# ACCELERATED MSOR METHOD 

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1. Abstract. Since the development of the "SOR" method by David Young [3], there has been a strong interest to use more than one parameter for SOR to improve the convergence [13], [14], [15] and [16].
D. Young himself considered a two parametric method called "MSOR". The two parameters weight the diagonal of positive-definite and consistently ordered 2-cyclic matrix [6], removing Young's hypothesis that the eigenvalues of Jacobi iteration matrix must all be real. We prove for certain cases that when "SOR" diverges, the two parametric method converges.
2. Introduction. To find the solution vector $x$ to the linear system

$$
\begin{equation*}
A x=b, \tag{2.1}
\end{equation*}
$$

where $A$ is a sparse $n \times n$ matrix and $b$ is a given $n$-vector of complex $n$-space. Stationary iterative methods, including SOR, solve the $n \times n$ linear system (2.1) by first splitting $A$ into two terms,

$$
\begin{equation*}
A=A_{0}-A_{1} \tag{2.2}
\end{equation*}
$$

where $A_{0}^{-1}$ is easy to compute. Relation (2.2) can be written as:

$$
\begin{equation*}
A=A_{0}\left(I-A_{0}^{-1} A_{1}\right)=A_{0}(I-B) \tag{2.3}
\end{equation*}
$$

where $B=A_{0}^{-1} A_{1}$ is called the iteration matrix. Therefore, the linear system (2.1) can be written as

$$
\begin{equation*}
x=B x+A_{0}^{-1} b . \tag{2.4}
\end{equation*}
$$

Then, by choosing any arbitrary starting vector $x_{0}$, the equation (2.4) is used to generate the vector sequence $\left\{x_{k}\right\}$, constructed as

$$
\begin{equation*}
x_{k+1}=B x_{k}+A_{0}^{-1} b \quad k=0,1,2, \ldots . \tag{2.5}
\end{equation*}
$$

By relation (2.3), it is clear that if $\left\{x_{k}\right\}$ converges at all, it must converge to $x_{\text {sol }}=A^{-1} b$ (vector solution), where $A x_{\text {sol }}=b$. Relation (2.3) shows that $\left\{x_{k}\right\}$ produced by (2.5) converges to $x_{\text {sol }}=A^{-1} b$ for any $x_{0}$ if and only if $\rho(B)<1$, where $\rho(B)$ is the spectral radius of $B[1]$. The smaller $\rho(B)$, the faster the sequence $\left\{x_{k}\right\}$ converges to $x_{\text {sol }}=A^{-1} b$ (asymptotically).

The above splitting is called stationary since there is no altering of parameter from iteration to iteration. It is called one part splitting since each $x_{k+1}$ depends only on one previous vector $x_{k}$.

Examples of one-part stationary splitting are represented in the following important iteration methods.
(i) Successive Overrelaxation (SOR) method was developed independently by Frankel [2] and Young [3], [4] in 1950.
S.O.R: Choose

$$
A_{0}=\frac{1}{\omega} D-L, \quad A_{1}=\left(\frac{1}{\omega}-1\right) D+U
$$

where $D$ is the diagonal part of $A$ and $-L,-U$, are strictly lower and upper triangular parts of $A$ respectively. Then, iteration matrix $B_{\omega}$ is given by

$$
B_{\omega}=(D-\omega L)^{-1}((1-\omega) D+\omega U)
$$

(ii) Modified Successive Overrelaxation (MSOR) method first considered by Devogelaere [5] in 1958. Here is how it works. Consider the matrix $A$ in the following form

$$
A=\left(\begin{array}{cc}
D_{1} & M \\
N & D_{2}
\end{array}\right)
$$

where $D_{1}$ and $D_{2}$ are square non-singular matrices. Use $\omega$ for the "red" equations corresponding to $D_{1}$ and $\omega^{\prime}$ for the "black" equations corresponding to $D_{2}$ then M.S.O.R: Choose

$$
A_{0}=\left(\begin{array}{cc}
\frac{1}{\omega} D_{1} & 0 \\
N & \frac{1}{\omega^{\prime}} D_{2}
\end{array}\right)
$$

