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NORMS OF CERTAIN RATIONAL FUNCTIONS OF A MATRIX AND SCHUR'S

## CRITERION FOR POLYNOMIALS

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Abstract: A theoren of Schur asserts that all the roots of a poly nomial $p$ are of modulus less than one if and only if a certain Hermitian form derived from $p$ is positive definite. It is shown here that there is a wide class of Hermitian forms with the same property. The proof depends on the following fact: if $g$ is a rational function of the form $g(z)=p(z) / z^{n} p(1 / z)^{-}$and $A$ is an $n \times n$ natrix satisfying rank $\left(I-A^{*} \mathbb{A}\right)=1,\|A\| \leqslant 1$ and having spectral radius less than one, then $\|g(A)\|<1$ or $\|g(A)\|>1$ depending on whether or not ail roots of $p$ lie in the open unit disc (\|l.\|) denotes the operator norm on n-dimensional Hilbert space).

Key words: Norm, spectral radius, polynomial, root, Hermitian Iorm.

AMS: Primary 15A42, 12D10
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In 1917 I. Schur wrote a paper [5] on a subject which at that time interested many mathematicians, namely, the relation between the $H^{\infty}$-norm of a function and its Taylor coefficients. Schur's paper is still cited for two reasons: firstly, it is the source of the formula for the determinant of a partitioned matrix which is known as Schur's formula, and secondly, it contains his well-known criterion for a polynomial to have all its roots in the open unit disc. These were, however, merely a lemma and a corollary of the
main results, and the paper must be judged to have led to a dead end. Apart from a slight further development by Nevanlinna [1], no further use seens to have been made of the algebraic techniques developed by Schur. And indeed, many facts obtained in the paper by laborious algebraic computation can be seen much more easily and naturally by means of the geonetric and analytic approaches now familiar to us, but which were in their infancy in 1917. For example, after fifteen pages of mighty wielding of determinants Schur arrives at a result (Theorem $X$ ) expressible in modern terminology as follows: if $g$ is a formal power series then the norm of the operation of multiplication by $g$, thought of as an operator on the Hardy space $\mathrm{H}^{2}$ of the open unit disc, is equal to the $H^{\infty}$-norm of $g$. Nowadays this fact is almost obvious: it can be proved in a few lines with the aid of the Poisson kernel. But all this is not to say that Schur has been superseded entirely, for his methods als o yield at the same time facts which are not at all so obvious to modern eyes. Such are, roughly speaking, finite dimensional versions of results of the above type holding for special classes of functions $g$. Here is an illustration. Let $g$ be a given by a formula

$$
g(z)=\frac{p(z)}{z^{n} p(l / \overline{\mathbf{z}})^{-}}
$$

where $p$ is a polynomial of degree $n$, and let $S_{n}$ be the shift operator on n-dimensional Hilbert space:

$$
S_{n}\left(x_{1}, \ldots, x_{n}\right)=\left(x_{2}, \ldots, x_{n}, 0\right)
$$

Then $g$ is bounded in the open unit disc if and only if $g\left(S_{n}\right)$ is a contraction. The criterion for polynomials mentioned
above is a corollary of this fact.
Generally speaking questions about operators are easier in finite than in infinite dimensions, but here we have to deal with questions which gain their interest from finite dimensionality, and are trivial in the infinite-dimensional case. In view of the fact that our understanding of Schur's main results gains so much in simplicity, clarity and generality fron a functional analytic approach, it is natural to look for a similarly advantageous treatment of problems of the latter sort. This article makes a contribution to this programme.

Problems of such a nature have in fact been handle d geometrically, but only relatively recently. One of the earliest examples is the paper of V. Pták [2], where it is proved that if $A$ is a contraction on n-dimensional Hilbert space having spectral radius less than one then $A^{n}$ has norm strictly less than one. Several other proofs of this fact have been published since: references to them can be found in [3]. We begin with a generalization of Pták's theorem. The symbol || . || will denote both the norn of an element of a Hilbert space $H$ and the operator norm of an operator on $H$, while $|A|_{\sigma}$ will denote the spectral radius of an operator $A$ on $H$.

Theorem 1 Let $A_{1}, A_{2}, \ldots, A_{n}$ be commuting linear operators on an $n$-dimensional Hilbert space $H$, and let $\left|A_{i}\right|_{\sigma}<1$ and $\left\|A_{i}\right\| \leq$ for each $i$. Then $\left\|A_{1} A_{2} \ldots A_{n}\right\|<1$.

