}
\left(\sum_{j=1}^{m} s_{j} B_{j}\right)^{\alpha}=\left(\sum_{j=1}^{m} s_{j} A_{j}^{q}\right)^{\alpha} \geq \sum_{j=1}^{m} s_{j} B_{j}^{\alpha}=\sum_{j=1}^{m} s_{j} A_{j}^{p} \tag{1.42}
\end{equation*}
$$

One only needs $0<p<q$ to derive inequality (1.42). Since $p \geq 1$, Lemma 1.3 implies that $B \rightarrow B^{1 / p}$ is order-preserving on $K$, so one obtains from (1.42)

$$
\begin{equation*}
\left(\sum_{j=1}^{m} s_{j} A_{j}^{q}\right)^{1 / q} \geq\left(\sum_{j=1}^{m} s_{j} A_{j}^{p}\right)^{1 / p} \tag{1.43}
\end{equation*}
$$

For positive real $\boldsymbol{A}_{\boldsymbol{j}}$, inequality (1.43) is a classical result [20].
Unfortunately, when $H$ is infinite dimensional Lemma 1.6 is inadequate for our purposes. We need to exploit strict concavity and strict convexity in a more quantitative way, analogous to the idea of uniform convexity for a norm. The next lemma illustrates the sort of uniform convexity we need for the case of the strictly convex map $\boldsymbol{A} \rightarrow \boldsymbol{A}^{-1}$ when $\boldsymbol{A}$ is positive definite.

LEMMA 1.7. Suppose that $A_{i}, 1 \leq i \leq m$, are p.d., bounded linear maps of a Hilbert space $H$ into itself and $\alpha I \leq A_{i} \leq \beta I$ for $1 \leq i \leq m$, where $\alpha$ and $\beta$ are positive reals. Assume that $\sigma_{k}, 1 \leq k \leq m$, are positive reals such that $\sum_{k=1}^{m} \sigma_{k}=1$. Then, for $1 \leq i, j \leq m$,

$$
\begin{equation*}
\sum_{k=1}^{m} \sigma_{k} A_{k}^{-1}-\left(\sum_{k=1}^{m} \sigma_{k} A_{k}\right)^{-1} \geq \sigma_{i} \sigma_{j}\left(\sigma_{i}+\sigma_{j}\right)^{-1} \alpha\left(A_{i}^{-1}-A_{j}^{-1}\right)^{2} \tag{1.44}
\end{equation*}
$$

where the inequality refers to the partial ordering induced by the cone of positive semidefinite bounded linear maps.

Proof. Because $A \rightarrow A^{-1}$ is convex, the left side of (1.44) is always greater than or equal to zero. To prove (1.44) it suffices (by relabelling) to prove it when $i=1, j=2$.

If $A \geq \alpha I$ and $B \geq \alpha I$, then the spectral mapping theorem implies $A^{-1} \leq \alpha^{-1} I, B^{-1} \leq \alpha^{-1} I$ and

$$
\begin{equation*}
\lambda A^{-1}+(1-\lambda) B^{-1} \leq \alpha^{-1} I \text { and }\left(\lambda A^{-1}+(1-\lambda) B^{-1}\right)^{-1} \geq \alpha I \tag{1.45}
\end{equation*}
$$

Using inequality (1.45) in (1.34) gives

$$
\begin{equation*}
(1-\lambda) A^{-1}+\lambda B^{-1}-((1-\lambda) A+\lambda B)^{-1} \geq \alpha \lambda(1-\lambda)\left(A^{-1}-B^{-1}\right)^{2} \tag{1.46}
\end{equation*}
$$

If we define $\lambda=\sigma_{2} /\left(\sigma_{1}+\sigma_{2}\right), A_{1}=A$ and $A_{2}=B$, inequality (1.46) gives

$$
\begin{equation*}
(1-\lambda) A_{1}^{-1}+\lambda A_{2}^{-1}-\left((1-\lambda) A_{1}+\lambda A_{2}\right)^{-1} \geq \alpha \lambda(1-\lambda)\left(A_{1}^{-1}-A_{2}^{-1}\right)^{2} \equiv R \tag{1.47}
\end{equation*}
$$

If we define $B_{1}=(1-\lambda) A_{1}+\lambda A_{2}$ with $\lambda=\sigma_{2}\left(\sigma_{1}+\sigma_{2}\right)^{-1}$, inequality (1.47) and the convexity of $A \rightarrow A^{-1}$ give

$$
\begin{equation*}
\sum_{j=1}^{m} \sigma_{j} A_{j}^{-1}-\left(\sum_{j=1}^{m} \sigma_{j} A_{j}\right)^{-1} \geq\left(\sigma_{1}+\sigma_{2}\right) B_{1}^{-1}+\sum_{j=3}^{m} \sigma_{j} A_{j}^{-1} \tag{1.48}
\end{equation*}
$$

$$
-\left[\left(\sigma_{1}+\sigma_{2}\right) B_{1}+\sum_{j=3}^{m} \sigma_{j} \boldsymbol{A}_{j}\right]^{-1}+\left(\sigma_{1}+\sigma_{2}\right) R \geq\left(\sigma_{1}+\sigma_{2}\right) R .
$$

Inequality (1.48) is precisely the statement of the lemma for $i=1$ and $j=2$.
The next theorem, when combined with Theorem 1.1, will enable us to prove convergence in the strong operator topology.

THEOREM 1.2. Let $K$ be the cone of positive semidefinite bounded linear maps of a Hilbert space $H$ into itself. Let $C$ denote the $m$-fold Cartesian product of $K$ with itself. Thus $C \subseteq Y$, where $\boldsymbol{Y}$ is the $m$-fold Cartesian product of $X=\mathcal{L}(H)$ with itself. Assume that $\left(B^{(k)}\right), \quad k \geq 0$, is a sequence in $C^{0}$, and write $B^{(k)}=\left(B_{1}^{(k)}, B_{2}^{(k)}, \cdots, B_{m}^{(k)}\right)$. Suppose that $\phi:(0, \infty) \rightarrow \mathbb{Z}$ is a $C^{1}$, real-valued function such that $\lim _{x \rightarrow+\infty} d \phi(x) / d x=0$ and such that $\phi$ has an analytic extension to $U=\left\{z \in \mathbb{C}: \operatorname{Im}(z)^{+\infty} \neq 0\right.$ or $(\operatorname{Im}(z)=0$ and $\left.\operatorname{Re}(z)>0)\right\}$ and $\operatorname{Im}(\phi(z))>0$ for all $z$ such that $\operatorname{Im}(z)>0$. Assume that there exist positive numbers $\lambda_{p}, 1 \leq p \leq m$, such that

$$
\begin{equation*}
w-\lim _{k \uparrow \infty}\left[\phi\left(\sum_{p=1}^{m} \lambda_{p} B_{p}^{(k)}\right)-\sum_{p=1}^{m} \lambda_{p} \phi\left(B_{p}^{(k)}\right)\right]=0 \tag{1.49}
\end{equation*}
$$

and that there are positive numbers $\alpha$ and $\beta$ such that

$$
\begin{equation*}
\alpha I \leq B_{p}^{(k)} \leq \beta I \quad \text { for } k \geq 1,1 \leq p \leq m \tag{1.50}
\end{equation*}
$$

Then for any $i$ and $j$ with $1 \leq i, j \leq m$,

$$
\begin{equation*}
s-\lim _{k \uparrow \infty}\left[\phi\left(B_{i}^{(k)}\right)-\phi\left(B_{j}^{(k)}\right)\right]=0 \tag{1.51}
\end{equation*}
$$

If there exist positive numbers $u_{i}, 1 \leq i \leq m$, and $E \in K^{0}$ such that $\sum_{i=1}^{m} u_{i}=1$ and

$$
s-\lim _{k \uparrow \infty} \sum_{i=1}^{m} u_{i} \phi\left(B_{i}^{(k)}\right)=E
$$

then

$$
s-\lim _{k \uparrow \infty} \phi\left(B_{i}^{(k)}\right)=E
$$

If the restriction of $\phi$ to an open neighborhood of $(0, \infty)$ in $\mathbb{C}$ is one-one, then

$$
s-\lim _{k \uparrow \infty} B_{i}^{(k)}=\phi^{-1}(E)
$$

PROOF. Loewner's theory (see [16]) implies that if $\lambda \neq 0$ and $\lambda$ is not a negative real, then

$$
\begin{equation*}
\phi(\lambda)=\alpha_{1}+\beta_{1} \lambda+\int_{-\infty}^{0}\left[-t-\left(1+t^{2}\right)(\lambda-t)^{-1}\right] d \mu(t) \tag{1.52}
\end{equation*}
$$

where $\alpha_{1}$ is real, $\beta_{1} \geq 0$ and $\mu$ is a nonnegative, finite Borel measure. Using this formula one easily proves that the condition $\lim _{x \rightarrow \infty} d \phi(x) / d x=0$ implies that $\beta_{1}=0$.

We claim first that for all $i$ and $j, 1 \leq i, j \leq m$, one has

$$
s-\lim _{k \uparrow \infty} \phi\left(B_{i}^{(k)}\right)-\phi\left(B_{j}^{(k)}\right)=0
$$

We shall prove this for $i=1$ and $j=2$, since the general argument is the same. We obtain from (1.52) that

$$
\begin{equation*}
\phi\left(\sum_{p=1}^{m} \lambda_{p} B_{p}^{(k)}\right)-\sum_{p=1}^{m} \lambda_{p} \phi\left(B_{p}^{(k)}\right)= \tag{1.53}
\end{equation*}
$$

$$
\int_{-\infty}^{0}\left(1+t^{2}\right)\left\{\sum_{p=1}^{m} \lambda_{p}\left(B_{p}^{(k)}-t I\right)^{-1}-\left(\sum_{p=1}^{m} \lambda_{p}\left(B_{p}^{(k)}-t I\right)\right)^{-1}\right\} \mu(d t)
$$

Inequality (1.50) implies that

$$
B_{p}^{(k)}-t I \geq(\alpha-t) I
$$

so Lemma 1.7 and (1.53) give

$$
\begin{equation*}
\lambda_{1} \lambda_{2}\left(\lambda_{1}+\lambda_{2}\right)^{-1} \int_{-\infty}^{0}\left(1+t^{2}\right)(\alpha-t)\left[\left(B_{1}^{(k)}-t I\right)^{-1}-\left(B_{2}^{(k)}-t I\right)^{-1}\right]^{2} \mu(d t) \tag{1.54}
\end{equation*}
$$

The left side of (1.54) is assumed to approach 0 in the weak operator topology as $k \rightarrow \infty$. Hence, for any $x \in H$,

$$
\begin{equation*}
\lim _{k \uparrow \infty} \int_{-\infty}^{0}\left(1+t^{2}\right)(\alpha-t)\left\|\left(B_{1}^{(k)}-t I\right)^{-1} x-\left(B_{2}^{(k)}-t I\right)^{-1} x\right\|^{2} \mu(d t)=0 \tag{1.55}
\end{equation*}
$$

Now we return to (1.52). Using for the first time the fact that $\beta_{1}=0$, we obtain

$$
\begin{gather*}
\phi\left(B_{1}^{(k)}\right)(x)-\phi\left(B_{2}^{(k)}\right)(x)  \tag{1.56}\\
=\int_{-\infty}^{0}\left(1+t^{2}\right)\left[\left(B_{2}^{(k)}-t I\right)^{-1} x-\left(B_{1}^{(k)}-t I\right)^{-1} x\right] \mu(d t) .
\end{gather*}
$$

Because $B_{i}^{(k)} \geq \alpha I$ for $k \geq 1$ and $1 \leq i \leq m$, it is easy to see that there exists $M_{1}$ such that

$$
\left\|\left(B_{i}^{(k)}-t I\right)^{-1}\right\| \leq M_{1}|t|^{-1} \quad \text { for all } t \leq-1, k \geq 1,1 \leq i \leq m
$$

It follows that for $t \leq-1$

$$
\begin{gather*}
\left\|\left(B_{1}^{(k)}-t I\right)^{-1}-\left(B_{2}^{(k)}-t I\right)^{-1}\right\|=  \tag{1.57}\\
\left\|\left(B_{1}^{(k)}-t I\right)^{-1}\left(B_{2}^{(k)}\right)\left(B_{2}^{(k)}-t I\right)^{-1}\right\| \leq(2 \beta) M_{1}^{2} t^{-2}
\end{gather*}
$$

If we use this estimate in (1.56) and recall that $\mu$ is a finite measure, we find that for any $\varepsilon>0$ and any $x \in H$, there exists a constant $M$ depending only on $\varepsilon, \beta,\|x\|$ and $\mu$ such that for all $k \geq 1$

$$
\begin{equation*}
\int_{-\infty}^{-M}\left(1+t^{2}\right)\left\|\left(B_{1}^{(k)}-t I\right)^{-1} x-\left(B_{2}^{(k)}-t I\right)^{-1} x\right\| \mu(d t)<\varepsilon / 2 \tag{1.58}
\end{equation*}
$$

On the other hand, the Cauchy-Schwartz inequality implies that

$$
\begin{equation*}
\int_{-M}^{0}\left(1+t^{2}\right)\left\|\left(B_{1}^{(k)}-t I\right)^{-1} x-\left(B_{2}^{(k)}-t I\right)^{-1} x\right\| \mu(d t) \leq \tag{1.59}
\end{equation*}
$$

$\left\{\int_{-M}^{0}\left(1+t^{2}\right)\left\|\left(B_{1}^{(k)}-t I\right)^{-1} x-\left(B_{2}^{(k)}-t I\right)^{-1} x\right\|^{2} \mu(d t)\right\}^{1 / 2}\left(\int_{-M}^{0}\left(1+t^{2}\right) \mu(d t)\right)^{1 / 2}$.
(1.55) implies that, for fixed $x \in H$, the right side of (1.59) approaches zero as $k \rightarrow \infty$ and hence is less than $\varepsilon / 2$ for $k$ sufficiently large. Combining inequalities (1.58) and (1.59) gives, for $k$ sufficiently large,

$$
\int_{-\infty}^{0}\left(1+t^{2}\right)\left\|\left(B_{1}^{(k)}-t I\right)^{-1} x-\left(B_{2}^{(k)}-t I\right)^{-1} x\right\| \mu(d t)<\varepsilon
$$

Using (1.56), we see that

$$
\lim _{k \uparrow \infty}\left\|\phi\left(B_{1}^{(k)}\right) x-\phi\left(B_{2}^{(k)}\right) x\right\|=0
$$

As already remarked, the same argument shows that $\phi\left(B_{i}^{(k)}\right)-\phi\left(B_{j}^{(k)}\right) \rightarrow 0$ for any $i$ and $j$.

Suppose now that there exist positive reals $u_{1}, u_{2}, \cdots, u_{m}$ such that $\sum_{p=1}^{m} u_{p}=1$ and

$$
\begin{equation*}
s-\lim _{k \uparrow \infty} \sum_{p=1}^{m} u_{p} \phi\left(B_{p}^{(k)}\right)=E . \tag{1.60}
\end{equation*}
$$

For any fixed $i, 1 \leq i \leq m$, (1.60) can be rewritten as

$$
s-\lim _{k \uparrow \infty}\left[\sum_{p=1}^{m} u_{p} \phi\left(B_{i}^{(k)}\right)+\sum_{p=1}^{m} u_{p}\left(\phi\left(B_{p}^{(k)}\right)-\phi\left(B_{i}^{(k)}\right)\right)\right]=E
$$

which implies that

$$
s-\lim _{k \uparrow \infty} \phi\left(B_{i}^{(k)}\right)=s-\lim _{k \uparrow \infty} \sum_{p=1}^{m} u_{p} \phi\left(B_{i}^{(k)}\right)=E
$$

Finally, if $\phi$ is one-one an open neighborhood of $(0, \infty)$ in $\mathbb{C}$, then $\phi^{-1}$ is defined and analytic on an open neighborhood of $\phi([\alpha, \beta])$ in $\mathbb{C}$. Lemma 1.2 implies that

$$
s-\lim _{k \uparrow \infty} \phi^{-1}\left(\phi\left(B_{i}^{(k)}\right)\right)=s-\lim _{k \uparrow \infty} B_{i}^{(k)}=\phi^{-1}(E)
$$

REMARK 1.3. The functions $\phi(z)=\log (z)$ and $\phi(z)=z^{p}, 0<p<1$, satisfy the hypotheses of Theorem 1.2.

Remark 1.4. Suppose that $K$ and $H$ are as in Theorem 1.2, that $f$ is as in Lemma 1.3 and that there do not exist real constants $\alpha$ and $\beta$ such that $f(x)=\alpha+\beta x$ for all $x>0$. Then Lemma 1.6 implies that $A \rightarrow f(A)$ is a strictly concave map from $K^{0}$ to $\mathcal{L}(H)$. Define $f=\phi$ and assume that (1.49) and (1.50) are satisfied. if $H$ is finite dimensional, it follows from the strict concavity of $f$ and a simple compactness argument that for all $i$ and $j, 1 \leq i, j \leq m$,

$$
\lim _{k \uparrow \infty}\left\|B_{i}^{(k)}-B_{j}^{(k)}\right\|=0
$$

Thus, if $H$ is finite dimensional, Theorem 1.2 follows trivially from Lemma 1.6. Theorem 1.2 only provides new information when $H$ is infinite dimensional.

Theorems 1.1 and 1.2 are examples of convergence results in particular cones. However, one can give versions of Theorem 1.1 which are valid for general classes of cones. Since we shall not use such results, we shall not prove them here, but it may be of interest to state the theorems.

THEOREM 1.3. Let $K$ be a cone with nonempty interior in a finitedimensional Banach space $\boldsymbol{X}$. Let $C$ denote the $n$-fold Cartesian product of $K$. Let $Y$ denote the $n$-fold Cartesian product of $\boldsymbol{X}$. Suppose that $f: C^{0} \rightarrow C^{0}$ and $\phi: K^{0} \rightarrow X$ are continuous maps. For any $y \in C^{0}$ assume that there exists $z \in C^{0}$ (dependent on $y$ ) and positive constants $\alpha$ and $\beta$ such that

$$
\begin{gathered}
\alpha z \leq f^{j}(y) \leq \beta z, \quad \text { and } \\
\Phi\left(f^{j}(y)\right) \leq \Phi(\beta z)
\end{gathered}
$$

for all $j \geq 0$, where $\Phi: C^{0} \rightarrow Y$ is defined by $\Phi\left(x_{1}, x_{2}, \cdots, x_{n}\right)=$ $\left(\phi\left(x_{1}\right), \phi\left(x_{2}\right), \cdots, \phi\left(x_{n}\right)\right.$.

Assume that $M$ is an $n \times n$, primitive column-stochastic matrix with nonnegative entries such that

$$
\Phi(f(x)) \geq \Phi(x) M \quad \text { for all } x \in C^{0}
$$

Let $u$ be the unique positive column vector such that $\sum_{i=1}^{n} u_{i}=1$ and $M u=u$. If, for a given $x \in C^{0}$, we write

$$
f^{k}(x)=\left(x_{1}^{(k)}, x_{2}^{(k)}, \cdots, x_{n}^{(k)}\right)
$$

then there exists $w \in X, w$ dependent on $x$, such that

$$
\lim _{k \uparrow \infty} \phi\left(x_{i}^{(k)}\right)=w \quad \text { for } 1 \leq i \leq n
$$

If $\phi$ is one-one, then $w \in \phi\left(K^{0}\right)$ and

$$
\lim _{k \uparrow \infty} x_{i}^{(k)}=\phi^{-1}(w) \quad \text { for } 1 \leq i \leq n
$$

The proof of Theorem 1.3 is basically the same as that of Theorem 1.1. One uses finite dimensionality to insure that the cone is "normal", i.e. that for each $w \in K$, the set $\{v \in K: v \leq w\}$ is bounded in norm. One also uses finite dimensionality to guarantee that for every $y \in C^{0},\left\{f^{j}(y): j \geq 1\right\}$ is precompact.