Therefore, iteration matrix $B_{\left(\omega, \omega^{\prime}\right)}$ is defined by

$$
B_{\left(\omega, \omega^{\prime}\right)}=A_{0}^{-1} A_{1}=\left(\begin{array}{cc}
(1-\omega) I_{1} & \omega F \\
\omega^{\prime}(1-\omega) G & \omega \omega^{\prime} G F+\left(1-\omega^{\prime}\right) I_{2}
\end{array}\right)
$$

where $F=-D_{1}^{-1} M$ and $G=-D_{2}^{-1} N$. Young [6] has proved that if $A$ is positive-definite, then

$$
\rho\left(B_{\omega_{b}}\right)<\bar{\rho}\left(B_{\left(\omega, \omega^{\prime}\right)}\right),
$$

where $\bar{\rho}\left(B_{\left(\omega, \omega^{\prime}\right)}\right)$ is virtual spectral radius of $B_{\left(\omega, \omega^{\prime}\right)}$. Young also showed that $B_{1}$ (GaussSeidel iteration matrix) converges faster than MSOR if $A$ is positive definite, $0<\omega \leq 1$ and $0<\omega^{\prime} \leq 1$. Moussavi generalized Young's Theorem by considering $0<\omega \leq 1$ or $0<\omega^{\prime} \leq 1$ [17]. Mcdowell [7] and Taylor [8] analyzed the convergence of the MSOR method and obtained slightly better convergence by considering $\rho\left(B_{\left(\omega, \omega^{\prime}\right)}\right)$ instead of $\bar{\rho}\left(B_{\left(\omega, \omega^{\prime}\right)}\right)$.

In this paper, a comparison of the spectral radii of iteration matrices $B_{\left(1, \omega^{\prime}\right)}$ and $B_{(\omega, 1)}$ with $B_{1}$ is done for the case when the eigenvalues of $B_{j}$ (Jacobi) are not all real (Theorem 3.1). It can also be shown that if $A$ is positive-definite, then iteration matrices $B_{\left(1, \omega^{\prime}\right)}$ and $B_{(\omega, 1)}$ induce faster convergence than $B_{1}$ (Gauss-Seidel) for $1<\omega<2$ and $1<\omega^{\prime}<2$ (Corollary 3.3). If $A$ is an irreducible, $L$-matrix with $\rho\left(B_{j}\right)<1$, then a relationship can be found between $\rho\left(B_{\left(\omega, \omega^{\prime}\right)}\right)$ and $\rho\left(B_{1}\right)$. A sufficient condition for $\rho\left(B_{\left(1, \omega^{\prime}\right)}\right)<1$ can also be found (Theorem 3.7). Finally it is shown that if the SOR method is not convergent and the eigenvalues of SOR are in a certain region in the plane, then iteration matrix $B_{(1,2)}$ induces rapid convergence of $\left\{x_{k}\right\}$ (Theorem 3.9).

## 3. Accelerated MSOR Method.

Theorem 3.1. Suppose

$$
A=\left(\begin{array}{cc}
D_{1} & M \\
N & D_{2}
\end{array}\right)
$$

where $D_{1}$ and $D_{2}$ are non-singular matrices and let $\rho\left(B_{j}\right)<1$. Then
(a) If the eigenvalues of $B_{1}$ with modulus $\rho\left(B_{1}\right)$ have the real part less than $\rho^{2}\left(B_{1}\right)$, then

$$
\rho\left(B_{(\omega, 1)}\right)>\rho\left(B_{1}\right) \text { and } \rho\left(B_{\left(1, \omega^{\prime}\right)}\right)>\rho\left(B_{1}\right)
$$

for all $1<\omega<2$ and $1<\omega^{\prime}<2$.
(b) If the eigenvalues of $B_{1}$ with modulus $\rho\left(B_{1}\right)$ have the real part greater than $\rho^{2}\left(B_{1}\right)$, then

$$
\rho\left(B_{(\omega, 1)}\right)>\rho\left(B_{1}\right) \text { and } \rho\left(B_{\left(1, \omega^{\prime}\right)}\right)>\rho\left(B_{1}\right)
$$

for all $0<\omega<1$ and $0<\omega^{\prime}<1$.
(c) If (a) and (b) hold together, then $\rho\left(B_{1}\right)$ is the smallest.