This result has already been proved and used [6], but the prod is very short and is worth repeating since it should make the sequel easier to follow.

Let us first recall the fact that if an operator $A$ on $H$ satisfies $\|\mathbb{A}\| \leqslant 1$, then $\{x:\|A x\|=\|x\|\}$ is a subspace of H; for we have I - $A^{*} \mathbf{A} \geq 0$, so that I - $A^{*} A$ has a Hermitian square root, and consequently

$$
\begin{aligned}
\|A x\|=\|x\| & \Longleftrightarrow\left(\left(I-A^{*} A\right) x, x\right)=0 \\
& \Longleftrightarrow\left\|\left(I-A^{*} A\right)^{1 / 2} x\right\|^{2}=0 \\
& \Longleftrightarrow x \in \operatorname{Ker}\left(I-A^{*} A\right)^{1 / 2} .
\end{aligned}
$$

We may thus introduce subspaces $V_{i}, 0 \leqslant i \leqslant n$, of $H$ defined by
(I) $\nabla_{i}= \begin{cases}H & \text { if } i=0 ; \\ \left\{x \in H:\left\|A_{1} \Delta_{2} \ldots A_{i} x\right\|=\|x\|\right\} \text { if } i>0 .\end{cases}$

It follows from the commatativity of the $A$ 's that $V_{i+1} \subseteq V_{i}$. We show that if $V_{i} \neq\{0\}$ then $V_{i+1}$ is in fact a proper subspace of $V_{i}$. The restriction of $A_{i}$ to $\nabla_{i}$ is an isometry for each $i$, and thus, for $x \in V_{i+1}$,

$$
\left\|A_{1} \ldots A_{i}\left(A_{i+1} x\right)\right\|=\|x\|=\left\|A_{i+1} x\right\|,
$$

which implies that $A_{i+1} \mathbf{x} \in \mathbf{V}_{i}$. In other words, $\boldsymbol{A}_{i+1} \boldsymbol{V}_{i+1} \subseteq \boldsymbol{V}_{\boldsymbol{i}}$. If, for some $i, \nabla_{i+1}=\nabla_{i} \neq\{0\}$, then $V_{i+1}$ is a non-trivial subspace of $H$, invariant with respect to $\mathbf{A}_{\mathbf{i}+1}$, on which $\mathbf{A}_{\mathbf{i}+1}$ is isonetric. It follows that $A_{i+1}$ has an eigenvalue of unit modulus, contrary to hypothesis. It follows that the sequence $H \supseteq V_{1} \supseteq V_{2} \supseteq \ldots$ descends strictly until it reaches $\{0\}$, and hence $V_{n}=\{0\}$. This implies the desired conclusion $\left\|A_{1} A_{2} \ldots \Delta_{n}\right\|<1$. The main result of the paper is in a kindred spirit.

Theoren 2. Let $p$ be a polynomial of degree $k \geq 1$ havink no two roots conjugate with respect to the unit circle and let $g$ be the rational function given by the formula:

$$
\begin{equation*}
g(z)=\frac{p(z)}{z^{k} p(1 / z)^{-}} \tag{2}
\end{equation*}
$$

Let A be a linear operator on an n-dimensional Hilbert space H satisfying $\|\Lambda\| \leqslant 1,|A|_{\Sigma}<1$ and rank $\left(I-A^{*} A\right)=r$. If $r k \leqslant n$, sone root of $p$ has modulus ereater than one and $g(A)$ is defined, then $\|\varepsilon(A)\|>1$.

It should be emphasized that the hypothesis on the roots of $p$ is suppesed to include the condition that no root be conjugate to itself, i.e. be of unit modulus. The hypothesis thus implies that the expression (2) for $g$ is in its lowest terns.
c(A) fails to be defined exactly when an eigenvalue of $\mathbf{A}$ coincides with a pole of $q$, or in other words, is conjugate to a root of p with respect to the unit circle. This can of course oceur since the eigenvalues of A lie inside the unit circle while $p$ is supposed to have a root outside it.