To generalize Theorem 1.3 to certain infinite dimensional cones, we need some terminology. Suppose that $\tau$ is a topology on a Banach space $X$ and that $\boldsymbol{X}$ becomes a Hausdorff, locally convex topological vector space in the $\tau$ topology and the $\tau$ topology is coarser than the norm topology (so every $\boldsymbol{\tau}$-open set is open in the norm topology). If $K$ is a cone in $X$ we shall say that a sequence $\left(x_{j}\right)$ of elements of $K$ is monotonic increasing (with respect to the ordering induced by $K$ ) if $x_{j} \leq x_{j+1}$ for all $j \geq 1$, and we shall say that $\left(x_{j}\right)$ is bounded above in the partial ordering induced by $K$ if there exists $w \in K$ such that $x_{j} \leq w$ for all $j \geq 1$. If $K, X$ and $\tau$ are as above, we shall say that $K$ has the "monotone convergence property in the $\tau$ topology" if every monotonic increasing sequence $\left(x_{j}\right)$ in $K$ such that $\left(x_{j}\right)$ is bounded above in the partial ordering induced by $K$ has a limit $z \in K$ in the $\tau$ topology. In the situation of Theorem 1.1, $X=\mathcal{L}(H), K$ is the cone of p.s.d. operators in $X, \mathcal{T}$ is the strong operator topology and $K$ has the monotonic convergence property in the $\tau$ topology.

THEOREM 1.4. Let notation and assumptions be as in Theorem 1.3, except do not assume that $X$ is finite dimensional. Suppose that $\tau$ is a topology on $\boldsymbol{X}$, coarser than the norm topology, such that $\boldsymbol{X}$ is a Hausdorff, locally convex topological vector space in the $\tau$ topology and $K$ has the monotone convergence property in the $\tau$ topology. If $u$ is the unique normalized positive eigenvector of $M$ and $f^{k}(x)=\left(x_{1}^{(k)}, x_{2}^{(k)}, \cdots, x_{n}^{(k)}\right)$, there exists $w \in X$ so that

$$
\lim _{k \uparrow \infty} \sum_{i=1}^{n} u_{i} \phi\left(x_{i}^{(k)}\right)=w
$$

where convergence is in the $\tau$ topology and $u_{i}$ is the ith component of $u$. If $L: X \rightarrow \mathbb{C}$ is any linear map such that $L(x) \geq 0$ for all $x \in K$ and $L$ is continuous in the $\tau$ topology, then

$$
\lim _{k \uparrow \infty} L\left(\phi\left(x_{i}^{(k)}\right)\right)=L(w) \quad \text { for } 1 \leq i \leq n
$$

The proof of Theorem 1.4 is very similar to that of Theorem 1.1. If $y \in H$, the role of $L$ in Theorem 1.4 is served by $L(A)=<A y, y>$ for $A \in \mathcal{L}(H)$.

## 2. - Convergence results for generalizations of the $A G M$

The original motivation for this paper was the problem of proving the convergence of $f^{k}(A, B)$ for the maps

$$
\begin{align*}
& f(A, B)=((A+B) / 2, \quad \exp ((\log A) / 2+(\log B) / 2) \text { or }  \tag{2.1}\\
& f(A, B)=(\alpha A+(1-\alpha) B, \exp (\beta \log A+(1-\beta) \log B))
\end{align*}
$$

where $A$ and $B$ are positive definite, bounded linear operators on a Hilbert space. However, there are many generalizations of the classical Gauss-LagrangeLegendre $A G M$, and to treat the extensions of these generalizations to the operator-valued case in a reasonably unified way it is necessary to consider much more general $f$ than those in (2.1).

Thus, if $H$ is a Hilbert space, let $K$ denote the cone of p.s.d. operators in $\mathcal{L}(H)$. Let $C$ denote the $n$-fold Cartesian product of $K$ with itself. If $\sigma \in \mathbb{R}^{n}$, call $\sigma$ a "probability vector" if all components $\sigma_{i}$ of $\sigma$ are nonnegative and $\sum_{i=1}^{n} \sigma_{i}=1$. If $r$ is a real number, $\sigma$ is a probability vector and $A=\left(A_{1}, A_{2}, \cdots, \boldsymbol{A}_{n}\right) \in C^{0}$, define a map $M_{r \sigma}: C^{0} \rightarrow K^{0}$ by

$$
\begin{equation*}
M_{r \sigma}(A)=\left(\sum_{j=1}^{n} \sigma_{j} A_{j}^{r}\right)^{1 / r} \tag{2.2}
\end{equation*}
$$

If $r=0$, (2.2) does not make sense and we define

$$
\begin{equation*}
M_{0 \sigma}(A)=\exp \left(\sum_{j=1}^{n} \sigma_{j} \log \left(A_{j}\right)\right) \tag{2.3}
\end{equation*}
$$

If $A \in C^{0}$, then

$$
\begin{equation*}
\lim _{r \rightarrow 0}\left\|M_{r \sigma}(A)-M_{0 \sigma}(A)\right\|=0 \tag{2.4}
\end{equation*}
$$

Theorem 2.1. Let $K, H$ and $C$ be as above. For each $i, 1 \leq i \leq n$, let $\Gamma_{i}$ be a finite collection of ordered pairs $(r, \sigma)$ such that $r$ is a nonnegative real and $\sigma$ is a probability vector. For each $(r, \sigma) \in \Gamma_{i}$, let $c_{i r \sigma}$ be a positive real number. Define $f: C^{0} \rightarrow C^{0}$ by

$$
\begin{equation*}
f_{j}(A)=\sum_{(r . \sigma) \in \Gamma_{j}} c_{j r \sigma} M_{r \sigma}(A) \tag{2.5}
\end{equation*}
$$

where $f_{j}(A)$ denotes the $j$ th component of $f(A)$. Assume that

$$
\begin{equation*}
\sum_{(r . \sigma) \in \Gamma,} c_{j r \sigma}=1 \quad \text { for } 1 \leq j \leq n \tag{2.6}
\end{equation*}
$$

If $\pi_{i}: n \rightarrow$ denotes the projection onto the ith component of a vector, define

$$
\begin{equation*}
m_{i j}=\sum_{(r . \sigma) \in \Gamma,} c_{j r \sigma} \pi_{i}(\sigma) \tag{2.7}
\end{equation*}
$$

and assume that the $n \times n$ matrix $M=\left(m_{i j}\right)$ is primitive. Let u (u a column vector) denote the unique probability vector such that $M u=u$ and let $u_{i}=\pi_{i}(u)$. If, for $A=\left(A_{1}, A_{2}, \cdots, A_{n}\right) \in C^{0}$, we write

$$
\begin{equation*}
f^{k}(A)=\left(A_{1}^{(k)}, A_{2}^{(k)}, \cdots, A_{n}^{(k)}\right)=A^{(k)} \tag{2.8}
\end{equation*}
$$

there exists $E \in K^{0}$ such that

$$
\begin{equation*}
s-\lim _{k \uparrow \infty} \sum_{i=1}^{n} u_{i} \log \left(A_{i}^{(k)}\right)=\log E \tag{2.9}
\end{equation*}
$$

and

$$
\begin{equation*}
w-\lim _{k \uparrow \infty} \log \left(A_{i}^{(k)}\right)=\log E \quad \text { for } 1 \leq i \leq n \tag{2.10}
\end{equation*}
$$

If $(r, \sigma)$ and $(\rho, \tau)$ are both elements of $\Gamma_{i}$ for some $i, 1 \leq i \leq n$, then

$$
\begin{equation*}
s-\lim _{k \uparrow \infty}\left[\log M_{r \sigma}\left(A^{(k)}\right)-\log M_{\rho \tau}\left(A^{(k)}\right)\right]=0 \tag{2.11}
\end{equation*}
$$

If $(r, \sigma) \in \Gamma_{i}$ for some $i$ and $r>0$, then for all $p$ and $q$ such that $\pi_{p}(\sigma)>0$ and $\pi_{q}(\sigma)>0$ one has

$$
\begin{equation*}
s-\lim _{k \uparrow \infty}\left[\log \left(A_{p}^{(k)}\right)-\log \left(A_{q}^{(k)}\right)\right]=0 \tag{2.12}
\end{equation*}
$$

If $H$ is finite dimensional,

$$
\begin{equation*}
\lim _{k \uparrow \infty}\left\|A_{i}^{(k)}-E\right\|=0 \text { for } 1 \leq i \leq n \tag{2.13}
\end{equation*}
$$

If there exists $(\bar{r}, \bar{\sigma}) \in \Gamma_{i}$ for some $i$ such that $\bar{r}>0$ and all components of $\bar{\sigma}$ are positive, then

$$
\begin{equation*}
s-\lim _{k \uparrow \infty} A_{i}^{(k)}=E \quad \text { for } 1 \leq i \leq n \tag{2.14}
\end{equation*}
$$

Proof. If $A=\left(A_{1}, A_{2}, \cdots, A_{n}\right) \in C^{0}$ and $B=\left(B_{1}, B_{2}, \cdots, B_{n}\right)=f(A)$ and if we define $S_{1 j}=S_{1 j}(A)$ and $S_{2 j}=S_{2 j}(A)$ by

$$
\begin{equation*}
S_{1 j}=\log \left(\sum_{(r . \sigma) \in \Gamma,} c_{j r \sigma} M_{r \sigma}(A)\right)-\sum_{(r . \sigma) \in \Gamma,} c_{j r \sigma} \log M_{r \sigma}(A) \tag{2.15}
\end{equation*}
$$

and

$$
\begin{equation*}
S_{2 j}=\sum_{r \neq 0} c_{j r \sigma} r^{-1}\left[\log \left(\sum_{p=1}^{n} \pi_{p}(\sigma) A_{p}^{r}\right)-\sum_{p=1}^{n} \pi_{p}(\sigma) \log \left(A_{p}^{r}\right)\right] \tag{2.16}
\end{equation*}
$$

we find that

$$
\begin{equation*}
\log B_{j}-\sum_{i} m_{i j} \log A_{i}=S_{1 j}+S_{2 j} \tag{2.17}
\end{equation*}
$$

The concavity of $D \rightarrow \log D=\phi(D)$, for $D \in K^{0}$, implies that the right side of (2.17) is positive semidefinite. It follows that if $\Phi(A)$ is defined as in Theorem 1.1 and $\phi=\log$ we have

$$
\begin{equation*}
\Phi(f(A)) \geq \Phi(A) M \quad \text { for all } A \in C^{0} \tag{2.18}
\end{equation*}
$$

If $A=\left(A_{1}, A_{2}, \cdots, A_{n}\right) \in C^{0}$ and $\alpha$ and $\beta$ are positive numbers such that

$$
\begin{equation*}
\alpha I \leq A_{j} \leq \beta I \tag{2.19}
\end{equation*}
$$

the spectral mapping theorem implies that for $r>0$

$$
\alpha^{r} I \leq A_{j}^{r} \leq \beta^{r} I
$$

and

$$
(\log \alpha) I \leq \log A_{j} \leq(\log \beta) I
$$

If $\sigma$ is any probability vector, it follows that

$$
\alpha^{r} I \leq \sum_{j=1}^{n} \sigma_{j} A_{j}^{r} \leq \beta^{r} I
$$

and

$$
(\log \alpha) I \leq \sum_{j=1}^{n} \sigma_{j} \log A_{j} \leq(\log \beta) I
$$

and by applying the spectral mapping theorem again we conclude that

$$
\begin{equation*}
\alpha I \leq M_{r \sigma}(A) \leq \beta I . \tag{2.20}
\end{equation*}
$$

A variant of this argument shows that (2.20) also holds if $r<0$. We obtain directly from (2.20) that

$$
\alpha I \leq \sum_{(r . \sigma) \in \Gamma,} c_{j r \sigma} M_{r \sigma}(A) \leq \beta I
$$

or

$$
\alpha I \leq A_{j}^{(1)} \leq \beta I \quad \text { for } 1 \leq j \leq n
$$

A simple induction now shows that if $A$ satisfies (2.19) and the notation is as in (2.8) then

$$
\begin{equation*}
\alpha I \leq A_{i}^{(k)} \leq \beta I \quad \text { for } 1 \leq i \leq n \text { and } k \geq 1 \tag{2.21}
\end{equation*}
$$

and (2.21) implies that

$$
\begin{equation*}
(\log \alpha) I \leq \log A_{i}^{(k)} \leq(\log \beta) I, \quad 1 \leq i \leq n \text { and } k \geq 1 \tag{2.22}
\end{equation*}
$$

Inequalities (2.20) - (2.22) verify the hypotheses of Theorem 1.1, so Theorem 1.1 gives (2.9) and (2.10). If $H$ is finite dimensional, weak convergence implies norm convergence and by applying the exponential map to (2.10) we obtain (2.13).

If $H$ is infinite dimensional, more care is necessary. If we replace $A$ by $A^{(k)}=f^{k}(A)$ and $B$ by $A^{(k+1)}$ in (2.17) then

$$
\begin{equation*}
\log A_{j}^{(k+1)}-\sum_{i} m_{i j} \log A_{i}^{(k)}=S_{1 j}\left(A^{(k)}\right)+S_{2 j}\left(A^{(k)}\right) \tag{2.23}
\end{equation*}
$$

By using (2.10) and the fact that $\sum_{i} m_{i j}=1$, we conclude that the left side of (2.23) converges to zero in the weak operator topology. Since $S_{1 j}\left(A^{(k)}\right)$ is p.s.d. and all summands of $S_{2 j}\left(A^{(k)}\right)$ are p.s.d., (2.23) implies that

$$
\begin{equation*}
w-\lim _{k \uparrow \infty} S_{1 j}\left(A^{(k)}\right)=0, \quad 1 \leq k \leq n \tag{2.24}
\end{equation*}
$$

and

$$
\begin{equation*}
w-\lim _{k \uparrow \infty}\left[\log \left(\sum_{p=1}^{n} \pi_{p}(\sigma)\left(A_{p}^{(k)}\right)^{r}\right)-\sum_{p=1}^{n} \pi_{p}(\sigma) \log \left(\left(A_{p}^{(k)}\right)^{r}\right)\right]=0 \tag{2.25}
\end{equation*}
$$

where (2.25) is satisfied if $(r, \sigma) \in \Gamma_{j}$ for some $j$ and $r>0$. We now apply Theorem 1.2 (recalling that $\phi(z)=\log z$ satisfies the hypotheses of Theorem 1.2). Using Theorem 1.2 and (2.24) we find that (2.11) holds, and Theorem 1.2 and (2.25) imply (2.12). If there exists $(\bar{r}, \bar{\sigma})$ as in the statement of the theorem, we obtain from (2.12) that

$$
\begin{equation*}
s-\lim _{k \nmid \infty}\left[\log A_{i}^{(k)}-\log A_{j}^{(k)}\right]=0 \quad \text { for } 1 \leq i, j \leq n \tag{2.26}
\end{equation*}
$$

Combining (2.9) and (2.26) we obtain as in Theorem 1.2,

$$
s-\lim _{k \uparrow \infty} \log A_{i}^{(k)}=\log E \quad \text { for } 1 \leq i \leq n
$$

and (2.14) now follows easily.
Theorem 2.1 provides insufficient information if $H$ is infinite dimensional and $r=0$ for all $(r, \sigma) \in \Gamma_{j}, 1 \leq j \leq n$. We consider this case separately in the next theorem.

THEOREM 2.2. Let the notation and assumptions be as in Theorem 2.1. For $1 \leq i \leq n$, assume that if $(r, \sigma) \in \Gamma_{i}$, then $r=0$, so $\Gamma_{i}$ can be considered a finite set of probability vectors and we can write

$$
f_{i}(A)=\sum_{\sigma \in \Gamma_{t}} c_{i \sigma} M_{0 \sigma}(A), \quad c_{i \sigma}=c_{i 0 \sigma}
$$

Let $u$ be the unique probability vector such that $M u=u$. Assume that there are $n-1$ pairs of probability vectors $\sigma^{(j)}$ and $\tau^{(j)}, 2 \leq j \leq n$, such that $\sigma^{(j)}$ and $\tau^{(j)} \in \Gamma_{i}$ for some $i$ depending on $j$ and such that the $n-1$ vectors, $\alpha^{(j)}=\sigma^{(j)}-\tau^{(j)}, 2 \leq j \leq n$, are linearly independent. Then for any $A \in C^{0}$, there exists $E \in K^{0}$ such that

$$
\begin{equation*}
s-\lim _{k \uparrow \infty} f^{k}(A)=(E, E, \cdots, E) \tag{2.27}
\end{equation*}
$$

If $n=2$ or if $H$ is finite dimensional, (2.27) remains valid without the assumption that there exist vectors $\alpha^{(j)}, 2 \leq j \leq n$, as above.

PROOF. Theorem 2.1 implies that if $\sigma$ and $\tau$ are any two probability vectors in $\Gamma_{j}, 1 \leq j \leq n$, then

$$
s-\lim _{k \uparrow \infty}\left[\log M_{0 \sigma}\left(A^{(k)}\right)-\log M_{0 \tau}\left(A^{(k)}\right)\right]=0
$$

or equivalently

$$
\begin{equation*}
s-\lim _{k \uparrow \infty} \sum_{i=1}^{n}\left(\sigma_{i}-\tau_{i}\right) \log \left(A_{i}^{(k)}\right)=0 \tag{2.28}
\end{equation*}
$$

If $u$ is the eigenvector of $M$ in the statement of Theorem 2.1 , let $N$ be the $n \times n$ matrix whose first column is $u$ and whose $j$ th column is $\alpha^{(j)}$ for $2 \leq j \leq n$. Equations (2.9) and (2.28) imply that, in the notation of Theorem 2.1,

$$
\begin{equation*}
s-\lim _{k \uparrow \infty} \Phi\left(A^{(k)}\right) N=(\log E, 0,0, \cdots, 0) \tag{2.29}
\end{equation*}
$$

where

$$
\Phi\left(A^{(k)}\right)=\left(\log A_{1}^{(k)}, \log A_{2}^{(k)}, \cdots, \log A_{n}^{(k)}\right)
$$

Because the components of $\alpha^{(j)}, 2 \leq j \leq n$, sum to zero and the components of $u$ sum to one, it is easy to see that if we define $\alpha^{(1)}=u$, the $n$ vectors
$\alpha^{(j)}, 1 \leq j \leq n$, are linearly independent. This implies that $N$ is invertible, and since

$$
\Phi(E) N=(\log E, 0,0, \cdots, 0)
$$

we conclude that

$$
s-\lim _{k \uparrow \infty} \Phi\left(A^{(k)}\right)=(\log E, 0,0, \cdots, 0) N^{-1}=(\log E, \log E, \cdots, \log E)
$$

which (with the aid of Lemma 1.2) gives (2.27).
If $H$ is finite dimensional, (2.27) follows directly from Theorem 2.1 without any knowledge of the vectors $\alpha^{(j)}$. If $n=2$ and $\Gamma_{1}$ or $\Gamma_{2}$ contains more than one element, there is a nonzero vector $\alpha^{(2)}=\sigma-\tau$ ( $\sigma$ and $\tau$ in $\Gamma_{i}$ for $i=1$ or 2 ) and the theorem follows from our previous remarks. Thus assume that $n=2$ and that $\Gamma_{1}$ and $\Gamma_{2}$ each contains only one element, say $\Gamma_{1}=\{\sigma\}$ and $\Gamma_{2}=\{\tau\}$. If $\sigma=\left(\sigma_{1}, \sigma_{2}\right)$ and $\tau=\left(\tau_{1}, \tau_{2}\right)$, we find that

$$
\left(\log A_{1}^{(k+1)}, \log A_{2}^{(k+1)}\right)=\left(\log A_{1}^{(k)}, \log A_{2}^{(k)}\right) M
$$

where

$$
M=\left(\begin{array}{ll}
\sigma_{1} & \tau_{1} \\
\sigma_{2} & \tau_{2}
\end{array}\right)
$$

and $M$ is assumed primitive. It follows that

$$
\left(\log A_{1}^{(k)}, \log A_{2}^{(k)}\right)=\left(\log A_{1}, \log A_{2}\right) M^{k}
$$

and

$$
n-\lim _{k \uparrow \infty} \log A_{1}^{(k)}=n-\lim _{k \uparrow \infty} \log A_{2}^{(k)}=u_{1} \log A_{1}+u_{2} \log A_{2}
$$

where the column vector $\left(u_{1}, u_{2}\right)^{T}$ is the unique probability vector which is an eigenvector for $M$. One obtains the norm convergence of $A_{1}^{(k)}$ and $A_{2}^{(k)}$ from the above equation.