Proof. According to Young [6],

$$
(\lambda+\omega-1)\left(\lambda+\omega^{\prime}-1\right)=\lambda \omega \omega^{\prime} \mu^{2}
$$

where

$$
\lambda \in \sigma\left(B_{\left(\omega, \omega^{\prime}\right)}\right) \text { and } \mu \in \sigma\left(B_{j}\right)
$$

It is clear that $B_{\left(1, \omega^{\prime}\right)}$ and $B_{(\omega, 1)}$ are Jacobi shifting of $B_{1}$, with parameters $\omega$ and $\omega^{\prime}$, respectively [12]. Hence if

$$
\xi \in \sigma\left(B_{(\omega, 1)}\right) \text { and } \psi \in \sigma\left(B_{\left(1, \omega^{\prime}\right)}\right)
$$

then

$$
\begin{align*}
\xi & =\omega \mu^{2}+(1-\omega) \cdot 1  \tag{3.1.1}\\
\psi & =\omega^{\prime} \mu^{2}+\left(1-\omega^{\prime}\right) \cdot 1 \tag{3.1.2}
\end{align*}
$$

(a) All the eigenvalues of $B_{1}$ with modulus $\rho\left(B_{1}\right)$ are on the arc $T B T^{\prime}$ (Figure 1), where $S T$ and $S T^{\prime}$ are tangent lines to the circle $C$ with center at the origin and radius $\rho\left(B_{1}\right)$. It is easy to show that $x$-coordinates of $T$ and $T^{\prime}$ is $\rho^{2}\left(B_{1}\right)$. Hence if $\omega>1$ or $\omega^{\prime}>1$, then $\mu^{2}$ shifts to $\xi$ or $\psi$ outside the circle $C$ on the line which passes through two points $\mu^{2}$ and $S:(1,0)$, respectively. This means that (3.1.1) or (3.1.2) gives a slower convergence. Of course in this case one could find $0<\omega<1$ or $0<\omega^{\prime}<1$ such that

$$
\rho\left(B_{(\omega, 1)}\right)<\rho\left(B_{1}\right) \text { and } \rho\left(B_{\left(1, \omega^{\prime}\right)}\right)<\rho\left(B_{1}\right)
$$



Figure 1.
(b) All the eigenvalues of $B_{1}$ with modulus $\rho\left(B_{1}\right)$, are on the arc $T A T^{\prime}$ (Figure 1). Then $\omega<1$ or $\omega^{\prime}<1$ shifts $\mu^{2}$ toward point $S:(1,0)$ on the line which passes through two points $\mu^{2}$ and $S:(1,0)$. Hence (3.1.1) and (3.1.2) will give a slower convergence. Of course in this case one could find $1<\omega<2$ or $1<\omega^{\prime}<2$ such that

$$
\rho\left(B_{(\omega, 1)}\right)<\rho\left(B_{1}\right) \text { or } \rho\left(B_{\left(1, \omega^{\prime}\right)}\right)<\rho\left(B_{1}\right) .
$$

(c) Clear by part (a) and part (b).

Lemma 3.2. $B_{(\omega, 1)}$ is a Jacobi shifting of $B_{\left(1, \omega^{\prime}\right)}$ or vice versa.
Proof. By (3.1.1) and (3.1.2)

$$
\mu^{2}=\frac{\psi+\omega^{\prime}-1}{\omega^{\prime}}=\frac{\xi+\omega-1}{\omega}
$$

or equivalently

$$
\psi=\frac{\omega^{\prime}}{\omega} \xi+\left(1-\frac{\omega^{\prime}}{\omega}\right) \cdot 1
$$

Corollary 3.3. Let

$$
A=\left(\begin{array}{cc}
D_{1} & M \\
N & D_{2}
\end{array}\right)
$$

where $D_{1}$ and $D_{2}$ are non-singular matrices. If $\mu_{1}=\rho\left(B_{j}\right)<1$ and all the eigenvalues of $B_{j}$ are real, then

$$
\rho\left(B_{(\omega, 1)}\right)<\rho\left(B_{1}\right) \text { or } \rho\left(B_{\left(1, \omega^{\prime}\right)}\right)<\rho\left(B_{1}\right)
$$

for $1<\omega<2$ or $1<\omega^{\prime}<2$.
Proof. Since all the eigenvalues of $B_{1}$ are on the line segment $\left[0, \mu_{1}^{2}\right]$, it is clear by part (a) of Theorem 3.1 that $\rho\left(B_{1}\right)$ is greater than $\rho\left(B_{(\omega, 1)}\right)$ or $\rho\left(B_{\left(1, \omega^{\prime}\right)}\right)$ for $1<\omega<2$ or $1<\omega^{\prime}<2$.