Proof of Theorem 2. We can suppose that $p$ is nonic. Let

$$
p(z)=\left(z-\alpha_{1}\right)\left(z-\alpha_{2}\right) \ldots\left(z-\alpha_{k}\right)
$$

and suppese that $g(A)$ is defined. This neans that $I-\bar{\alpha}_{j} A$ is non-singular, $l \leqslant j \leqslant k$, so that we can introduce the operator $A_{j}$ given by

$$
\begin{equation*}
\Lambda_{j}=\left(\Lambda-\infty_{j} I\right)\left(I-\bar{\alpha}_{j} \Lambda\right)^{-1} \tag{3}
\end{equation*}
$$

We have then

$$
g(A)=\Lambda_{1} A_{2} \cdots \Lambda_{k},
$$

and the $A_{j}$ 's all commute. We wish to introduce a descending sequence of subspaces of $H$, uch as in the proof of Theoren 1, but the definition (1) will no longer serve since we are dealing with $\mathbf{A}_{j}$ which need not satisfy $\left\|\mathbf{A}_{j}\right\| \leq 1$. It turns out that a sonewhat nore complicated definition is appropriate.

For any subset $\sigma$ of $\{1, \ldots, k\}$ let $\Lambda_{\sigma}=\prod_{j \in \sigma} \Lambda_{j}$ ( $\Lambda_{\phi}$ being defined to be the identity operator), and let
(4) $V_{i}= \begin{cases}H & \text { if } i=0 \\ \left\{x \in H:\left\|\Lambda_{\sigma} x\right\|=\|x\|\right. & \text { for every } \sigma \subseteq\{1, \ldots, i\}\} \\ & \text { if } 1 \leq i \leq k .\end{cases}$

Lemma. (i) $V_{i}$ is a subspace of $H$ of dimension at least $n-r i, 0 \leqslant i \leqslant k$.
(ii) $V_{i}$ is a proper subspace of $V_{i-1}, i \leqslant i \leqslant n$.
(iii) If $x \in V_{i-1} \backslash V_{i}, 1 \leqslant i \leqslant n$, then $\left\|A_{1} \Lambda_{2} \ldots A_{i} x\right\| \neq$ $\neq\|x\|$.

Proof of Lemma. It is an immediate consequence of the definition of $v_{i}$ that the left hand side of the following identity is contained in the right hand side:
(5) $V_{i}=v_{i-1} \cap\left\{x \in H:\left\|A_{i} x\right\|=\|x\|\right\} \cap \mathbb{A}_{i}^{-1} V_{i-1}$. Now consider any element $x$ of the right hand side and any subset $\sigma$ of $\{1, \ldots, i\}$. If i $\notin \sigma$ then $\left\|A_{\sigma} x\right\|=\|x\|$ since $x \in V_{i-1}$. Otherwise we may write $\sigma=\tau \cup\{i\}, \tau £\{1, \ldots$ $\ldots, i-1\}$, and (since the $A_{j}$ 's conmute) $\Lambda_{\sigma}=\Lambda_{\tau} A_{i}$. Since ${A_{i}}^{x} \subseteq \mathrm{~V}_{\mathbf{i}-1}$ we have $\left\|\mathbf{A}_{\boldsymbol{\sigma}} \mathbf{x}\right\|=\left\|\boldsymbol{\Lambda}_{\boldsymbol{\tau}} \mathbf{A}_{i} \mathbf{x}\right\|=\left\|\mathbf{A}_{\mathbf{i}} \mathbf{x}\right\|=\|\mathbf{x}\|$. Hence $x \in V_{i}$ and (5) is established.

Despite the fact that $A_{i}$ need not be a contraction, it is still true that $\left\{x:\left\|\Lambda_{i} \mathbf{x}\right\|=\|x\|\right\}$ is a subspace of.H. This can be inferred from the Mrbius identity
(6) $I-A_{i}^{*} A_{i}=\left(I-\bar{\alpha}_{i} \alpha_{i}\right)\left(I-\alpha_{i} A^{*}\right)^{-1}\left(I-A^{*} A\right)\left(I-\bar{\alpha}_{i} A\right)^{-1}$
which shows that $I-A_{i}^{*} A_{i}$ is either positive or negative se-mi-definite, depending on whether $\left|\propto_{i}\right|<1$ or $\left|\propto_{i}\right|>1$, and hence that either $I-A_{i}^{*} \mathbf{A}_{i}$ or $\mathbb{A}_{i}^{*} \mathbf{A}_{i}-I$ has a Hernitian
square root. The standard argument given in the proof of Theoren 1 therefore applies. It is now easif seen with the help of an induction argument that $\mathrm{V}_{\mathrm{i}}$ is a subspace of H for each i.