REMARK 2.1. Theorems 2.1 and 2.2 are not sharp results if $H$ is infinite dimensional. It is possible that strong convergence of $\left(\boldsymbol{A}_{i}^{(k)}\right), 1 \leq i \leq n$, is valid with only the assumption that $M$ is primitive, but we have not been able to prove this.

It is worth noting that the conclusions of Theorem 2.2 remain valid if there exist $n-1$ linearly independent vectors $\alpha^{(j)}, 2 \leq j \leq n$, such that

$$
s-\lim _{k \uparrow \infty} \sum_{i=1}^{n} \alpha_{i}^{(j)} \log A_{i}^{(k)}=0
$$

and

$$
\sum_{i=1}^{n} \alpha_{i}^{(j)}=0
$$

The vectors $\alpha^{(j)}$ do not have to arise as in Theorem 2.2.
REMARK 2.2. Let notation and assumptions be as in Theorem 2.2 but do not assume the existence of probability vectors $\sigma^{(j)}$ and $\tau^{(j)}$ as in Theorem 2.2. Instead suppose that

$$
\begin{equation*}
m_{i j}=\sum_{\sigma \in \Gamma,} c_{j \sigma} \pi_{i}(\sigma)=u_{i} \quad \text { for } 1 \leq i, j \leq n \tag{2.30}
\end{equation*}
$$

Then the conclusion of Theorem 2.2 (in particular, (2.27)) still holds.
To see this, note that $M=\left(m_{i j}\right)$ is the matrix in Theorem 2.1 and that the column vector $u$ whose $i$ th entry equals $u_{i}\left(u_{i}\right.$ as in (2.30)) satisfies $M u=u$. Theorem 2.1 implies that for all $\sigma, \tau \in \Gamma_{j}, 1 \leq j \leq n$,

$$
\begin{equation*}
s-\lim _{k \nmid \infty}\left[\log M_{0 \sigma}\left(A^{(k)}\right)-\log M_{0 \tau}\left(A^{(k)}\right)\right]=0 \tag{2.31}
\end{equation*}
$$

and there exists $E \in K^{0}$ so that

$$
s-\lim _{k \uparrow \infty}\left[\sum_{j=1}^{n} u_{j} \log \left(A_{j}^{(k)}\right)\right]=\log E .
$$

Because $m_{i j}=u_{i}$ we have

$$
\sum_{\sigma \in \Gamma,} c_{j \sigma} \log M_{0 \sigma}\left(A^{(k)}\right)=\sum_{i=1}^{n} u_{i} \log \left(A_{i}^{(k)}\right)
$$

so

$$
\begin{equation*}
s-\lim _{k \uparrow \infty}\left[\sum_{\sigma \in \Gamma_{j}} c_{j \sigma} \log M_{0 \sigma}\left(A^{(k)}\right)\right]=\log E \tag{2.32}
\end{equation*}
$$

Combining (2.31) and (2.32) we obtain

$$
s-\lim _{k \uparrow \infty} \log M_{0 \sigma}\left(A^{(k)}\right)=\log E
$$

so Lemma 1.2 implies

$$
\begin{equation*}
s-\lim _{k \uparrow \infty} M_{0 \sigma}\left(A^{(k)}\right)=E \tag{2.33}
\end{equation*}
$$

Using (2.33) we see that

$$
\begin{equation*}
s-\lim _{k \nmid \infty} A_{j}^{(k+1)}=s-\lim _{k \uparrow \infty} \sum_{\sigma \in \Gamma,} c_{j \sigma} M_{0 \sigma}\left(A^{(k)}\right)=E \tag{2.34}
\end{equation*}
$$

which is the desired result.

Although the assumptions on $M$ in Remark 2.2 are restrictive, they are satisfied in some important applications. Remark 2.2 also provides further evidence that Theorem 2.2 is far from best possible.

It is useful in some applications to allow functions which are the composition of functions like those in Theorems 2.1 or 2.2. One can give analogues of Theorems 2.1 and 2.2 for such functions even if $H$ is infinite dimensional, but for simplicity we restrict ourselves to the finite dimensional case.

Theorem 2.3. Let $K, H, C, \phi$ and $\Phi$ be as in Theorem 2.1 and suppose that $H$ is finite dimensional. Assume that $f: C^{0} \rightarrow C^{0}$ and $g: C^{0} \rightarrow C^{0}$ are continuous maps and define $h=\mathrm{g} \circ \mathrm{f}$. Assume that for every $A \in C^{0}$ there exist $B \in C^{0}$ and positive reals $\alpha$ and $\beta$ such that

$$
\alpha B \leq h^{j}(A) \leq \beta B
$$

and

$$
\Phi\left(h^{j}(A)\right) \leq \Phi(\beta B)
$$

for all $j \geq 0$. Assume that there exist $n \times n$ column-stochastic matrices $M$ and $N$ with nonnegative entries such that for every $A \in C^{0}$

$$
\Phi(f(A)) \geq \Phi(A) M
$$

and

$$
\Phi(g(A)) \geq \Phi(A) N
$$

and $M N$ is primitive. Then there exists $E \in K^{0}$ such that, if $h^{k}(A)=$ $\left(\boldsymbol{A}_{1}^{(k)}, \boldsymbol{A}_{2}^{(k)}, \cdots, A_{n}^{(k)}\right)$,

$$
\lim _{k \uparrow \infty}\left\|\phi\left(A_{i}^{(k)}\right)-\phi(E)\right\|=0 \quad \text { for } 1 \leq i \leq n
$$

If $\phi$ is one-one, one also obtains

$$
\lim _{k \uparrow \infty}\left\|A_{i}^{(k)}-E\right\|=0
$$

Proof. By using the hypotheses on $f$ and $g$ one finds

$$
\Phi(h(A)) \geq \Phi(f(A)) N \geq \Phi(A) M N .
$$

Thus $h$ satisfies the hypotheses of Theorem 1.1 , with $M N$ replacing $M$ in Theorem 1.1, and Theorem 2.3 follows immediately from Theorem 1.1.

COROLLARY 2.1. Let $K, C$ and $H$ be as in Theorem 2.1 and assume that $H$ is finite dimensional. Let $f$ be as in Theorem 2.1, but do not assume that the matrix $M$ defined by (2.7) is primitive. Suppose that $g: C^{0} \rightarrow C^{0}$ is like
$f$. More precisely, for $1 \leq i \leq n$, let $T_{i}$ be a finite collection of ordered pairs $(s, r)$ such that $s$ is a nonnegative real and $\tau$ is a probability vector. For each $(s, \tau) \in T_{i}$, let $d_{i s \tau}$ be a positive real number. Define $g: C^{0} \rightarrow C^{0}$ by

$$
g_{j}(A)=\sum_{(s, \tau) \in T,} d_{j s \tau} M_{s \tau}(A)
$$

Assume that

$$
\sum_{(s . \tau) \in T,} d_{j s \tau}=1 \quad \text { for } 1 \leq j \leq n
$$

and define an $n \times n$ column-stochastic matrix $P$ by

$$
p_{i j}=\sum_{(s, \tau) \in T_{j}} d_{j s \tau} \pi_{i}(\tau)
$$

If $M P$ is primitive and $h=\mathrm{g} \circ \mathrm{f}$, then for any $A \in C^{0}$ there exists $E \in K^{0}$ such that

$$
\lim _{k \uparrow \infty}\left\|A_{i}^{(k)}-E\right\|=0 \quad \text { for } 1 \leq i \leq n
$$

where

$$
h^{k}(A)=\left(A_{1}^{(k)}, A_{2}^{(k)}, \cdots, A_{n}^{(k)}\right)
$$

Proof. If $\phi=\log$, the proof of Theorem 2.1 shows that

$$
\Phi(f(A)) \geq \Phi(A) M \text { and } \Phi(g(A)) \geq \Phi(A) P
$$

for all $A \in C^{0}$. Thus Corollary 2.1 follows easily from Theorem 2.3. Details are left to the reader.

Theorems 2.1 and 2.2 provide no information if $f$ is given as in (2.5) and $r<0$ for some $(r, \sigma) \in \Gamma_{j}$. However, Theorems 1.1 and 1.2 provide information about certain functions of this type also.

Theorem 2.4. Let $K, C$ and $H$ be as in Theorem 2.1. For $1 \leq j \leq n$, let $\Gamma_{j}$ be a finite collection of probability vectors and for each $\sigma \in \Gamma_{j}$ let $c_{j \sigma}$ be a positive real number such that

$$
\sum_{\sigma \in \Gamma_{j}} c_{j \sigma}=1
$$

Define a map $f: C^{0} \rightarrow C^{0}$ by

$$
f_{j}(A)=\sum_{\sigma \in \Gamma,} c_{j \sigma} M_{-1 \sigma}(A)
$$

where $f_{j}(A)$ denotes the $j$ th component of $f(A)$ and

$$
M_{-1 \sigma}(A)=\left(\sum_{i=1}^{n} \sigma_{i} A_{i}^{-1}\right)^{-1}
$$

for $\sigma=\left(\sigma_{1}, \sigma_{2}, \cdots, \sigma_{n}\right)$ and $A=\left(A_{1}, A_{2}, \cdots, A_{n}\right)$. Define $m_{i j}$ by

$$
m_{i j}=\sum_{\sigma \in \Gamma,} \pi_{i}(\sigma) c_{j \sigma}
$$

and assume the $n \times n$ column-stochastic matrix $M=\left(m_{i j}\right)$ is primitive. Finally assume that there are $n-1$ pairs of probability vectors $\sigma^{(j)}$ and $\tau^{(j)}, 2 \leq j \leq n$, such that $\sigma^{(j)}$ and $\tau^{(j)}$ are in $\Gamma_{i}$ for some $i=i(j)$ and the $n-1$ vectors $\alpha^{(j)}=\sigma^{(j)}-\tau^{(j)}, 2 \leq j \leq n$, are linearly independent. Then for any $A \in C^{0}$, there exists $E \in K^{0}$ such that

$$
s-\lim _{k \uparrow \infty} f^{k}(A)=(E, E, \cdots, E)
$$

Proof. Define $\phi(z)=-z^{-1}$ and notice that $\phi(z)$ satisfies the conditions of Theorem 1.2 and that $B \rightarrow-B^{-1}\left(B \in K^{0}\right)$ is concave (see Section 1). Using the concavity of $\phi$ one easily sees that if $A \in C^{0}$ and $B=f(A)$,

$$
\begin{gather*}
\phi\left(B_{j}\right)-\sum_{i=1}^{n} m_{i j} \phi\left(A_{i}\right) \\
=\phi\left(\sum_{\sigma \in \Gamma_{j}} c_{j \sigma} M_{-1 \sigma}(A)\right)-\sum_{\sigma \in \Gamma_{j}} c_{j \sigma} \phi\left(M_{-1 \sigma}(A)\right) \tag{2.35}
\end{gather*}
$$

and the right side of the above equation is positive semidefinite. The rest of the proof follows from (2.35) by using Theorems 1.1 and 1.2 as in Theorem 2.2 and is left to the reader.

REMARK 2.3. It is important to note that if $\sigma$ is a probability vector, one can write (for $A \in C^{0}$ )

$$
M_{1 \sigma}(A)=\sum_{i=1}^{n} \sigma_{i} A_{i}=\sum_{i=1}^{n} \sigma_{i}\left(A_{i}^{-1}\right)^{-1}=\sum_{i=1}^{n} \sigma_{i} M_{-1 \tau(i)}(A)
$$

where $\tau(i)$ is the probability vector with 1 at the $i$ th position. Thus if $f$ is given as in (2.5) and $r=1$ or $r=-1$ for each $(r, \sigma) \in \Gamma_{j}$, then by relabelling and redefining $\Gamma_{i}$ one can assume that $f$ is as in Theorem 2.4.

By using the above remark and Theorem 2.4 we obtain the following corollary.

COROLLARY 2.2. Let the notation and assumptions be as in Theorem 2.4 except do not assume the existence of $n-1$ pairs $\sigma^{(j)}$ and $\tau^{(j)}$ as in Theorem 2.4. Assume that there exist $n$ positive numbers $d_{j}, 1 \leq j \leq n$, such that $f_{1}(A)$, the first component of $f(A)$, satisfies

$$
f_{1}(A)=f_{1}\left(A_{1}, A_{2}, \cdots, A_{n}\right)=\sum_{j=1}^{n} d_{j} A_{j}
$$

Then for any $A \in C^{0}$, there exists $E \in K^{0}$ such that

$$
s-\lim _{k \uparrow \infty} f^{k}(A)=(E, E, \cdots, E)
$$

Proof. As noted in Remark 2.3, write

$$
f_{1}(A)=\sum_{\sigma \in \Gamma_{1}} c_{\sigma} M_{-1 \sigma}(A)
$$

where $\Gamma_{1}$ comprises the $n$ probability vectors $\tau(j), 1 \leq j \leq n$, and $\tau(j)$ has 1 in the $j$ th position. Then

$$
\alpha^{(j)}=\tau(1)-\tau(j), \quad 2 \leq j \leq n
$$

are $n-1$ linearly independent vectors and the corollary follows from Theorem 2.4.

A very special case of Corollary 2.2 is an arithmetic-harmonic mean (the case $\alpha=\beta=1 / 2$ below) which has been considered by Fujii [19].

COROLLARY 2.3. Let $H$ be a Hilbert space, $K$ the cone of positive semidefinite operators in $\mathcal{L}(H)$, and $C=K \times K$. If $\alpha$ and $\beta$ are real numbers such that $0<\alpha, \beta<1$, define $f: C^{0} \rightarrow C^{0}$ by

$$
f(A, B)=\left(\alpha A+(1-\alpha) B,\left(\beta A^{-1}+(1-\beta) B^{-1}\right)^{-1}\right)
$$

If $(A, B) \in C^{0}$, there exists $E \in K^{0}$ such that

$$
s-\lim _{k \uparrow \infty} f^{k}(A, B)=(E, E)
$$

Although we shall not prove this here, one can prove convergence in the operator norm under the assumptions of Corollary 2.3.

Similarly, as a direct corollary of Theorem 2.2 we obtain an operator valued extension of the $A G M$ of the type suggested by (2.1).

Corollary 2.4. Let $K, C$ and $H$ be as in Corollary 2.3. If $\alpha$ and $\beta$ are real numbers such that $0<\alpha, \beta<1$, define $f: C^{0} \rightarrow C^{0}$ by

$$
f(A, B)=(\alpha A+(1-\alpha) B, \exp (\beta \log A+(1-\beta) \log B))
$$

Then for any $(A, B) \in C^{0}$, there exists $E \in K^{0}$ such that

$$
s-\lim _{k \uparrow \infty} f^{k}(A, B)=(E, E)
$$

Proof. Since $A$ can be written as $\exp (\log A)$ and similarly for $B$, the mapping $f$ is of the form considered in Theorem 2.2 and (in the notation of

Theorem 2.2) $n=2$. One easily checks that

$$
M=\left[\begin{array}{cc}
\alpha & \beta \\
1-\alpha & 1-\beta
\end{array}\right]
$$

so $M$ is primitive and the corollary follows from Theorem 2.2.
If $A$ and $B$ are positive real numbers in Corollary 2.4 and $\alpha=\beta=1 / 2$, it was known classically (see [13]) that one can express the limit of $f^{k}(A, B)$ in terms of an elliptic integral. However, as D. Borwein and P.B. Borwein have pointed out [11], for general $\alpha$ and $\beta$ (even if $\alpha=\beta$ ) there is no known integral formula for the limit of $f^{k}(A, B)$.

There are many other classical variants of the $A G M$. Carlson [13] gives a unified treatment of some of these results. In one example, Carlson defines (for $a$ and $b$ positive reals)

$$
\begin{gathered}
f_{1}(a, b)=(a+b) / 2, f_{2}(a, b)=(a b)^{1 / 2}, f_{3}(a, b)=[a(a+b) / 2]^{1 / 2} \\
\text { and } f_{4}(a, b)=[b(a+b) / 2]^{1 / 2}
\end{gathered}
$$

He then defines a map $f_{i j}(a, b)$ by

$$
f_{i j}(a, b)=\left(f_{i}(a, b), f_{j}(a, b)\right)
$$

The case $i=1$ and $j=3$ is usually attributed to Borchardt (see [6], [9]). One can prove that

$$
\lim _{k \uparrow \infty} f_{i j}^{k}(a, b)=(c, c), \quad c=L_{i j}(a, b)
$$

and Carlson gives explicit integral formulas for $L_{i j}(a, b)$.
We wish to generalize the above convergence results to pairs of p.d. bounded linear operators $A$ and $B$ on a Hilbert space. If $A$ and $B$ commute or if, as in Section 4 below, one uses a different analogue of the square root of the product of two positive numbers, one can also generalize the integral formulas $L_{i j}(a, b)$. However, we shall only carry this out for the case of the $A G M$ in Sections 3 and 4 below. Thus let $H$ be a Hilbert space and $K$ the cone of p.s.d. operators in $\mathcal{L}(H)$. If $A, B \in K^{0}$ and $r$ is a real number such that $0 \leq r \leq 1$, define

$$
\begin{equation*}
h(A, B, r)=\exp (r \log (A)+(1-r) \log (B)) \tag{2.36}
\end{equation*}
$$

Define a map $f$ of $K^{0} \times K^{0}$ into itself by

$$
\begin{equation*}
f(A, B)=\left(A_{1}, B_{1}\right) \tag{2.37}
\end{equation*}
$$

where

$$
\begin{equation*}
A_{1}=c_{0} A+c_{1} B+c_{2} h\left(A, B, \gamma_{2}\right)+c_{3} h\left(A, C, \gamma_{3}\right)+c_{4} h\left(B, C, \gamma_{4}\right) \tag{2.38}
\end{equation*}
$$

$$
\begin{gather*}
B_{1}=d_{0} A+d_{1} B+d_{2} h\left(A, B, \delta_{2}\right)+d_{3} h\left(A, C, \delta_{3}\right)+d_{4} h\left(B, C, \delta_{4}\right)  \tag{2.39}\\
C=(A+B) / 2
\end{gather*}
$$

It will always be assumed that

$$
\begin{equation*}
\sum_{j=0}^{4} c_{j}=\sum_{j=0}^{4} d_{j}=1 \tag{2.41}
\end{equation*}
$$

Corollary 2.5. Let $H$ be a Hilbert space and $K$ the cone of p.s.d. operators in $\mathcal{L}(H)$. Let $c_{j}$ and $d_{j}, 0 \leq j \leq 4$, be nonnegative real numbers which satisfy (2.41) and let $\gamma_{j}$ and $\delta_{j}, 2 \leq j \leq 4$, be real numbers such that $0<\gamma_{j}, \delta_{j}<1$ for $2 \leq j \leq 4$. In addition assume that $c_{j}<1$ and $d_{j}<1$ for $j=0$ and $j=1$. Define a map $f$ of $K^{0} \times K^{0}$ into itself by (2.36) - (2.40). Then for any $(A, B) \in K^{0} \times K^{0}$, there exists $E \in K^{0}$ such that

$$
s-\lim _{k \uparrow \infty} f^{k}(A, B)=(E, E)
$$

PROOF. Define $D=K \times K \times K$ and define a map $g: D^{0} \rightarrow D^{0}$ by

$$
\begin{equation*}
g(A, B, C)=\left(A_{1}, B_{1}, C_{1}\right) \tag{2.42}
\end{equation*}
$$

where $C$ is now an arbitrary element $K^{0}, A_{1}$ and $B_{1}$ are given by (2.38) and (2.39) respectively and

$$
\begin{equation*}
C_{1}=(1 / 2)\left(A_{1}+B_{1}\right) \tag{2.43}
\end{equation*}
$$

If $\pi$ projects $D^{0}$ onto $K^{0} \times K^{0}$ so that

$$
\pi(A, B, C)=(A, B)
$$

an easy induction argument shows that if $(A, B, C) \in D^{0}$ and $C=(A+B) / 2$ then

$$
\begin{equation*}
\pi\left(g^{k}(A, B, C)\right)=f^{k}(A, B) \quad \text { for all } k \geq 1 \tag{2.44}
\end{equation*}
$$

Thus to prove Corollary 2.5 it suffices to prove that for any $(A, B, C) \in D^{0}$ there exists $E \in K^{0}$ such that

$$
\begin{equation*}
s-\lim _{k \uparrow \infty} g^{k}(A, B, C)=(E, E, E) \tag{2.45}
\end{equation*}
$$

The map $g$ is of the form in Theorem 2.2 (whereas $f$ is not), and one could try to apply Theorem 2.2 and Remark 2.1. Such an approach requires slightly stronger assumptions than we have made, so we shall use a somewhat different argument.