Theorem 3.4. Let

$$
A=\left(\begin{array}{cc}
D_{1} & M \\
N & D_{2}
\end{array}\right)
$$

where $D_{1}$ and $D_{2}$ are non-singular matrices. If $A$ is an irreducible $L$-matrix and $\rho\left(B_{j}\right)<1$, then

$$
\rho\left(B_{\left(1, \omega^{\prime}\right)}\right)<\rho\left(B_{\left(\omega, \omega^{\prime}\right)}\right)
$$

for $0<\omega<1$ and $0<\omega^{\prime}<1$.
Proof. To get $B_{\left(1, \omega^{\prime}\right)}$ and $B_{\left(\omega, \omega^{\prime}\right)}$, split matrix $A$ in the following ways. $A=A_{0}-A_{1}$, where

$$
A_{0}=\left(\begin{array}{cc}
D_{1} & 0 \\
N & \frac{1}{\omega^{\prime}} D_{2}
\end{array}\right)
$$

and

$$
A_{1}=\left(\begin{array}{cc}
0 & -M \\
0 & \left(\frac{1}{\omega^{\prime}}-1\right) D_{2}
\end{array}\right)
$$

and $A=A_{0}^{\prime}-A_{1}^{\prime}$, where

$$
A_{0}^{\prime}=\left(\begin{array}{cc}
\frac{1}{\omega} D_{1} & 0 \\
N & \frac{1}{\omega^{\prime}} D_{2}
\end{array}\right)
$$

and

$$
A_{1}^{\prime}=\left(\begin{array}{cc}
\left(\frac{1}{\omega}-1\right) D_{1} & -M \\
0 & \left(\frac{1}{\omega^{\prime}}-1\right) D_{2}
\end{array}\right)
$$

Since $A$ is an $L$-matrix, $D_{1}$ and $D_{2}$ are positive, and $N$ and $M$ are non-positive matrices, thus $-M$ is non-negative. Since $0<\omega<1$ and $0<\omega^{\prime}<1$,

$$
\left(\frac{1}{\omega}-1\right) D_{1} \text { and }\left(\frac{1}{\omega}-1\right) D_{2}
$$

are positive, hence $A_{1}^{\prime} \geq A_{1}>0$. Since $A$ is an $L$-matrix and $\rho\left(B_{j}\right)<1, A$ is an $M$-matrix [6]. That is, $A^{-1} \geq 0$. But because $A$ is also irreducible, $A^{-1}>0$ [9]. By Varga's Theorem 3.15,

$$
\rho\left(B_{\left(1, \omega^{\prime}\right)}\right)<\rho\left(B_{\left(\omega, \omega^{\prime}\right)}\right)
$$

for $0<\omega<1$ and $0<\omega^{\prime}<1$ [9].
Corollary 3.5. Under the assumption of Theorem 3.4, if all the eigenvalues of $B_{1}$ with modulus $\rho\left(B_{1}\right)$ have the real part greater than $\rho^{2}\left(B_{1}\right)$, then

$$
\rho\left(B_{\left(\omega, \omega^{\prime}\right)}\right)>\rho\left(B_{1}\right)
$$

for $0<\omega<1$ and $0<\omega^{\prime}<1$.
Proof. Clear by Theorem 3.4 and part (b) of Theorem 3.1.
Lemma 3.6. Suppose that

$$
A=\left(\begin{array}{cc}
I_{1} & M \\
N & I_{2}
\end{array}\right)
$$

Then

$$
\left(I-B_{\left(\omega, \omega^{\prime}\right)}\right)^{-1}
$$

exists if and only if $(I-N M)^{-1}$ exists.
Proof. Since

$$
\begin{gathered}
B_{\left(\omega, \omega^{\prime}\right)}=\left(\begin{array}{cc}
(1-\omega) I_{1} & \omega M \\
\omega^{\prime}(1-\omega) N & \omega \omega^{\prime} N M+\left(1-\omega^{\prime}\right) I_{2}
\end{array}\right), \\
\left.I-B_{( } \omega, \omega^{\prime}\right)=\left(\begin{array}{cc}
\omega I_{1} & -\omega M \\
-\omega^{\prime}(1-\omega) N & -\omega \omega^{\prime} N M+\omega^{\prime} I_{2}
\end{array}\right)
\end{gathered}
$$