To prove the stament about dimensions we establish an altern tive description of $\nabla_{i}$. For $1 \leqslant i, j \leqslant k$ let $\nu(i, j)$ denote the number of indices $\ell$ in $\{1,2, \ldots, i\}$ for which $\alpha_{\ell}=$ $=\alpha_{j}$; thus $\nu(i, j)$ can be zero if $j>i$ but $\nu(i, j) \geq 1$ if $j \leq i$. We shall show that, for $1 \leq i \leqslant k$,

$$
\begin{gather*}
V_{i}=\left\{x \in H:\left(I-A^{*} A\right)\left(I-\bar{\alpha}_{j} A\right)^{-\nu_{x}}=0,\right.  \tag{7}\\
I \leq j \leq i, 1 \leq 2 \leq 2(i, j)\} .
\end{gather*}
$$

In the case $i=1, \nu(1,1)=1$ and, since $I-A_{1}^{*} \Delta_{1}$ is semi-definite,

$$
\begin{aligned}
v_{1} & =\left\{x:\left(\left(I-A_{1}^{*} A_{1}\right) x, x\right)=0\right\} \\
& =\left\{x:\left(I-A_{1}^{*} A_{1}\right) x=0\right\} \\
& =\left\{x:\left(I-A^{*} A\right)\left(I-\bar{o}_{1} \Delta\right)^{-1} x=0\right\}
\end{aligned}
$$

(the last step being by virtue of (6)), so that (7) holds. Denote the right hand side of (7) by $w_{i}$ and observe that
(8) $w_{i}=w_{i-1} \cap \operatorname{Ker}\left(I-A^{*} A\right)\left(I-\bar{\alpha}_{i} A\right)^{-\nu(i, i)}$
so that, to establish (7) by induction, we must show that, for $x \in W_{i-1}, 2 \leqslant i \leqslant k,\left\|\Delta_{i} \Lambda_{\sigma} x\right\|=\|x\|$ for every subset $\boldsymbol{\sigma}$ of $\{1, \ldots, i-1\}$ if and only if
$\left(I-A^{*} A\right)\left(I-\bar{\alpha}_{i} A\right)^{-\nu(i, i)_{x}=0 \text {. } . ~ . ~}$
Consider, then, $x \in W_{i-1}=\nabla_{i-1}$ and $\sigma \in\{1, \ldots, i-1\}$. By definition of $V_{i-1},\left\|\mathbf{A}_{\sigma} x\right\|=\|x\|$ and so $\left\|A_{i} A_{\&} x\right\|=\|x\|$ if and only if $\left\|A_{i} A_{\delta} x\right\|=\left\|A_{6} x\right\|$, or equivalently,

$$
\left(\mathbf{A}_{\delta}^{*}\left(I-A_{i}^{*} \mathbf{A}_{i}\right) \mathbf{A}_{\sigma} x, x\right)=0
$$

or, by virtue of ( 6 ),

$$
\begin{equation*}
\left(\Lambda_{\sigma}^{*}\left(I-\alpha_{i} \Lambda^{*}\right)^{-1}\left(I-\Delta^{*} A\right)\left(I-\bar{\alpha}_{i} \Lambda\right)^{-1} \lambda_{\delta} x, x\right)=0 . \tag{9}
\end{equation*}
$$

Now $\left(I-\bar{\alpha}_{i} A\right)^{-1} \mathbf{\Lambda}_{\sigma}$ is a rational function of $\Lambda$; its denominator is of degree one more than its numerator, and the factor ( $I-\bar{\alpha}_{j} \Lambda$ ) eccurs in the denominator with an exponent which depends on the number of indices $\mathcal{E} \in \mathbb{G} \cup\{i\}$ for which $\alpha_{\ell}=\alpha_{j}$. It will be seen that the partial fractions expresoion for $\left(I-\bar{\alpha}_{i} \Lambda\right)^{-1} \mathbf{A}_{\sigma}$ can be written

$$
\begin{equation*}
\left(I-\bar{\alpha}_{i} \Lambda\right)^{-1} \lambda_{\sigma}=\sum_{j=1}^{i} e_{j}\left(I-\bar{\alpha}_{j} \Lambda\right)^{-\nu(j, j)} \tag{10}
\end{equation*}
$$

where $c_{1}, \ldots, c_{i}$ are scalars depending on $\sigma$. Substituting (10) in (9) we deduce that $\left\|\boldsymbol{A}_{i} \mathbf{\Lambda}_{\boldsymbol{\gamma}} \mathbf{x}\right\| \leq\|\mathbf{x}\|$ if and only if (11) $j_{j,} \sum_{l=1}^{i} \bar{c}_{j}{ }^{c} e^{\left(\left(I-\bar{\alpha}_{j} \Lambda^{*}\right)^{-\nu(j, j)}\right.}$

$$
\left.\left(I-A^{*} A\right)\left(I-\bar{\alpha}_{\ell} A\right)^{-\nu(l, l)} x, x\right)=0
$$