We first eliminate some trivial cases. If we have

$$
\begin{equation*}
c_{0}+c_{1}+c_{2}=1=d_{0}+d_{1}+d_{2} \tag{2.46}
\end{equation*}
$$

so $d_{3}=d_{4}=c_{3}=c_{4}=0,\left(\boldsymbol{A}_{1}, \boldsymbol{B}_{1}\right)$ is a function of $(\boldsymbol{A}, \boldsymbol{B})$, the function being of the type considered in Theorem 2.2. Thus we are in the case $n=2$ of Theorem 2.2. The assumption that $c_{j}<1$ and $d_{j}<1$ for $j=0$ and 1 insures that the corresponding $2 \times 2$ matrix $M$ has all positive entries. Thus we are done if (2.46) is satisfied.

Similarly, if

$$
\begin{equation*}
d_{0}+d_{3}=1=c_{0}+c_{3} \tag{2.47}
\end{equation*}
$$

$\left(A_{1}, C_{1}\right)$ is a function of $(A, C)$ and the corollary follows easily from the case $n=2$ in Theorem 2.2. Also, if

$$
\begin{equation*}
d_{1}+d_{4}=1=c_{1}+c_{4} \tag{2.48}
\end{equation*}
$$

$\left(B_{1}, C_{1}\right)$ is a function of $(B, C)$ and we return to the case $n=2$ in Theorem 2.2.

Thus we can assume that (2.46), (2.47) and (2.48) are all not satisfied. Let $M$ be the $3 \times 3$ column-stochastic matrix defined as in Theorem 2.2 for our map $g$. One can easily check from the defining equations for $g$ that the third column of $M$ is the arithmetic average of the first two columns. Because $c_{j}<1$ and $d_{j}<1$ for $j=0$ and $j=1$, each of the first two columns of $M$ has at most one zero entry. If the first entries of columns one and two of $M$ are both zero, then (2.48) is satisfied, contrary to assumption. Similarly, if the second entries of both columns one and two of $M$ are both zero, (2.47) is satisfied, contrary to assumption. Finally, if the third entry of column one of $M$ equals zero and the third entry of column two of $M$ equals zero, (2.46) is satisfied, contrary to assumption. Thus the zero entry of column one of $M$ (if it exists) is never in the same position as the zero entry of column two, so the third column of $M$, being the arithmetic average of columns one and two has all positive entries. Using this information, one easily checks that $M^{2}$ has all positive entries.

If $u$ is the probability column vector such that $M u=u$ and if we write

$$
g^{k}(A, B, C)=\left(A_{k}, B_{k}, C_{k}\right)
$$

Theorem 2.1 implies that there exists $E \in K^{0}$ such that

$$
\begin{equation*}
s-\lim _{k \nmid \infty}\left(u_{1} \log A_{k}+u_{2} \log B_{k}+u_{3} \log C_{k}\right)=\log E \tag{2.49}
\end{equation*}
$$

and $\log A_{k}, \log B_{k}$ and $\log C_{k}$ converge weakly to $\log E$. Because

$$
w-\lim _{k \uparrow \infty} \log C_{k}=w-\lim _{k \uparrow \infty} \log \left(\left(A_{k}+B_{k}\right) / 2\right)=\log E
$$

we conclude that

$$
\begin{equation*}
w-\lim _{k \uparrow \infty}\left[\log \left(\left(A_{k}+B_{k}\right) / 2\right)-\left(\log A_{k}\right) / 2-\left(\log B_{k}\right) / 2\right]=0 \tag{2.50}
\end{equation*}
$$

Using (2.50) and Theorem 1.2 we conclude that

$$
\begin{equation*}
s-\lim _{k \uparrow \infty}\left[\log A_{k}-\log B_{k}\right]=0 \tag{2.51}
\end{equation*}
$$

Because $C_{k}=\left(A_{k}+B_{k}\right) / 2$ for $k \geq 1$, we have for $k \geq 1$
$\left(\log A_{k+1}, \log B_{k+1}, \log C_{k+1}\right) \geq\left(\log A_{k}, \log B_{k}, \log C_{k}\right) M$ $\geq\left(\log A_{k}, \log B_{k},\left(\log A_{k}\right) / 2+\left(\log B_{k}\right) / 2\right) M$.

Using the above inequality we find that for $k \geq 1$

$$
\left(\log A_{k+1}, \log B_{k+1}\right) \geq\left(\log A_{k}, \log B_{k}\right) P
$$

where $P=\left(p_{i j}\right)$ has elements

$$
p_{i j}=m_{i j}+m_{3 j} / 2, \quad 1 \leq i, j \leq 2
$$

Because each of the first two columns of $M$ has at most one zero entry, all entries of $P$ are positive, and $P$ is obviously column-stochastic. If $v$ is the unique column probability vector such that $P v=v$, Theorem 2.1 implies

$$
\begin{equation*}
s-\lim _{k \uparrow \infty}\left(v_{1} \log A_{k}+v_{2} \log B_{k}\right)=\log E \tag{2.52}
\end{equation*}
$$

for some $E \in K^{0}$. One obtains directly from (2.51) and (2.52) that

$$
s-\lim _{k \uparrow \infty} \log A_{k}=s-\lim _{k \uparrow \infty} \log B_{k}=\log E
$$

so Lemma 1.2 implies

$$
s-\lim _{k \uparrow \infty} A_{k}=s-\lim _{k \uparrow \infty} B_{k}=E=s-\lim _{k \uparrow \infty} C_{k}
$$

There are many other classical generalizations of the $A G M$. For example, G. Borchadt (see [6] and [9] for references) studied the map

$$
f(a, b, c, d)=\left(a_{1}, b_{1}, c_{1}, d_{1}\right)
$$

where

$$
\begin{gather*}
a_{1}=(a+b+c+d) / 4, b_{1}=\left([a b]^{1 / 2}+[c d]^{1 / 2}\right) / 2,  \tag{2.53}\\
\left.\left.c_{1}=(\mid a c]^{1 / 2}+[b d]^{1 / 2}\right) / 2, d_{1}=(\mid a d]^{1 / 2}+[b c]^{1 / 2}\right) / 2
\end{gather*}
$$

and $a, b, c$ and $d$ are (initially) positive reals. If $f^{n}$ denotes the $n$th iterate of $f$ and

$$
f^{n}(a, b, c, d)=\left(a_{n}, b_{n}, c_{n}, d_{n}\right)
$$

Borchardt proved (without great difficulty) that

$$
\begin{equation*}
\lim _{n \uparrow \infty} f^{n}(a, b, c, d)=(\gamma, \gamma, \gamma, \gamma), \gamma>0 \tag{2.54}
\end{equation*}
$$

The number $\gamma=\mu(a, b, c, d)$ can be considered a generalized $A G M$. Borchardt established many properties of this $A G M$. If $a=c$ and $b=d$, then $a_{n}=c_{n}$ and $b_{n}=d_{n}$ for all $n \geq 1$ and (2.53) reduces to the original $A G M$.

As is pointed out in [6], Borchardt's algorithm is a special case of a more general construction, called a "monster algorithm" in [6]. Let $G$ be a finite group of order $n$. If $\Theta_{1}$ and $\Theta_{2}$ are real-valued functions from $G$ to $\mathbb{R}$, the convolution of $\Theta_{1}$ and $\Theta_{2}, \Theta_{1} * \Theta_{2}$, is defined by

$$
\begin{equation*}
\left(\Theta_{1} * \Theta_{2}\right)(t)=n^{-1} \sum_{s \in G} \Theta_{1}(s) \Theta_{2}\left(t s^{-1}\right) \tag{2.55}
\end{equation*}
$$

(More generally, convolution can be defined with respect to a measure on a locally compact topological group). Define $C$ to be the set of functions $\Theta: G \rightarrow[0, \infty)$, so $C$ can be identified with the standard cone in $\mathbb{R}^{n}, n=|G|$. Define $F: C^{0} \rightarrow C^{0}$ by

$$
\begin{equation*}
F(\Theta)=\Theta^{1 / 2} * \Theta^{1 / 2}, \text { where }\left(\Theta^{1 / 2}\right)(s)=[\Theta(s)]^{1 / 2} \tag{2.56}
\end{equation*}
$$

It is proved in [6] (at least for $G$ abelian) that, if $F^{k}$ denotes the $k$ th iterate of $F$, then for any $\Theta \in C^{0}$

$$
\lim _{k \uparrow \infty} F^{k}(\Theta)=\Theta_{\infty}
$$

where $\Theta_{\infty}$ is a positive, constant function. Borchardt's algorithm corresponds to $G=C_{2} \times C_{2}$, where $C_{2}$ is a group of order 2 . Another interesting example corresponds to a cyclic group of order 3 (see [6]).

We generalize this construction to operator-valued functions. Let $K$ be the cone of p.s.d., bounded linear operators on a Hilbert space $H$. Let $G$ be a finite group of order $n$. Let $C$ denote the cone of maps $\Theta: G \rightarrow K$, so $C=\prod_{g \in G} K$. Define $F: C^{0} \rightarrow C^{0}$ by

$$
\begin{equation*}
(F(\Theta))(t)=n^{-1} \sum_{s \in G} \exp \left((1 / 2) \log (\Theta(s))+(1 / 2) \log \left(\Theta\left(t s^{-1}\right)\right)\right) \tag{2.57}
\end{equation*}
$$

This reduces to (2.56) when $\Theta$ is a real-valued.
COROLLARY 2.6. Let the notation be as in the immediately preceding paragraph. Then for any $\Theta \in C^{0}$,

$$
s-\lim _{k \uparrow \infty} F^{k}(\Theta)=\Theta_{\infty} \in C^{0}
$$

where $\Theta_{\infty}$ depends on $\Theta$ and $\Theta_{\infty}: G \rightarrow K^{0}$ is a constant function.
Proof. The cone $C$ can be considered as the $n$-fold Cartesian product of $K$, and with this identification the map $F$ is a special case of the maps considered in Theorem 2.2. We shall derive Corollary 2.6 from Remark 2.2.

Define $\phi: K^{0} \rightarrow X=\mathcal{L}(H)$ by $\phi(A)=\log A$ and if $Y$ denotes the Banach space of maps from $G$ to $X$, define $\Phi: C^{0} \rightarrow Y$ by

$$
(\Phi(\Theta))(t)=\log (\Theta(t)) \quad \text { for all } t \in G
$$

Applying $\phi$ to (2.57) and using the facts that

$$
\begin{equation*}
\sum_{s \in G} \log (\Theta(s))=\sum_{s \in G} \log \left(\Theta\left(t s^{-1}\right)\right) \tag{2.58}
\end{equation*}
$$

and $\phi$ is concave gives

$$
\begin{equation*}
\log (F(\Theta)(t)) \geq n^{-1} \sum_{s \in G} \log (\Theta(s)) \tag{2.59}
\end{equation*}
$$

or

$$
\Phi(F(\Theta)) \geq \Phi(\Theta) M
$$

where $M$ is the doubly stochastic matrix with all entries equal to $n^{-1}$. Thus we are in the situation of Remark 2.2 and the corollary follows.

As already noted, we immediately obtain the operator analogue of Borchardt's algorithm from Corollary 2.6.

Corollary 2.7. Let $K$ denote the cone of p.s.d., bounded self-adjoint linear operators on a Hilbert space $H$. Define $U=K \times K \times K \times K$. Define $f: U^{0} \rightarrow U^{0}$ by

$$
f(A, B, C, D)=\left(A_{1}, B_{1}, C_{1}, D_{1}\right)
$$

where

$$
\begin{gathered}
A_{1}=(1 / 4)(A+B+C+D) \\
B_{1}=(1 / 2)[\exp ((1 / 2) \log A+(1 / 2) \log B)+\exp ((1 / 2) \log C+(1 / 2) \log D)] \\
C_{1}=(1 / 2)[\exp ((1 / 2) \log A+(1 / 2) \log C) \\
+\exp ((1 / 2) \log B+(1 / 2) \log D)], \text { and } \\
D_{1}=(1 / 2)[\exp ((1 / 2) \log A+(1 / 2) \log D)+\exp ((1 / 2) \log B+(1 / 2) \log C)]
\end{gathered}
$$

Then if $f^{m}(A, B, C, D)=\left(A_{m}, B_{m}, C_{m}, D_{m}\right)$, there exists $E \in K^{0}$ such that

$$
s-\lim _{m \uparrow \infty}\left(A_{m}, B_{m}, C_{m}, D_{m}\right)=(E, E, E, E)
$$

Many other generalized means have operator-valued versions that can be analyzed by our methods. We mention only two more examples. Borchardt and Schwab (see [6]) considered the map

$$
f(a, b)=\left(a_{1}, b_{1}\right), a_{1}=(a+b) / 2, \quad b_{1}=\left(a_{1} b\right)^{1 / 2}
$$

and the corresponding mean given by

$$
\lim _{n \uparrow \infty} f^{n}(a, b)=(\delta, \delta)
$$

Carlson [12] observed that the Borchardt-Schwab algorithm is naturally embedded in an algorithm involving three variables. Given positive numbers $a, b$ and $c$, define $\alpha, \beta$ and $\gamma$ by

$$
\alpha=(b+c) / 2, \quad \beta=(a+c) / 2, \gamma=(a+b) / 2
$$

and define $f(a, b, c)=\left(a_{1}, b_{1}, c_{1}\right)$, where

$$
a_{1}=(\beta \gamma)^{1 / 2}, \quad b_{1}=(\alpha \gamma)^{1 / 2} \text { and } c_{1}=(\alpha \beta)^{1 / 2}
$$

If $f^{n}(a, b, c)=\left(a_{n}, b_{n}, c_{n}\right)$, Carlson proved that

$$
\lim _{n \uparrow \infty}\left(a_{n}, b_{n}, c_{n}\right)=(\delta, \delta, \delta) .
$$

He related the limit to certain integrals. If $b=c$, then $b_{n}=c_{n}$ for all $n \geq 1$ and one recovers the Borchardt-Schwab algorithm.

To generalize Carlson's algorithm to operator-valued maps, let $K$ denote the cone of p.s.d., bounded self-adjoint linear maps of a Hilbert space $H$ to itself. Define $E=K \times K \times K$. Define $f: E^{0} \rightarrow E^{0}$ by

$$
\begin{equation*}
f(A, B, C)=\left(A_{1}, B_{1}, C_{1}\right) \tag{2.60}
\end{equation*}
$$

where

$$
A_{1}=\exp ((1 / 2) \log (\beta)+(1 / 2) \log (\gamma)), \quad B_{1}=\exp ((1 / 2) \log (\alpha)+(1 / 2) \log (\gamma))
$$

$$
\begin{equation*}
C_{1}=\exp ((1 / 2) \log (\alpha)+(1 / 2) \log (\beta)) \tag{2.61}
\end{equation*}
$$

and

$$
\begin{equation*}
\alpha=(B+C) / 2, \quad \beta=(A+C) / 2 \text { and } \gamma=(A+B) / 2 . \tag{2.62}
\end{equation*}
$$

COROLLARY 2.8. If $f: E^{0} \rightarrow E^{0}$ is defined by (2.60) - (2.62), then for any $(A, B, C) \in E^{0}$,

$$
\begin{equation*}
s-\lim _{k \uparrow \infty} f^{k}(A, B, C)=(D, D, D) \tag{2.63}
\end{equation*}
$$

where $D \in K^{0}$.
Proof. Define $\Phi(A, B, C)=(\phi(A), \phi(B), \phi(C))$ for $(A, B, C) \in E^{0}$, where $\phi(A)=\log A$ for $A \in K^{0}$. It follows easily from the concavity of $\log$ that

$$
\Phi(f(A, B, C)) \geq \Phi(A, B, C) M
$$

where

$$
M=\left(\begin{array}{lll}
1 / 2 & 1 / 4 & 1 / 4 \\
1 / 4 & 1 / 2 & 1 / 4 \\
1 / 4 & 1 / 4 & 1 / 2
\end{array}\right)
$$

The reader can easily verify that the other hypotheses of Theorem 1.1 hold, so if $f^{k}(A, B, C)=\left(A_{k}, B_{k}, C_{k}\right)$, there exists $D \in K^{0}$ such that

$$
\begin{align*}
& w-\lim _{k \uparrow \infty} \log A_{k}=w-\lim _{k \uparrow \infty} \log B_{k}  \tag{2.64}\\
& =w-\lim _{k \uparrow \infty} \log C_{k}=\log D
\end{align*}
$$

and

$$
\begin{equation*}
s-\lim _{k \uparrow \infty}(1 / 3)\left(\log A_{k}+\log B_{k}+\log C_{k}\right)=\log D \tag{2.65}
\end{equation*}
$$

The defining equation for $f$ gives

$$
\begin{align*}
& \log A_{k+1}-(1 / 2) \log A_{k}-(1 / 4) \log B_{k}-(1 / 4) \log \left(C_{k}\right)= \\
& \quad(1 / 2)\left[\log \left((1 / 2)\left(A_{k}+C_{k}\right)\right)-(1 / 2) \log A_{k}-(1 / 2) \log \left(C_{k}\right)\right]  \tag{2.66}\\
& +(1 / 2)\left[\log \left((1 / 2)\left(A_{k}+B_{k}\right)\right)-(1 / 2) \log A_{k}-(1 / 2) \log B_{k}\right]
\end{align*}
$$

The left side of (2.66) converges to zero in the weak operator topology, so Theorem 1.2 and (2.66) imply

$$
\begin{equation*}
s-\lim _{k \uparrow \infty}\left[\log A_{k}-\log C_{k}\right]=s-\lim _{k \uparrow \infty}\left[\log B_{k}-\log C_{k}\right]=0 \tag{2.67}
\end{equation*}
$$

(2.65) and (2.67) give

$$
\begin{equation*}
s-\lim _{k \uparrow \infty} \log A_{k}=s-\lim _{k \uparrow \infty} \log B_{k}=s-\lim _{k \uparrow \infty} \log C_{k}=\log D \tag{2.68}
\end{equation*}
$$

and the corollary follows from (2.68) with the aid of Lemma 1.2.
A final example is an algorithm of Meissel (see [6]). For positive real numbers $a, b, c$ define

$$
\begin{equation*}
f(a, b, c)=\left([a+b+c] / 3, \quad([a b+a c+b c] / 3)^{1 / 2},(a b c)^{1 / 3}\right) . \tag{2.69}
\end{equation*}
$$

If $K$ is the cone of p.s.d., self-adjoint linear operators on a Hilbert space $H$ and $E=K \times K \times K$, one can define a map $f: E^{0} \rightarrow E^{0}$ which is an analogue of the map in (2.69), namely,

$$
\begin{equation*}
f(A, B, C)=\left(A_{1}, B_{1}, C_{1}\right) \tag{2.70}
\end{equation*}
$$

where

$$
\begin{align*}
A_{1} & =(A+B+C) / 3, \quad B_{1}=[(1 / 3) \exp (\alpha+\beta) \\
& +(1 / 3) \exp (\alpha+\gamma)+(1 / 3) \exp (\beta+\gamma)]^{1 / 2}  \tag{2.71}\\
C_{1} & =\exp ((\alpha+\beta+\gamma) / 3), \quad \alpha=\log A, \beta=\log B \text { and } \gamma=\log C .
\end{align*}
$$

COROLLARY 2.9. If $H$ and $K$ are as above and $f: E^{0} \rightarrow E^{0}$ is defined by (2.70)-(2.71), then for any $(A, B, C) \in E^{0}$, there exists $D \in K^{0}$ such that

$$
\begin{equation*}
s-\lim _{n \rightarrow \infty} f^{n}(A, B, C)=(D, D, D) \tag{2.72}
\end{equation*}
$$

PROOF. Corollary 2.9 follows by essentially the same argument used to prove Corollary 2.8 and is left to the reader.