Suppose that the matrix

$$
\left(\begin{array}{ll}
X & U \\
Y & V
\end{array}\right)
$$

is the inverse of

$$
\left(I-B_{\left(\omega, \omega^{\prime}\right)}\right) .
$$

Then

$$
\begin{align*}
\omega X-\omega M Y & =I_{1}  \tag{3.6.3}\\
-\omega^{\prime}(1-\omega) N X-\omega \omega^{\prime} N M Y+\omega^{\prime} Y & =0 \\
\omega U-\omega M V & =0 \\
-\omega^{\prime}(1-\omega) N U-\omega \omega^{\prime} N M V+\omega^{\prime} V & =I_{2} \tag{3.6.4}
\end{align*}
$$

By (3.6.3) and (3.6.4) one gets

$$
\begin{aligned}
\left(I-B_{\left(\omega, \omega^{\prime}\right)}\right)^{-1} & =\left(\begin{array}{ll}
X & U \\
Y & V
\end{array}\right) \\
& =\left(\begin{array}{cc}
\frac{1}{\omega} I_{1}+\frac{1-\omega}{-\omega} M(I-N M)^{-1} N & \frac{1}{\omega^{\prime}} M(I-N M)^{-1} \\
\frac{1-\omega}{\omega}(I-N M)^{-1} & \frac{1}{\omega^{\prime}}(I-N M)^{-1}
\end{array}\right) .
\end{aligned}
$$

Note that

$$
\left(I-B_{\left(\omega, \omega^{\prime}\right)}\right)^{-1} \text { and }\left(I+B_{\left(\omega, \omega^{\prime}\right)}\right)
$$

are commutative.
Theorem 3.7. Let

$$
A=\left(\begin{array}{cc}
I_{1} & M \\
N & I_{2}
\end{array}\right)
$$

and $\gamma$ be the eigenvalue of $(I-N M)^{-1}$ with the smallest real part, i.e., $0<\operatorname{Re} \gamma \leq \operatorname{Re} \lambda$ for all $\lambda \in \sigma\left((I-N M)^{-1}\right)$. Let $\rho\left(B_{j}\right)<1$. Then $\rho\left(B_{\left(1, \omega^{\prime}\right)}\right)<1$ if and only if $0<\omega^{\prime}<2 \operatorname{Re} \gamma$.

Proof. If $\mu$ is an eigenvalue of $B_{j}$, then $\mu^{2}$ is an eigenvalue of $N M$. Hence $\rho(N M)<1$, which implies that $(I-N M)^{-1}$ exists [9]. First we show that $(I-N M)^{-1}$ is $N$-stable,
which means all the eigenvalues of $(I-N M)^{-1}$ have positive real parts. Suppose that $\gamma$ is an eigenvalue of $(I-N M)^{-1}$, then one can write it in the form

$$
\gamma=\frac{1}{1-\mu^{2}}
$$

where $\mu \in \sigma\left(B_{j}\right)$. Let $\mu=x+y i$. Then

$$
\begin{equation*}
\operatorname{Re} \gamma=\frac{1-x^{2}+y^{2}}{\left(1-x^{2}+y^{2}\right)+4 x^{2} y^{2}} \tag{3.7.5}
\end{equation*}
$$

since $x^{2}+y^{2}<1,(3.7 .5)$ is positive. Let

$$
H=\left(I-B_{\left(\omega, \omega^{\prime}\right)}\right)^{-1}\left(I+B_{\left(\omega, \omega^{\prime}\right)}\right)=2\left(I-B_{\left(\omega, \omega^{\prime}\right)}\right)^{-1}-I
$$

Then

$$
H_{\left(1, \omega^{\prime}\right)}=\left(\begin{array}{cc}
I_{1} & \frac{2}{\omega^{\prime}} M(I-N M)^{-1}  \tag{3.7.6}\\
0 & \frac{2}{\omega^{\prime}}(I-N M)^{-1}-I_{2}
\end{array}\right)
$$

Therefore, the eigenvalues of $H_{\left(1, \omega^{\prime}\right)}$ are the same as the eigenvalues of its diagonal submatrices. Hence

$$
\sigma\left(H_{\left(1, \omega^{\prime}\right)}\right)=1 \cup\left\{\left.\frac{2}{\omega^{\prime}} \nu-1 \right\rvert\, \nu \in \sigma\left((I-N M)^{-1}\right)\right\}
$$