Since $x \in \mathbf{W}_{i-1}$ all the terms in (11) for which $j<i$ or $\ell<i$ vanish (this follows from the induction hypothesis), and (11) reduces to
$\left|c_{i}\right|^{2}\left(\left(I-\alpha_{i} \Lambda^{*}\right)^{-\nu(i, i)}\left(I-\Lambda^{*} \Lambda\right)\left(I-\bar{\alpha}_{i} A\right)^{-2(i, i)} x, x\right)=0$
and hence te
(12) $c_{i}\left(I-A^{*} A\right)\left(I-\bar{x}_{i} A\right)^{-\nu(i, i)_{x}}=0$.

It is now imnediate that if $x \in \boldsymbol{W}_{i}$ then $\left\|\boldsymbol{\Lambda}_{i} \mathcal{A}_{\mathbf{6}} x\right\|=\|x\|$ for all $\sigma \subseteq\{1, \ldots, i-1\}$, and hecne $x \in \nabla_{i}$. To prove the converse, let $x \in V_{i}$ and choose $\sigma$ to be $\{1, \ldots, i-1\}$. Then in the resolution (10) of ( $\left.I-\bar{\alpha}_{i} \Lambda\right)^{-I_{\Lambda_{\sigma}}}$ into partial fractions we have $c_{i} \neq 0$. To see this regard $A$ as a scalar variable for
the moment and notici that the left hand side of (10) has a pole of order $\nu(i, i)$ at $A=1 / \bar{\alpha}_{i}$ (this depends on the hypothesis about the non-conjugacy of the $\alpha$ 's, which ensures that no cancellation can take plase between numerator and denominator). The same must be true of the right hand side, and this can only be so if $c_{i} \neq 0$. Since $x \in V_{i}$, $\left\|A_{i} A_{\sigma} x\right\|=\|x\|$ and therefore (12) holds, and since $c_{i} \neq 0$ we have $\left(I-\mu^{*} A\right)\left(I-\bar{\alpha}_{i} A\right)^{-2(i, i)_{x}=0 \text {. It follows from (8) }}$ that $x \in V_{i}$. This completes the proof of the identity (7)

Statement (iii) of the Lemna also follows from the abo~ ve argument. If $x \in V_{i-1}$ and $\left\|A_{i} \ldots A_{i} x\right\|=\|x\|$ then, as we have just shown, $\left(I-A^{*} A\right)\left(I-\bar{\alpha}_{i} A\right)^{-\nu(i, i)}{ }_{x}=0$. But now that (7) is proved, we can rewrite (8)

$$
\begin{equation*}
\nabla_{i}=\nabla_{i-1} \cap \operatorname{Ker}\left(I-\Lambda^{*} A\right)\left(I-\bar{o}_{i} \Delta\right)^{-\nu(i, i)} . \tag{13}
\end{equation*}
$$

Thus $x \in V_{i}$.
The relation (13) shows that the codimension of $V_{i}$ in $\mathbf{V}_{i-1}$ is no greater than the rank of $I-d^{*} \mathbb{A}$, which is $r$. Thus din $\nabla_{i} \geq$ din $\nabla_{i-1}-r, 1 \leq i \leq k$. Statement ( $i$ ) follows at once.

The remaining assertion of the Lemna is justified much as in the proof of Theorem l. By (i),

$$
\operatorname{din} V_{i-1} \geq n-r(i-1) \geq n-r k+r \geq r
$$

Now $r$ cannot be zero, else $A$ would be unitary, contrary to the supposition $|\mathbb{A}|_{\sigma}<1$. Thus $V_{i-1}$ is a non-null subspace, $I \Leftrightarrow i \Leftrightarrow k$. From (5) we have $A_{i} V_{i} \subseteq \nabla_{i-1}$, so that if $\nabla_{i}=V_{i-1}$, $\nabla_{i}$ is a non-trivial invariant subspace for $A_{i}$ and $A_{i} \mid \nabla_{i}$ is an isometry. Thus $A_{i}$ has an eigenvalue of unit modulus. However, referring to the definition (3) of $A_{i}$ we perceive that the eigenvalues of $\Lambda_{i}$ are of the form $\left(\lambda-\alpha_{i}\right)\left(1-\bar{\alpha}_{i} \lambda\right)^{-1}$
where $\lambda$ (being an eigenvalue of $A$ ) is not of unit modulus. It is well known (and easily checked) that $\left(\lambda-\alpha_{j}\right)\left(1-\bar{\alpha}_{i} \lambda\right)^{-1}$ cannot then be of unit modulus either (when $\left|\alpha_{i}\right| \neq 1$, as here). This contradiction shows that $\nabla_{i} \neq \nabla_{i-1}$.