## 3. - Some elementary properties of the arithmetic-geometric mean

In this section we shall establish some basic properties of

$$
\begin{equation*}
M(A, B)=\lim _{k \uparrow \infty} f^{k}(A, B) \tag{3.1}
\end{equation*}
$$

where

$$
\begin{equation*}
f(A, B)=((A+B) / 2, \exp ((1 / 2) \log A+(1 / 2) \log B)) \tag{3.2}
\end{equation*}
$$

Much of what we say extends to the more general examples considered in Section 2, but for simplicity we shall restrict ourselves to the above case or, occasionally, the "monster algorithm" of Corollary 2.6.

We begin with some generalities. If $\boldsymbol{X}$ is a complex Banach space, $G$ is an open subset of $X$ and $f: G \rightarrow X$ is continuous, $f$ is called analytic if, whenever $B_{r}(u)=\{y:\|y-u\|<r\} \subseteq G, \Psi \in X^{*}$ is a complex linear
functional and $v \in X$ is such that $\|v\|=1$, then $\lambda \rightarrow \Psi(f(u+\lambda v))$ is complex analytic for all $\lambda \in \mathbb{C}$ such that $|\lambda|<r$.

If $X$ is a complex Banach space, let $Y$ denote the $n$-fold Cartesian product of $X$ and for $y=\left(y_{1}, y_{2}, \cdots, y_{n}\right) \in Y$, define a seminorm $p(y)$ by

$$
\begin{equation*}
p(y)=\max \left\{\left\|y_{i}-y_{j}\right\|: 1 \leq i, j \leq n\right\}, \quad y=\left(y_{1}, y_{2}, \cdots, y_{n}\right) \tag{3.3}
\end{equation*}
$$

Then $p(y)=0$ if and only if $y \in S$, where

$$
\begin{equation*}
S=\left\{y=\left(y_{1}, y_{1}, \cdots, y_{1}\right): y_{1} \in X\right\} \tag{3.4}
\end{equation*}
$$

Lemma 3.1. Let $X, Y, p$ and $S$ be as above. For $y^{0} \in S$ and $\delta>0$ let $B_{\delta}\left(y^{0}\right)=\left\{y \in Y:\left\|y-y^{0}\right\|<\delta\right\}=U$ and suppose that $f: U \rightarrow Y$ is an analytic map such that $f\left(y^{0}\right)=y^{0}$. Assume that there exists $c<1$ and a constant $k$ such that

$$
\begin{equation*}
p(f(y)) \leq c p(y) \tag{3.5}
\end{equation*}
$$

and

$$
\begin{equation*}
\|f(y)-y\| \leq k p(y) \tag{3.6}
\end{equation*}
$$

for all $y \in U$. Then there exists $r>0$ such that $f^{m}(y) \in U$ for all $m \geq 1$ whenever $y \in \overline{B_{r}\left(y^{0}\right)}$ and $f^{m}(y)$ converges uniformly in $y \in \overline{B_{r}\left(y^{0}\right)}$ as $m \rightarrow \infty$ to a limit $g(y) \in S$ such that $f(g(y))=g(y)$. The map $y \rightarrow g(y)$ is complex analytic on $B_{r}\left(y^{0}\right)$.

PROOF. For definiteness, define $\|y\|=\max _{1 \leq i \leq n}\left\|y_{i}\right\|$. If a sequence of analytic functions converges uniformly on an open $\operatorname{set}^{1 \leq i \leq n} G$ to a limit $g, g$ is analytic on $G$. Thus, in our case, to prove $g$ is analytic on $B_{r}\left(y^{0}\right)$ for some $r>0$, it suffices $\frac{\text { to prove that } f^{m} \text { (which is analytic) is defined and converges uniformly on }}{B_{r}\left(y^{0}\right)}$.

Take $r_{0}>0$ such that

$$
r_{0}+2 k r_{0} /(1-c)<\delta
$$

and note that $p(y) \leq 2 r_{0}$ for all $y \in B_{r_{0}}\left(y^{0}\right)$. Define

$$
r_{j}=r_{0}+2 k r_{0}\left(\sum_{i=0}^{j=1} c^{i}\right)<\delta, j \geq 1
$$

and assume that if $y \in B_{r_{0}}\left(y^{0}\right)$, then $f^{j}(y) \in B_{r_{j}}\left(y^{0}\right)$ for $0 \leq j \leq m$ and $p\left(f^{j}(y)\right) \leq c^{j}\left(2 r_{0}\right)$ for $0 \leq j \leq m$. Then (3.5) and (3.6) imply that

$$
\left\|f^{m+1}(y)-f^{m}(y)\right\| \leq k p\left(f^{m}(y)\right) \leq k c^{m}\left(2 r_{0}\right)
$$

so

$$
\begin{aligned}
\left\|f^{m+1}(y)-y^{0}\right\| & \leq\left\|f^{m+1}(y)-f^{m}(y)\right\|+\left\|f^{m}(y)-y^{0}\right\| \\
& \leq r_{0}+2 k r_{0} \sum_{i=0}^{m-1} c^{i}+2 k r_{0} c^{m}=r_{m+1}
\end{aligned}
$$

By mathematical induction we find that $f^{j}(y) \in B_{r},\left(y^{0}\right)$ for all $j \geq 1$. If $y \in B_{r_{0}}\left(y^{0}\right)$ and $m$ and $v$ are positive integers, $m<v$, then

$$
\begin{align*}
\left\|f^{m}(y)-f^{v}(y)\right\| & \leq \sum_{i=1}^{v-m}\left\|f^{m+i}(y)-f^{m+i-1}(y)\right\| \\
& \leq k \sum_{i=1}^{v-m} p\left(f^{m+i-1}(y)\right)  \tag{3.7}\\
& \leq k\left(2 r_{0}\right) \sum_{i=1}^{v-m} c^{m+i-1} \\
& \leq k\left(2 r_{0}\right) c^{m} /(1-c)
\end{align*}
$$

(3.7) shows that $\left(f^{m}(y)\right)$ is a Cauchy sequence with limit $g(y)$. If $v \rightarrow \infty$ in (3.7), then

$$
\left\|f^{m}(y)-g(y)\right\| \leq 2 r_{0} k c^{m} /(1-c)
$$

so the convergence is uniform in $y \in B_{r_{0}}\left(y^{0}\right)$. Obviously $g(y) \in U$ and $f(g(y))=g(y)$, and because $p(f(g(y)))=p(g(y)) \leq c p(g(y)), \quad p(g(y))=0$ and $g(y) \in S$.

Under the hypotheses of Lemma 3.1, convergence is "linear", whereas for the examples of interest to us, convergence is actually "quadratic" (see [35], Chapter 12 for definitions) and hence extremely rapid. The next lemma describes the situation we shall actually encounter.

LEMMA 3.2. Let the notation and assumptions be as in Lemma 3.1 except instead of assuming that $f$ satisfies inequality (3.5), suppose that there exists a constant $M$ such that

$$
\begin{equation*}
p(f(y)) \leq M(p(y))^{2} \tag{3.8}
\end{equation*}
$$

for all $y \in U$. Then the conclusions of Lemma 3.1 are still valid.
Furthermore, if $\delta$ in Lemma 3.1 is so small that

$$
\begin{equation*}
\sup _{y \in U} M p(y) \leq c<1 \tag{3.9}
\end{equation*}
$$

then, setting $u=u_{m}=2^{m}$,

$$
\begin{equation*}
\left\|f^{m}(y)-g(y)\right\| \leq\left(k M^{-1}\right) c^{u}\left(1-c^{u}\right)^{-1} \tag{3.10}
\end{equation*}
$$

for all $y \in \overline{B_{r_{0}}\left(y^{0}\right)}$ and $m \geq 0$.
Proof. By decreasing $\delta$ we can assume that (3.9) is satisfied on $B_{\delta}\left(y^{0}\right)$. (3.8) then implies that (3.5) is satisfied, so the conclusions of Lemma 3.1 hold. An easy induction shows that if we define $u_{j}=2^{3}$, then

$$
\begin{equation*}
p\left(f^{m}(y)\right) \leq M^{-1}(M p(y))^{u_{m}} \leq M^{-1} c^{u_{m}} \tag{3.11}
\end{equation*}
$$

If we use (3.6) and (3.11) we find

$$
\begin{align*}
& \left\|f^{m}(y)-g(y)\right\|=\lim _{v \uparrow \infty}\left\|f^{m}(y)-f^{v}(y)\right\| \\
& \leq \sum_{j=m}^{\infty}\left\|f^{j+1}(y)-f^{j}(y)\right\| \leq k M^{-1} \sum_{j=m}^{\infty} c^{u} . \tag{3.12}
\end{align*}
$$

If we define $\rho=c^{2^{m}}$, it is easy to see that

$$
\begin{equation*}
\sum_{j=m}^{\infty} c^{u^{j}} \leq \sum_{i=1}^{\infty} \rho^{i}=\rho(1-\rho)^{-1} \tag{3.13}
\end{equation*}
$$

and (3.12) and (3.13) give (3.10).
Next we establish a theorem which, as we shall see later, is applicable to the "monster algorithm" of Corollary 2.6.

Theorem 3.1. Let $X$ be a complex Banach space and $Y$ the $n$-fold Cartesian product of $X$ with itself. Let $p$ and $S$ be as defined in (3.3) and (3.4).

Suppose that $V$ is an open subset of $Y$ and $f: V \rightarrow Y$ is a complex analytic map. Define $W \subseteq V$ by

$$
\begin{align*}
W= & \left\{y \in V: f^{m}(y) \in V \text { for all } m \geq 1\right. \text { and there exists }  \tag{3.14}\\
& \left.u \in V \cap S \text { such that } \lim _{m \uparrow \infty}\left\|f^{m}(y)-u\right\|=0\right\} .
\end{align*}
$$

(The element $u$ in (3.14) depends on $y$ ). For each $u \in V \cap S$, assume there exist positive constants $c, \delta$ and $k$ (dependent on $u$ ) such that $c<1$ and

$$
\begin{equation*}
p(f(y)) \leq c p(y) \quad \text { and } \quad\|f(y)-y\| \leq k p(y) \tag{3.15}
\end{equation*}
$$

for all $y \in B_{\delta}(u)=\{y:\|y-u\|<\delta\}$. Then $W$ is an open set and

$$
g(y)=\lim _{m \uparrow \infty} f^{m}(y), \quad y \in W
$$

is an analytic function on $W$.

Proof. If $y \in W$, select $u \in V \cap S$ such that $f^{m}(y)$ converges to $u$. If $c, \delta$ and $k$ are as above, select $r_{0}$ as in Lemma 3.1, so that if $w \in \overline{B_{r_{0}}(u)}$, then $f^{j}(w) \in B_{\delta}(u)$ for all $j \geq 1$ and

$$
\lim _{j \uparrow \infty} f^{j}(w)=g(w)
$$

where $g$ is an analytic function on $B_{r_{0}}(u)$. There exists an integer $N$ so that $f^{N}(y) \in B_{r_{0}}(u)$, and by continuity of $f^{N}$, there exists $\delta_{1}>0$ so that $f^{N}(z) \in B_{r_{0}}(u)$ for all $z$ such that $\|z-y\|<\delta_{1}$. It follows that if $\|z-y\|<\delta_{1}$,

$$
\begin{equation*}
g(z)=\lim _{m \uparrow \infty} f^{m}(z)=g\left(f^{N}(z)\right) \in S \tag{3.16}
\end{equation*}
$$

Thus we see that $W$ is open. Also, because the restriction of $g$ to $B_{r_{0}}(u)$ is analytic (by Lemma 3.1) and $f^{N}$ is analytic, (3.16) implies that $g$ is analytic on an open neighborhood of $y$.

The argument of lemma 3.2 shows that to verify (3.15) in Theorem 3.1, it suffices to verify (3.6) and (3.8). To accomplish this for the examples of interest to us we need the next fact.

Lemma 3.3. Let $H$ be a Banach space and $X=\mathcal{L}(H)$. Suppose that $A_{0} \in X$ and $\sigma\left(A_{0}\right) \cap(-\infty, 0]$ is empty. Then there exists $\delta>0$ and $M>0$ such that (1) for all $A \in B_{\delta}\left(A_{0}\right)$, the open $\delta$ ball about $A_{0}, \sigma(A) \cap(-\infty, 0]$ is empty and (2) if $A_{1}, A_{2}, \cdots, A_{m}$ are any elements of $B_{\delta}\left(A_{0}\right)$ and $\alpha_{1}, \alpha_{2}, \cdots, \alpha_{m}$ are positive reals such that $\sum_{j=1}^{m} \alpha_{j}=1$, then

$$
\begin{equation*}
\left\|\log \left(\sum_{j=1}^{m} \alpha_{j} A_{j}\right)-\sum_{j=1}^{m} \alpha_{j} \log A_{j}\right\| \leq M \sum_{1 \leq i \leq j \leq m}\left\|A_{i}-A_{j}\right\|^{2} \tag{3.17}
\end{equation*}
$$

Proof. By assumption, $(1-t) A_{0}+t I$ is invertible for $0 \leq t \leq 1$, and by continuity of the map $A \rightarrow A^{-1}$ on the set of invertible linear operators, $\left\|\left[(1-t) A_{0}+t I\right]^{-1}\right\|$ is uniformly bounded for $0 \leq t \leq 1$, say by a constant $M_{1}$. If $\left\|A-A_{0}\right\|<\delta$ and $\delta<\left(2 M_{1}\right)^{-1}$, then writing $A_{0 t}=t A_{0}+(1-t) I$ and $A_{t}=t A+(1-t) I$ for $0 \leq t \leq 1$, one has

$$
A_{t}=A_{0 t}\left[I+A_{0 t}^{-1} t\left(A-A_{0}\right)\right]
$$

so $A_{t}$ is invertible (the product of invertible operators) and

$$
\left\|A_{t}^{-1}\right\| \leq\left\|A_{0 t}^{-1}\right\| \sum_{s=0}^{\infty}\left(\delta M_{1}\right)^{j} \leq 2\left\|A_{0 t}^{-1}\right\| \leq 2 M_{1}, \quad 0 \leq t \leq 1
$$

To prove (3.17), first assume that $m=2$ and take $A$ and $B$ in $B_{\delta}\left(A_{0}\right)$ and $\alpha$ so that $0<\alpha<1$. If $C \in B_{\delta}\left(A_{0}\right)$ and we write $C_{t}=I+t(C-I)$ for $0 \leq t \leq 1$, we have

$$
\begin{align*}
\log C & =\int_{0}^{1}(d / d t) \log \left(C_{t}\right) d t=\int_{0}^{1} C_{t}^{-1}(C-I) d t  \tag{3.18}\\
& =\int_{0}^{1} t^{-1}\left[I-C_{t}^{-1}\right] d t
\end{align*}
$$

If we take $C=\alpha A+(1-\alpha) B, C=A$ and $C=B$ in (3.18), we find after simplification that

$$
\begin{gather*}
\log (\alpha A+(1-\alpha) B)-\alpha \log A-(1-\alpha) \log B \\
=\int_{0}^{1} t^{-1}\left\{\alpha A_{t}^{-1}+(1-\alpha) B_{t}^{-1}-\left[\alpha A_{t}+(1-\alpha) B_{t}\right]^{-1}\right\} d t \tag{3.19}
\end{gather*}
$$

where

$$
\begin{equation*}
A_{t}=t A+(1-t) I \text { and } B_{t}=t A+(1-t) I \tag{3.20}
\end{equation*}
$$

By using Lemma 1.5 and (1.34) we obtain

$$
\begin{equation*}
\alpha A_{t}^{-1}+(1-\alpha) B_{t}^{-1}-\left[\alpha A_{t}+(1-\alpha) B_{t}\right]^{-1} \tag{3.21}
\end{equation*}
$$

$$
=\alpha(1-\alpha) B_{t}^{-1}\left(A_{t}-B_{t}\right) A_{t}^{-1}\left[(1-\alpha) A_{t}^{-1}+\alpha B_{t}^{-1}\right]^{-1} B_{t}^{-1}\left(A_{t}-B_{t}\right) A_{t}^{-1}
$$

Using the identity

$$
\left[(1-\alpha) A_{t}^{-1}+\alpha B_{t}^{-1}\right]^{-1}=A_{t}\left[(1-\alpha) B_{t}+\alpha A_{t}\right]^{-1} B_{t}
$$

in (3.21) and simplifying we find

$$
\begin{gather*}
\alpha A_{t}^{-1}+(1-\alpha) B_{t}^{-1}-\left[\alpha A_{t}+(1-\alpha) B_{t}\right]^{-1} \equiv D_{t}=  \tag{3.22}\\
\alpha(1-\alpha) t^{2} B_{t}^{-1}(B-A)\left[(1-\alpha) B_{t}+\alpha A_{t}\right]^{-1}(B-A) A_{t}^{-1}
\end{gather*}
$$

Since $\left\|B_{t}^{-1}\right\| \leq 2 M_{1},\left\|A_{t}^{-1}\right\| \leq 2 M_{1}$ and

$$
\left\|\left[(1-\alpha) B_{t}+\alpha A_{t}\right]^{-1}\right\| \leq 2 M_{1}
$$

we obtain from (3.22) (using also that $\alpha(1-\alpha) \leq(1 / 4)$ ) that

$$
\left\|D_{t}\right\| \leq 2 M_{1}^{3} t^{2}\|B-A\|^{2}
$$

Using this estimate in (3.19) yields

$$
\begin{gather*}
\|\log (\alpha A+(1-\alpha) B)-\alpha \log A-(1-\alpha) \log B\| \\
\leq 2 M_{1}^{3}\|B-A\|^{2} \int_{0}^{1} t d t=M_{1}^{3}\|B-A\|^{2} \tag{3.23}
\end{gather*}
$$

Our argument actually shows that if $U$ is an open neighborhood of $A_{0}$ such that $((1-t) A+t I)^{-1}$ exists for all $A \in U$ and $0 \leq t \leq 1$ and

$$
\left\|[(1-t) A+t I]^{-1}\right\| \leq 2 M_{1}
$$

then (3.23) is satisfied.
We now proceed by induction. We have proved the lemma for $m=2$, the constant $M$ in (3.17) being $M_{1}^{3}$. Assume for some $m>2$ that we have proved the lemma for $m-1$ and that the constant $M$ in (3.17) can be taken to be $M_{1}^{3}$. Let $A_{j}$ and $\alpha_{j}, 1 \leq j \leq m$, be as in the statement of Lemma 3.3 and define $A=A_{1}$,

$$
B=\left(1-\alpha_{1}\right)^{-1} \sum_{j=2}^{m} \alpha_{j} A_{j}
$$

and $\alpha=\alpha_{1}$. Then $A, B \in B_{\delta}\left(A_{0}\right)$ and (3.23) gives

$$
\begin{gather*}
\left\|\log \left(\sum_{j=1}^{m} \alpha_{j} A_{j}\right)-\alpha_{1} \log A_{1}-\left(1-\alpha_{1}\right) \log B\right\|  \tag{3.24}\\
\leq M_{1}^{3}\|B-A\|^{2}=M_{1}^{3}\left\|\sum_{j=2}^{m}\left(\alpha_{j} /\left(1-\alpha_{1}\right)\right)\left(A_{j}-A_{1}\right)\right\|^{2} .
\end{gather*}
$$

Using the Cauchy-Schwarz inequality,

$$
\begin{align*}
&\left\|\sum_{j=2}^{m}\left(\alpha_{j} /\left(1-\alpha_{1}\right)\right)\left(A_{j}-A_{1}\right)\right\|^{2} \leq\left[\sum_{j=2}^{m}\left(\alpha_{j} /\left(1-\alpha_{1}\right)\right)\left\|A_{j}-A_{1}\right\|^{2}\right. \\
& \leq \sum_{j=2}^{m}\left\|A_{j}-A_{1}\right\|^{2} \tag{3.25}
\end{align*}
$$

On the other hand, the inductive assumption implies

$$
\begin{gather*}
\left\|\left(1-\alpha_{1}\right) \log B-\left(1-\alpha_{1}\right) \sum_{j=2}^{m}\left(\alpha_{j} /\left(1-\alpha_{1}\right)\right) \log A_{j}\right\| \\
\leq\left(1-\alpha_{1}\right) M_{1}^{3}\left(\sum_{2 \leq i \leq j \leq m}\left\|A_{i}-A_{j}\right\|^{2}\right) \tag{3.26}
\end{gather*}
$$

Combining (3.24) - (3.26) we find that

$$
\begin{equation*}
\left\|\log \left(\sum_{j=1}^{m} \alpha_{j} A_{j}\right)-\sum_{j=1}^{m} \alpha_{j} \log \left(A_{j}\right)\right\| \leq M_{1}^{3} \sum_{1 \leq i \leq j \leq m}\left\|A_{i}-A_{j}\right\|^{2} \tag{3.27}
\end{equation*}
$$

so the lemma has been proved by induction.
With these preliminaries we return to the "monster algorithm" of Corollary 1.5.