$H_{\left(1, \omega^{\prime}\right)}$ is $N$-stable if and only if all the real parts of its eigenvalues are positive, that is,

$$
\frac{2}{\omega} \operatorname{Re} \gamma-1>0
$$

or equivalently $0<\omega^{\prime}<2 \operatorname{Re} \gamma$. Then it is clear that $\rho\left(B_{\left(1, \omega^{\prime}\right)}\right)<1$ (see Theorem 1.5 in [6]).
Lemma 3.8. Let

$$
A=\left(\begin{array}{cc}
D_{1} & M \\
N & D_{2}
\end{array}\right)
$$

where $D_{1}$ and $D_{2}$ are non-singular matrices. Let $\xi$ be an eigenvalue of $B_{\left(1, \omega^{\prime}\right)}$ and $\lambda$ be an eigenvalue of $B_{\omega^{\prime}}$, then eigenvalues $\xi$ of $B_{\left(1, \omega^{\prime}\right)}$ and $\lambda$ of $B_{\omega^{\prime}}$ are related by the following relation

$$
\begin{equation*}
\xi=\frac{1}{\omega^{\prime}} \lambda+\frac{\left(\omega^{\prime}-1\right)^{2}}{\omega^{\prime}} \frac{1}{\lambda}-\frac{\left(\omega^{\prime}-1\right)\left(\omega^{\prime}-2\right)}{\omega^{\prime}} . \tag{3.8.7}
\end{equation*}
$$

Moreover, $\xi=\left(l_{1}\left(g\left(l_{2}(\lambda)\right)\right)\right)$, where

$$
l_{1}(\lambda)= \pm\left(\frac{1}{\omega^{\prime}-1}\right) \lambda, \quad g(\lambda)=\lambda+\frac{1}{\lambda}
$$

and

$$
l_{2}(\lambda)= \pm \frac{\left(\omega^{\prime}-1\right)}{\omega^{\prime}} \lambda-\frac{\left(\omega^{\prime}-1\right)\left(2-\omega^{\prime}\right)}{\omega^{\prime}}
$$

Proof. Suppose that $\psi$ is an eigenvalue of $B_{\left(\omega, \omega^{\prime}\right)}$ and $\lambda$ is an eigenvalue of $B_{\omega^{\prime}}$. According to Young [9],

$$
(\psi+\omega-1)\left(\psi+\omega^{\prime}-1\right)=\psi \omega \omega^{\prime} \mu^{2} \text { and }\left(\lambda+\omega^{\prime}-1\right)^{2}=\lambda \omega^{\prime} \mu^{2}
$$

which implies

$$
\begin{equation*}
\psi \lambda\left(\omega^{\prime} \psi-\omega \lambda\right)+\lambda \omega\left(\omega^{\prime}-\omega\right)\left(\omega^{\prime}-2\right)+\left(\omega^{\prime}-1\right)\left(\omega^{\prime} \lambda(\omega-1)-\omega \psi\left(\omega^{\prime}-1\right)\right)=0 \tag{3.8.8}
\end{equation*}
$$

Then if $\xi$ is an eigenvalue of $B_{\left(1, \omega^{\prime}\right)}$,

$$
\lambda \xi\left(\omega^{\prime} \xi-\lambda\right)+\left(\omega^{\prime}-1\right)\left(\omega^{\prime}-2\right) \lambda \xi+\left(\omega^{\prime}-1\right)\left(-\xi\left(\omega^{\prime}-1\right)\right)=0
$$

by (3.8.8). If $\xi \neq 0$, then (3.8.7) holds. Moreover, suppose that $\xi=\left(l_{1}\left(g\left(l_{2}\right)\right)\right)(\lambda)$, where $l_{1}$ and $l_{2}$ are linear functions, say $l_{1}(\lambda)=k \lambda+l, l_{2}(\lambda)=b \lambda+c$ and $g(\lambda)=\lambda+\frac{1}{\lambda}$. Then

$$
\begin{equation*}
\xi=\left(l_{1}\left(g\left(l_{2}(\lambda)\right)\right)\right)=(b k) \lambda+\frac{b}{k \lambda+l}+(b l+c) . \tag{3.8.9}
\end{equation*}
$$