We can now conclude the proof of Theorem 2. We can suppose that $\left|\alpha_{k}\right|>1$. By (ii) of the Lemna there exists $x \in V_{k-1} \backslash$ $\backslash v_{k}$. For such an $x$
$\left\|A_{1} \mathbf{A}_{2} \ldots \mathbf{A}_{k-1} \mathbf{x}\right\|=\|\mathbf{x}\|$ but, by (iii).

$$
\left\|\mathbf{A}_{k} \mathbf{A}_{1} A_{2} \cdots \mathbf{A}_{k-1} \mathbf{x}\right\| \neq\|x\| .
$$

Thus, if we write $\Lambda_{1} \Lambda_{2} \ldots \Lambda_{k-1} x=y$, we have $\left\|\Lambda_{k} y\right\| \neq\|y\|$. Now the urbius identity (6) (with $i=k$ ) shows that $I-A_{k}^{*} A_{k}$ is negative seni-definite, which implies that $\left\|\Lambda_{k} u\right\| \geq\|u\|$ for every $u \in H$. It follows that $\left\|\Lambda_{k} y\right\|>\|y\|=\|x\|$. That is,

$$
\left\|A_{k} A_{1} A_{2} \ldots A_{k-1} x\right\|>\|x\|
$$

Hence $\left\|\Lambda_{1} \Lambda_{2} \ldots \Lambda_{k}\right\|>1$, as required.
Corollary. Let $p$ be a polynonial of degree $k$ which is relatively prime to the polynomial q defined by

$$
q(z)=z^{k} p(1 / z)^{-} .
$$

Let $A$ be an $n \times n$ matrix such that $|A|_{\sigma}<1$ and $I-A^{*} A$ is positive semi-definite and of rank $r$. If $n \geq r k$, the Hernitian form $H(\cdot)$ on ${\underset{\sim}{c}}^{n}$

$$
H(x)=\|q(A) x\|^{2}-\|p(A) x\|^{2}
$$

is positive semi-definite if and only if the roots of $p$ all have nodulus less than one.

Proof. Suppose H( P ) is positive semi-definite. Under these assumptions $q(A)$ must be non-singular, for etherwise there is an eigenvalue $\lambda$ of $A$ such that $q(\lambda)=0$. If we
take $x$ to be corresponding eigenvector we find that
$H(x)=\|q(\lambda) x\|^{2} \ldots\|p(\lambda) x\|^{2}=-|p(\lambda)|^{2}\|x\|^{2}$.
Since $H(x) \geq 0, p(\lambda)=0$ and so $p$ and $q$ have a root in comnon, contrary to hypothesis. Another way of stating the assumption on $H(\cdot)$ is to say that
$q(A)^{*} q(A)-p(A)^{*} p(A) \geq 0$, or equivalently, $I-g(A)^{*} g(A) \geq 0$ where $g(A)=p(A) q(A)^{-1}$. This is in turn equivalent to $\|g(A)\| \leqslant 1$, and Theorem 2 shows that this can only hold if no root of $p$ has absolute value greater than one. The assumption that $p$ and $q$ be relatively prine clearly inplies that no root of p can lie on the unit circle, and hence all roots of $p$ have modulus strictly less than one.

Suppose, conversely, that the roots of $p$ all have modulus less than 1 , and suppose further, for the nonent, that $q(A)$ is non-singular, so that we can form $g(A)=p(A) q(A)^{-1}$. In the notation of the proof of Theorem 2, $g(A)=A_{1} \mathbf{A}_{2} \ldots$. $\ldots \Lambda_{k}$, while $\left\|\lambda_{y}\right\| \leqslant 1$ for each $j$ (this follows from (6)). Hence, by the subnultiplicativity of the norm, $\|g(A)\| \leqslant 1$, and so $I-g(A)^{*} g(A) \geq 0$. Multiplying fore and aft by $q(A)^{*}$, $q(A)$, respectively, we infer that $q(A) * q(A)-p(A) * p(A) \geq 0$. Since the cone of positive semi-definite matrices is closed the restriction on $q(A)$ can be removed by a simole continuity argument.