THEOREM 3.2. Let $H$ be a complex Banach space, let $X=\mathcal{L}(H)$ and let $U=\{A \in X: \sigma(A) \cap(-\infty, 0]$ is empty $\}$, where $\sigma(A)$ denotes the spectrum of A. If $G$ is a finite group of order $n \geq 2$, let $Y$ denote the Banach space of maps from $G$ to $X$, so $\boldsymbol{Y}$ can be identified with the $n$-fold Cartesian products of $X$. Define $V=\{\Theta \in Y: \Theta(s) \in U$ for all $s \in G\}$ and define $F: V \rightarrow Y$ by

$$
\begin{equation*}
(F \Theta)(t)=n^{-1} \sum_{s \in G} \exp \left((1 / 2) \log \Theta(s)+(1 / 2) \log \Theta\left(t s^{-1}\right)\right) \tag{3.28}
\end{equation*}
$$

Define $S$ to be the set of constant functions in $Y$ and define $W$ by

$$
\begin{aligned}
W= & \left\{\Theta \in V: F^{m}(\Theta) \in V \text { for all } m \geq 1\right. \text { and there exists } \\
& \left.\Psi \in V \cap S \text { such that } \lim _{m \uparrow \infty}\left\|F^{m}(\Theta)-\Psi\right\|=0\right\}
\end{aligned}
$$

where $F^{m}$ denotes the mth iterate of $F$ and $\Psi$ depends on $\Theta$. Then $W$ is an open subset of $Y$. If $g(\Theta)$ is defined by

$$
\begin{equation*}
g(\Theta)=\lim _{m \uparrow \infty} F^{m}(\Theta) \tag{3.29}
\end{equation*}
$$

for $\Theta \in W$, the map $\Theta \rightarrow g(\Theta)$ is analytic. The convergence in (3.29) is quadratic. If $H$ is a finite dimensional Hilbert space, $W$ contains $W_{1}$, where

$$
W_{1}=\{\Theta \in Y: \Theta(s) \text { is positive definite and self-adjoint for all } s \in G\} .
$$

Proof. Select $\Psi \in V \cap S$. By Theorem 3.1 it suffices to prove that there exists $\delta>0$ such that (3.6) and (3.8) are satisfied for all $\Theta \in B_{\delta}(\Psi)=$ $\left\{\Theta \in Y: \sup _{s \in G}\|\Theta(s)-\Psi(s)\|<\delta\right\}$. The map $A \rightarrow \exp (A)$ is $C^{1}$ with bounded Fréchet derivative on bounded sets in $X$, so the map $A \rightarrow \exp (A)$ is Lipschitzian on bounded sets. Similarly, for $\delta$ small enough, $A \rightarrow \log A$ is Lipschitzian on $\overline{B_{\delta}(\Psi)}$. Thus to prove that $F$ satisfies inequality (3.8) on $B_{\delta}(\Psi)$ for some $\delta>0$, it suffices to prove that there exists $M$ such that

$$
\begin{equation*}
p(\log F(\Theta)) \leq M(p(\log \Theta))^{2} \tag{3.30}
\end{equation*}
$$

for all $\Theta \in B_{\delta}(\Psi)$, where, for $\alpha \in Y$,

$$
\begin{equation*}
p(\alpha)=\sup _{s_{1}, s_{2} \in G}\left\|\alpha\left(s_{1}\right)-\alpha\left(s_{2}\right)\right\| . \tag{3.31}
\end{equation*}
$$

We know that for $\theta \in V$

$$
\begin{gather*}
n^{-1} \sum_{s \in G}(1 / 2) \log \Theta(s)+(1 / 2) \log \Theta\left(t s^{-1}\right) \\
=n^{-1} \sum_{\sigma \in G} \log \Theta(\sigma) \tag{3.32}
\end{gather*}
$$

Using Lemma 3.3 we find that there exists $\delta>0$ and a constant $M_{1}$ such that for all $\Theta \in B_{\delta}(\Psi)$ one has

$$
\begin{align*}
& \| \log F(\Theta(t))-n^{-1} \sum_{s \in G}\left[(1 / 2) \log \Theta(s)+(1 / 2) \log \Theta\left(t s^{-1}\right)\right] \| \\
& \leq M_{1} \sum_{s_{1} \cdot s_{2} \in G}\left\|\beta\left(s_{1}, t\right)-\beta\left(s_{2}, t\right)\right\|^{2} \tag{3.33}
\end{align*}
$$

where

$$
\begin{equation*}
\beta(s, t) \equiv \exp \left((1 / 2) \log \Theta(s)+(1 / 2) \log \Theta\left(t s^{-1}\right)\right) \tag{3.34}
\end{equation*}
$$

Because the map $\boldsymbol{A} \rightarrow \exp \boldsymbol{A}$ is Lipschitzian on bounded sets, we find from (3.32) - (3.34) that there exists a constant $M_{2}$ such that

$$
\begin{equation*}
\left\|\log F(\Theta(t))-n^{-1} \sum_{s \in G} \log \Theta(s)\right\| \leq M_{2}(p(\log \Theta))^{2} \tag{3.35}
\end{equation*}
$$

for all $\Theta \in B_{\delta}(\Psi)$. It follows from (3.35) and the triangle inequality that for all $t_{1}, t_{2} \in G$,

$$
\begin{equation*}
\left\|\log F\left(\Theta\left(t_{1}\right)\right)-\log F\left(\Theta\left(t_{2}\right)\right)\right\| \leq 2 M_{2}(p(\log \Theta))^{2} \tag{3.36}
\end{equation*}
$$

so

$$
p(\log F(\Theta)) \leq 2 M_{2}(p(\log \Theta))^{2}
$$

It remains to prove (3.6). By using the Lipschitz nature of $A \rightarrow \exp A$ and $A \rightarrow \log A$ on appropriate sets, it suffices to prove that there exist $\delta>0$ and a constant $k$ such that

$$
\|\log F(\Theta)-\log \Theta\| \leq k p(\log \Theta)
$$

for all $\Theta \in B_{\delta}(\Psi)$. By using (3.35) and the triangle inequality we see that

$$
\begin{align*}
& \|\log F(\Theta(t))-\log \Theta(t)\| \leq\left\|\log F(\Theta(t))-n^{-1} \sum_{s \in G} \log \Theta(s)\right\|  \tag{3.37}\\
& +\left\|n^{-1} \sum_{s \in G}(\log \Theta(s)-\log \Theta(t))\right\| \leq M_{2}(p(\log \Theta))^{2}+p(\log \Theta)
\end{align*}
$$

If $M_{3}$ is chosen so that $p(\log \Theta) \leq M_{3}$ for all $\Theta \in B_{\delta}(\Psi)$, (3.37) implies that

$$
\|\log F(\Theta)-\log \Theta\| \leq\left(M_{2} M_{3}+1\right) p(\log \Theta)
$$

The final assertion of Theorem 3.2 follows immediately from Corollary 2.6 and the fact that strong convergence implies norm convergence in finite dimensions.

The $A G M$ of Section 1 is a special case of the monster algorithm when the group is of order 2. Thus:

COROLLARY 3.1. Let $H$ be a complex Banach space, $X=\mathcal{L}(H)$ and $Y=X \times X$. Let $U=\{A \in X: \sigma(A) \cap(-\infty, 0)=\phi\}$ and let $V=\{(A, B) \in Y: A \in U$ and $B \in U\}$. For $(A, B) \in V$, define $f(A, B)$ by

$$
f(A, B)=((A+B) / 2, \quad \exp ((1 / 2) \log A+(1 / 2) \log B))
$$

Define $W$ by

$$
\begin{aligned}
& W=\left\{(A, B) \in V: f^{m}(A, B) \in V \text { for all } m \geq 1\right. \text { and there exist } \\
& \left.A_{\infty} \in U \text { and } B_{\infty} \in U \text { such that } \lim _{m \uparrow \infty}\left\|f^{m}(A, B)-\left(A_{\infty}, B_{\infty}\right)\right\|=0\right\}
\end{aligned}
$$

Then $W$ is open and if $g(A, B)$ is defined by

$$
g(A, B)=\lim _{m \uparrow \infty} f^{m}(A, B)
$$

for $(A, B) \in W$, then $(A, B) \rightarrow g(A, B)$ is analytic. If $H$ is a finite dimensional Hilbert space, $W$ contains $W_{1}$, where
$W_{1}=\{(A, B): A$ and $B$ are positive definite, bounded and self adjoint $\}$.
Proof. Corollary 3.1 follows from Theorem 3.2 if one observes that $A_{\infty}=B_{\infty}$ whenever

$$
\lim _{m \uparrow \infty}\left\|f^{m}(A, B)-\left(A_{\infty}, B_{\infty}\right)\right\|=0
$$

for some $A_{\infty} \in U, B_{\infty} \in U$. This is because the form of $f$ implies

$$
A_{\infty}=\left(A_{\infty}+B_{\infty}\right) / 2
$$

REMARK 3.1. It would be interesting to obtain more information about the set $W$ in Corollary 3.1, even for $H$ finite dimensional. For instance, is it true that almost every pair $(A, B) \in Y$ (with respect to Lebesgue measure) belongs to $W$ ? Numerical studies for $\operatorname{dim} H=2,3,4,5$ suggest this may be true.

If $H$ is a Hilbert space and $A \in \mathcal{L}(H)=\boldsymbol{X}, \boldsymbol{A}$ is called accretive if $\operatorname{Re}<A x, x>\geq 0$ for all $x \in H$, and $A$ is strictly accretive if there exists
$\alpha>0$ so that $\operatorname{Re}<A x, x>\geq \alpha\|x\|^{2}$ for all $x \in H$. It is natural to conjecture that for almost every pair $(A, B)$ such that $A$ and $B$ are strictly accretive one has $(A, B) \in W$. However, one can give an example of $2 \times 2$ upper triangular accretive matrices $A$ and $B$ such that if $\left(A_{1}, B_{1}\right)=f(A, B), B_{1}$ is not accretive. This, of course, does not disprove the conjecture.

Because $f$ is homogeneous of degree 1 and

$$
f\left(S^{-1} A S, S^{-1} B S\right)=S^{-1} f(A, B) S
$$

it is obvious that if $(A, B) \in W$, then $(\lambda A, \lambda B) \in W$ for all $\lambda>0$ and $\left(S^{-1} A S, S^{-1} B S\right) \in W$ for all invertible $S$. If $H$ is finite dimensional and $A$ and $B$ are both upper triangular matrices with spectrum strictly in the right half plane, one can also prove that $(A, B) \in W$, though we omit the proof.

There remains one easy case in which one can prove $(A, B) \in W$, that is, when $A B=B A$. Stickel [32] has discussed the commutative case when $A$ and $B$ are matrices, but his argument seeems incomplete. We shall sketch an approach which works when $H$ is a complex Banach space.

Suppose that $H$ is a complex Banach space and $A, B \in \mathcal{L}(H)=X$ are commuting linear operators. Consider the algebra $A$ of complex-valued functions $g$ which are defined and analytic on $U_{g} \times V_{g}$, where $U_{g}$ is an open neighborhood of $\sigma(A)$ and $V_{g}$ is an open neighborhood of $\sigma(B)$. Two such functions $g_{1}$ and $g_{2}$ are identified if they agree on $U \times V$, where $U$ and $V$ are some open sets containing $\sigma(A)$ and $\sigma(B)$ respectively. If $g \in A$ is defined on $U \times V$, let $\Gamma_{1} \subseteq U$ be a finite union of simple, closed rectifiable curves which contain $\sigma(A)$ in the union of their interiors, and similarly for $\Gamma_{2} \subseteq V$. Define $g(A, B) \in \mathcal{L}(H)$ by

$$
\begin{equation*}
g(A, B)=(2 \pi i)^{-2} \int_{\Gamma_{2}} \int_{\Gamma_{1}} g\left(z_{1}, z_{2}\right)\left(z_{1}-A\right)^{-1}\left(z_{2}-B\right)^{-1} d z_{1} d z_{2} \tag{3.38}
\end{equation*}
$$

The operator $g(A, B)$ defined by (3.38) does not depend on the particular choice of $\Gamma_{1}$ and $\Gamma_{2}$ as above. Furthermore, the map $g \rightarrow g(A, B) \in \mathcal{L}(H)$ is an algebra homomorphism from $A$ to $\mathcal{L}(H)$. The proof of this fact is a minor variant of the argument (see [33], [36]) for defining the functional calculus for a single operator and will not be given here. In fact, the functional calculus summarized by (3.38) is a special case of a much more subtle functional calculus developed by Shilov, Waelbrock, Arens-Calderon and Arens: see [7] for references. If $g_{n}: U \times V \rightarrow \mathbb{C}, n \geq 1$, is a sequence of analytic functions and $g_{n}$ converges uniformly on $U \times V$ to $g$, then by using (3.38) one can easily see that $g_{n}(A, B)$ converges in norm to $g(A, B)$.

If $U$ and $V$ are as above and $g: U \times V \rightarrow \mathbb{C}$ is analytic, then by using the fact that $g \rightarrow g(A, B)$ is an algebra homomorphism and that the function identically equal to 1 goes to the identity in $\mathcal{L}(H)$ one can see that

$$
\begin{equation*}
\sigma(g(A, B)) \subseteq g(\sigma(A) \times \sigma(B)) \tag{3.39}
\end{equation*}
$$

Applying (3.39) to $g(z, w)=(z+w) / 2$ and $g(z, w)=z w$ gives

$$
\begin{gather*}
\sigma((A+B) / 2) \subseteq\{(u+v) / 2: u \in \sigma(A), v \in \sigma(B)\} \text { and }  \tag{3.40}\\
\sigma(A B) \subseteq\{u v: u \in \sigma(A), v \in \sigma(B)\}
\end{gather*}
$$

We can also use the composition of analytic functions. If $g \in A$ and $h$ is analytic on an open neighborhood of $g(\sigma(A) \times \sigma(B)) \subseteq \mathbb{C}$, then $j=\mathrm{h} \circ \mathrm{g} \in A$ and (3.39) implies that $h(g(A, B))$ is defined and one can prove, as for the functional calculus in one variable, that

$$
\begin{equation*}
j(A, B)=h(g(A, B)) \tag{3.41}
\end{equation*}
$$

If $g(z, w)=\left(g_{1}(z, w), g_{2}(z, w)\right)$, where $g_{j} \in A$ for $j=1,2$, one can define

$$
g(A, B)=\left(g_{1}(A, B), g_{2}(A, B)\right)=\left(A_{1}, B_{1}\right)
$$

and it is easy to see that $A_{1}$ and $B_{1}$ commute. If $W_{j}$ is an open neighborhood of $g_{j}(\sigma(A) \times \sigma(B))$ and $h: W_{1} \times W_{2} \rightarrow \mathbb{C}^{2}$ is analytic, then $j=\mathrm{h} \circ \mathrm{g}$ is defined on $U_{0} \times V_{0}$, where $U_{0}$ is an open neighborhood of $\sigma(A)$ and $V_{0}$ an open neighborhood of $\sigma(B)$ and

$$
\begin{equation*}
j(A, B)=(\mathrm{h} \circ \mathrm{~g})(A, B)=h(g(A, B)) \tag{3.42}
\end{equation*}
$$

Now define open subsets $G \subseteq \mathbb{C}$ and $\mathcal{O} \subseteq \mathcal{L}(H)$ by

$$
\begin{align*}
& G=\{z \in \mathbb{C}: \operatorname{Re}(z)>0\} \text { and }  \tag{3.43}\\
& \mathcal{O}=\{A \in \mathcal{L}(H): \sigma(A) \subseteq G\}
\end{align*}
$$

Let $z^{1 / 2}$ denote the standard single-valued branch of the square root function:

$$
\begin{equation*}
z^{1 / 2}=\rho^{1 / 2} e^{i \Theta / 2}, \text { where } z=\rho e^{i \Theta}, \rho>0 \text { and }-\pi<\Theta<\pi \tag{3.44}
\end{equation*}
$$

It is easy to see that for $z$ and $w \in G$

$$
\exp ((1 / 2) \log z+(1 / 2) \log w)=z^{1 / 2} w^{1 / 2}=(z w)^{1 / 2}
$$

Thus, using the properties of the functional calculus, we see that if $(A, B) \in$ $0 \times 0$ and $A B=B A$, then

$$
\begin{equation*}
\exp ((1 / 2) \log A+(1 / 2) \log B)=(A B)^{1 / 2} \tag{3.45}
\end{equation*}
$$

If $G_{1}=\{z \in \mathbb{C}: z \neq 0$ and $z$ is not a negative real $\}$, then if $(A, B) \in \mathcal{O} \times \mathcal{O}$, (3.40) implies that $\sigma(A B) \subseteq G_{1}$, and the spectral mapping theorem implies that

$$
\begin{equation*}
\sigma\left((A B)^{1 / 2}\right) \subseteq\left\{z^{1 / 2}: z \in G_{1}\right\}=G \tag{3.46}
\end{equation*}
$$

(3.40) also implies that $\sigma((A+B) / 2) \subseteq G$ if $(A, B) \in O \times 0$. It follows that if we define $W=\{(A, B) \in \mathcal{O} \times \mathcal{O}: A B=B A\}$ and if $F: W \rightarrow \mathcal{L}(H) \times \mathcal{L}(H)$ is defined by

$$
\begin{equation*}
F(A, B)=\left((A+B) / 2,(A B)^{1 / 2}\right) \tag{3.47}
\end{equation*}
$$

then $F(W) \subseteq W$. Thus if $(A, B) \in W$ we can consider $F^{m}(A, B)$, where $F^{m}$ is the $m$ th iterate of $F: W \rightarrow W$.