Comparing (3.8.9) with (3.8.7),

$$
\begin{equation*}
b k=\frac{1}{\omega^{\prime}}, \frac{b}{k \lambda+l}=\frac{\left(\omega^{\prime}-1\right)^{2}}{\omega^{\prime}} \frac{1}{\lambda}, b l+c=\frac{\left(\omega^{\prime}-1\right)\left(2-\omega^{\prime}\right)}{\omega^{\prime}} \tag{3.8.10}
\end{equation*}
$$

By choosing $l=0$ in (3.8.10),

$$
k= \pm\left(\frac{1}{\omega^{\prime}-1}\right) \quad \text { and } \quad b= \pm \frac{\left(\omega^{\prime}-1\right)}{\omega^{\prime}}
$$

that implies

$$
l_{1}(\lambda)= \pm\left(\frac{1}{\omega^{\prime}-1}\right) \lambda, \quad l_{2}(\lambda)= \pm \frac{\left(\omega^{\prime}-1\right)^{2}}{\omega^{\prime}} \lambda-\frac{\left(\omega^{\prime}-1\right)\left(2-\omega^{\prime}\right)}{\omega^{\prime}}
$$

and $g(\lambda)=\lambda+\frac{1}{\lambda}$.
Theorem 3.9. Let

$$
A=\left(\begin{array}{cc}
D_{1} & M \\
N & D_{2}
\end{array}\right)
$$

where $D_{1}$ and $D_{2}$ are non-singular matrices. Suppose that eigenvalues of SOR lie inside the shaded area of Figure 2, where the circles $C_{1}$ and $C_{3}$ both have radius

$$
\frac{1+R}{2}
$$

with centers at

$$
\left(\frac{1-R}{2}, 0\right), \quad\left(\frac{-1+R}{2}, 0\right)
$$

respectively. Moreover, the circles $C_{2}$ and $C_{4}$ both have radius

$$
\frac{1+\frac{1}{h}}{2}
$$

with centers at

$$
\left(\frac{1-\frac{1}{R}}{2}, 0\right), \quad\left(\frac{-1+\frac{1}{R}}{2}, 0\right)
$$

respectively. Then the eigenvalues of $B_{(1,2)}$ are inside the shaded area of Figure 3.


Figure 2.
Furthermore, if $1<R<3+2 \sqrt{2}$, then $\rho\left(B_{(1,2)}\right)<1$.


Figure 3.
Proof. Since $\omega^{\prime}=2$, by (3.8.7)

$$
\begin{equation*}
\xi=\frac{1}{2}\left(\lambda+\frac{1}{\lambda}\right) \tag{3.9.11}
\end{equation*}
$$

(3.9.11) can be written in the following form [10],

$$
\frac{\xi-1}{\xi+1}=\left(\frac{\lambda-1}{\lambda+1}\right)^{2}
$$

by the following auxiliary transformations

$$
\text { (i) } Z_{1}=\frac{\lambda-1}{\lambda+1} \quad \text { (ii) } Z_{2}=Z_{1}^{2} \quad \text { (iii) } \frac{\xi-1}{\xi+1}=Z_{2}
$$

The image of circle which passes through two points $(1,0)$ and $(-R, 0)$, (i.e., circle $\left.C_{1}\right)$ under the transformation (i) is a circle say $C$, which goes through two points $(0,0)$ and

$$
\left(\frac{-R-1}{-R+1}, 0\right) .
$$

The image of circle $C$ under the transformation (ii) is a cardioid with the following equation

$$
\rho=\frac{(R-1)^{2}}{(R+1)^{2}}(1+\cos \varphi) .
$$

Finally, the image of this cardioid under the transformation (iii) is a symmetric Joukowski airfoil with respect to the real axis, which passes through two points $(1,0)$ and $\left(-\frac{1}{2}\left(R+\frac{1}{R}\right), 0\right)[11]$ (Figure 4).


Figure 4.
Obviously the image of the circle $C_{2}$ under transformation (i) is the circle $C$. Then the image of the circle $C_{1}$ and $C_{2}$ under transformation (3.9.11) coincide. Also it is clear that all the points outside the circle $C_{2}$ and inside of the circle $C_{1}$ (i.e., all points belong to $C_{1}-C_{2}$ ) map inside the Joukowski airfoil. The same argument holds for circles $C_{3}$ and $C_{4}$, i.e., all points belong to $C_{3}-C_{4}$ map inside the symmetric Joukowski airfoil about the real axis which passes through the points $(-1,0)$ and $\left(-\frac{1}{2}\left(R+\frac{1}{R}\right), 0\right)$.

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