Making a suitable choice of A we derive a test akin to Schur's test for polynomials, but suffering from the disadvantage that we have to check $p$ and $q$ for common factors. If we stick to the case $r=1$, as is a natural thing to do in practice, we can circunvent this awkward feature, and obtain
a result which contains that of Schur.
Theorem 3. Let p be a polynomial of degree $n, n \geq 1$, and let $A$ be an $n \times n$ matrix such that $|A|_{\sigma}<1$ and $I-A^{*} A$ is positive semi-definite and of rank one. The roots of $p$ are of modulus less than one if and only if the Hermitian form H( ) on ${\underset{\sim}{c}}^{\text {n }}$ given by

$$
H(x)=\|q(A) x\|^{2}-\|p(A) x\|^{2}
$$

is positive definite, where $q$ is the polynomial defined by

$$
q(z)=z^{n} p(1 / z)^{-}
$$

Proof. Suppose that the roots of pare of absolute value less than one. The roots of $q$ must then all lie outside the unit circle, and since $|A|_{6}<1$, it follows that $q(A)$ is non-singular. In the notation of the proof of Thoerem 2 (with $k=n$ ) we have $p(A) q(A)^{-1}=g(A)=A_{1} A_{2} \ldots A_{n}$. Now the $A_{j}$ 's commute and satisfy $\left\|A_{j}\right\| \leqslant 1$ (recall (6)); moreover, since $|\mathcal{A}|_{\sigma}<1$, it follows from properties of Mbbius transformations that $\left|A_{j}\right|_{\sigma}<1$. Hence, by Theorem 1 , $\|\mathbf{E}(\mathbf{A})\|<1$. This is to say that $I-g(A)^{*} g(A)$ is positive definite, and hence that $q(A)^{*} q(A)-p(A)^{*} p(A)$ is positive definite, as required.

Conversely, let $H($ - ) be positive definite. A fortiori $q(A) * q(A)$ is positive definite, and hence $q(A)$ is non-singular. Let the highest common factor of $p$ and $q$ be $f:$ we wish to show that $f=1$. Write $p=p_{1} f, q=q_{1} f$ where $p_{1}$, $q_{1}$ are relatively prime polynomials of degree $k$ and $p_{1}(z)=$ $=\left(z-\alpha_{1}\right) \ldots\left(z-\alpha_{k}\right)$, and note that $q_{1}(z)=z^{k} p_{1}(m)^{-}$, as can readily be seen when $p$ and $q$ are written as products of linear factors. The identity
$q(A)^{*} q(A)=p(A)^{*} p(A)=q(A)^{*}\left[I-g_{1}(A) * g_{1}(A)\right] q(A)$, where $g_{1}(A)=p_{1}(A) q_{1}(A)^{-1}$, and the fact that $q(A)$ is non-singular show that $I-\boldsymbol{g}_{1}(\boldsymbol{A})^{*} \mathbf{g}_{1}(\mathbf{A})$ is positive definite, and hence has kernel $\{0\}$. We have

$$
\begin{aligned}
\{0\} & =\operatorname{Ker}\left(I-g_{1}(A) *_{g_{1}}(A)\right)=\operatorname{Ker}\left(I-g_{1}(A)^{*} g_{1}(A)\right)^{1 / 2} \\
& =\left\{x:\left\|g_{1}(A) x\right\|=\|x\|\right. \\
& =\left\{x:\left\|A_{1} \cdots A_{k} x\right\|=\|x\|\right. \\
& \supseteq \nabla_{r},
\end{aligned}
$$

where we are using once again the notation of Theorem 2
(with $p$ replaced by $p_{1}$ ). Thus $V_{k}=\{0\}$. Since now $r=1$, the Lemme (i) shows that $\operatorname{dim} v_{k} \geqslant n-k$, and hence $n=k$. Thus $p=$ $=p_{1}$ and $f=1$. Now that we know that $p$ and $q$ are relatively prime we may apply the Corollary to Theorem 2 to deduce that the roots of $p$ all lie in the open unit disc. This completes the proof of Theorem 3.