On the other hand, if we define $f(z, w)=\left((z+w) / 2,(z w)^{1 / 2}\right)$, then $f(G \times G) \subseteq G \times G$. For $(A, B) \in W$, we have $f(A, B)=F(A, B)$, where $f(A, B)$ is defined by (3.38). If $f^{m}$ denotes the $m$ th iterate of $f: G \times G \rightarrow G \times G$, the previously mentioned properties of the functional calculus (particularly the rules of composition) imply that

$$
\begin{equation*}
f^{m}(A, B)=F^{m}(A, B) \tag{3.48}
\end{equation*}
$$

where the left side of (3.48) is defined by (3.38) with $g=f^{m}$.
The classical theory of the $A G M$ implies that for all $(z, w) \in G \times G$ one has

$$
\begin{equation*}
\lim _{m \uparrow \infty} f^{m}(z, w)=(\pi / 2)\left(g(z, w)^{-1}, g(z, w)^{-1}\right) \tag{3.49}
\end{equation*}
$$

where

$$
\begin{equation*}
g(z, w)=\int_{0}^{\infty}\left(s^{2}+z^{2}\right)^{-1 / 2}\left(s^{2}+w^{2}\right)^{-1 / 2} d s \tag{3.50}
\end{equation*}
$$

and that the convergence in (3.49) is uniform on compact subsets of $G \times G$. It follows from (3.49) and (3.50) that

$$
\begin{equation*}
\lim _{m \uparrow \infty} F^{m}(A, B)=(\pi / 2)(g(A, B))^{-1} \tag{3.51}
\end{equation*}
$$

where $g(A, B)$ in (3.51) is defined by (3.38) and $g$ is as in (3.50). Finally, a relatively simple argument (which we omit) shows that

$$
\begin{equation*}
g(A, B)=\int_{0}^{\infty}\left(s^{2}+A^{2}\right)^{-1 / 2}\left(s^{2}+B^{2}\right)^{-1 / 2} d s \tag{3.52}
\end{equation*}
$$

where the right side of (3.52) is interpreted as an improper Riemann integral with values in $\mathcal{L}(H)$. Thus:

Proposition 3.1. Let $H$ be a complex Banach space,

$$
\begin{gathered}
G=\{z \in \mathbb{C}: \operatorname{Re}(z)>0\} \\
G_{1}=\{z \in \overparen{S}: z \neq 0 \text { and } z \text { is not a negative real }\}
\end{gathered}
$$

and

$$
W=\{(A, B) \in \mathcal{L}(H) \times \mathcal{L}(H): A B=B A, \sigma(A) \subseteq G \text { and } \sigma(B) \subseteq G\}
$$

If $(A, B) \in W$, then $\sigma((A+B) / 2) \subseteq G, \sigma(A B) \subseteq G$, and if, for $(A, B) \in W$, $F(A, B)$ is defined by
$F(A, B)=((A+B) / 2, \exp ((1 / 2) \log A+(1 / 2) \log B))=\left((A+B) / 2,(A B)^{1 / 2}\right)$,
then $F(W) \subseteq W$ and

$$
\begin{equation*}
n-\lim _{m \uparrow \infty} F^{m}(A, B)=(\pi / 2)\left(g(A, B)^{-1}, g(A, B)^{-1}\right) \tag{3.53}
\end{equation*}
$$

where $g(A, B)$ is defined by (3.52).

## 4. - Alternate definitions of the $A G M$

In the previous section we have considered one type of generalization of the classical $A G M$. However, there are many possible "reasonable" generalizations of the $A G M$ to pairs of bounded linear operators $(A, B)$. In fact, as we shall see later, there is a continum of arithmetic-geometric means, all of which are defined when $A$ and $B$ are p.d. and self-adjoint, all of which give the same value when $A B=B A$, but all of which in general give different values when $A B \neq B A$.

We begin with an observation which was made to the authors by the referee of our earlier paper [14]. If $A$ and $B$ are $n \times n$ Hermitian, positive definite matrices, define $f(A, B)$ by

$$
\begin{equation*}
f(A, B)=\left((A+B) / 2, \quad B\left(B^{-1} A\right)^{1 / 2}\right) \tag{4.1}
\end{equation*}
$$

where $\left(B^{-1} A\right)^{1 / 2}$ is defined by (1.1). The referee remarked that, despite appearances, the expression $B\left(B^{-1} A\right)^{1 / 2}$ is symmetric in $A$ and $B$ and is positive definite and self-adjoint. Furthermore, he observed that for $A$ and $B$ p.d. and self-adjoint, it is relatively easy to prove that

$$
\begin{equation*}
\lim _{k \uparrow \infty} f^{k}(A, B)=(C, C) \tag{4.2}
\end{equation*}
$$

where $C$ is p.d. and self-adjoint.
If $A$ and $B$ are p.d., self-adjoint bounded linear maps of a Hilbert space $H$ to itself, J. Fujii [19] has defined a map $g$ by

$$
\begin{equation*}
g(A, B)=((A+B) / 2, \quad A \# B) \tag{4.3}
\end{equation*}
$$

where $A \# B$ is a "geometric mean" introduced by Pusz and Woronowicz [31]. Pusz and Woronowicz proved (see also Theorem 2 in [3]) that

$$
\begin{equation*}
A \# B=A^{1 / 2}\left(A^{-1 / 2} B A^{-1 / 2}\right)^{1 / 2} A^{1 / 2} \tag{4.4}
\end{equation*}
$$

Fujii proves that if $A$ and $B$ are p.d., self-adjoint operators in $\mathcal{L}(H)$, then

$$
\begin{equation*}
s-\lim _{k \uparrow \infty} g^{k}(A, B)=(C, C) \tag{4.5}
\end{equation*}
$$

where $C$ is p.d. and self-adjoint.
We shall show first that

$$
f(A, B)=g(A, B)
$$

where $f(A, B)$ is defined by (4.1) and $g(A, B)$ by (4.3) and (4.4). This will imply of course that the limits defined in (4.2) and (4.5) are equal. We begin with a trivial lemma.

Lemma 4.1. Let $H$ be a Hilbert space and suppose the $A, B \in \mathcal{L}(H)$ and $A$ and $B$ are p.d. and self-adjoint. Then $\sigma(A B)=$ spectrum of $A B \subseteq(0, \infty)$.

Proof. Because $\sigma(A B)=\sigma\left(A^{-1 / 2}(A B) A^{1 / 2}\right)=\sigma\left(A^{1 / 2} B A^{1 / 2}\right)$, and $A^{1 / 2} B A^{1 / 2}$ is p.d. and self-adjoint, the lemma is proved.

We now recall some basic fact about the functional calculus for linear operators. If $H$ is a Hilbert space, $A \in \mathcal{L}(H)$ and $f$ is analytic on an open neighborhood $U$ of $\sigma(A) \cup \sigma\left(A^{*}\right)$ and $f(\bar{z})=\overline{f(z)}$ for all $z$ in $U$, then $f\left(A^{*}\right)=(f(A))^{*}$. If $H$ is a Banach space, $A \in \mathcal{L}(H), f$ is analytic on an open neighborhood of $\sigma(A)$ and $S$ is invertible, then

$$
\begin{equation*}
f\left(S^{-1} A S\right)=S^{-1} f(A) S \tag{4.6}
\end{equation*}
$$

We shall always use the standard single-valued branches of $z^{\lambda}$ and $\log z$. Thus if

$$
\begin{equation*}
G_{1}=\{z \in \mathbb{C}: z \neq 0 \text { and } z \text { is not a negative real }\} \tag{4.7}
\end{equation*}
$$

and $z=r e^{i \Theta},|\Theta|<\pi, r>0$, then

$$
z^{\lambda}=r^{\lambda} e^{i \lambda \Theta} \quad \text { and } \quad \log z=\log r+i \Theta
$$

It follows that if $H$ is a complex Banach space, $A \in \mathcal{L}(H)$ and $\sigma(A) \subseteq G_{1}$, then for any real numbers $\lambda$ and $\mu$,

$$
\begin{equation*}
A^{\lambda} A^{\mu}=A^{\lambda}+\mu \tag{4.8}
\end{equation*}
$$

and if $\lambda$ and $\mu$ are real numbers such that $|\lambda| \leq 1$ (so $z^{\lambda} \in G_{1}$ for all $z \in G_{1}$ )

$$
\begin{equation*}
\left(A^{\lambda}\right)^{\mu}=A^{\lambda \mu},|\lambda| \leq 1 \tag{4.9}
\end{equation*}
$$

Lemma 4.2. Suppose $H$ is a complex Banach space and $A$ and $B$ are in $\mathcal{L}(H), B$ is invertible and $\sigma\left(B^{-1} A\right) \subseteq G_{1}$, where $G_{1}$ is as in (4.7). Then $A$ and $B A^{-1}$ are invertible and for any real number $\lambda$,

$$
\begin{gather*}
\sigma\left(B^{-\lambda} A B^{\lambda-1}\right)=\sigma\left(B^{-1} A\right) \subseteq G_{1} \text { and }  \tag{4.10}\\
\sigma\left(A^{-\lambda} B A^{\lambda-1}\right)=\sigma\left(A^{-1} B\right) \subseteq G_{1}
\end{gather*}
$$

Furthermore, for all real $\lambda$,

$$
\begin{gather*}
B^{\lambda}\left(B^{-\lambda} A B^{\lambda-1}\right)^{1 / 2} B^{1-\lambda}=B\left(B^{-1} A\right)^{1 / 2}  \tag{4.11}\\
A^{\lambda}\left(A^{-\lambda} B A^{\lambda-1}\right)^{1 / 2} A^{1-\lambda}=A\left(A^{-1} B\right)^{1 / 2}  \tag{4.12}\\
A\left(A^{-1} B\right)^{1 / 2}=B\left(B^{-1} A\right)^{1 / 2} \tag{4.13}
\end{gather*}
$$

If $H$ is a Hilbert space and $A$ and $B$ are p.d. and self-adjoint, $B\left(B^{-1} A\right)^{1 / 2}$ is p.d. and self-adjoint.

Proof. Because $B$ is invertible and $B^{-1} A$ is invertible, $A=B\left(B^{-1} A\right)$ is invertible and $A^{-1} B$ and $B A^{-1}$ are invertible. By using (4.8), we can write

$$
B^{-\lambda} A B^{\lambda-1}=S\left(B^{-1} A\right) S^{-1}, \quad S=B^{1-\lambda}
$$

so

$$
\sigma\left(B^{-1} A\right)=\sigma\left(S\left(B^{-1} A\right) S^{-1}\right) \subseteq G_{1}
$$

By interchanging the roles of $A$ and $B$, we obtain the other part of (4.10).
If $S=B^{1-\lambda}$, (4.6) and (4.8) yield

$$
\begin{gathered}
B^{\lambda}\left(B^{-\lambda} A B^{\lambda-1}\right)^{1 / 2} B^{1-\lambda}=B S^{-1}\left(B^{-\lambda} A B^{\lambda-1}\right)^{1 / 2} S \\
=B\left(S^{-1} B^{-\lambda} A B^{\lambda-1} S\right)^{1 / 2}=B\left(B^{-1} A\right)^{1 / 2}
\end{gathered}
$$

which is (4.11). (4.12) is obtained by a similar argument.
(4.13) is equivalent to

$$
\begin{equation*}
\left(B^{-1} A\right)\left(A^{-1} B\right)^{1 / 2}=\left(B^{-1} A\right)^{1 / 2} \tag{4.14}
\end{equation*}
$$

However, (4.9) implies that

$$
\left(A^{-1} B\right)^{1 / 2}=\left[\left(B^{-1} A\right)^{-1}\right]^{1 / 2}=\left(B^{-1} A\right)^{-1 / 2}
$$

which yields (4.14).
If $H$ is a Hilbert space and $A$ and $B$ are p.d. and self-adjoint, Lemma 4.1 implies that $\sigma\left(B^{-1} A\right) \subseteq(0, \infty)$. Thus the first part of the lemma is applicable, and taking $\lambda=1 / 2$ in (4.11) and (4.12) gives

$$
B \# A=B\left(B^{-1} A\right)^{1 / 2}=A\left(A^{-1} B\right)^{1 / 2}=A \# B
$$

It is easy to see that $B \# A$ is self-adjoint and p.d., so $B\left(B^{-1} A\right)^{1 / 2}$ is also.
Lemma 4.2 implies that the functions given by (4.1) and (4.3) respectively are equal when $A$ and $B$ are p.d., so the fact that (4.2) is valid (in the strong operator topology) follows from Fujii's theorem.

We shall now show that a much stronger result than (4.2) is valid. Let $H$ be a complex Banach space and $X=\mathcal{L}(H)$. For $k$ a positive integer, define

$$
\begin{equation*}
G_{k}=\left\{z=r e^{i \Theta}: r>0 \text { and }|\Theta|<\pi / 2^{k-1}\right\} \tag{4.15}
\end{equation*}
$$

and

$$
\begin{equation*}
U_{k}=\left\{(A, B) \in X \times X: B \text { is invertible and } \sigma\left(B^{-1} A\right) \subseteq G_{k}\right\} \tag{4.16}
\end{equation*}
$$

Lemma 4.3. Let $H$ be a complex Banach space and let $U_{k}$ be as in (4.16) for positive integers $k$. For $(A, B) \in U_{1}$, define $f(A, B)$ by

$$
\begin{equation*}
f(A, B)=\left((A+B) / 2, \quad B\left(B^{-1} A\right)^{1 / 2}\right) \tag{4.17}
\end{equation*}
$$

If $(A, B) \in U_{k}$, then $f(A, B)=\left(A_{1}, B_{1}\right) \in U_{k+1}$.
PROOF. By assumption, $\sigma\left(B^{-1} A\right) \subseteq G_{k}$, so the spectral mapping theorem implies that

$$
\sigma\left(\left(B^{-1} A\right)^{1 / 2}\right) \subseteq\left\{z^{1 / 2}: z \in G_{k}\right\}=G_{k+1}
$$

and $\left(B^{-1} A\right)^{1 / 2}$ is invertible. It is assumed that $B$ is invertible, so

$$
B_{1}=B\left(B^{-1} A\right)^{1 / 2}
$$

is invertible. Then

$$
\begin{equation*}
B_{1}^{-1} A_{1}=\left(B^{-1} A\right)^{-1 / 2} B^{-1}((A+B) / 2)=\left[\left(B^{-1} A\right)^{1 / 2}+\left(B^{-1} A\right)^{-1 / 2}\right] / 2 \tag{4.18}
\end{equation*}
$$

It follows from (4.18) and the spectral mapping theorem that

$$
\begin{align*}
\sigma\left(B_{1}^{-1} A_{1}\right) & \subseteq\left\{\left(z^{1 / 2}+z^{-1 / 2}\right) / 2: z \in \sigma\left(B^{-1} A\right)\right\} \\
& \subseteq\left\{\left(z^{1 / 2}+z^{-1 / 2}\right) / 2: z \in G_{k}\right\} \tag{4.19}
\end{align*}
$$

If $z \in G_{k}, z^{1 / 2} \in G_{k+1}$, and if $w \in G_{k}, w^{-1} \in G_{k}$; also, $G_{k+1}$ is convex for $k \geq 1$. Thus we conclude from (4.19) that $\sigma\left(B_{1}^{-1} A_{1}\right) \subseteq G_{k+1}$.

Lemma 4.4. Let $U_{1}, f$ and $H$ be as in Lemma 4.3. If $C$ and $D$ are elements of $\mathcal{L}(H)$ and $(A, B) \in U_{1}$, define

$$
C(A, B)=(C A, C B) \text { and }(A, B) D=(A D, B D)
$$

If $C$ and $D$ are invertible and $(A, B) \in U_{1}$, then $C(A, B) D \in U_{1}$ and

$$
\begin{equation*}
f^{m}(C(A, B) D)=C f^{m}(A, B) D \tag{4.20}
\end{equation*}
$$

for all $m \geq 1$.
Proof. If $(A, B) \in U_{1}, B$ is invertible, so $C B D$ is also invertible. Then

$$
C(A, B) D=(C A D, C B D)
$$

so $(C B D)^{-1}(C A D)=\left(D^{-1} B^{-1} C^{-1}\right)(C A D)=D^{-1}\left(B^{-1} A\right) D$ and

$$
\sigma\left((C B D)^{-1}(C A D)\right)=\sigma\left(D^{-1}\left(B^{-1} A\right) D\right)=\sigma\left(B^{-1} A\right) \subseteq G_{1}
$$

This shows that $C(A, B) D \subseteq U_{1}$.
The above calculation also shows that $\left((C B D)^{-1}(C A D)\right)^{1 / 2}=$ $D^{-1}\left(B^{-1} A\right)^{1 / 2} D$, so $(C B D)\left((C B D)^{-1}(C A D)\right)^{1 / 2}=C B\left(B^{-1} A\right)^{1 / 2} D$. It follows that

$$
f(C(A, B) D)=C f(A, B) D
$$

and a simple mathematical induction (left to the reader) gives (4.20).
With these preliminaries we can prove an extension of Fujii's theorem.
THEOREM 4.1. Let $H$ be a complex Banach space, $\boldsymbol{X}=\mathcal{L}(H), G_{1}=\{z=$ $r e^{i \Theta} \in \mathbb{C}: r>0$ and $\left.|\Theta|<\pi\right\}$, and $U_{1}=\{(A, B) \in X \times X: B$ is invertible and $\left.\sigma\left(B^{-1} A\right) \subseteq G_{1}\right\}$. For $(A, B) \in U_{1}$, define

$$
f(A, B)=\left((A+B) / 2, \quad B\left(B^{-1} A\right)^{1 / 2}\right)
$$

Then $f\left(U_{1}\right) \subseteq U_{1}$ and for every $(A, B) \in U_{1}$ there exists $E \in X, E$ invertible, such that

$$
\begin{equation*}
n-\lim _{k \uparrow \infty} f^{k}(A, B)=(E, E) \tag{4.21}
\end{equation*}
$$

If $H$ is a Hilbert space and $A$ and $B$ are p.d. and self-adjoint, then $(A, B) \in U_{1}$, and if $\left(A_{1}, B_{1}\right)=f(A, B), A_{1}$ and $B_{1}$ are p.d. and self-adjoint.

PROOF. Lemma 4.3 implies that $f^{k}(A, B) \in U_{k+1} \subseteq U_{1}$, where $U_{k}$ is given by (4.16). If we write $f^{k}(A, B)=\left(A_{k}, B_{k}\right)$, we obtain from Lemma 4.4 that

$$
\begin{equation*}
f^{k}(A, B)=A_{1} f^{k-1}\left(I, A_{1}^{-1} B_{1}\right), k \geq 2 \tag{4.22}
\end{equation*}
$$

If $G_{2}$ is as in (4.15) and $C$ and $D$ are any commuting bounded linear operators such that $\sigma(C) \subseteq G_{2}$ and $\sigma(D) \subseteq G_{2}$, the properties of the functional calculus for commuting operators (see the end of Section 3) imply that

$$
f(C, D)=\left((C+D) / 2, D\left(D^{-1} C\right)^{1 / 2}\right)=\left((C+D) / 2,(C D)^{1 / 2}\right)=\left(C_{1}, D_{1}\right) .
$$

Thus, for such $C$ and $D$, Proposition 3.1 implies that there exists an invertible $E \in X$ such that

$$
\begin{equation*}
n-\lim _{k \uparrow \infty} f^{k-1}(C, D)=(E, E) \tag{4.23}
\end{equation*}
$$

Now take $C=I$ and $D=A_{1}^{-1} B_{1}$. Lemma 4.3 implies that $\sigma(D) \subseteq G_{2}$ and of course these $C$ and $D$ commute. Thus we conclude from (4.22) and (4.23) that

$$
\lim _{k \uparrow \infty}\left\|f^{k}(A, B)-(E, E)\right\|=0
$$

The statements in Theorem 4.1 concerning the self-adjoint case follow directly from lemmas 4.1 and 4.2 .

REmARK 4.1. The above argument shows that

$$
\begin{equation*}
\lim _{k \uparrow \infty} f^{k}(A, B)=(\pi / 2) A_{1}\left(g\left(I, A_{1}^{-1} B_{1}\right)^{-1}, g\left(I, A_{1}^{-1} B_{1}\right)^{-1}\right) \tag{4.24}
\end{equation*}
$$

where $\left(A_{1}, B_{1}\right)=f(A, B)$ and $g(A, B)$ is defined in (3.52). If $\sigma\left(A^{-1} B\right)$ is contained in the right half plane, one can replace $A_{1}$ by $A$ and $B_{1}$ by $B$ in (4.24).