Theorem 3 gives us infinitely many ways of testing whether the roots of a polynomial have modulus less than one one way for each choice of $A$ satisfying $|A|_{\sigma}<1,\|A\| \leqslant 1$ and rank $\left(I-A^{*} A\right)=1$. The simplest such choice is $A=S$ where

$$
s=\left[\begin{array}{lllll}
0 & 1 & 0 & \cdots & 0 \\
0 & 0 & 1 & \ldots & 0 \\
0 & \cdots & \cdots & \cdots & \cdot \\
0 & 0 & 0 & \cdots & 1 \\
0 & 0 & 0 & ,, & 0
\end{array}\right]
$$

(this is the matrix with respect to the natural basis of $\mathcal{C}^{n}$ of the shift operator $S_{n}$ mentioned in the introduction). Then if

$$
p(z)=a_{n} z^{n}+a_{n-1} z^{n-1}+\ldots+a_{0},
$$

so that

$$
q(z)=\bar{a}_{0} z^{n}+\bar{a}_{1} z^{n-1}+\ldots+\bar{a}_{n},
$$

we have

$$
p(A)=\left[\begin{array}{cccc}
a_{0} & a_{1} & \cdots & a_{n-1} \\
0 & a_{0} & \cdots & a_{n-2} \\
0 & \cdots & \cdots & , \\
0 & 0 & \cdots & a_{0}
\end{array}\right]
$$

so that, if $\left.x=x_{0}, \ldots, x_{n-1}\right)^{\top}$,

$$
\begin{aligned}
p(A) x= & \left(a_{0} x_{0}+\ldots+a_{n-1} x_{n-1}, a_{0} x_{1}+\ldots+a_{n-2} x_{n-1}, \ldots\right. \\
& \left.\ldots, a_{0} x_{n-1}\right)^{T},
\end{aligned}
$$

and se

$$
\|p(A) \times\|^{2}=\sum_{j=0}^{n-1}\left|a_{0} x_{1 j}+a_{1} x_{j+1}+\ldots+a_{n-j-1} x_{n-1}\right|^{2}
$$

On performing a similar calculation for $q$ we find that the Hermitian form $H()$ is in this case

$$
\begin{aligned}
H(x)= & \sum_{j=0}^{n-1}\left|a_{n} x_{j}+a_{n-1} x_{j+1}+\ldots+a_{j+1} x_{n-1}\right|^{2} \\
& -\sum_{j=0}^{n-1}\left|a_{0} x_{j}+a_{1} x_{j+1}+\ldots+a_{n-j-1} x_{n-1}\right|^{2} .
\end{aligned}
$$

This yields precisely Schur's criterion [5, § 13].
It is conceivable that there might on occasion be something to be gained by making a different choice of $A$. Now the class $\mathcal{E}$ of matrices $A$ satisfying the conditions we need $\left(\|\mathbb{A}\| \leqslant 1,|\boldsymbol{A}|_{\sigma}<1\right.$, rank $\left(I-A^{*} \mathbf{A}\right)=1$ ) has arisen in some closely related investigations of the author and V. Ptak, and has been shown to have some very interesting properties. For example, $\mathcal{E}$ consists of all matrices unitarily equivalent to a matrix of the form $(A S+B)(C S+D)^{-1}$ where $A, B, C, D$ are
$\mathrm{n} \times \mathrm{n}$ matrices such that

$$
\left[\begin{array}{ll}
\mathbf{A} & \mathbf{B} \\
\mathbf{C} & \mathrm{D}
\end{array}\right]^{*}\left[\begin{array}{rr}
\mathrm{I} & 0 \\
\mathrm{O} & -\mathbf{I}
\end{array}\right]\left[\begin{array}{ll}
\mathbf{A} & \mathbf{B} \\
\mathbf{C} & \mathbf{D}
\end{array}\right]=\left[\begin{array}{rr}
\mathbf{I} & 0 \\
0 & -I
\end{array}\right]
$$

(see [4, Theorem 1]). This description of $\mathcal{E}$ enables us to write down quite a large class of members of $\mathcal{E}$ explicitly ([4, § 2]): $\varepsilon$ contains all matrices of the form
$A=\left[\begin{array}{ccccc}p_{1} & s_{1} s_{2} & -s_{1} \bar{p}_{1} s_{2} & \ldots & (-1)^{n_{s}} \bar{p}_{2} \cdots \bar{p}_{n-1} s_{n} \\ 0 & p_{2} & s_{2} s_{3} & \ldots & (-1)^{n-1} \\ s_{2} \bar{p}_{3} \ldots \bar{p}_{n-1} s_{n} \\ 0 & \cdots & \ldots & \cdots\end{array}\right]$
where $p_{1}, \ldots, p_{n}$ are arbitrary complex numbers of mogulus less than 1 , and $s_{j}=\left(1-p_{j} \bar{p}_{j}\right)^{1 / 2}, 1 \leqslant j \leqslant n$.

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