REMARK 4.2. Even if $A$ and $B$ are p.d. and self-adjoint, Theorem 4.1 gives new information: the convergence in (4.21) is in the operator norm, whereas in Fujii's theorem convergence is in the strong operator topology.

Theorem 4.1 gives a reasonable definition of the $A G M$ which does not in general agree with the definition in Section 1. We now show that there is a family of reasonable definitions of the $A G M$, parametrized by $\lambda \geq 1$, which reduce to the definition in Theorem 4.1 for $\lambda=1$ and give the definition in Section 1 for $\lambda=\infty$.

It is convenient to prove another lemma first.
Lemma 4.5. Let $H$ be a complex Hilbert space and suppose that $A, B \in \mathcal{L}(H)$ and $A$ and $B$ are p.d. and self-adjoint. Then for every $\lambda \geq 1$

$$
\begin{equation*}
B\left(B^{-1} A\right)^{1 / 2} \leq(A+B) / 2 \leq\left(\left(A^{\lambda}+B^{\lambda}\right) / 2\right)^{1 / \lambda} \tag{4.25}
\end{equation*}
$$

If $B \leq A$, then also

$$
\begin{equation*}
B \leq B\left(B^{-1} A\right)^{1 / 2} \tag{4.26}
\end{equation*}
$$

PROOF. The right inequality in (4.25) is inequality (1.43) and has already been proved. By multiplying on the left and the right by $B^{-1 / 2}$, one sees that $B\left(B^{-1} A\right)^{1 / 2} \leq(A+B) / 2$ if and only if

$$
\begin{equation*}
B^{1 / 2}\left(B^{-1} A\right)^{1 / 2} B^{-1 / 2}=\left(B^{-1 / 2} A B^{-1 / 2}\right)^{1 / 2} \leq\left(B^{-1 / 2} A B^{-1 / 2}+I\right) / 2 \tag{4.27}
\end{equation*}
$$

If we define

$$
L=\left(B^{-1 / 2} A B^{-1 / 2}\right)^{1 / 2}
$$

$L$ is a p.d., self-adjoint operator, and (4.27) is equivalent to $2 L \leq L^{2}+I$, which is certainly true, because $0 \leq(L-I)^{2}=L^{2}-2 L+I$.

If $B \leq A$, then $I=B^{-1 / 2} B B^{-1 / 2} \leq B^{-1 / 2} A B^{-1 / 2}$ and

$$
\begin{equation*}
I \leq\left(B^{-1 / 2} A B^{-1 / 2}\right)^{1 / 2} \tag{4.28}
\end{equation*}
$$

Multiplying (4.28) on the left and right by $B^{1 / 2}$ and using Lemma 4.2 gives

$$
B \leq B^{1 / 2}\left(B^{-1 / 2} A B^{-1 / 2}\right)^{1 / 2} B^{1 / 2}=B\left(B^{-1} A\right)^{1 / 2}
$$

If $H$ is a complex Hilbert space and $A, B$ are p.d., self-adjoint bounded linear operators, define, for fixed $\lambda \geq 1$,

$$
\begin{equation*}
f_{\lambda}(A, B)=\left((A+B) / 2,\left[B^{1 / \lambda}\left(B^{-1 / \lambda} A^{1 / \lambda}\right)^{1 / 2}\right]^{\lambda}\right)=\left(A_{1}, B_{1}\right) \tag{4.29}
\end{equation*}
$$

Since $A^{1 / \lambda}$ and $B^{1 / \lambda}$ are p.d. and self-adjoint, Lemma 4.2 implies that $B_{1}$ is p.d. and self- adjoint.

TheOrem 4.2. Let $H$ be a complex Hilbert space. For $A, B \in \mathcal{L}(H), A$ and $B$ p.d. and self-adjoint, define $f_{\lambda}(A, B)$ by (4.29). Then for any pair $(A, B)$ of p.d., bounded, self-adjoint linear operators,

$$
\begin{equation*}
s-\lim _{k \uparrow \infty} f_{\lambda}^{k}(A, B)=(E, E) \tag{4.30}
\end{equation*}
$$

where $E$ is p.d. and self-adjoint and $f_{\lambda}^{k}$ denotes the kth iterate of $f_{\lambda}$.
Proof. Define

$$
\begin{aligned}
g_{\lambda}(A, B)= & \left(\left[\left(A^{\lambda}+B^{\lambda}\right) / 2\right]^{1 / \lambda}, B\left(B^{-1} A\right)^{1 / 2}\right) \\
& \Psi_{\lambda}(A, B)=\left(A^{1 / \lambda}, B^{1 / \lambda}\right)
\end{aligned}
$$

so

$$
\Psi_{\lambda}^{-1}(A, B)=\left(A^{\lambda}, B^{\lambda}\right)
$$

A simple calculation shows that

$$
f_{\lambda}=\Psi_{\lambda}^{-1} g_{\lambda} \Psi_{\lambda}
$$

so

$$
f_{\lambda}^{k}=\Psi_{\lambda}^{-1} g_{\lambda}^{k} \Psi_{\lambda} .
$$

Thus it suffices to prove that if $A$ and $B$ are p.d. and self-adjoint, there exists $E$ such that $E$ is p.d. and self-adjoint and $g_{\lambda}^{k}(A, B) \rightarrow(E, E)$. If we define

$$
g_{\lambda}^{k}(A, B)=\left(A_{k}, B_{k}\right)
$$

Lemma 4.5 implies that

$$
\begin{equation*}
A_{k} \geq B_{k} \quad \text { for } k \geq 1 \tag{4.31}
\end{equation*}
$$

Using (4.31) and Lemma 4.5 again,

$$
\begin{equation*}
B_{k+1}=B_{k}\left(B_{k}^{-1} A_{k}\right)^{1 / 2} \geq B_{k} \tag{4.32}
\end{equation*}
$$

so ( $B_{k}$ ) is monotonic increasing for $k \geq 1$.
Select positive constants $\alpha$ and $\beta$ so that

$$
\alpha I \leq A_{1}, \quad B_{1} \leq \beta I
$$

Assume by mathematical induction that

$$
\begin{equation*}
\alpha I \leq A_{k}, \quad B_{k} \leq \beta I \tag{4.33}
\end{equation*}
$$

for some $k \geq 1$. The spectral mapping theorem implies that

$$
\alpha^{\lambda} I \leq A_{k}^{\lambda}, \quad B_{k}^{\lambda} \leq \beta^{\lambda} I
$$

and

$$
\alpha I \leq A_{k+1}=\left(\left(A_{k}^{\lambda}+B_{k}^{\lambda}\right) / 2\right)^{1 / \lambda} \leq \beta I .
$$

Also

$$
\alpha I \leq B_{k} \leq B_{k+1} \leq A_{k+1} \leq \beta I
$$

so (4.33) is satisfied for all $k \geq 1$ by mathematical induction.
From (4.32) and (4.33), there exists $E \in \mathcal{L}(H), E$ p.d. and self-adjoint so that for all $x \in H$,

$$
\lim _{k \uparrow \infty}\left\|B_{k} x-E x\right\|=0
$$

In particular, for all $x \in H$

$$
\begin{equation*}
\lim _{k \nmid \infty}\left(B_{k+1}-B_{k}\right) x=0 \tag{4.34}
\end{equation*}
$$

Using (4.11) to write $B_{k+1}$ in a different form, we see that (4.34) implies

$$
\begin{equation*}
\lim _{k \uparrow \infty} B_{k}^{1 / 2}\left[\left(B_{k}^{-1 / 2} A_{k} B_{k}^{-1 / 2}\right)^{1 / 2}-I\right] B_{k}^{1 / 2} x=0 \tag{4.35}
\end{equation*}
$$

for all $x \in H$. On the other hand, one can prove by using (4.33) that for $k \geq 1$

$$
\begin{equation*}
\left\|B_{k}^{1 / 2}\left[\left(B_{k}^{-1 / 2} A_{k} B_{k}^{-1 / 2}\right)^{1 / 2}+I\right] B_{k}^{-1 / 2}\right\| \leq M \tag{4.36}
\end{equation*}
$$

where $M$ is a constant independent of $k \geq 1$. Combining (4.35) and (4.36) we see that for all $x \in H$

$$
\begin{equation*}
\lim _{k \uparrow \infty}\left(A_{k}-B_{k}\right) x=\lim _{k \uparrow \infty}\left(H_{k} G_{k}\right)(x)=0 \tag{4.37}
\end{equation*}
$$

where

$$
G_{k}=B_{k}^{1 / 2}\left[\left(B_{k}^{-1 / 2} A_{k} B_{k}^{-1 / 2}\right)^{1 / 2}-I\right] B_{k}^{1 / 2}
$$

and

$$
H_{k}=B_{k}^{1 / 2}\left[\left(B_{k}^{-1 / 2} A_{k} B_{k}^{-1 / 2}\right)^{1 / 2}+I\right] B_{k}^{-1 / 2}
$$

(4.37) implies that $A_{k}$ also converges to $E$ in the strong operator topology.

The reason for considering the maps $f_{\lambda}$ is:
Theorem 4.3. Let $H$ be a complex Hilbert space and let $A, B \in \mathcal{L}(H)$ be p.d., self-adjoint operators such that

$$
\begin{equation*}
\alpha I \leq A, B \leq \beta I \tag{4.38}
\end{equation*}
$$

where $\alpha$ and $\beta$ are positive reals. Then, for $f_{\lambda}$ as in (4.29),

$$
\begin{equation*}
n-\lim _{\lambda \rightarrow \infty} f_{\lambda}(A, B)=\left(\frac{A+B}{2}, \exp \left(\left(\frac{1}{2}\right) \log A+\left(\frac{1}{2}\right) \log B\right)\right) . \tag{4.39}
\end{equation*}
$$

The convergence in (4.39) is uniform for pairs $A, B$ which satisfy (4.38) for fixed positive numbers $\alpha$ and $\beta$.

Proof. We shall use the standard "big oh" notation. Thus if $\boldsymbol{R}_{\lambda} \in \mathcal{L}(H)$ is defined for all large $\lambda$, we shall write

$$
R_{\lambda}=O\left(\lambda^{-p}\right)
$$

if there exists a constant $M$ such that

$$
\begin{equation*}
\left\|R_{\lambda}\right\| \leq M \lambda^{-p} \tag{4.40}
\end{equation*}
$$

for all $\lambda \geq \lambda_{0}$. In our case $R_{\lambda}$ will always depend on operators $A$ and $B$ which satisfy (4.38), and the constant $M$ in (4.40) and the number $\lambda_{0}$ can always be chosen to depend only on $\alpha$ and $\beta$.

The properties of the functional calculus imply that

$$
\begin{aligned}
A^{1 / \lambda} & =\exp ((1 / \lambda) \log A) \\
B^{-1 / \lambda} & =\exp ((-1 / \lambda) \log B)
\end{aligned}
$$

By using the Taylor series for the exponential one obtains

$$
\begin{align*}
B^{-1 / \lambda} & =I-(1 / \lambda) \log B+O\left(\lambda^{-2}\right)  \tag{4.41}\\
A^{1 / \lambda} & =I+(1 / \lambda) \log A+O\left(\lambda^{-2}\right)
\end{align*}
$$

(4.41) gives

$$
\begin{equation*}
B^{-1 / \lambda} A^{1 / \lambda}=I+(1 / \lambda)(\log A-\log B)+O\left(\lambda^{-2}\right) \tag{4.42}
\end{equation*}
$$

The binomial theorem is applicable to $(I+C)^{1 / 2}$ when $C \in \mathcal{L}(H)$ satisfies $\|C\|<1$, so for $\lambda$ large enouch (4.42) implies that

$$
\begin{equation*}
\left(B^{-1 / \lambda} A^{1 / \lambda}\right)^{1 / 2}=I-(1 /[2 \lambda])(\log B-\log A)+O\left(\lambda^{-2}\right) \tag{4.43}
\end{equation*}
$$

By using (4.43) and the formula

$$
B^{1 / \lambda}=I+(1 / \lambda) \log B+O\left(\lambda^{-2}\right)
$$

we obtain

$$
\begin{equation*}
B^{1 / \lambda}\left(B^{-1 / \lambda} A^{1 / \lambda}\right)^{1 / 2}=I+(1 /[2 \lambda])(\log A+\log B)+O\left(\lambda^{-2}\right) \tag{4.44}
\end{equation*}
$$

If $C \in \mathcal{L}(H)$ and $\|C\|<1$, one has the Taylor series

$$
\log (I+C)=\sum_{k=1}^{\infty}(-1)^{k-1} C^{k} / k
$$

By applying this formula to (4.44), one finds that for large $\lambda$

$$
\begin{equation*}
\log \left(B^{1 / \lambda}\left(B^{-1 / \lambda} A^{1 / \lambda}\right)^{1 / 2}\right)=(1 / 2 \lambda)(\log A+\log B)+O\left(\lambda^{-2}\right) \tag{4.45}
\end{equation*}
$$

The functional calculus implies that

$$
\begin{equation*}
\left[B^{1 / \lambda}\left(B^{-1 / \lambda} A^{1 / \lambda}\right)^{1 / 2}\right]^{\lambda}=\exp \left[\lambda \log \left(B^{1 / \lambda}\left(B^{-1 / \lambda} A^{1 / \lambda}\right)^{1 / 2}\right)\right] \tag{4.46}
\end{equation*}
$$

and combining (4.45) and (4.46) yields

$$
\begin{equation*}
\left[B^{1 / \lambda}\left(B^{-1 / \lambda} A^{1 / \lambda}\right)^{1 / 2}\right]^{\lambda}=\exp ((1 / 2) \log A+(1 / 2) \log B+O(1 / \lambda)) \tag{4.47}
\end{equation*}
$$

$$
=\exp ((1 / 2) \log A+(1 / 2) \log B)+O(1 / \lambda)
$$

which completes the proof.
REMARK 4.3. It is not hard to prove a direct analogue of Theorem 3.1 for the maps $f_{\lambda}, \lambda \geq 1$. In particular, one can prove that there exists a positive
number $\varepsilon$, independent of $\lambda \geq 1$, such that if $C, D \in \mathcal{L}(H),\|C\|<\varepsilon$ and $\|D\|<\varepsilon$ and $A=I+C$ and $B=I+D$, then $f_{\lambda}(A, B)$ is defined and

$$
\begin{equation*}
n-\lim _{k \uparrow \infty} f_{\lambda}^{k}(A, B)=\left(M_{\lambda}(A, B), \quad M_{\lambda}(A, B)\right) \tag{4.48}
\end{equation*}
$$

and $M_{\lambda}(A, B)$ is an analytic function of $(A, B)$. If we use the Lie bracket notation,

$$
[A, B]=A B-B A
$$

for operators $A$ and $B$ in $\mathcal{L}(H)$ and if we define $\mu=\lambda^{-1}$, then an unpleasant calculation (which we omit) proves that for $A=I+C$ and $B=I+D$ and $\varepsilon$ sufficiently small ( $\varepsilon$ independent of $\lambda \geq 1$ )

$$
\begin{align*}
& {\left[B^{1 / \lambda}\left(B^{-1 / \lambda} A^{1 / \lambda}\right)^{1 / 2}\right]^{\lambda}=I+(1 / 2)(C+D)-(1 / 8)(C-D)^{2}} \\
& +(1 / 16)(C-D)(C+D)(C-D)+((\mu-1) / 24)([[C, D], C-D])  \tag{4.49}\\
& +\left((\mu-1)^{2} / 48\right)([[C, D], C-D])+R_{4}(\mu, C, D)
\end{align*}
$$

There exists a constant $M$ independent of $\lambda \geq 1$ such that

$$
\begin{equation*}
\left\|R_{4}(\mu, C, D)\right\| \leq M(\|C\|+\|D\|)^{4} \tag{4.50}
\end{equation*}
$$

where $R_{4}(\mu, C, D)$ is as in (4.49). By using (4.49), one can prove fairly easily that

$$
\begin{align*}
M_{\lambda}(A, B) & =I+(C+D) / 2-(1 / 16)(C-D)^{2} \\
& +(1 / 32)(C-D)(C+D)(C-D)  \tag{4.51}\\
& +((\mu-1) / 48)([[C, D], C-D]) \\
& +\left((\mu-1)^{2} / 96\right)([[C, D], C-D])+R(\mu, C, D)
\end{align*}
$$

where

$$
\|R(\mu, C, D)\| \leq M_{1}(\|C\|+\|D\|)^{4}
$$

and $M_{1}$ is independent of $\lambda \geq 1$. By using (4.51) one can see that in general the operators $M_{\lambda}(A, B)$ are different for all $\lambda \geq 1$.

REMARK 4.4. If $\sigma=\left(\sigma_{1}, \sigma_{2}, \cdots, \sigma_{n}\right)$ is a probability vector and $A_{1}, A_{2}, \cdots, A_{n}$ are positive reals, one can consider $\prod_{i=1}^{n} A_{i}^{\sigma_{i}} \equiv A^{\sigma}$. This remark concerns what is a reasonable analogue of $A^{\sigma}$ when ${ }_{i=1}^{\boldsymbol{A}_{1}}, \boldsymbol{A}_{2}, \cdots, A_{n}$ are positive definite linear operators. One possibility is (0.7). However, if $n=2$, another possibility is

$$
\begin{equation*}
\left(A_{1}, \sigma_{1}\right) \#\left(A_{2}, \sigma_{2}\right) \equiv A_{1}\left(A_{1}^{-1} A_{2}\right)^{\sigma_{2}} \tag{4.51}
\end{equation*}
$$

Using the methods of this section, one can easily show that the right side of (4.51) is positive definite and

$$
\left(A_{1}, \sigma_{1}\right) \#\left(A_{2}, \sigma_{2}\right)=\left(\boldsymbol{A}_{2}, \sigma_{2}\right) \#\left(A_{1}, \sigma_{1}\right)
$$

However, if $n=3$, there are at least three possible reasonable analogues of $A_{1}^{\sigma_{1}} A_{2}^{\sigma_{2}} A_{3}^{\sigma_{3}}$, where $\sigma=\left(\sigma_{1}, \sigma_{2}, \sigma_{3}\right)$ is a probability vector. One is

$$
\begin{equation*}
\left(B^{\sigma_{1}+\sigma_{2}}, \sigma_{1}+\sigma_{2}\right) \#\left(A_{3}, \sigma_{3}\right) \tag{4.52}
\end{equation*}
$$

where $B=\left(A_{1}, \sigma_{1} /\left(\sigma_{1}+\sigma_{2}\right)\right) \#\left(A_{2}, \sigma_{2} /\left(\sigma_{1}+\sigma_{2}\right)\right)$. Another is

$$
\begin{equation*}
\left(A_{1}, \sigma_{1}\right) \#\left(C^{\sigma_{2}+\sigma_{3}}, \sigma_{2}+\sigma_{3}\right) \tag{4.53}
\end{equation*}
$$

where $C=\left(\boldsymbol{A}_{2}, \sigma_{2} /\left(\sigma_{2}+\sigma_{3}\right)\right) \#\left(\boldsymbol{A}_{3}, \sigma_{3} /\left(\sigma_{2}+\sigma_{3}\right)\right)$. The formulas (4.52) and (4.53) define positive definite operators if $\boldsymbol{A}_{1}, \boldsymbol{A}_{2}$, and $\boldsymbol{A}_{3}$ are positive definite, but numerical examples show that they are different in the absence of commutativity assumptions.

There does not appear to be a single "right" generalization of the AGM to three or more positive definite operators.

## Acknowledgments

J.E.C. was supported in part by U.S. National Science Foundation grant BSR 84-07461, a Fellowship from the John D. and Catherine T. MacArthur Foundation, and the hospitality of Mr. and Mrs. William T. Golden.

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[^0]:    * Partially supported by National Science Foundation grant DMS 85-03316.
    + Partially supported by National Science Foundation grant BSR 84-07461. Pervenuto alla Redazione il 20 Giugno 1987.

